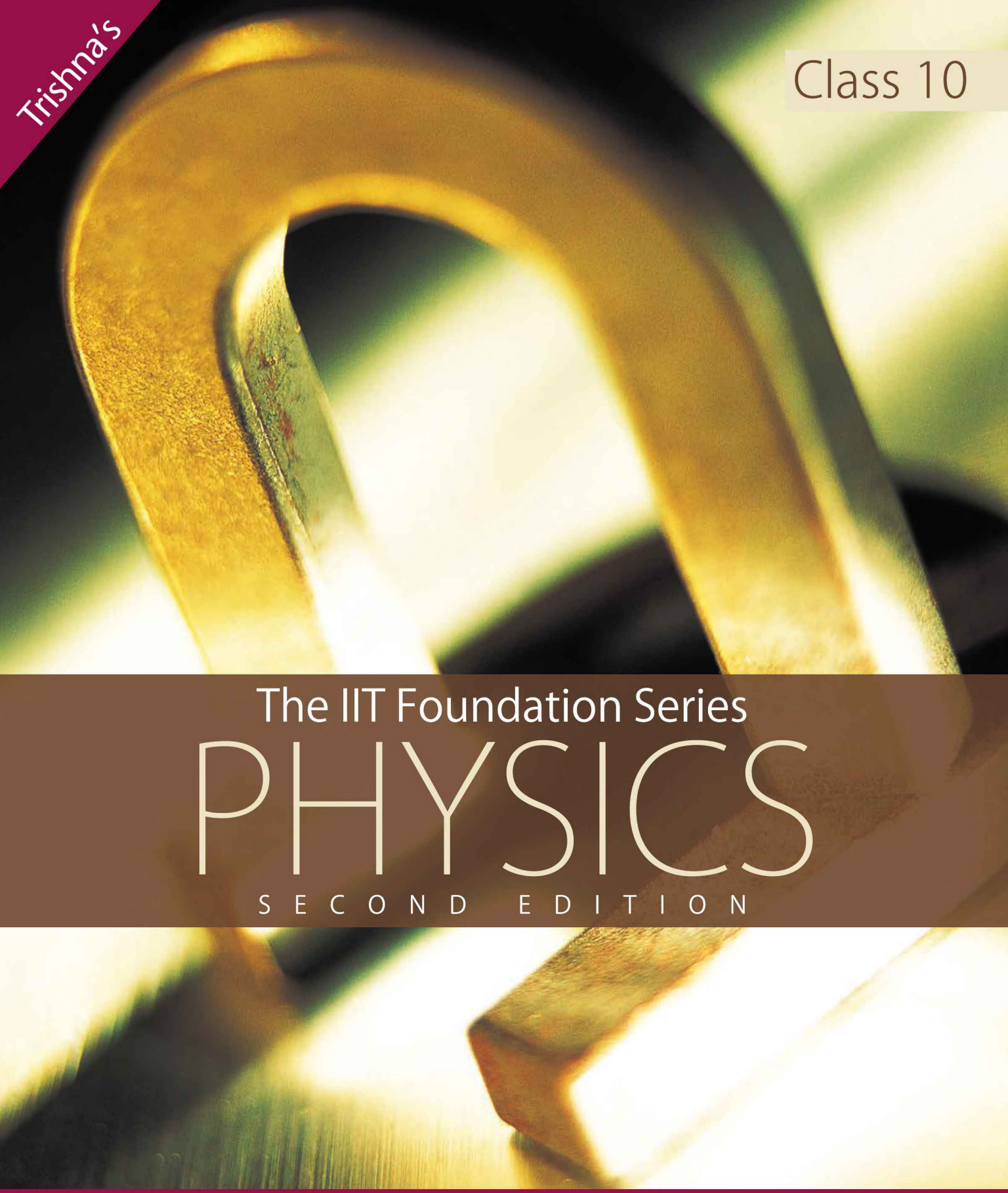
Trishna's

Class 10



The IIT Foundation Series

PHYSICS

SECOND EDITION

ALWAYS LEARNING

PEARSON

PHYSICS CLASS 10

The IIT Foundation Series

(Second Edition)

Trishna Knowledge Systems
A division of
Triumphant Institute of Management Education Pvt. Ltd.



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Preface to the Second Edition



It is with enthusiasm that we present the second edition of the IIT Foundation series of books. The success of our first edition indicates that there exists a large group of students who wish to delve into Mathematics, Physics and Chemistry, beyond the restrictions of their school syllabus. We take this opportunity to thank all those who have contributed to the success and to reaffirm our commitment to inculcating scientific temper among the student community.

Over the last two years we have received feedback from many students and teachers who have used our books, and it has given us great satisfaction to know that they have benefitted immensely. The teachers have taught from these books and have appreciated the approach adopted in the presentation of concepts. Their feedback has encouraged us to extend the IIT-Foundation series of books to include books for students of Class 7 as well. In most school curricula, the fundamentals of Mathematics and Science that students study in Classes 8 to 10 are introduced in Class 7. While a proper understanding of these fundamentals is essential for good performance in higher classes, it is not possible for Class 7 students to develop indepth understanding of concepts unless these are presented lucidly.

With this edition, therefore, we present books for Class 7. We are confident that these will help students develop conceptual clarity at their early age, and that the student community will take full advantage of this new inclusion.

As with all our publications, the objective of this series is to provide students with a comprehensive understanding of fundamental concepts, to teach them the application of these concepts, and to help hone their problem-solving skills. We trust all our young students will find these books relevant and enlightening.

Preface



As the old adage goes, "nothing succeeds like success." The truth in this maxim cannot be overstated in today's competitive world. The present-day student is under immense pressure to thrive and emerge triumphantly in examinations. Students aspire to get into pre-eminent educational institutes to pursue the best courses–be it in engineering, medicine, arts or sciences–to enable them to prepare for careers at the global level. Their performance in entrance examinations are often the cornerstones that determine if they would be admitted into these hallowed halls of learning. With most of these exams being designed to challenge the innate talent and ingenuity of students, it is only natural that they find these tests most demanding and that they find themselves competing with the country's best minds for those few coveted seats. Only those students with a thorough understanding of the fundamental concepts and exceptional problem-solving skills pass out with flying colours in these tests.

The "IIT Foundation Series" books are designed to provide students with a comprehensive understanding of the fundamental concepts, to teach them the application of these concepts and to hone their problem-solving skills.

The objective of the IIT Foundation Series books is to ensure that students are able to delve beyond the restrictions of their regular school syllabus and get a fundamental understanding of Mathematics, Physics and Chemistry. The books are designed to kindle student interest in these subjects and to encourage them to ask questions that lead to a firm grip on the principles governing each concept.

Irrespective of the field of study that the student may choose to take up later, it is imperative that he or she develops a sound understanding of Mathematics and Science, since it forms the basis for most modern-day activities. Lack of a firm background in these subjects may not only limit the capacity of the student to solve complex problems but also lessen his or her chances to make it into top-notch institutes that provide quality education.

This book is intended to serve as the backbone of the student's preparation for a range of competitive exams, going beyond the realms of the usual school curriculum to provide that extra edge so essential in tackling a typical question paper.

A distinctive feature of this book is that it has been written by a team of well-qualified teachers experienced in imparting the fundamentals of Mathematics and Science, and their applications to active learners at T.I.M.E. (Triumphant Institute of Management Education Pvt. Ltd). They have also been instrumental in developing high-quality study material for IIT Foundation courses for Classes 7 to 10. We are sure that you will find this book, prepared by such stalwarts, to be very useful in your preparation for entrance examinations.

About the IIT Foundation Series



This book is a perfect companion not only for the students of 7th Grade, but also for higher grades. It will help them achieve the much-needed conceptual clarity in the topics which form the basis for their higher study.

Some of the important features of the book are listed below:

- Builds skills that will help students succeed in school and various competitive examinations.
- The methodology is aimed at helping students thoroughly understand the concepts in Mathematics, Physics and Chemistry.
- Helps develop a logical approach to Mathematics, Physics and Chemistry, thereby enabling more effective learning.
- Lays stress on questions asked by board/school examinations as well as application of concepts.
- The concepts are explained in a well structured and lucid manner, using simple language. This aids learning.
- A large number of examples have been included to help reinforce the concepts involved.
- Different levels of practice exercises have been provided which help students develop the necessary application and problem-solving skills.
- The exercises have been designed keeping in mind the various board/school examinations and competitive examinations, such as the NTSE, NLSTSE, Science Olympiad and Cyber Olympiad.
- The book will not only help the students in better understanding of what is taught in regular school classes (and hence enable them to do well in board examinations) but will also help in developing the acumen, resulting in a distinctive edge ovr their peers.
- Given below are a few examples that demonstrate how the course will help students in understanding the fundamentals:

How does a kingfisher catch fish?

The kingfisher flies vertically over the position of the fish, then plunges into the water at a 90^0 angle. The concept here is that the normally incident rays do not undergo refraction, hence the fish lies exactly where it appears to be. At any other angle, the apparent location of the fish would be different from its real location.

Why do we normally swing our arms while walking, and why not when we carry a load in our hands?

The center of gravity of a body depends on the distribution of mass in the body. As we walk, the movement of the legs tends to cause a shift in the centre of gravity. To compensate for this shift we swing our arms. When we are carrying a load in the hands, however, the effective C.G is lower, making it easier to maintain balance.

Why does salt become damp when kept exposed during the rainy season and not when kept exposed during summer?

In the rainy season humidity in the atmosphere is very high, i.e., there is a lot of moisture in the atmosphere. Thus, calcium chloride, which is the impurity present in common salt, absorbs this moisture and makes the salt damp. In summer, however, as the temperature is high, calcium chloride tends to loose moisture through the process of evaporation, and the salt is left free-flowing.

Structure of the IIT Foundation Series



The IIT Foundation Series is available in Mathematics, Physics and Chemistry. Each chapter in the book is divided into three parts, namely, theory, test your concepts and concept application.

➤ **Theory:**

The theory part deals with the various concepts in Physics/Chemistry/Mathematics, which is a part of the syllabus prescribed by major boards for Class X. The concepts are explained in a lucid manner, and diagrams have been provided, wherever necessary, to illustrate these concepts.

➤ **Test your Concepts:**

This exercise is provided at the end of the theory section of each chapter. These exercises are a collection of very short answer, short answer and essay type of descriptive questions. It is intended to provide students with model questions that they may face in the board examination.

Students are expected to prepare for these questions before they attempt any examination based on that particular chapter. Towards the end of the book, the students will find key points for selected questions of the exercise. These key points provide students with an idea of the points that should be a part of an answer for such a question.

➤ **Concept Application:**

This is a collection of exercises in four different classes: Class 7, Class 8, Class 9 and Class 10.

Class 8 consists of basic objective questions. These questions test the basic knowledge of students and enable them to gauge their understanding of concepts when they start solving this exercise. The key for this exercise is provided at the end of the respective chapter.

Classes 9 and 10 consist of descriptive questions of a higher level of difficulty. These questions help students to *apply the concepts* that they have learnt. Key points for selected questions of these exercises have been provided at the end of each chapter in order to help students solve these questions.

These books are available for 7th, 8th, 9th and 10th classes separately for Mathematics, Physics and Chemistry.

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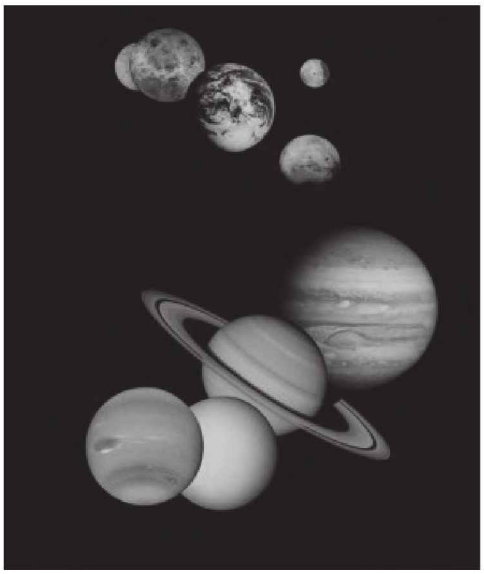
CHAPTER 9
Modern Physics

CHAPTER 10
Sources of Energy

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Electronics

1

Kinematics



INTRODUCTION

One can study the motion of a body in a given time interval if its initial velocity, final velocity, acceleration, and displacement are known.

The relation between $\vec{v}$, $\vec{u}$, $\vec{a}$, t, and $\vec{s}$ for a body moving with uniform acceleration in a straight path are well known to us. Equations which relate these quantities are known as equations of motion.

The equations of motion are:

(i) $v = u + at$

(ii) $s = ut + \frac{1}{2} at^2$

(iii) $v^2 = u^2 + 2as$

(iv) $s_n = u + a\,(n - 1/2)$

Do the above equations hold good for a body which is dropped from a height, or projected vertically upwards?

Answer to the above can be found out by the example given below.

Drop two stones, of masses 1 kg and 5 kg respectively simultaneously, from the same height. Note the time taken by the two stones to reach the ground.

Which stone reaches the ground first?

We observe that both the stones reach the ground in approximately the same time.

The above fact was proved experimentally by Galileo. He dropped a heavy iron ball and a light wooden ball simultaneously from the leaning tower of Pisa. Both the balls reached the ground at the same time though their masses were not the same.

Further experiments proved that wherever air resistance is negligible, all bodies, irrespective of their masses would take the same time to reach the ground when dropped from the same height. In fact, a body dropped from a height would accelerate as it travels downwards, i.e., its velocity increases as it moves down. On the other hand, the velocity of a body projected vertically upwards would decrease as it moves up. Since this acceleration is due to the gravitational force exerted by the earth, it is generally referred to as "acceleration due to gravity" and is denoted by the letter 'g'.

The value of this 'g' at mean sea level is approximately 9.8 m s^{-2}

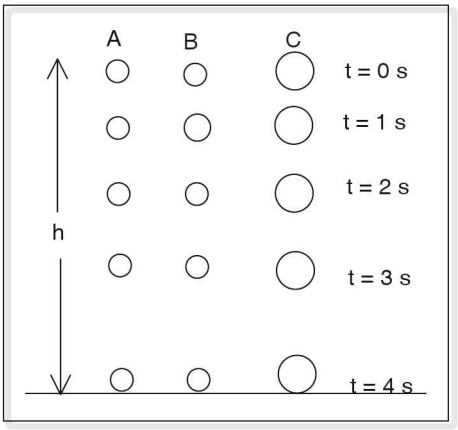


Figure 1.1

For obvious reasons, the acceleration due to gravity acts downwards and would take a '+' sign or '–' sign depending on the sign convention chosen. And as a body moves along a straight line path either when dropped from a height or when projected vertically up, the same equations of motion may be used in both the cases by taking appropriate signs for 'g'. Thus the equations of motion for a body moving vertically under the influence of gravity are

1. $v = u \pm gt$
2. $s = ut \pm \frac{1}{2} gt^2$
3. $v^2 = u^2 \pm 2gs$
4. $s_n = u \pm g(n - \frac{1}{2})$

The value of 'g' at mean sea level is taken as 9.8 m s^{-2}.

Freely falling body

When an apple drops from a tree, the initial velocity of the apple is zero due to inertia of rest. The apple moves towards the earth under the influence of gravity. Though the motion is opposed by friction of air, the resistance of air may be considered negligible in comparison to the weight of the body. Such a motion which takes place under the influence of only gravity is called free-fall.

Equations of motion for a freely falling body

Let us consider a special case of a freely falling body released from rest (i.e dropped) which means the initial velocity 'u' is equal to zero. The body moves towards the earth due to the gravitational force of the earth.

Considering the downward direction as positive (you can also take it as negative) the equations of motion of a freely falling body, dropped from a height 'h' from the ground, are obtained by substituting zero for u, h for s, and g for a in the equations of motion. Then we get,

1. $v = gt$
2. $h = \frac{1}{2} gt^2$
3. $v^2 = 2gh$
4. $s_n = g (n - \frac{1}{2})$

Equations of motion for a body thrown vertically upwards

Let a ball be thrown vertically upwards. What do we observe?

The following facts are observed:

1. The impulsive force given to the ball provides an initial velocity to the ball for it to move upwards.
2. The ball reaches a certain maximum height where it stop (i.e $v = 0$) and then returns to the ground.

Conclusions

1. It is impossible for a body to move away from the earth on its own due to the earth's gravitational pull. Hence an initial velocity should be given to the body.
2. At low speeds, a body cannot keep going up indefinitely but will return to the ground after reaching a certain height.

Here the acceleration due to gravity acts in a direction opposite to the direction of initial velocity.

Considering the upward direction as positive the equations of motion for a body projected vertically upwards are obtained by substituting h for s and –g for a as follows.

1. $v = u - gt$
2. $h = ut - \frac{1}{2} gt^2$
3. $v^2 = u^2 - 2gh$
4. $s_n = u - g (n - {}^1/_2)$

☛ *Example*

A ball is dropped from the top of a building 19.6 m high. Find the time taken by the ball to reach the ground.

Solution

This is a case of a freely falling body. Since the ball is dropped the initial velocity will be zero. Thus we have,

$u = 0$

$s = 19.6$ m

$a = + g = + 9.8$ m s^{-2} ('+' sign since the body is moving downwards).

Using the second equation of motion, $s = ut + \frac{1}{2}at^2$

$19.6\ m = 0 + \frac{1}{2} \times (9.8\ m\ s^{-2}) \times (t)^2$

$t^2 = \frac{19.6 \times 2}{9.8} = 4$

$\therefore\ t = 2\ s.$

☛ *Example*

A Diwali rocket, when ignited at the ground level, rises vertically upwards to the level of a window 10 m from the ground. Find the magnitude of velocity of the rocket at the time of its projection.

Solution

Since the body is moving up, the acceleration due to gravity 'g' is taken with a '–' sign in the equations of motion. The body rises up to a height of 10 m, i.e., its velocity at this height is zero, substituting these values in the 3rd equation of motion, $v^2 = u^2 + 2as$, we get

$0 = u^2 - 2(9.8\ m\ s^{-2})\ (10\ m)$

$\therefore u^2 = 196$

$u = 14\ m\ s^{-1}$

☛ *Example*

A stone is released from a hot-air balloon which is rising steadily with a velocity of 4 m s^{-1}. Find the velocity of the stone at the end of 3 s after it is released.

Solution

Let the downward direction be positive. As the stone was moving with the balloon, the initial velocity of the stone is the same as that of the balloon. Thus we have,

$u = -4\ m\ s^{-1}$

$a = +9.8\ m\ s^{-2}$

$t = 3\ s$

Considering the first equation of motion,

$v = u + at$, and substituting the values of u, a and t

$v = -4 + 9.8 \times 3$

$v = -4 + 29.4$

$v = 25.4\ m\ s^{-1}.$

☛ *Example*

A stone is dropped by a person from the top of a building, which is 200 m tall. At the same time, another stone is thrown upwards, with a velocity of 50 m s^{-1} by a person standing at the foot of the building. Find the time after which the two stones meet.

Solution

Let the two stones meet at a distance of x m from the top of the building, and 't' be the time taken.

Let us assume the downward direction as positive,

For the stone, that is dropped, its initial velocity u = 0 m s^{-1}; displacement s = x and acceleration = acceleration due to gravity (g).

Using $s = ut + \frac{1}{2}at^2$, we get $x = (0)\,t + \frac{1}{2}gt^2$ ---------- (1)

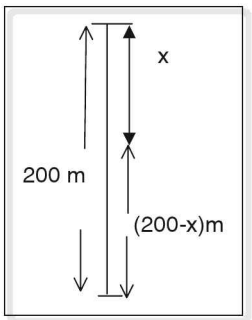


Figure 1.2

For the stone that is projected vertically upwards, its initial velocity, u = –50 m s^{-1}; displacement s = – (200 – x) and acceleration a = g. Using s = $ut + \frac{1}{2}at^2$ we get

$-(200 - x) = -50 \times t + \frac{1}{2}gt^2$

$200 = 50\,t - \frac{1}{2}gt^2 + x$ --- (2)

From the equations (1) and (2), we have

$200 = 50t - \frac{1}{2}gt^2 + \frac{1}{2}gt^2$

$200 = 50t$

$\therefore t = 4$ s

Maximum height reached by a body projected vertically upwards

Consider two balls *A* and *B*, and let *A* be thrown upwards with a higher initial velocity than *B*. Observe which reaches greater height.

Ball *A* reaches a greater height than ball *B*.

Thus the maximum height reached by a body depends on its initial velocity and is independent of its mass.

When a body is thrown upwards, the velocity of the body decreases continuously and finally becomes zero at a certain height. This height where the final velocity of the body is zero is called the maximum height. (denoted by h_{max} or H).

At the maximum height, v = 0

$\therefore$ From $v^2 = u^2 - 2gh$ we get,

$0 = u^2 - 2g\,h_{max}$

$\Rightarrow h_{max}$ or $H = \frac{u^2}{2g}$

Here, h_{max} refers to the maximum height attained by the body thrown vertically upwards.
From this expression it is evident that the maximum height depends on the velocity of projection.

Time of ascent (t_a)

It is the time taken by a body, projected vertically upwards, to reach the maximum height.
At maximum height the final velocity of the body v = 0.
From the equation,
v = u + at, we get
$\Rightarrow 0 = u - gt_a$
$\Rightarrow t_a = \frac{u}{g}$

Thus the time of ascent of a body, projected vertically upwards, depends on the velocity of projection.

Hence greater the initial velocity, greater will be the maximum height attained by a body and its time of ascent.

Time of descent (t_d)

It is the time taken by a body, projected vertically upwards, to reach the point of projection from its position of maximum height.

Let t_d be the time of descent of a body projected vertically upwards. Let 'u' be the velocity of projection of the body and h_{max} be its maximum height. In the return journey of the body from its position of maximum height, its initial velocity is zero. Hence, using $s = ut + \frac{1}{2}at^2$, we get

$h_{max} = 0 + \frac{1}{2} g t_d^2 \Rightarrow t_d = \sqrt{\frac{h_{max} \times 2}{g}}$

As $h_{max} = \frac{u^2}{2g}$, we get

$t_d = \sqrt{\frac{u^2}{2g} \times \frac{2}{g}} \Rightarrow t_d = \sqrt{\frac{u^2}{g^2}} \Rightarrow t_d = \frac{u}{g}$

Thus we find that the time of ascent is always equal to the time of descent.

Time of flight (T)

It is the total time taken by a body projected vertically upwards to reach the position of maximum height and then return to the point of projection.

It is denoted by T.
It is the total time for which the body remains in the air, and is equal to the sum of the time of ascent and the time of descent.
Time of flight = $t_a + t_d$
$\Rightarrow T = u/g + u/g \Rightarrow T = \frac{2u}{g}$

Velocity on reaching the ground

For a freely falling body dropped from rest, the initial velocity is zero.

From $v^2 = u^2 + 2as$, substituting $u = 0$, $a = g$ and $s = h$, we get

$v^2 = 2gh$

$\therefore v = \sqrt{2gh}$ --------- (1)

Here, v is the final velocity on reaching the ground and h is the height from which the body is dropped. When a body is projected vertically upwards with initial velocity u, then at the maximum height, $v = 0$

$0 = u^2 - 2gh$ (From the equation $v^2 = u^2 + 2as$)

$u^2 = 2gh$

$u = \sqrt{2gh}$ ---------- (2)

From equations (1) and (2) it is seen that the magnitude of the velocity with which the body reaches the ground is equal to the magnitude of the velocity with which it is projected vertically upwards.

Two-dimensional motion

☛ *Example*

An aeroplane is flying horizontally at a height of 500 m with a velocity of 72 km h^{-1} over a flood affected area. Food packets are to be dropped for the people standing at a particular place on the ground. At what horizontal distance from that place should the food packets be released from the plane, so as to reach the intended place.

Solution

When a food packet is released from an aeroplane moving horizontally, due to inertia the food packets would have an initial velocity equal the velocity of the plane, i.e., 72 km h^{-1} in the horizontal direction, while the initial velocity in the vertical direction would be zero. Thus the motion of the body is in two dimensions – vertical, under the influence of gravity (acceleration = +g m s^{-2}) and horizontal due to its initial velocity.

Furthermore, let us neglect the effects of wind and air drag. Considering the vertical motion alone, the time taken by a packet to reach the ground is, time of descent, $t = \sqrt{\frac{2h}{g}} = \sqrt{\frac{2 \times 500}{10}} = \sqrt{100} = 10$ s

Simultaneously the packet is also moving horizontally. In the horizontal direction, acceleration of the food packets $a = 0$ m s^{-2} since there is not any force acting in the horizontal direction and the packet travels with uniform velocity $u = 72$ km $h^{-1} = 72 \times \frac{5}{18}$ m $s^{-1} = 20$ m s^{-1}, in the horizontal direction.

Let *x* m be the horizontal distance travelled by the body in the time it take to reach the ground.

Using $s = ut + \frac{1}{2}at^2$ in the horizontal direction, we get $x = (20\ m\ s^{-1}) \times (10\ s) = 200\ m$.

Thus, the food packets should be dropped from the plane at a horizontal distance of 200 m from the intended place for the people to be able to receive them.

Sign convention

While solving numerical questions it would be convenient to adopt one single sign convention rather than remembering two sets of equations of motion for bodies moving along a vertical path under the influence of gravity, whether freely falling bodies or those projected vertically upwards.

According to this sign convention, displacement velocity and acceleration, being vector quantities are taken as positive quantities when directed upwards and as negative quantities when directed downwards.

According to this sign convention, displacement, velocity and acceleration, being vector quantities, are taken as positive quantities when directed upwards and as negative quantities when directed downwards.

Initial velocity	$-u$	$+u$
Final velocity	$-v$	$+v$
Acceleration	$-g$	$-g$
Displacement	$-h$	$+h$
$V = u + at$	$-v = -u - gt$ $\Rightarrow v = u + gt$	$V = u - gt$
$V^2 - u^2 = 2as$	$v^2 - u^2 = 2(-g)(-h)$ $\Rightarrow v^2 - u^2 = 2gh$	$v^2 - u^2 = 2(-g)(+h)$ $\Rightarrow v^2 - u^2 = -2gh$
$s = ut + 1/2at^2$	$-h = -ut - 1/2gt^2$ $\Rightarrow h = ut + 1/2gt^2$	$+h = ut + 1/2(-g)(t^2)$ $\Rightarrow h = ut - 1/2gt^2$
$S_n = u + a/2\ (2n - 1)$	$-Hn = -u + (-g)/2\ (2n - 1)$ $\Rightarrow Hn = u + g/2\ (2n - 1)$	$+Hn = u + (-g)/2\ (2n - 1)$ $\Rightarrow Hn = u - g/2\ (2n - 1)$

☛ *Example*

A body at rest falls from a height and covers 34.3 m in the last second of its fall. Find the height from which the body is dropped.

Solution

Let the body take 'n' seconds to reach the ground. Thus nth second is its last second of fall. Initial velocity of the body, $u = 0\ m\ s^{-1}$. The distance covered by the body in its nth second $s_n = 34.3$ m. Using $s_n = u + g\ (n - ½)$ we get $34.3 = 0 + 9.8\left(n - \frac{1}{2}\right) \Rightarrow n = 4$

Thus the duration of the fall for the body is t = 4 seconds.

$\therefore$ The height from which the body is dropped, $h = \frac{1}{2} gt^2 = \frac{1}{2} (9.8)\ (4)^2 = 78.4$ m.

☛ *Example*

A stone is dropped to the ground from the top of a tower and it reaches the ground in 8 s. How much time will it take to cover the first quarter of the distance from the top?

Solution

Let the height from which the stone is dropped be x. Since the stone is dropped its initial velocity, $u = 0\ m\ s^{-1}$, $s = -x$, $a = -g$

Time of descent of the stone, t = 8 s

$\therefore$ From the equation $s = ut + \frac{1}{2} at^2$, we have

$x = ½\ g \times 8^2$

$x = 32\ g$ -------- (1)

Let t_1 be the time taken to cover first quarter of the distance

$\therefore \frac{x}{4} = ½\ gt_1^{\ 2}$

$\Rightarrow \frac{32g}{4} = ½\ gt_1^{\ 2}$

$\Rightarrow t_1^{\ 2} = 16$

$\Rightarrow t_1 = 4$ s

☛ *Example*

A ball is thrown vertically upwards from the top of a tower with a velocity 10 m s^{-1}. The ball reaches the ground with the velocity 30 m s^{-1}. What is the height of the tower?
(Take g = 10 m s^{-2})

Solution

According to the sign convention taking the vectors in the upward direction as positive
initial velocity, $u = +10\ m\ s^{-1}$,
final velocity, $v = -30\ m\ s^{-1}$,
displacement, s = –h, where h is the height of the tower and
acceleration, $a = -10\ m\ s^{-2}$
Using $v^2 = u^2 + 2as$ we get
$(-30)^2 = (10)^2 + 2\ (-10)\ (-h)$

$\therefore h = \frac{(-30)^2 - 10^2}{2 \times 10} = \frac{900 - 100}{20} = 40\,m$

test your concepts

Very short answer type questions

1. Motion of an aeroplane is an example of _____ dimensional motion.
2. Define maximum height.
3. A particle is projected up with a velocity of $\sqrt{29}$ m s^{-1} from the tower of height 10 m. Its velocity on reaching the ground is _____m s^{-1}
4. In a velocity time graph, the negative slope indicates _____.
5. Define time of ascent and time of descent.
6. A body is falling freely under gravity. The distances covered by it in the first, second and third seconds of its motion are in the ratio of _____.
7. The equations of motion are applicable only when the body moves with _____.
8. A hollow iron ball and a solid iron ball are dropped from the same height. Which ball reaches the ground first?
9. Show that for a body the magnitude of velocity of projection in the upward direction is the same as the magnitude of velocity of the body on reaching the point of projection on its downward motion.
10. The sum of the time of ascent and the time of descent is called _____.
11. What is acceleration due to gravity?
12. A stone is projected up with a velocity of 4.9 m s^{-1} from the top of tower and it reaches the ground after 3 s. Then the height of that tower is _____.
13. Two balls are dropped from height h_1 and h_2 respectively. The ratio of their velocities on reaching the ground is _____.
14. What is the acceleration of a body when it is projected vertically upwards?
15. Two bodies of different masses m_1 and m_2 are dropped from two different heights h_1 and h_2. The ratio of the times taken by the two to reach the ground is _____.
16. If an object is thrown vertically up with a velocity of 19.6 m s^{-1}, it strikes the ground after _____s.
17. Write the equations of motion for a body projected vertically downwards.
18. If a body is projected horizontally from top of a building. In 2 second, it falls vertically through a distance of _____ m. ($g = 10$ m s^{-2})
19. What is time of flight?
20. If two balls are thrown vertically upwards with the same velocity, what is the difference in the maximum height gained by them?

Short answer type questions

21. Deduce an expression for the maximum height reached by a body.
22. Obtain an expression for time of flight.

23. Why does the velocity decrease when the body is projected upwards?

24. Show that the time of ascent is equal to time of descent for a body projected vertically upward.

25. Derive an expression for the velocity when a body dropped from a height reaches the ground.

Essay type questions

26. Two kids standing on different floors of an apartment block drop balls from the balconies. The second kid standing at a lower level drops the ball 2 seconds after the first kid does. Both the balls reach the ground simultaneously, 5 seconds from the time the first ball was dropped. What is the difference between the heights of the two balconies? (Take $g = 10\ m\ s^{-2}$)

27. A ball is dropped from a certain height. If it takes 0.2 s to cross the last 6.2 m before hitting the ground, find the height from which it was dropped. (Take $g = 10\ m\ s^{-2}$)

28. A ball is thrown vertically upwards with a speed of $30\ m\ s^{-1}$. Draw a graph showing the velocity of the ball as a function of time. (Take $g = 10\ m\ s^{-2}$)

29. A man standing at a distance of 6 m from a building 21.6 m high observes a kid slipping from the top floor. With what speed should he run to catch the kid at a height of 2 m from the ground? (Assume speed of the person to be uniform; $g = 10\ m\ s^{-2}$)

30. A stone is dropped from a height of 45 m. What will be the distance travelled by it during the last second of its motion? (Take $g = 10\ m\ s^{-2}$)

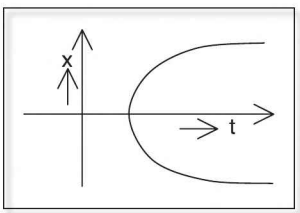
CONCEPT APPLICATION

Concept Application Level—1

Direction for questions 1 to 7: State whether the following statements are true or false.

1. The area under acceleration time graph gives change in velocity.

2. A particle cannot have its displacement (x) versus time (t) graph, as shown in the figure below.
Graph



3. The direction of acceleration due to gravity depends on the direction of motion of a body.

4. A projectile has constant speed in horizontal direction.

5. The displacement of a body has the same direction as that of its average velocity.

6. A body can have constant velocity and still have varying speed.

7. The velocity of a body can change even if its acceleration is zero.

Direction for questions 8 to 14: Fill in the blanks.

8. For a projectile, the magnitude of acceleration in the horizontal direction is _____ and that in the vertical direction is _____.

9. If a particle moves from *A* to *B* along a straight line with velocity v_1 and then *B* to *A* with velocity v_2 its average velocity is _____.

10. The time taken by a body which is projected horizontally from a height 'h' from the ground to reach the ground is _____.

11. Two balls are dropped from heights h_1 and h_2 respectively. The ratio of their time of descents is _____.

12. A body is projected vertically upward with velocity u m s^{-1}. When it returns to the point of projection, its velocity is _____.

13. The _____ of motion changes during the motion of a projectile.

14. During upward motion of a body projected vertically upward, the angle between velocity and 'g' is _____.

Direction for question 15: Match the entries in column A with appropriate ones from column B.

15.

A.	Constant speed with varying velocity	( )	a.	A body dropped from a certain height
B.	Zero displacement and finite distance.	( )	b.	Body on reaching point of projection.
C.	Final velocity (v) = 0, finite acceleration	( )	c.	Motion of a car on a busy road.
D.	Non-zero vertical velocity and non-zero acceleration	( )	d.	Parachute descending down from an aeroplane.
E.	Vertical component of velocity is zero.	( )	e.	Uniform circular motion
F.	Non uniform velocity	( )	f.	At maximum height of a body under projectile motion.
G.	Zero initial velocity	( )	g.	At any intermediate point of a body dropped from certain height
H.	Retarded motion	( )	h.	At highest point of body projected vertically up.

Direction for questions 16 to 30: For each of the questions, four choices have been provided. Select the correct alternative.

16. A body projected vertically up has displacement of 16 m in the first n seconds while it was moving up. Its magnitude of displacement in the last n second while falling down is

(1) 8 m (2) 4 m (3) 16 m (4) 2 m

17. The acceleration of a moving body can be found from

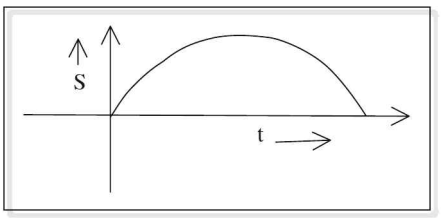
(1) slope of velocity time graph.
(2) area under velocity time graph.
(3) slope of distance time graph.
(4) area under distance time graph.

18. Two bodies are projected from the ground with the same speed. If the angles of their projection from the ground are 45° and 15° respectively, the ratio of their range is

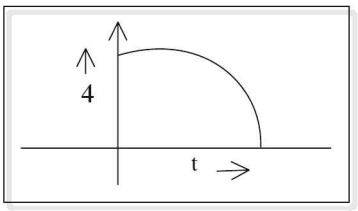
(1) 1 : 2 (2) 2 : 1 (3) $\sqrt{3} : 2$ (4) $1 : \sqrt{2}$

19. For a body thrown vertically up and falling back to the point of projection, the displacement–time graph is _____.

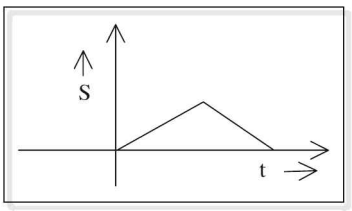
(1)



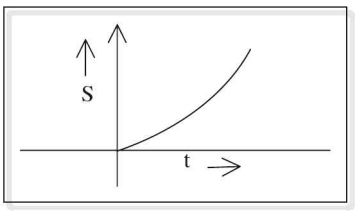
(2)



(3)



(4)

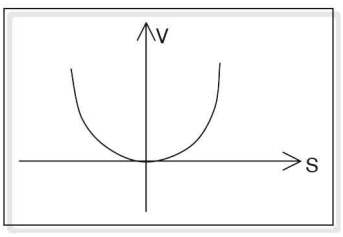


20. A stone dropped from the top of a tower travels 4.9 m in the last second, then the velocity of the stone on reaching the ground is

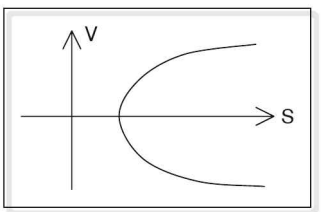
(1) 19.6 m s^{-1} (2) 9.8 m s^{-1} (3) 4.9 m s^{-1} (4) 29.4 m s^{-1}

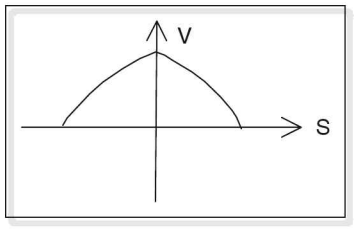
21. The graph of velocity versus displacement of a body which is thrown vertically up and falling back to the point of projection is _____.

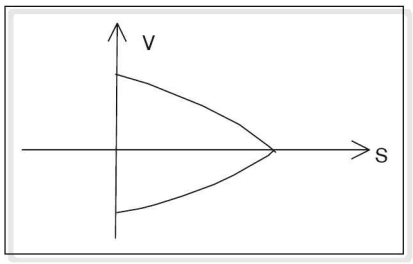
(1)



(2)



(3) 

(4) 

2. When a body is projected vertically up from the ground its velocity is reduced to $\frac{1}{4}$ th of its velocity at ground. Then the maximum height reached by the body is

(1) $\frac{1}{4}$ (2) $\frac{4}{3}$ (3) $\frac{2}{3}$ (4) $\frac{3}{5}$

23. If a body is projected with certain velocity making an angle 30° with the horizontal, then
(1) its horizontal velocity remains constant.
(2) its vertical velocity changes.
(3) on falling to the ground, its vertical displacement is zero.
(4) All of the above

24. A freely falling body from rest acquires velocity V in falling through a distance h. If here after the body falls through a further distance h velocity acquired by it is

(1) 2V (2) V (3) $\sqrt{2}$V (4) 4V

25. Two cars A and B move such that car A moving with a uniform velocity of 15 m s^{-1} over takes car B moving from rest with an acceleration of 3 m s^{-2}. After how much time do they meet again?

(1) 5 s (2) 15 s (3) $\sqrt{3}$ s (4) 10 s

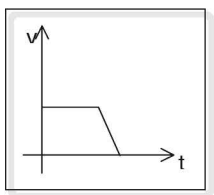
26. A stone dropped from the top of a tower travels 15 m in the last second of its motion. If g = 10 m s^{-2} then the time of fall is

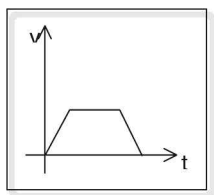
(1) 2 s (2) 2.5 s (3) 5 s (4) 3 s

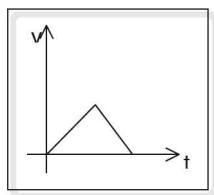
27. A body is projected vertically upward with certain velocity. The magnitude of its displacement in the last second of its upward motion is _____.

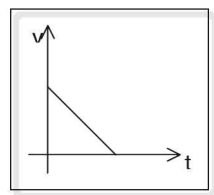
(1) 2 g (2) $\frac{g}{3}$ (3) $\frac{g}{2}$ (4) $\frac{3g}{2}$

28. In the following v-t graphs, identify the graph that represents a body moving with uniform velocity and then with uniform retardation until it stops.

(1) 

(2) 

(3) 

(4) 

29. Three bodies A, B and C are thrown simultaneously with the same speed from the top of a building. A is thrown vertically upward, B horizontally and C, vertically downward. All the three, on reaching the ground,

(1) have equal velocities. (2) have equal vertical displacements.

(3) take equal time to reach the ground. (4) Both (1) and (2)

30. An ant moves along the sides of a square room of length 4 m starting from A and reaches the opposite corner C by travelling from A to B and from B to C. If the time taken is 2 s, the average velocity of the particle is

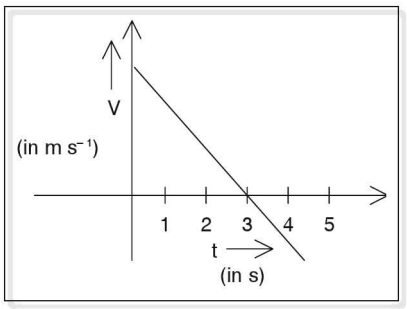
(1) 4 m s^{-1} (2) $2\sqrt{2}$ m s^{-1} (3) 2 m s^{-1} (4) $4\sqrt{2}$ m s^{-1}

Concept Application Level—2

31. A lift ascends from rest with uniform acceleration of 4 m s^{-2}, then it moves with uniform velocity and finally comes to rest with a uniform retardation of 4 m s^{-2}. If the total distance covered during ascending by the lift is 28 m and the total time for ascending 8 s respectively, then find the time for which the lift moves with uniform velocity. Also find its uniform velocity.

32. The following graph represents the velocity-time graph of a body projected vertically upward under gravity.

Find the maximum height attained by the body. (Take g = 10 m s^{-2})



33. A body moves with a uniform speed in a circular path of radius r. What is the displacement of the body on covering $\frac{1}{4}$th of the revolution?

34. A body is projected vertically upward. Show that the

(i) distance covered by it in the first 'n' seconds of ascent is equal to the distance covered by it in the first 'n' seconds of its descent.

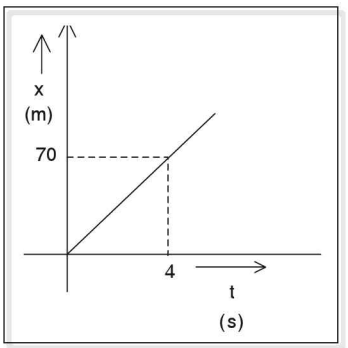
(ii) the distance covered by it in the last 'n' seconds of ascent is equal to the distance covered by it in the first 'n' seconds of its descent.

35. A missile is launched from the ground making 45° with the horizontal to hit a target at a horizontal distance of 300 km. If it is required to hit a target at a horizontal distance of 675 km launched at same angle with horizontal, find the percentage change in its velocity of projection.

36. A freely falling body acquires a velocity 'v' m s^{-1} in falling through a distance of 80 m. How much further distance should it fall, so as to acquire a velocity of '2v' m s^{-1}? (Take g = 10 m s^{-2})

37. A body is dropped from certain height H. If the ratio of the distances traveled by it in (n – 3) seconds to (n – 3)rd second is 4 : 3, find H. (Take g = 10 m s^{-2})

38. The variation of horizontal displacement (x) of a projectile with time (t) is as shown in the figure below.
If the angle of projection is 60°, find the velocity of projection.

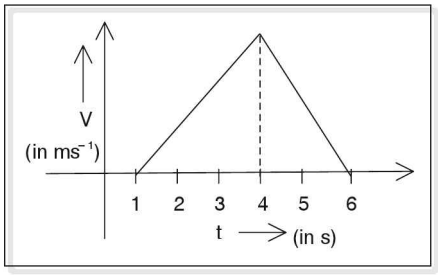


39. A juggler sets 4 balls in motion in succession to a height of 19.6 m from his hand. What is the velocity of projection of each ball? Once set in motion, what are the displacements of the other three balls when the 2nd ball just leaves his hand?

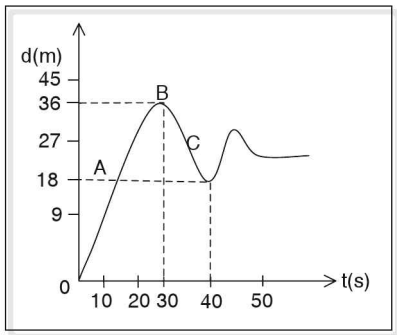
40. A person watching through the window of an apartment sees a ball that rises vertically up and then vertically down for a total time of 0.5 s. If the height of the window is 2 m, find the maximum height above the window reached by the ball. ($g = 10\ m\ s^{-2}$).

41. A freely falling body crosses points P, Q and R with velocities V, 2V and 3V respectively. Find the ratio of the distances PQ to QR.

42. The velocity-time graph of a car is as shown in the figure below.
Find the acceleration and deceleration of the car if its total displacement is 90 m.



43. In the displacement versus time graph below find the average velocity in the time interval of 0 to 30 second. Comment on the signs of velocities at points *A*, *B* and *C*.



44. Are the equations of motion applicable to bodies projected vertically up with any velocity, say 8 km s^{-1}, for determining the maximum height?

45. A stone is thrown horizontally from a tower with certain velocity and at the same time, another stone is dropped freely from the same height. Which stone reaches earlier? Explain.

Concept Application Level—3

46. The displacement (x) variation of a particle with time (t) is given by expression $x = -15t^2 + 20t + 30$ find the position of the body at $t = 0$ s and $t = 6$ s. Is this particle moving with uniform acceleration? Explain.

47. The horizontal distance traveled by a bullet fired at 45° is 360 m. If it is fired from a lorry moving in the direction of bullet with the uniform velocity 18 km hr^{-1} and with same elevation, what is the horizontal distance traveled by the bullet? (Take $g = 10$ m s^{-2}).

48. The graph of vertical displacement (y) versus horizontal displacement (x) of a projectile is as shown below.

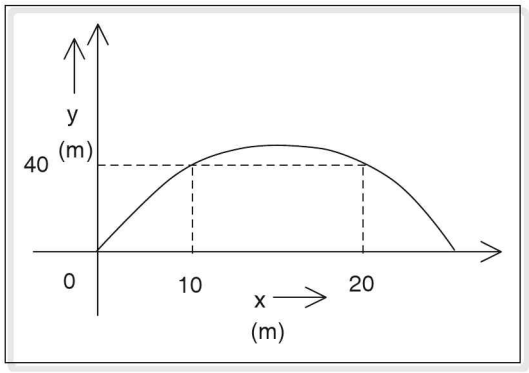


Figure 1.8

From the values given in the graph, find the time at which the projectile has the displacements, as indicated in the graph.

49. A body moving linearly with uniform acceleration covers distances p, q, r and s after successive intervals of time t s. Find the ratio of $\frac{(s-p)}{(r-q)}$.

50. A small steel ball is dropped from a height of 2.5 m into a tall glycerin jar. It hits the surface of glycerin with certain velocity and sinks to the bottom by losing 20% of its velocity at the surface. If the time taken to reach the bottom after it is dropped is 2 s and velocity in glycerin is lost at constant rate, then find the average velocity of the ball over the whole journey.

Very short answer type questions

1. three dimensional

2. The point at which the velocity of a vertically projected body becomes zero.

3. 15

4. deceleration

5. (i) Time taken by a body projected vertically up to reach maximum height.
(ii) Time taken by a freely falling body to reach ground.

6. 1 : 3 : 5

7. uniform acceleration

8. Both reach at a time

10. time of flight

11. The uniform acceleration experienced by a freely falling body.

12. 29.4 m

13. $\sqrt{\frac{h_1}{h_2}}$

14. – 9.8 m s^{-2}

15. $\sqrt{h_1} : \sqrt{h_2}$

16. 4

17. $v = u - gt$, $s = ut - \frac{1}{2}gt^2$, $v^2 - u^2 = -2gh$

18. 20

19. The time taken by a body projected vertically up to reach ground.

20. No difference

Short answer type questions

21. (i) Let the body be projected with an initial velocity u on reaching the maximum height (H) the velocity of the body becomes zero.
(ii) Use the formula $v^2 - u^2 = 2as$.
$s = H, v = 0; a = +g; u = u$
(iii) $v^2 - u^2 + 2as$

$\Rightarrow v^2 = u^2 + 2gh \Rightarrow H = \frac{u^2}{2g}$

22. (i) time of ascent(t_a) = time of descent (t_d) = $\frac{u}{g}$
(ii) Hence, time of flight (T) = $t_a + t_d$.

23. The acceleration due to gravity (g) is acting downwards.

24. (i) Time of ascent, $t_a = u/g$
(ii) Time of descent $t_d = v/g$
(iii) Final velocity (v) = initial velocity (u)

25. (i) Let the body be dropped from a height (h) with an initial velocity = u. The acceleration due to gravity (g) is acting downwards.
(ii) Let the final velocity of the body on reaching the ground be = v. Substitute the data in $v^2 - u^2 = 2as$
(iii) Take a = + g; s = h

Essay type questions

26. $h = \frac{1}{2}gt^2$

Ans: 80 m

27 $s = ut + ½ gt^2$

$u = v \Rightarrow$ velocity at beginning of 0.2 s = velocity after it travelled (h – 6.2) m

Ans: 51.2 m

28. (i) $v = u - gt$

(ii)

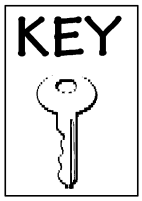
v (m s^{-1})					
t (s)	0	1	2	3	4

29. $h = \frac{1}{2}gt^2, s = ut$

Ans: 3 m s^{-1}

30. $h = \frac{1}{2}gt^2, h_n = g\left(n - \frac{1}{2}\right)$

Ans: 25 m



Concept Application Level—1

True or false

1. True
2. True
3. False
4. True
5. True
6. False
7. False

Fill in the blanks

8. zero, g
9. 0
10. $\sqrt{\frac{2h}{g}}$
11. $\sqrt{h_1} : \sqrt{h_2}$
12. $-u$ m s^{-1}
13. direction
14. 180°

Match the following

15. A : e
    B : b
    C : h
    D : g
    E : f
    F : c
    G : a
    H : d

Multiple choice questions

16. Choice (3)
17. Choice (1)
18. Choice (2)
19. Choice (1)
20. Choice (2)
21. Choice (4)
22. Choice (2)
23. Choice (4)
24. Choice (3)
25. Choice (4)
26. Choice (1)
27. Choice (3)
28. Choice (1)
29. Choice (2)
30. Choice (2)

Concept Application Level—2,3

Key points

31. (i) Find the time for acceleration, retardation and uniform velocity by using equations of motion.
    (ii) Find the uniform velocity.
    (iii) 4 m s^{-1}

32. (i) Find the velocity at t = 3 s. This gives the velocity v.
    (ii) From $v = u - gt$, find the value of u by substituting the value of v, g and t.
    (iii) Find the maximum height by substituting the values of u, v and g in equation $v^2 = u^2 - 2gh$.
    (iv) 45 m

33. (i) Find the relation between the number of revolutions and the corresponding angle described.
    (ii) Using the definition of displacement, identify the initial and final points
    (iii) Using trigonometric ratios find the displacement for $\frac{1}{4}$th of the revolution.
    (iv) $\sqrt{2}\,r$

34. (i) Find the velocity v_1 of a body projected upwards after t = $(t_1 - n)$ seconds t_1 = time of ascent = $\frac{u_1}{g}$ using equation v_1 = $u_1 - gt$
    (ii) v_1 is the initial velocity in the beginning of last n seconds

(iii) Find s_1 by substituting v = 0 and u = v_1 in equation $v^2 = u^2 - 2gs_1$

(iv) Find the displacement s_2 for the body falling from rest for n seconds using equation $s = ut + \frac{1}{2} gt^2$

(v) Compare s_1 and s_2

35. (i) Consider the relation between the horizontal distance and the initial velocity.

(ii) 50% increase

36. (i) Let h_1 be the displacement during which the velocity changes from v to 2v. Find h_1 by substituting u = v, v = 2v in equation $v^2 = u^2 + 2gh_1$

(ii) Find v^2 for a body dropped from a height of 80 m from the equation $v^2 = u^2 + 2gh$

(iii) Substitute the value of v^2 in equation obtained in 1 and find h_1

(iv) 240 m

37. (i) Obtain the equation for the distance travelled in (n – 3) seconds using equation $s_1 = \frac{1}{2} gt^2$

(ii) Obtain equation for the distance travelled in (n – 3)rd second using equation $s_n = u + \frac{g}{2} (2n - 1)$

(iii) Divide equation obtained from 1 by equation obtained in 2

(iv) Equate the ratio to 4 : 3

(v) Find the value of n

(vi) Substitute the value of n in equation $s = \frac{1}{2} gt^2$ (t = n) and find s which gives the value of H.

(vii) 125 m

38. (i) Find the value of x and t in the graph

(ii) Find the value of velocity of projection, u, using the equation $x = u \cos \theta \, . \, t$

(iii) 35 m s^{-1}

39. (i) Find h for all the balls using u and t.

(ii) 14.7↑, 19.6, 14.7↓

40. (i) Let h be height above the window by which the ball rises and t_1 be the time.

(ii) The height of the ball from its fall is given by $h = \frac{1}{2} gt^2_1$.

(iii) Time taken for the ball to cross the window = (t_1 + ¼) seconds and distance covered to cross the window = h + 2.

(iv) Substitute s = h + 2, u = 0 and time $t = t_1 + \frac{1}{4}$ in the equation $s = ut + \frac{1}{2} gt^2$.

(v) Equate R.H.S. of 2 and 4 and find t.

(vi) Substitute the value of t in 2 and find h.

(vii) 160 m

41. (i) Consider the equation of motion $v^2 = u^2 + 2gs$ for a freely falling body.

(ii) Velocity at P is the initial velocity and velocity at Q is the final velocity for the displacement PQ.

(iii) Substitute initial and final velocity in equation given in 1 and find S_1.

(iv) Similarly velocity at Q and velocity at R are initial and final velocity for the displacement QR = S_2.

(v) Substitute the initial velocity and final velocity for displacement QR in equation as given in 1 and then find S_2.

(vi) Find the ratio S_1 and S_2.

(vii) 3 : 5

42. (i) Area under v - t graph gives the total displacement.

(ii) Area of triangle = ½ × base × height

(iii) Substitute the base and area of the triangle and find the height of the triangle.

(iv) The height of the triangle gives the maximum velocity attained by the car.

(v) Find the time taken to attain maximum velocity, t_1 and the time taken to come to rest, t_2 from maximum velocity from the graph.

(vi) Find acceleration using equation $v = at_1$ and deceleration using equation $v = u + at_2$ (v = 0 and substitute u, the maximum velocity of the car from 4)

(vii) Acceleration = 12 m s^{-2}

Deceleration = 18 m s^{-2}

43. (i) Average velocity $= \dfrac{\text{Total displacement}}{\text{total time}}$

(ii) 1.2 m s^{-1}

44. (i) When can the equations of motion be applied?

(ii) As the height of the body above the earth's surface increases, does acceleration due to gravity remain constant?

45. (i) Apply the conditions for a body thrown horizontally with certain velocity as well as vertical displacement.

(ii) Finding the time of descent

(iii) Compare the times of descent of the two stones

46. (i) Change in position is displacement

(ii) Change in displacement in 1 and is velocity

(iii) Change in velocity in 1 and is acceleration.

47. (i) Using the formula for horizontal distance calculate the initial velocity to find the time of flight. Using this time of flight calculate the additional the horizontal distance traveled by bullet fired from the lorry due to the motion of the lorry.

(ii) 402.42 m

48. (i) Find the value of t_1 in terms of u by substituting x = 10 in the equation $x = u \cos \theta . t_1$

(ii) Similarily, find the value of t_2 by substituting x = 20 m in the above equation

(iii) Divide 1 and 2 to obtain the ratio t_1 and t_2.

(iv) Divide vertical displacement, $y = u \sin \theta \times t - \frac{1}{2} gt^2$ by g/2.

(v) Write the above equation in a quadratic form with variable t, i.e., in the form of $At^2 + Bt + C = 0$

(vi) The roots of the above equation are t_1 and t_2.

(vii) The product of roots, ie $t_1 t_2 = C/A$.

(viii) Substitute the values of C and A and find $t_1 t_2$.

(ix) Solve equation obtained from 3 and 8 which gives the value of t_1 and t_2, given y = 40 m corresponding to x = 10 m and x = 20 m.

(x) 2 s and 4 s

49. (i) Find p, q, r, s using $s = ut + \dfrac{1}{2} at^2$

(ii) Finding the ratio of $\dfrac{(s-p)}{(r-q)}$.

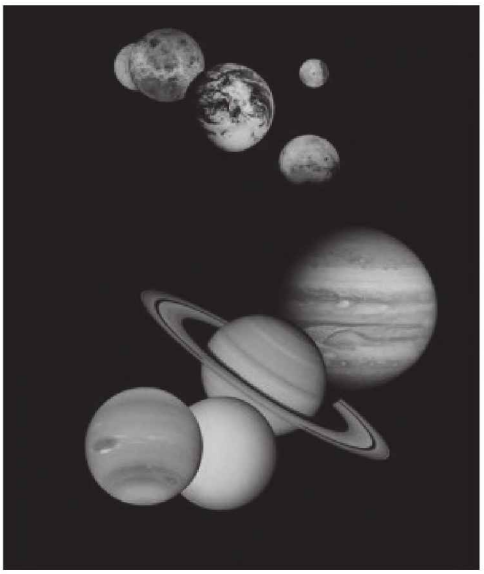
(iii) 3 : 1

50. (i) Find the velocity with which the ball moves in glycerine. Then find time the for which it travels.

(ii) Find the distance traveled in glycerin.

(iii) Find average velocity.

(iv) 5.3 m s^{-1}



2

Dynamics

INTRODUCTION

Dynamics is an important branch of physics which deals with the effect of forces on bodies. Newton's laws of motion which are of great importance relate the type of motion of the body that results from the action of the forces. To understand Newton's laws, we need to define and understand the importance of a physical quantity called momentum.

Linear momentum

Activity 1

Drop a cricket ball on the pan of a beam balance from a certain height. Note down how far the needle swings. Now remove the cricket ball and drop a tennis ball on to the pan from the same height as that in the case of the cricket ball and note how far the needle swings in this case. It is found that the needle swings to a greater extent when the cricket ball is dropped on the pan of the balance.

Activity 2

Similar to activity 1, drop a cricket ball on the pan of the beam balance from different heights separately and note down the extent of swings of the needle of the balance in each case.

It is found that the swing of the needle is greater when the cricket ball falls from a greater height. Greater the height of the ball from which it is dropped, higher is its velocity on hitting the pan of the balance and greater is the swing of the needle.

Thus from the above activity it is found that greater the mass or velocity, greater is the impact of the ball on the pan of the balance.

This combined effect of mass and velocity gives rise to a physical quantity known as linear momentum. Linear momentum of a particle is denoted by 'p' and that of a system of particles is denoted by P. Mathematically linear momentum can be written as linear momentum = mass × velocity $\Rightarrow P = m \times v$

Since the product of a scalar and vector is always a vector, momentum is a vector quantity and the direction of linear momentum is the same as that of velocity.

Units of momentum

Unit of momentum = (unit of mass) (unit of velocity)

$\Rightarrow$ S.I. unit of P = kg m s^{-1} or N s and

C.G.S. unit of P = g cm s^{-1} or dyne second

Dimensional formula

$P = m \times v \Rightarrow$ dimensional formula of P = (dimensional formula of mass) × (dimensional formula of velocity) = $[M\ L\ T^{-1}]$

Thus dimensional formula of P is $[M^1L^1T^{-1}]$.

Change of linear momentum

Linear momentum of a body changes in two cases. They are as follows.

(a) When there is a change in the velocity of the body, its mass remaining constant:

Change in linear momentum $\Delta P = m\ \Delta v$, Δv being the change in velocity

The above equation is valid only if $v << c$ or v is less than 10^6 m s^{-1}, where c is the velocity of light in vacuum.

(b) When there is a change in mass: When sub-atomic particles move with a velocity greater than the velocity of light, there is an increase in mass. Thus the change in momentum is better represented by $\Delta p = \Delta (mv)$

☛ *Example*

Calculate the velocity of a body of a mass 2 kg whose linear momentum is 5 N s.

Solution

Linear momentum of the body = P = mass × velocity

Mass, m = 2 kg

Momentum, P = 5 N s

Velocity, v = ?

$P = mv$

$5 = 2 \times v$

$v = \frac{5}{2} = 2.5$ m s^{-1}

☛ *Example*

A car of mass 1000 kg is moving uniformly at 10 m s^{-1}. If the engine of the car develops an extra linear momentum of 1000 kg m s^{-1}, calculate the new velocity with which the car runs.

Solution

Initial momentum of the car $P_1 = m \times v_1$

$P_1 = 1000 \times 10$

$P_1 = 10{,}000$ N s or kg m s^{-1}

Increase in momentum of the car = 1000 N s

Thus new momentum of the car $P_2 = P_1$ + increase in momentum

$P_2 = 10{,}000 + 1000$

$P_2 = 11000$ N s

$P_2 = m \times v_2$ Where v_2 is the new velocity of the car

$\therefore 11000 = 1000 \times v_2$

$$v_2 = \frac{11000}{1000} = 11 \text{ m s}^{-1}$$

$v_2 = 11$ m s^{-1}

Newton's First Law of Motion

Statement

Every body remains in a state of rest or of uniform motion along a straight line unless it is acted upon by an external force.

Thus from the above statement acceleration is zero if and only if net force acting on the body is zero $\Rightarrow \vec{a} = 0$ if $\vec{F} = 0$

This law states that every body resists the change in its state of rest or that of uniform motion. This property is known as inertia. Thus Newton's first law is also called the law of inertia. If no force acts on a body, there always exists a reference frame in which the body has no acceleration. Such a reference frame is called an inertial reference frame.

Newton's Second Law of Motion

Statement

Rate of change of linear momentum is directly proportional to the net force acting on the body and takes place in the direction of the force.

Derivation

Consider a body of mass 'm' moving with initial velocity 'u'. Let a force F act on it for 't' seconds, so that the velocity changes to v.

Initial linear momentum of the body $P_1 = mu$

Final linear momentum of the body $P_2 = mv$

Change in linear momentum in t second $= P_2 - P_1$

$= mv - mu$

Rate of change of linear momentum $= \frac{mv - mu}{t}$

According to Newton's second law of motion $F \propto \frac{mv - mu}{t}$

$F \propto \frac{m(v - u)}{t}$

From equations of motion $\frac{(v - u)}{t} = a$, the acceleration of the body

$\therefore F \propto ma \Rightarrow F = kma$ where k is constant of proportionality

If a force acting on a body of unit mass produces in it unit acceleration, then the force acting on the body is one unit.

Thus $F = kma \Rightarrow$ (unit force) = k (unit mass) (unit acceleration)

$1 = k \times 1 \times 1$

$\therefore k = 1$

Thus $F = ma$

Alternate method

Let ΔP be the change in the linear momentum of a body of mass 'm' due to a force F acting on the body for a small interval of time Δt.

Then, rate change of linear momentum $= \frac{\Delta p}{\Delta t}$

$\frac{\Delta p}{\Delta t} = \frac{\Delta(mv)}{\Delta t}$ --- (1) ($\because P = mv$)

or

$\frac{\Delta p}{\Delta t} = m\frac{\Delta v}{\Delta t}$ --- (2) provided the mass of the body is constant.

From the second law of motion we get $F \propto \frac{\Delta p}{\Delta t}$ or $F \propto m\frac{\Delta v}{\Delta t}$

But $\frac{\Delta v}{\Delta t}$ is the rate of the change of velocity which is equal to acceleration

$\therefore F \propto ma \quad F = kma$

Unit force is that force which, acting on a body of unit mass produces in it unit acceleration. Then k would be equal to 1, i.e., k = 1 Thus F = ma

Thus the net force acting on the body produces acceleration in it.

This law is the fundamental law of nature. Mathematically force is the product of mass and acceleration.

When we hammer a nail to a wall we apply a large force in a short interval of time. Thus there is a change in the momentum of the hammer. This change in momentum is equal to the product of force and time for which the force is applied. This product is called impulse.

Thus,

impulse = force × time.

Unit of impulse is N s or kg m s^{-1}

Dimensional formula of impulse = [Force] × [time]

$= [M^1L^1T^{-2}] \times [T^1]$

$= [M^1L^1T^{-1}]$

According to Newton's second law, acceleration produced in a given mass varies directly to the force applied.

$a \propto F$ if mass is constant

Thus the graph of acceleration versus force is a straight line.

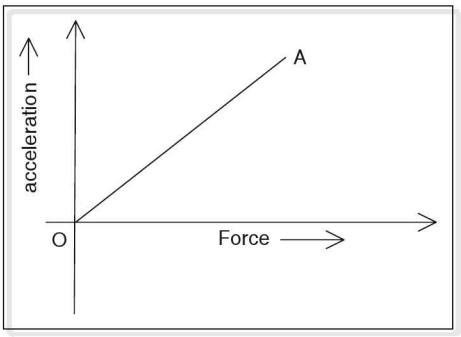


Figure 2.1

From the second law, acceleration produced by a given force is inversely proportional to the mass $a \propto \frac{1}{m}$ provided force is constant.

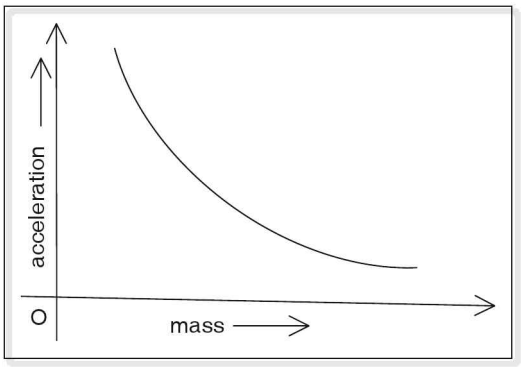


Figure 2.2

Units of force

Absolute unit of force

If m = 1 unit and a = 1unit then F = 1 unit

Absolute unit of force is that force acting on a body of unit mass produces unit acceleration.

S.I. unit of force

The S.I. unit of force is newton (N)

Definition of newton (N): If a force acting on a body of mass 1 kg produces in it an acceleration of 1 m s^{-2}, then the force is said to be 1 newton.

Thus 1 newton = 1 kg × 1 m s^{-2}.

C.G.S. unit of force

The C.G.S. unit of force is dyne

Definition of dyne: One dyne is the force that acts on a body of mass 1 g and produces in it an acceleration of 1 cm s^{-2}.

$\therefore$ 1 dyne = 1 gram × 1 cm s^{-2}.

Relation between newton and dyne

1 newton = 1 kg × 1 m s^{-2}

1 newton = 1000 g × 100 cm s^{-2}

1 newton = $10^3 \times 10^2$ g cm s^{-2}

$\therefore$ 1 newton = 10^5 dynes ($\because$ 1 dyne = 1 g cm s^{-2})

Gravitational unit of force

If a force acts on a body of unit mass and produces in it an acceleration equal to the acceleration due to gravity of earth, then the force is said to be gravitational unit of force.

In S.I. system gravitational unit of force is kilogram force, represented by kg_f.
$1\ kg_f = 1\ kg \times 9{\cdot}8\ m\ s^{-2} = 9{\cdot}8\ N$
In C.G.S. system one gravitational unit of force is gram –force, represented by g_f.
$1\ g_f = 1\ g \times 980\ cm\ s^{-2} = 980$ dynes.

Newton's Third Law of Motion

Statement: For every action there is an equal and opposite reaction.
If a body A exerts a force $\vec{F}$ on body B, then B exerts a force $-\vec{F}$ on A.

☛ *Example*

A force of 500 N acts on a body of mass 1000 kg and the body is brought to rest within a distance of 64 m. Find the initial velocity and the time taken by the body to come to rest.

Solution

Force = F = 500 N

Mass of the body, m = 1000 kg

From Newton's second law of motion,

F = ma, where F, m and a are the force, mass and acceleration respectively,

$$\therefore a = \frac{F}{m} = \frac{500}{1000} = 0.5 \text{ m s}^{-2}$$

To find initial velocity, consider the equation of motion $v^2 = u^2 + 2as$

Since the body is retarding $a = -0{\cdot}5 \text{ m s}^{-2}$ and since it comes to rest $v = 0$

$\therefore 0 = u^2 + (2 \times (-0{\cdot}5) \times 64)$

$0 = u^2 - (1 \times 64)$

$\Rightarrow u^2 = 64 \Rightarrow u = \sqrt{64} = 8 \text{ m s}^{-1}$

To find the time taken, consider the equation of motion

$v = u + at$

$0 = 8 - 0{\cdot}5 \times t$

$$t = \frac{8}{0.5} = 16 \text{ s}$$

$\therefore u = 8 \text{ m s}^{-1}$, $t = 16$ s and $a = -0{\cdot}5 \text{ m s}^{-2}$

☛ *Example*

A ball of mass 50 g at rest is hit by a bat and the ball covers 400 m in 2 s. If the ball was in contact with the bat for 0·1 s, find the magnitude of force applied on it. Assume that no other force acts on the ball after it is hit by the bat.

Solution

We are given that the distance covered by ball is 400 m

$\therefore s = 400$ m

And, time in which the ball covers a distance of 400 m = 2 s

$$\therefore \text{Final velocity of the ball} = \frac{400}{2} = 200 \text{ m s}^{-1}$$

This is the velocity of the ball after being hit. Initial velocity of the ball i.e., the velocity of the ball before being hit by the bat = 0

Time for which the force acts = 0·1 s

From equation of motion

$v = u + at$

$200 = 0 + a \times 0{\cdot}1$

$$a = \frac{200}{0.1} = 2000 \text{ m s}^{-3}$$

From Newton's second law F = ma.

Mass of the ball $m = 50 \times 10^{-3}$ kg

$\Rightarrow F = 50 \times 10^{-3} \times 2000 = 100$ N

Force acting on the ball = 100 N

Work

When a force acts on a body causing the body to be displaced in the direction of the force, then work is said to be done by the force.

When we try to push a wall the wall does not move even a bit. i.e., there is no displacement and no work is done. Thus, work is done only if a force acts on a body and if the body is displaced in the direction of the force.

Mathematical expression for work

Work done is directly proportional to the force applied on the body

$W \propto \vec{F}$ --- (1)

Work done is directly proportional to displacement of the body

$W \propto \vec{s}$ --- (2)

From equations (1) and (2), and keeping in mind that work is a scalar quantity.

$$W \propto \vec{F} \times \vec{s}$$

$$W = k \vec{F} \times \vec{s}$$

where k is proportionality constant. By defining the unit of work, k is taken as one.

Unit work

When a unit force acting an a body displaces it through one unit, then the work done is one unit.

Thus, k = 1

If F = 1 unit and s = 1 unit then W = 1 unit and k = 1,

If the direction of applied force on a body makes an angle θ with the direction of its displacement then the component of the force along the direction of displacement is F cosθ, and the work done is

$W = F \cos\theta \times s$ or

$W = F \times s \cos\theta$

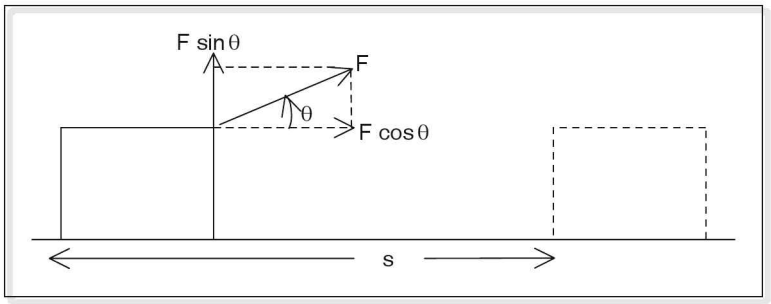


Figure 2.3

Thus the work is said to done when a force or its component acting on a body displaces it in its direction.

$W = F \times s \cos\theta$

or $W = \vec{F} \cdot \vec{s}$

Where $\vec{F} \cdot \vec{s} = Fs\cos\theta$

Thus, work is the dot product of force and displacement since work is a scalar quantity.
(Dot product of two vectors is a scalar)

Special cases

Case (i)

If the displacement is in the direction of force, then θ = 0° and cos0° = 1

$W = Fs\cos\theta$

$W = Fs$

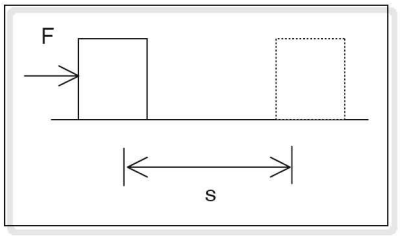


Figure 2.4

☛ *Example*

A horse pulling a cart. Force exerted by the horse on the cart and the displacement are in the same direction.

Case (ii)

If the displacement vector is at right angles to the applied force, then θ = 90°.

$W = Fs\cos90°$

$W = 0$ (zero) $\quad \because \cos 90° = 0$

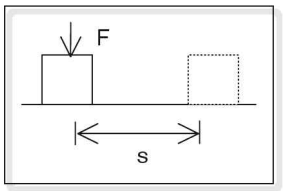


Figure 2.5

☛ *Example*

A porter walking with a load on his head. Work done by the porter is equal to zero.

Case (iii)

If the displacement is along a direction opposite to that of the force, $\theta = 180°$

$W = Fs \cos 180°$ ($\because \cos 180° = -1$)

$W = -Fs$

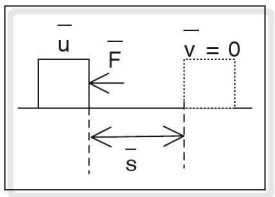


Figure 2.6

Work done by the force is negative in this case. This is the case where the force opposes the motion, e.g., the truck driver applies brakes to his moving vehicle.

If the force varies with displacement then the work done is obtained by plotting a graph. Force is plotted along the Y axis and displacement along the X-axis. The area enclosed by the curve gives the work done. If a constant force acts on the body then the graph is a straight line parallel to X-axis. Work done by constant force is equal to the area of rectangle formed in the F-s graph.

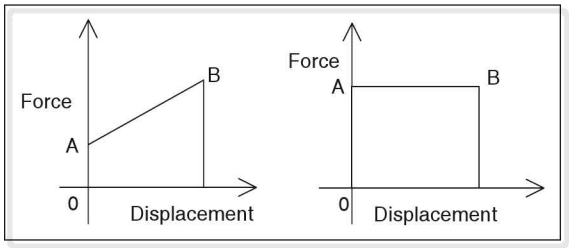


Figure 2.7

Work done by the force of gravity

All bodies are attracted towards the centre of earth due to force of gravity. If a body moves short distance horizontally over the surface of the earth or in a circular path around the earth, like satellites, work done by the force of gravity is zero because the displacement of the body at any instant is perpendicular to the direction of force of gravity. If a body of mass 'm' falls vertically down through a height h, then the work done will be equal to mgh, i.e., the change in the gravitational potential energy.

$\therefore$ Force due to gravity on the body

$F = mg$ and $\theta = 0°$

$\therefore \cos 0° = 1$

If h is the displacement, i.e., the height through which the body falls,

$W = Fs \cos\theta$

$W = (mg)\ (h)\ (\cos 0°) = mgh$

This can also be obtained from work-energy theorem, work done = change in potential energy = mgh

NOTE:

(i) No work is done when $\theta = 90°$ ($\cos 90° = 0$) i.e., when force and displacement are perpendicular to each other.

(ii) No work is done when a body undergoes circular motion.

Units of work

Work = Force × displacement

S.I. unit of work is joule (J).

1 J = 1 N × 1 m

Definition of joule

One joule of work is said to be done when a force of 1 newton acting on a body displaces it through 1m in its direction.

C.G.S. unit of work

C.G.S. unit of work is erg

1 erg = 1 dyne × 1 cm

If a force of 1 dyne acting on a body displaces it through 1 cm in its direction, then the work done in said to be one erg.

Dimensional formula for work

Work = Force × displacement. Thus, [work] = [force] [displacement]

$$= [M^1L^1T^{-2}] \times [L^1]$$

$$= [M^1L^{-2}T^{-2}]$$

Relation between joule and erg

1 J = 1 N × 1 m

1 J = 10^5 dyne × 10^2 cm

1 J = 10^7 dyne cm

But 1 dyne × 1 cm = 1 erg

$\therefore$ 1 J = 10^7 erg

Power

Work done does not depend on time. For example, If a person *A* lifts 10 kg load through 5 ft in 2 minutes and person *B* lifts the same load through the same height in 1 minute, the work done by both the persons is the same, but the rate at which the work is done is not the same. We say that the second person is stronger and has more power. Thus power is defined as rate of work done.

$$\text{Power} = \frac{\text{work}}{\text{time}}$$

$$\text{Power} = \frac{\text{force} \times \text{displacement}}{\text{time}}$$

$\because$ Work = force × displacement

$$P = \frac{F \times s}{t}$$

$$\left(\frac{s}{t} = \text{average velocity}\right)$$

$P = F \times v$

If the displacement makes an angle 'θ' with the force, then the power

$P = F \times v \cos\theta = \vec{F} \cdot \vec{v}$

$P = \vec{F} \cdot \vec{v}$ Thus power is a dot product of force and velocity, and a scalar quantity

Unit of power

S.I. unit of power is watt or J s^{-1}.

watt is denoted by W.

$$\text{power} = \frac{\text{work}}{\text{time}}$$

$$1\ W = \frac{1\ J}{1\ s}$$

Definition of watt

If one joule of work is done in one second then the power is said to be one watt.

$$1\ W = \frac{1\ J}{1\ s} = \frac{1\ N \times 1\ m}{1\ s} = \frac{1\ kg \times 1\ m\ s^{-2} \times 1\ m}{1\ s} = kg\ m^2\ s^{-3}$$

C.G.S. unit of power is erg per second $\frac{1\ erg}{1\ second} = \frac{1\ dyne \times 1\ cm}{1\ second}$

$$= \frac{1 g \times cm\ s^{-2} \times cm}{s}$$

$= 1\ g\ cm^2\ s^{-3}$

Multiple units or bigger units of power are

(a) kW – kilo watt = 10^3 W

(b) MW – mega watt = 10^6 W

(c) GW – giga watt = 10^9 W

(d) hp – Horse power = 746 W

Dimensional formula for power

$$\text{Power} = \frac{\text{work}}{\text{time}}. \text{Hence } [\text{power}] = \frac{[\text{work}]}{[\text{time}]}$$

$$= \frac{[M^1L^2T^{-2}]}{[T]} = [M^1L^2T^{-3}]$$

Energy

A body which is capable of doing work possesses energy. Thus energy is the ability or the capacity to do work.

Unit of energy

The unit of energy is the same as that of work.

Thus S.I. unit of energy is joule (J) and its C.G.S. unit is erg

Energy appears in various forms namely, heat, light, sound, electricity etc.

Heat energy

It is the energy which causes a sensation of hotness or coldness.

☛ *Example*

Energy obtained by burning wood. Heat energy is measured in calories or kilo calories.

Calorie: It is amount of heat energy required to raise the temperature of one gram of water through 1°C. It is denoted by 'cal'.

1 calorie = 4·2 J

kilo calorie: It is the amount of heat energy required to raise the temperature of one kilogram of water through 1°C.

1 kilo calorie = 4·2 kJ

Electrical energy

Energy possessed by an electron in motion is referred to as electrical energy.

Units of electrical energy.

(a) watt hour (W h)

It is the electric energy, and is usually expressed as watt-hour (W h)

1 W h = 1 watt × 1 hour

$$= \frac{1\text{ J}}{1\text{ s}} \times 3600\text{ s}$$

Thus, 1 W h = 3600 J

Kilowatt hour is a multiple of Wh and is a bigger unit of energy

1 kilo watt hour = 10^3 watt hour

= 3600×10^3 J

= 3·6 MJ

(b) Electron volt

It is the work required to move an electron or a proton (in an electric field) through a potential difference of exactly one volt. It is denoted by eV

1 electron volt = charge of 1 electron × 1 volt

$= 1{\cdot}6 \times 10^{-19}\,C \times 1\,V$

$= 1{\cdot}6 \times 10^{-19}\,J$

Light energy

It is the energy possessed by a luminous source which causes sensation of sight.

☛ *Example* light energy emitted by sun.

Sound energy

It is the energy possessed by vibrating bodies which produces sensation of hearing.

☛ *Example* the ringing of a bell.

Magnetic energy

Energy possessed by a magnet is called magnetic energy.

☛ *Example* Energy possessed by motors.

Chemical energy

Energy released or absorbed by atoms or molecules during chemical reaction is called chemical energy.

☛ *Example* Energy released during combustion. In nuclear fusion or fission nuclear mass is converted to energy ($E = mc^2$ where m is the mass, c is velocity of light in air).

Nuclear energy

Energy released during fission or fusion of nucleus is called nuclear energy.

☛ *Example* Atom bomb and hydrogen bomb release energy in the form of heat and radiation. Nuclear energy is used in nuclear reactor to generate electric power.

Mechanical energy

Energy possessed by a body due to its position or motion is known as mechanical energy.

Mechanical energy is classified into potential energy and kinetic energy.

Potential energy

It is the energy possessed by a body due to its position or configuration.

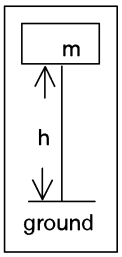
☛ *Example* Wound spring of a clock possesses potential energy. When the spring unwinds, it uses this energy for the movement or rotation of the hands of the clock. This potential energy possessed by the spring is called elastic potential energy. Stretched bow, stretched elastic band and compressed gas molecules possess elastic potential energy.

Elastic potential energy

Elastic potential energy is the potential energy possessed by an elastic body in a deformed state, i.e., it is the amount of work done in deforming the body within its limit of elasticity and is stored in the body as elastic potential energy.

Water stored at a height possesses potential energy due to its position. This energy is used to rotate a turbine which in turn generates electric energy. Energy possessed by water stored at height is called gravitational potential energy. Thus gravitational potential energy is the amount of work done in lifting the body against gravity.

Expression for gravitational potential energy



When a body of mass m is lifted to height 'h' work is done against gravity. This work done is stored as gravitational potential energy in the body.

Work done in lifting body = Force × displacement

W = mass × acceleration due to gravity × displacement

W = mgh

Thus gravitational potential energy = mgh

Note: Surface of the earth is taken as reference for height in calculations of gravitational potential energy. By this assumption gravitational potential energy of a body on the surface of the earth at mean sea level is considered to be zero.

Kinetic energy

Energy possessed by a moving particle is called kinetic energy.

☛ *Example* Wind possesses kinetic energy and this energy is used to rotate the blades of a wind mill. For a body of mass m and velocity v, kinetic energy is defined as

$$KE = \frac{1}{2}mv^2$$

Kinetic energy of a body increases, if the velocity or speed of the body increases. To increase the kinetic energy of a body, net force has to be applied. Thus the work done by this force is equal to the change in the kinetic energy of the body. This is known as work-energy theorem.

Work-kinetic energy theorem

The work done on a body by a force is equal to the change in its kinetic energy when change in speed of the body is the only effect of the force on the body.

Consider a body of mass m moving with a velocity 'v'. Let a retarding force 'F' be applied on it causing it to come to rest after a distance 's'.

From equation of motion

$v^2 = u^2 + 2as$ we get,

$0 = v^2 - 2as$

(negative sign is used because the body is retarding)

$v^2 = 2as$ or

$$\frac{v^2}{2} = as \text{ --- (1)}$$

Work done by the body before coming to rest = Force × displacement

$= ma \times s$

$= mas$ --- (2)

substituting equation (1) in equation (2), we get work done,

$$W = \frac{mv^2}{2}$$

kinetic energy = work done by the retarding force.

$$\therefore \text{ K.E} = \frac{1}{2}mv^2$$

If 'u' is the initial velocity of the body and 'v' is the final velocity then the change in kinetic energy,

$$K_f - K_i = \frac{1}{2}mv^2 - \frac{1}{2}mu^2$$

$$= \frac{1}{2}m(v^2 - u^2) = \frac{1}{2}m\,2as \quad (\because v^2 - u^2 = 2as)$$

= (ma)s = (force) × (displacement) = work done

This is the mathematical expression of work – energy theorem.

Different forms of kinetic energy

The three types of motion that a body may display are:

1. translatory motion
2. rotatory motion
3. vibrational motion

Thus there are three forms of kinetic energy.

(a) Translational kinetic energy: It is the kinetic energy possessed by a body moving along a straight line.

☛ *Example* Energy possessed by a freely falling body.

(b) Rotational kinetic energy: It is the kinetic energy possessed by a rotating body.

☛ *Exmaple* Wheels of a car in motion have rotational kinetic energy along with translational kinetic energy whereas the blades of a rotating fan have rotational kinetic energy only.

(c) Vibrational kinetic energy: It is the energy possessed by a body, which executes to and fro motion about its mean position.

☛ *Example*

(i) Energy possessed by an oscillating pendulum.

(ii) Energy possessed by a vibrating stretched string.

Relationship between kinetic energy and momentum

Kinetic energy of a body, $K = \frac{1}{2}mv^2$ --- (1)

Momentum $P = mv$ --- (2)

From (1) and (2)

$K = \frac{1}{2}\frac{p^2}{m}$

$P^2 = 2m.K$

where K is the kinetic energy

$\therefore P = \sqrt{2mK}$ or $K = \frac{P^2}{2m}$

Law of Conservation of Energy

Statement: Energy can neither be created nor destroyed but it can be transformed from one form to another form i.e., the total energy of a system remains constant.

Examples of interconversion of energy

(a) In solar cells, light energy is converted into electrical energy.

(b) In fans electrical energy is converted in to mechanical energy.

(c) In plants light energy is converted into chemical energy.

(d) In burning fuel like wood etc. chemical energy is converted into light and heat energy.

(e) In microphone sound energy is converted into electrical energy.

Conservation of mechanical energy

Statement: Total mechanical energy of an insolated system always remains constant.

Mechanical energy is the sum of potential energy and kinetic energy. According to law of conservation of energy if potential energy of a body increases its kinetic energy decreases and vice versa such that the total mechanical energy always remains constant.

NOTE: It is always the kinetic energy which is converted into other forms, i.e., during the conversion of mechanical energy into other forms.

Application of law of conservation of energy

(a) In the case of a freely falling body.

Consider a body of mass m at A falling freely under gravity dropped from A at height h. At the point 'A', energy of the body is purely potential because it is momentarily at rest.

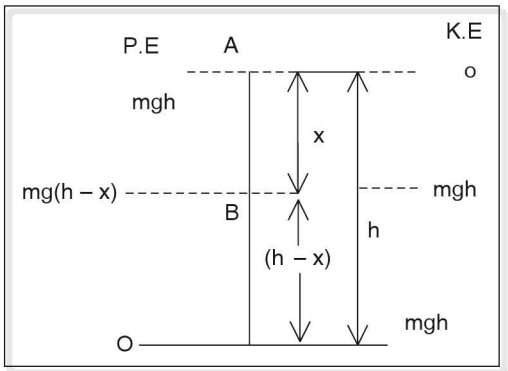


Figure 2.8

∴ Total energy at A = potential energy + kinetic energy

Total energy at A = mgh + 0 --- (1)

As the body falls through a distance x, its potential energy decreases because its height from the ground decreases but the kinetic energy increases as its velocity increases.

∴ Total energy at B = potential energy + kinetic energy = $mg(h - x) + \frac{1}{2}mv_B^2$ --- (2)

where v_B is the velocity at B

Substituting $u = 0, v = v_B, a = g$ and $s = x$ in the equation of motion $v^2 = u^2 + 2as$, we get

$v_B^2 = 2gx$ --- (3)

Substituting equation (3) in (2), we get

Total energy at B $= mg(h - x) + \frac{1}{2}m\ 2gx$

$= mgh - mgx + mgx$

∴ total energy at B = mgh --- (4)

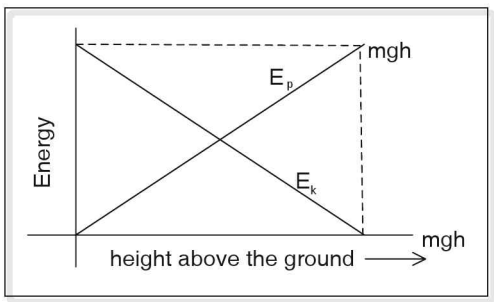


Figure 2.9

At the instant when the body strikes the ground, entire energy of the body is kinetic. Potential energy is equal to zero at the ground.

∴ Total energy at C = potential energy + kinetic energy

Total energy at C = $0 + \frac{1}{2}mv_c^2$ --- (5) where v_C is velocity at C

From equations,

$v_C^2 = u^2 + 2as$

$v_C^2 = 0 + 2gh$

Equation (5) can be written as total energy at C = $\frac{1}{2} m \times 2gh = mgh$

Thus the kinetic energy of the body is equal to initial potential energy. The total energy remains constant and is independent of height of fall.

NOTE: Resistance of air is neglected in verifying the above law.

(b) Conservation of energy in the case of a simple pendulum

O is the point of suspension of the simple pendulum. When the pendulum is made to swing it reaches equal heights on either side of the mean position A. B and C are called the extreme positions of the pendulum.

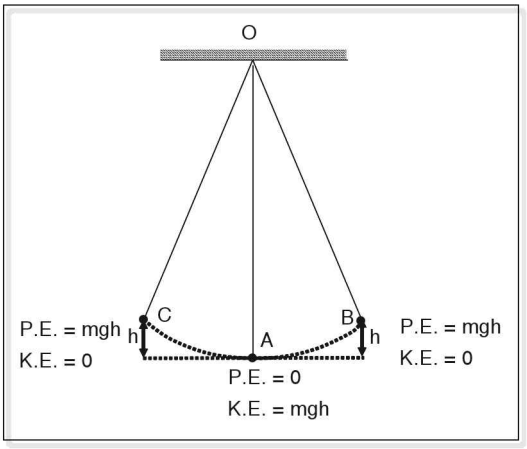


Figure 2.10 Motion of a simple pendulum

At an extreme position, the bob of the pendulum is momentarily at rest, hence its energy is entirely potential

∴ Total energy at extreme position = P.E + K. E. = mgh --- (1)

As it swings from extreme position to mean position, potential energy decreases and its kinetic energy increases such that the total energy remains constant. At the mean position, energy of the bob is purely kinetic and is equal to mgh.

Total energy at mean position = P.E + K.E.

$= 0 + \frac{1}{2} mv^2$ --- (2)

But $v^2 = u^2 + 2as$ ($u = 0, s = h, a = g$) $v^2 = 0 + 2gh$

∴ Total energy at mean position = $0 + \frac{1}{2} m \times 2gh$

$= mgh$ --- (3)

Thus K.E. at the mean position is equal to initial potential energy at the extreme position. As it moves from mean position to the other extreme position, its kinetic energy decreases but potential energy increases as the height increases.

Hence at extreme position, energy of the bob is entirely potential and at any intermediate position, total energy of the bob is the sum of potential energy and kinetic energy.

Thus, total energy of the bob remains constant throughout the swing.

NOTE: Air friction is neglected in the above discussion.

☛ *Example*

What is the work done by a force of 2 N in displacing a body by 2 m in the direction of the force?

Solution

Work done $= \vec{F} \cdot \vec{s}$

$= Fs\cos\theta$

F is the force,

s is the displacement and

θ is the angle between F and s

In the given example,

$\theta = 0°$, F = 2 N, s = 2 m

$\Rightarrow$ Work done $= 2 \times 2 \times \cos 0°$

$= 2 \times 2 \times 1 = 4$ J

☛ *Example*

What is the work done in lifting a mass of 10 kg through a height of 5 m from the ground? ($g = 10$ m s^{-2})

Solution

Work done against the force of gravity $= F \times h = mg \times h$

where m is the mass, g is acceleration due to gravity and 'h' is the height from the ground level.

Thus work done $= mgh = 10 \times 10 \times 5 = 500$ J

☛ *Example*

A bullet of mass 0.02 kg moving with a velocity of 30 m s^{-1} is brought to rest by a wooden plank. Calculate the work done on the bullet before it comes to rest.

Solution

According to work energy theorem,

Work done = change in kinetic energy

$\Rightarrow W = \frac{1}{2}mv^2 - \frac{1}{2}mu^2$ where m is the mass, v is the final velocity and u is the initial velocity of the body.

Since v = 0 ($\because$ bullet comes to rest)

$$W = \frac{-1}{2}mu^2$$

$$= \frac{-1}{2} \times 0.02 \times (30)^2$$

$= -9$ joules

The negative sign indicates that the force acting on the bullet to stop it and its displacement are opposite to each other.

☛ *Example*

What is the work done by a porter in carrying a load of 50 kg
1. while walking along the platform, and
2. while climbing up a slope of height 10 m ($g = 10\ m\ s^{-2}$)?

Solution

While carrying a load the porter applies a force equal to mg vertically upwards.
(i) Work done by a porter along the horizontal

$W = \vec{F} \cdot \vec{s} = Fs\cos\theta$

F is the force, s is displacement and θ is angle between F and s
Since $\theta = 90°$

Work done $W = F\ .\ s\cos 90° = 0$ (zero)

∴ Work done when the porter walks along the platform = 0

(ii) Work done when the porter walks along the slope

$W = \vec{F} \cdot \vec{s} = Fs\cos\theta$

Work is done by force of gravity F = mg and $s\cos\theta = h$

∴ W = mgh

∴ $W = 50 \times 10 \times 10 = 5000$ joules = 5 kJ.

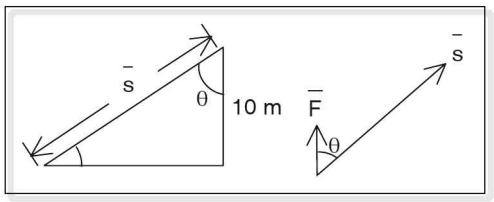


Figure 2.11

☛ *Example*

What is the power developed by an engine in lifting water of 3000 litre from a well 20 m deep in one minute (1 litre water = 10^3 kg; $g = 10\ m\ s^{-2}$)?

Assume that there is no loss of energy in the conversion of energy.

Solution

$$\text{Power} = \frac{\text{work}}{\text{time}}$$

Work = force × displacement

$W = \vec{F} \cdot \vec{s}$ work is done by the force of gravity

∴ F = mg and s = h $\theta = 0°$

$W = mgh\cos 0$

$= 3 \times 10^3 \times 10 \times 20 = 6 \times 10^5$ joules

$$\text{Power} = \frac{\text{work}}{\text{time}}$$

time = 1 minute = 60 seconds

$$\therefore \text{ Power} = \frac{6 \times 10^5}{60} = 10^4\ W = 10{,}000\ W$$

The power of the engine is 10 kW.

☛ *Example*

In a hydroelectric power station water is allowed to fall at the rate of 2000 kg per second on a turbine kept 100 m below the water level.

(i) Calculate the potential energy of water falling every one second at the highest point.

(ii) Find kinetic energy of this water when it falls through a height of 25 m.

(iii) If 80% of initial potential energy is converted in to electrical energy. Calculate the power output in 1 second ($g = 10\ m\ s^{-2}$).

Solution

(i) From definition of potential energy (P.E)

P.E = mgh

P.E = $2000 \times 10 \times 100$

$= 2 \times 10^6 J.$

From law of conservation of energy

Total energy at the highest point

E_1 = potential energy + kinetic energy

$E_1 = mgh + 0$ (K.E at the highest point = 0)

$\Rightarrow E_1 = 2 \times 10^6\ J$

(ii) When water falls through a height of 25 m its potential energy decreases but the kinetic energy (K.E) increases.

Total energy = P.E + K.E

$E_1 = mgh_1$ + K.E

where $h_1 = 100 - 25 = 75$ m

(height from the ground level)

$\therefore P_1$ = P.E at $h_1 = mgh_1$

$= 2000 \times 10 \times 75$

$= 150 \times 10^4$

$= 15 \times 10^5 J$

$= 1.5 \times 10^6 J$

From law of conservation of energy, total energy remains constant.

$\therefore E_1 = P_1$ + K.E

$E = 2 \times 10^6 J$

$E_1 = 1.5 \times 10^6$ + K.E

$2 \times 10^6 J = 1.5 \times 10^6$ + K.E

K.E = $2 \times 10^6 - 1.5 \times 10^6 = 0.5 \times 10^6 J$

(iii) 80% of initial potential energy is converted into electrical energy

$$80\% \text{ of } E_1 = \frac{80}{100} \times 2 \times 10^6$$

$= 16 \times 10^5$ J is converted in to electrical energy

$$\text{Power output} = \frac{\text{work}}{\text{time}} = \frac{\text{energy}}{\text{time}}$$

$$= \frac{16 \times 10^5}{1} = 16 \times 10^5 \text{ W}$$

Thus the power output is 16×10^5 W = 1.6 MW.

Parallel forces

A 5 N force acting along east can be represented by a ray 5 cm long with the arrow pointing towards east.

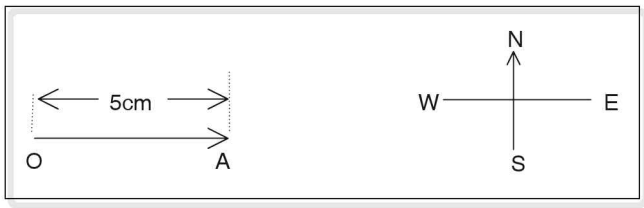


Figure 2.12

(a) The point of application of the force is represented by the point O.

(b) The length of line OA is proportional to the magnitude of the force.

(c) The direction of the arrow indicates the direction of the force. Thus a force can be represented graphically.

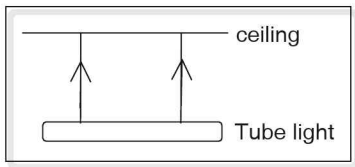


Figure 2.13

When a tube light fixture is suspended two points symmetrical about its centre from the ceiling of a room as shown, two equal forces act along the strings by which the fixture is suspended. These two forces act along the parallel lines and are hence called as parallel forces.

When two lads sit on either side of a see-saw, the weights of the lads act vertically downwards along the parallel lines. These parallel forces are called like parallel forces.

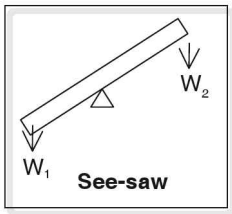


Figure 2.14

In the adjacent figure (i), A and B represent a set of like parallel forces. Similarly in figure (ii), C and D also represent another set of like parallel forces.

Thus the forces acting along parallel lines in the same direction are called like parallel forces.

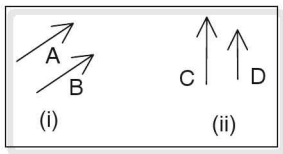


Figure 2.15

When we want to a lift a load using a wooden plank as shown, the weight of the load acts vertically downwards and the force exerted acts vertically upwards. These two forces act along the parallel lines in opposite direction. These types of parallel forces are called unlike parallel forces.

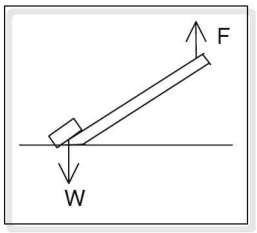


Figure 2.16

In the adjacent figure (iii), E and F represent a set of unlike parallel forces. Similarly in figure (iv), G and H also represent another set of unlike parallel forces.

Thus unlike parallel forces are forces acting along parallel lines in opposite directions.

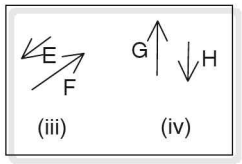


Figure 2.17

Thus parallel forces are classified into like parallel forces and unlike parallel forces.

Consider a ring pulled by two persons P and Q in two diametrically opposite directions with forces F_1 and F_2 respectively.

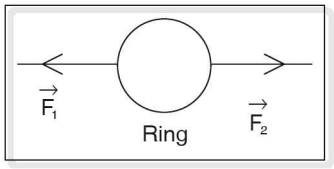


Figure 2.18

The ring moves in the direction of F_2 provided F_2 is greater than F_1. Then the resultant of the forces of F_1 and F_2 is $F_2 - F_1$ and acts in the direction of F_2.

If the magnitude of F_2 is equal to F_1 then the body does not move. Then the resultant of the two forces is zero.

Thus **resultant** of two or more parallel forces is that single force whose effect is the same as that of the combination of all forces.

When unequal weights F_1 and F_2 are placed on the pans of a balance, the resultant of the weights acts along R. If the balance is suspended by a string, the tension in the string F_3 is equal in magnitude and opposite in direction to that of the resultant. This force F_3 is called equilibrant.

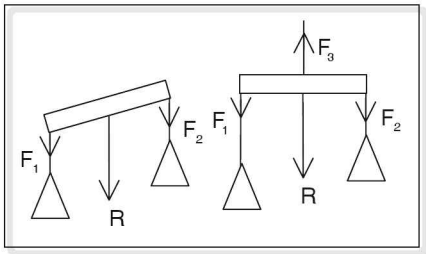


Figure 2.19

When 20 g_{wt} and 10g_{wt} are suspended to a beam or a rod, the two forces act along the parallel lines vertically downwards. The resultant of these two forces is equal to 20 + 10 = 30 g_{wt} and it acts vertically downwards near the greater force such that the point of action of the resultant divides the line joining the point of action of the two forces internally in the inverse ratio of the forces.

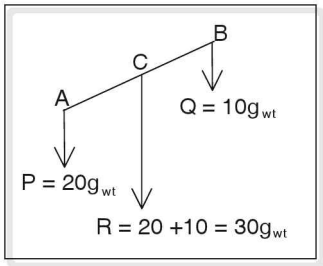


Figure 2.20

Thus the magnitude of the resultant of two parallel forces is equal to the sum of individual forces i.e., R = P + Q. The direction of the resultant is the same as the direction of either force. The resultant force R divides the line joining the point of action of P and Q in the inverse ratio of the forces.

$\Rightarrow P \times AC = Q \times BC$

When a body is lifted by a wooden plank by applying a force F at B, which acts vertically upwards, half the weight of the body W, acts vertically downwards on the plank at A. These two forces constitute unlike parallel forces and the resultant of these forces acts in the direction of the greater force. Its magnitude is equal to (W/2) ~ (F).

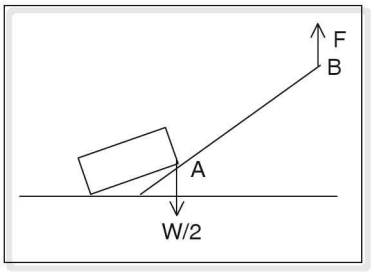


Figure 2.21

The point of action of the resultant divides the line joining the point of action of the two forces externally in an inverse ratio of the forces.

Thus the magnitude of the resultant of the two unlike parallel forces is equal to difference of the magnitude of the individual forces. Direction of the resultant is in the direction of the greater force and the point of action of the resultant divides the line joining the point of action of the two forces externally in the inverse ratio of the forces.

$P \times AC = Q \times BC$

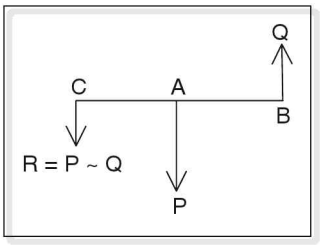


Figure 2.22

Moment of a force

When a force is applied on a door handle, it tends to turn (open or close) the door about the hinges.

This rotating or turning effect of a force about a fixed point or fixed axis is called moment of the force about that point. This moment of the force i.e., the turning effect is maximum if the applied force is perpendicular to the plane of the door and the point of application is farthest from the hinges.

Thus, the moment of force is measured by the product of the magnitude of the force and the perpendicular distance between the fixed point or the fixed axis and the line of action of the force.

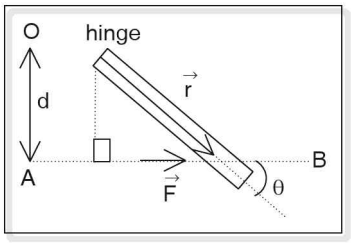


Figure 2.23

This moment of force, also called torque, is a vector quantity and is given as a cross product of the force vector, $\vec{F}$ and the radius vector, $\vec{r}$ as,

$\vec{\tau} = \vec{r} \times \vec{F}$ and the magnitude of torque is given by

$\tau = rF \sin\theta$

$= (F)\ (r \sin\theta)$

$= F.\ d$

The direction of the torque is perpendicular to the plane containing the vectors $\vec{r}$ and $\vec{F}$.

Thus, if O is the fixed point and AB is the line of action of the force and d is the perpendicular distance between O and line AB, the magnitude of the moment of force = $F \times d$

If the perpendicular distance between the line of action of force and fixed point is greater, greater will be the turning effect.

☛ *Example*

Door handles are provided far away from the hinges so that less force is required to obtain greater turning effect.

Similarly it is easier to tighten or unwind a screw using a long spanner or a screw driver with a bigger handle. By convention, moment of force is taken as positive if it rotates the body in an anti-clockwise direction and negative if it rotates the body in a clockwise direction. Moment of force is a vector quantity.

NOTE: Turning effect is zero when forces are applied on the fixed point or on the fixed axis.

Unit of moment of force

Moment of force = force × perpendicular distance

Hence unit of torque = (unit of force) (unit of distance)

Thus the S.I. unit of moment of force is N m and in C.G.S. system it is dyne cm.

Dimension formula of moment of force

Moment of force = force × perpendicular distance

Thus [torque] = [Force] [distance]

$= [M^1L^1T^{-2}] \times [L]$

$= [M^1L^2T^{-2}]$

NOTE: The dimensional formula of torque is the same as that of work.

Principle of moments or theorem of moments

Statement

When a number of parallel forces act on a rigid body and the body is in equilibrium then the algebraic sum of the moments of the individual forces about any point is equal to zero i.e., sum of anti-clockwise moments is equal to sum of clockwise moments.

If $\vec{F}_1$, $\vec{F}_2$ and $\vec{F}_3$ are three parallel forces acting about O under which the body is in equilibrium, then according to the principle of moments,

$F_1d_1 + F_3d_3 = F_2\,d_2$ or $F_1d_1 + F_3d_3 - F_2d_2 = 0$

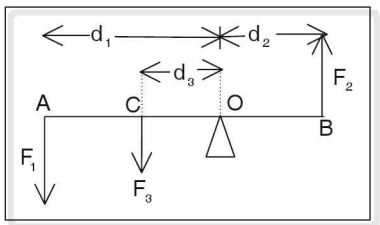


Figure 2.24

Consider a wooden plank of length 4 m supported at its ends by two supports whose tensions are 40 kg_{wt} and 30 kg_{wt}. Weight of 50 kg is suspended at the centre of the plank and 20 kg_{wt} is suspended at a distance 1m from the end of 40 kg_{wt}.

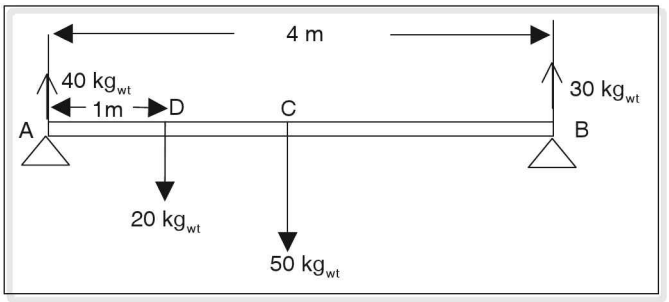


Figure 2.25

If the plank is in equilibrium, then algebraic sum of the moment at any point should be equal to zero. Taking moments about A we get

$(40 \times 0) - (20 \times 1) - (50 \times 2) + (30 \times 4) = -20 - 100 + 120 = 0$

Experiment to verify the principle of moments

1. Metre scale is supported at its centre by a knife edge so that the scale remains horizontal.
2. Slotted weights W_1 and W_2 are added to the weight hangers, which are suspended to the scale, such that the scale remains in equilibrium.
3. Distances between slotted weights and the centre of the scale are measured. Let the distances from the knife edge to the weights W_1 and W_2 be d_1 and d_2 respectively.
4. Moment of the force is calculated using the formula.

   Moment of force = Force × distance

   Clockwise moment = $W_2 \times d_2$

   Anti-clockwise moment = $W_1 \times d_1$

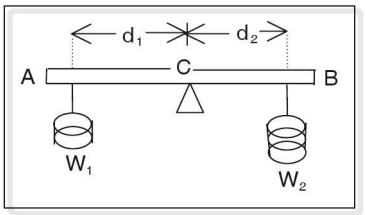


Figure 2.26

5. Experiment is repeated for different weights and different distances and the result is tabulated.

Observation

In all the cases it is found that the anticlockwise moment is equal to the clockwise moment.

NOTE: Experiment can be conducted by having more than two weights. In all the cases it can be verified that the sum of anticlockwise moments is equal to sum of clockwise moments.

Experiment to find the weight of a body using the principle of moments

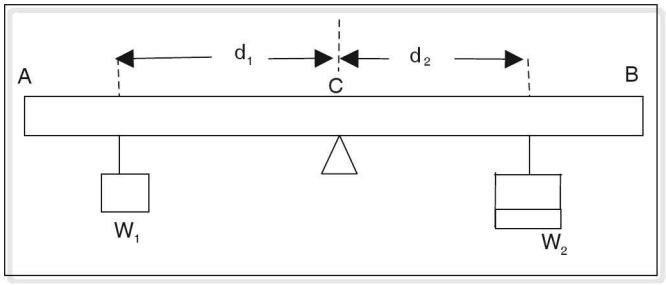


Figure 2.27

1. Metre scale is supported at its centre by a knife edge, so that the scale is horizontal
2. The unknown weight (W_1) is suspended on one side of the scale as shown and known weight (W_2) is suspended on the other side.
3. The two weights are adjusted and positioned on the scale, until the scale remains horizontal (equilibrium position).
4. Distance d_1 between unknown weight and the centre of the scale is noted down. Similarly distance d_2 between the known weight and the centre of scale is noted down.
5. From the principle of moments $W_1 \times d_1 = W_2 \times d_2$ or $W_1 = \dfrac{W_2 \times d_2}{d_1}$

   Thus the weight of the unknown body is determined.

Couple

Two equal and unlike parallel forces constitute a couple. A couple does not produce translatory motion since the resultant of the forces is zero. Rotating effect of the couple is known as moment of couple.

☛ *Example*

Couple is applied in turning (opening or closing) a tap and winding or unwinding a screw etc

Moment of couple is measured by product of either force and the perpendicular distance between the two forces (arm of couple).

In the above figure F_1 and F_2 would constitute a couple if $F_1 = F_2 = F$ (say)

$\therefore$ Moment of couple $= F \times AB$

$= F \times d$

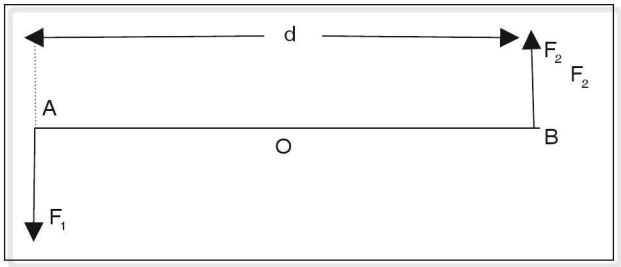


Figure 2.28

Moment of couple is taken as positive if the couple rotates the body in an anticlockwise direction and negative if it rotates the body in a clockwise direction.

Unit of moment of couple is N m in S.I. system and dyne cm in C.G.S. system.

NOTE: A couple cannot be balanced by a single force, it has to be balanced by another couple.

☛ *Example*

AB is a rigid rod of uniform cross section and length 6 m. Two forces P = 5 N and Q = 3 N act at the ends A and B of the rod. Find the magnitude, direction and point of application of the resultant if the

(a) two forces are like parallel forces,

(b) two forces are unlike parallel forces.

Solution

(a) For like parallel forces

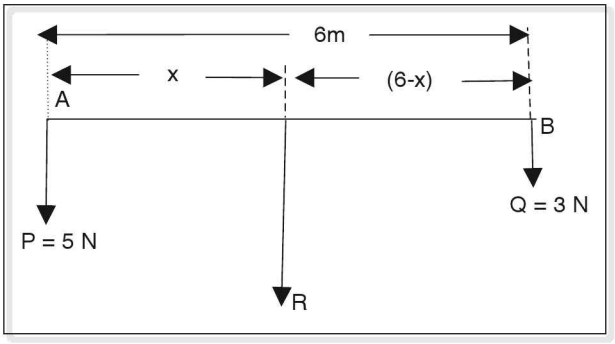


Figure 2.29

Magnitude of the resultant R = P + Q

R = 5 + 3 = 8 N

Direction of the resultant is in the direction of P and Q

Let the resultant act at a distance x from 5 N, then

P × AC = Q × BC

5 × x = 3 (6 – x)

⇒ 5x = 18 – 3x ⇒ 5x + 3x = 18

⇒ 8x = 18 or x = 2.25 m

∴ The resultant acts at a distance 2.25 m from 5 N.

(b) Unlike parallel forces

Magnitude of the resultant of two unlike parallel forces is R = P ~ Q

⇒ R = 5 – 3 = 2 N

Direction of the resultant is along the greater force.

⇒ direction of the resultant R is in the direction of P.

If the resultant acts at a distance x from 5 N, then

P × AC = Q × BC

5 × x = 3 (6 + x)

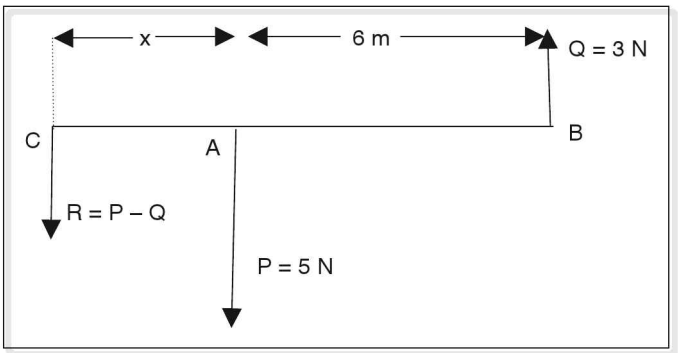


Figure 2.30

$5x = 18 + 3x$

$5x - 3x = 18$

$2x = 18 \Rightarrow x = 9$ m

i.e., resultant acts at a distance 9 m from 5 N.

☛ *Example*

Two lads of 40 kg_{wt} and 60 kg_{wt} sit on either side of a see-saw of length 3 m which is supported at its centre. Where should a third lad of 30 kg_{wt} sit so as to balance the see-saw?

Solution

Let the third lad sit at a distance 'x' from the support to balance the see-saw. From the principle of moments the sum of anticlockwise moments is equal to the sum of clockwise moments. (moment = force × perpendicular distance) Anti-clockwise moment is taken as positive

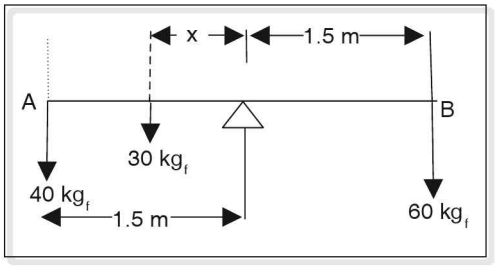


Figure 2.31

$(40 \times 1{\cdot}5) + (30 \times x) - (60 \times 1{\cdot}5) = 0$

$60 + 30x - 90 = 0$

$30x - 30 = 0$

$30x = 30 \Rightarrow x = 1$ m

☛ *Example*

A painter of 50 kg_{wt} stands on a wooden plank of length 5 m. The plank is suspended at its ends by ropes. The painter stands at a distance 2 m from one end of the plank. Find the tensions (force) on the support.

Solution

Let T_1 and T_2 be the tensions in the ropes.

Painter P stands at a distance 2 m from the end A.

From the principle of moments the algebraic sum of the moments about any point is equal to zero.

Anti-clockwise moment is taken positive. Applying theorem of moments about point A, we get

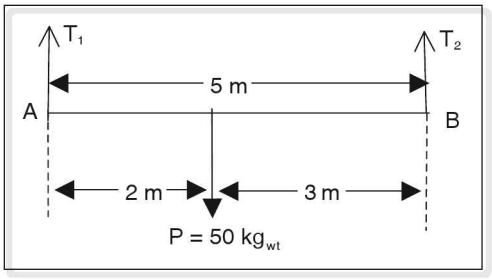


Figure 2.32

$(T_1 \times 0) - (50 \times 2) + (T_2 \times 5) = 0$

$-100 + 5T_2 = 0$

$5T_2 = 100 \Rightarrow T_2 = 20\ kg_{wt}$

(b) For equilibrium of the plank, condition is upward force = downward force

$\therefore T_1 + T_2 = P$

$\Rightarrow T_1 + 20 = 50$

$\Rightarrow T_1 = 50 - 20 = 30\ kg_{wt}$

Alternately, applying principle of moments about B;

$(T_1 \times 5) - (50 \times 3) + (T_2 \times 0) = 0$

$5T_1 - 150 = 0$

$\Rightarrow T_1 = 30\ kg_{wt}$

☛ *Example*

Find the turning effect if a force 2 N is applied on a door, perpendicular to its plane at a distance of 40 cm from its hinge.

Solution

Moment of force = force × perpendicular distance from fixed axis
Perpendicular distance, d = 40 cm =0.4 m
Thus, moment of force = $2 \times 40 \times 10^{-2}$
= 0.8 N m

Static equilibrium

A body is said to be in equilibrium, if it does not change its state of rest or uniform motion. If the body does not possess translatory motion and rotatory motion then the body is said to be in static equilibrium. Hence the condition for a body to be in equilibrium are

(i) the algebraic sum of the forces acting on the body should be equal to zero, and
(ii) the algebraic sum of the moments of forces about any point should be equal to zero.

Consider a wooden plank suspended by two parallel ropes at its ends A and B as shown. The wooden plank is in equilibrium under the action of forces. The upward forces acting on the plank are T_1 and T_2. T_1 and T_2 are the tension in the ropes. The weight of the plank W acts vertically downwards. The plank is in equilibrium because. $T_1 + T_2 = W$.

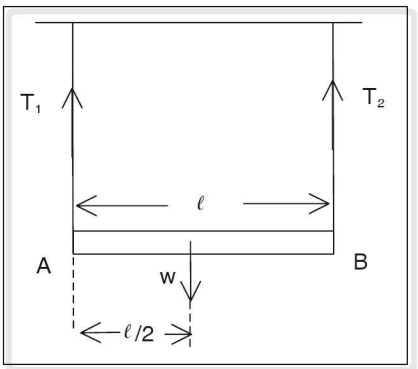


Figure 2.33

$T_1 + T_2 - W = 0$

Thus the upward force = downward force

The plank does not possess any rotatory motion

∴ The algebraic sum of moment of the forces about A = 0.

i.e., $T_1 \times 0 - W \times \frac{\ell}{2} - T_2 \times \ell = 0$

Where ℓ is the length of the plank

$W \times \frac{\ell}{2} - T_2 \times \ell$

Centre of gravity

Consider a rigid body. It consists of a number of particles and these particles are attracted towards the earth due to gravity,The forces acting on the particle are F_1, F_2, F_3 etc. These forces act downwards along parallel lines. These are called like parallel forces. These like parallel forces can be replaced by a single force whose effect is the same as the combination. This is called as the resultant. The resultant force acts at a particular point. This point can be assumed as the point where the whole weight, W of the body acts. This point is known as centre of gravity (C.G) Thus $\vec{W} = \vec{F_1} + \vec{F_2} + \vec{F_3} + \ldots\ldots\ldots$

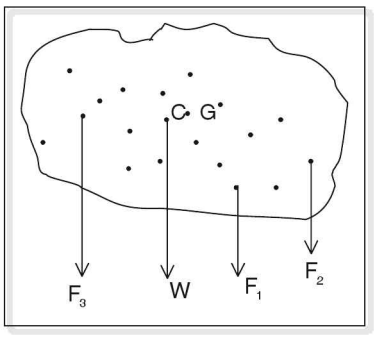


Figure 2.34

☛ *Example*

Consider a foot ball. The whole weight of the ball acts through the centre of the ball. This point is the centre of gravity of the ball. Thus in general centre of gravity (C.G) may lie inside or outside a body.

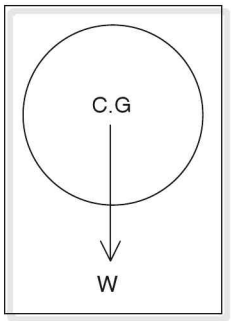


Figure 2.35

While studying the state of equilibrium of a body we often refer to the position of C.G. of the body to specify the position of the body. Thus when we say the position of the body is changed, we mean that the position of the C.G. is changed.

A funnel is placed on a table top with its base in contact with the horizontal surface because the tip of the funnel has smaller area of cross-section and when the funnel is made to stand on the tip it topples down. Thus for the funnel to be stable it should be rested with its broader base in contact with the horizontal surface.

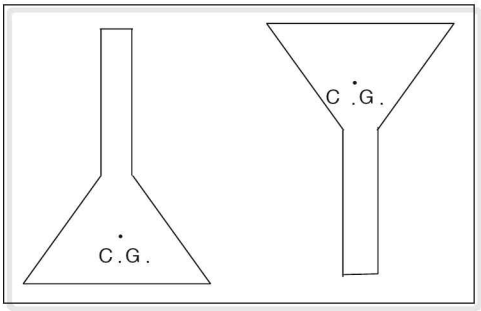


Figure 2.36 funnel

Formula race cars are shorter in height and have broader and heavier base. Thus their position of centre of gravity or the height of their centre of gravity from their base is lesser. Even if the car tilts due to high speed it can regain its original position because the vertical line from the centre of gravity falls within the base and the centre of gravity of the car is almost at the base.

The height of centre of gravity of an empty bottle from its base is higher when compared to the height of centre of gravity in the case when the bottle is half filled with a liquid. Hence the stability of the bottle is more when it is half filled compared to that when it is empty.

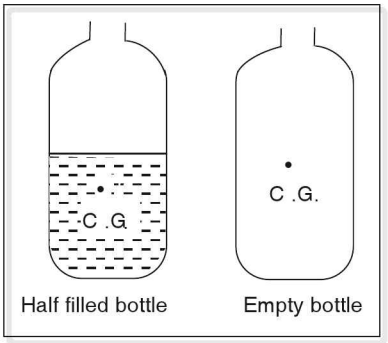


Figure 2.37

The famous leaning tower of Pisa is stable even though it is slightly tilted from the vertical, as the line joining its centre of gravity and the centre of the earth falls within its base.

Thus for a body to be stable,

(i) it should have a broader base,

(ii) the height of the centre of gravity should be lowered (C.G can be lowered by having heavy base) and

(iii) the vertical line drawn from the centre of gravity of the body to the centre of the earth should fall within the base of the body.

Stable equilibrium

When the base of a water bottle is in contact with the horizontal surface, the centre of gravity of the bottle falls within the base. If the bottle is disturbed slightly centre of gravity of the bottle is raised. As a result the stability of the bottle decreases. If the vertical line drawn from the centre of gravity to the centre of the earth passes between A and B, the moment of weight restores it to its initial position. Then the bottle is said to be in stable equilibrium.

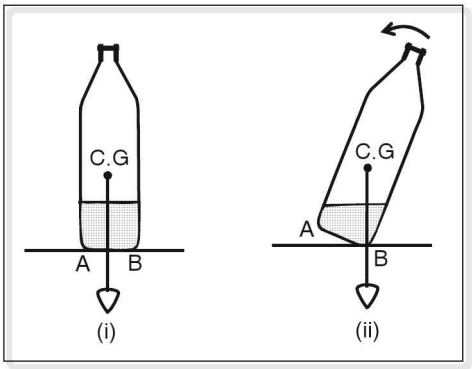


Figure 2.38

Thus a body is said to be in stable equilibrium if the body returns to its original state after the external force causing the change of position of the body is withdrawn.

Unstable equilibrium

A cone resting on its apex would topple when the slightest of disturbances causes the vertical line drawn from the C.G. to centre of the earth to fall outside the base.

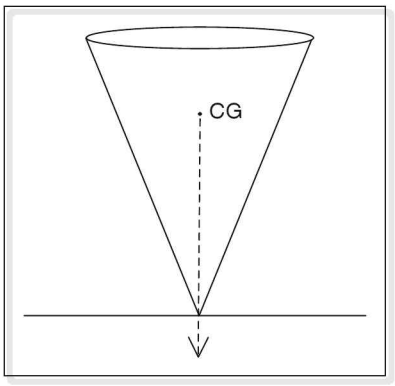


Figure 2.39

Thus the cone cannot regain its original position. This is known as unstable equilibrium. Thus a body is said to be in unstable equilibrium if the body does not return to its initial position after the removal of the external force causing a change in its position.

In the example of the bottle, if the bottle is tilted such that the vertical line from the C.G. passes to the right of B, the moment of weight turns the bottle clockwise and the bottle topples.

Neutral equilibrium

When a ball is rolled on a horizontal surface, it takes a new position but the position of centre of gravity remains intact. This state of the ball is known as neutral equilibrium.

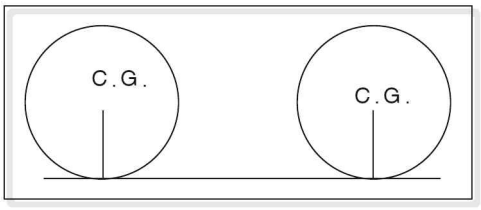


Figure 2.40

Thus a body is said to be in neutral equilibrium if it takes a new position without the change in its position of centre of gravity.

Practical examples of centre of gravity

(i) Kids or infants stand with their legs apart, so that the centre of gravity falls between the legs. This gives them the stability.

(ii) Vessels like jugs, mugs, water bottles, candles, candle stand etc have a broad base to prevent toppling.

(iii) The base of the ships pedestal fans and formula cars are made heavy in order to lower centre of gravity which gives them stability.

NOTE: The position of centre of gravity remains unchanged, whether the body is at rest or in motion.

Simple Machine

In order to cut paper or cloth, we make use of scissors. To squeeze lemon we use a squeezer. Slopes or inclined planes connecting higher floors to the lower floors are usually seen in tall buildings or hospitals. These slopes are used to carry load easily. These devices make our work easier and convenient.

In all these devices we apply a small force at some convenient point and a greater force acts on the body at another point in the desired direction. These devices are known as simple machines.

Nut cracker on its own cannot crack a nut. Energy or the force has to be applied to the nut cracker to crack the nuts. This force applied in the machine to overcome resistance is called effort or power, which is denoted by P.

In the above case of nut cracker, the force exerted on the machine or the cracker to overcome resistance is called the effort and the force required to crack the nut is called the load.

For any machine, the effort applied or the energy supplied to the machine is used to move machine parts, is partly to overcome friction and the remaining part of the energy is used to do useful work. Thus all the energy supplied to the machine is not used in doing useful work. Hence the useful work done by the machine is always less than the actual work done on the machine.

The machine whose parts are weightless and frictionless is called as an ideal machine. In an ideal machine the output energy is equal to the input energy or in other words all the work done on the machine is used in overcoming resistance or the load.

Thus in an ideal machine or perfect machine, output is equal to input.

Work done by the machine = Work done on the machine.

This is known as the principle of a machine. Ideal machine cannot be realised in practice because it is impossible to have a machine whose parts are weightless and frictionless.

Mechanical advantage: (M.A)

The ratio of load to effort is known as mechanical advantage.

$\therefore$ Mechanical advantage (M.A) $= \dfrac{\text{Load (W)}}{\text{Effort (P)}}$

If a large load is lifted with small effort then the mechanical advantage is greater than one. The effort applied on the machine is used to run the machine parts, to overcome friction and to do useful work. Thus the total load is the sum of the load due to moveable parts, load due to friction and the useful load.

This ratio of total load to effort is called ideal mechanical advantage (I.M.A.)

Thus I.M.A. $= \dfrac{\text{Total load}}{\text{effort}}$

The ratio of useful load to effort is called actual mechanical advantage (A.M.A.)

Thus A.M.A. $= \dfrac{\text{useful load}}{\text{effort}}$

Velocity ratio

Consider a load W lifted using a crow bar as shown in the figure. The effort P moves a certain distance 'x', then the load moves through a distance 'y' in the same interval of time in which the effort moves. The ratio of displacement of effort to the displacement of load is called velocity ratio (V.R.)

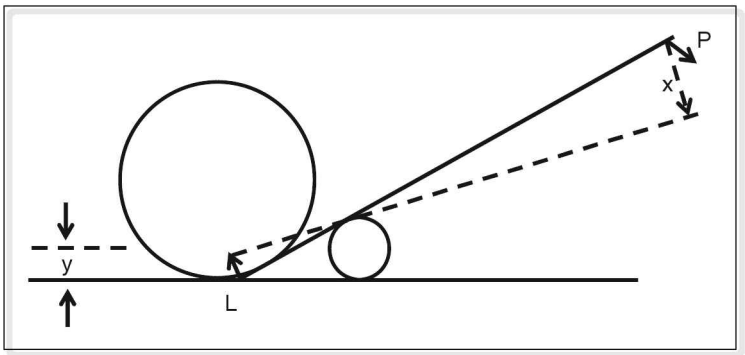


Figure 2.41

$\text{V.R} = \dfrac{\text{displacement of the effort}}{\text{displacement of the load}} \Rightarrow \text{V.R} = \dfrac{x}{y}$. Dividing the displacement of effort and load by time, we get V.R. = (x/t) ÷ (y/t) where t is time in which the load or the effort moves.

$\therefore$ Velocity ratio $= \dfrac{\text{velocity of the effort}}{\text{velocity of the load}}$

Efficiency (η)

It has already been pointed out that the actual work done by the machine is always less than the work done on the machine.

The ratio of useful work done by the machine to the work done on the machine is known as efficiency.

$$\text{Efficiency} = \frac{\text{Useful work done by the machine}}{\text{Work done on the machine}}$$

Useful work done by the machine is the work done to overcome useful load (the output) and the work done on the machine is called the input.

$$\therefore \text{ efficiency} = \eta = \frac{\text{output}}{\text{input}}$$

In practice since the output is less than input, efficiency of a machine is always less than one or 100%. For ideal machines output is equal to input. Hence the efficiency of an ideal machine is 100%.

$$\eta = \frac{\text{Work done in overcoming load}}{\text{Work done on the machine}}$$

If P is the effort which displaces through x in overcoming a load W, and y is the displacement of the load, then output = W × y and input = P × x. Hence

$$\text{Efficiency}, \eta = \frac{W \times y}{P \times x}$$

(∵ work = force × displacement)

$$\eta = \frac{W / P}{x / y}$$

$$\eta = \frac{\left(\frac{\text{Load}}{\text{effort}}\right)}{\left(\frac{\text{displacement of effort}}{\text{displacement of load}}\right)} \Rightarrow \eta = \frac{\text{mechanical advantage (M.A)}}{\text{velocity ratio (V.R)}}$$

This gives the relation between efficiency, mechanical advantage and velocity ratio.

If W_1 is the useful load and R is the resistance due to moveable parts and the friction, then the total load $W = W_1 + R$

In the case of a machine having friction and weight, work done in overcoming the total load is equal to work done by the effort.

$$\Rightarrow (W_1 + R) \times y = P \times x$$

$$\Rightarrow \frac{W_1}{P} + \frac{R}{P} = \frac{x}{y}$$

$$\Rightarrow \frac{R}{P} = \frac{x}{y} - \frac{W_1}{P} \Rightarrow R = P\left(\frac{x}{y} - \frac{W_1}{P}\right)$$

$$\Rightarrow R = \text{effort (V.R.} - \text{A.M.A.)}$$

A simple machine can be a straight or a bent rigid bar called lever capable of rotating about a fixed point, or it might be a rigid flat surface inclined at an angle called inclined plane. Thus simple machines can be classified into levers and inclined planes.

Pulley and wheel-and-axle are modifications of lever. Inclined plane is modified into wedge and screw.

Levers

The fixed point about which the lever rotates is called fulcrum.

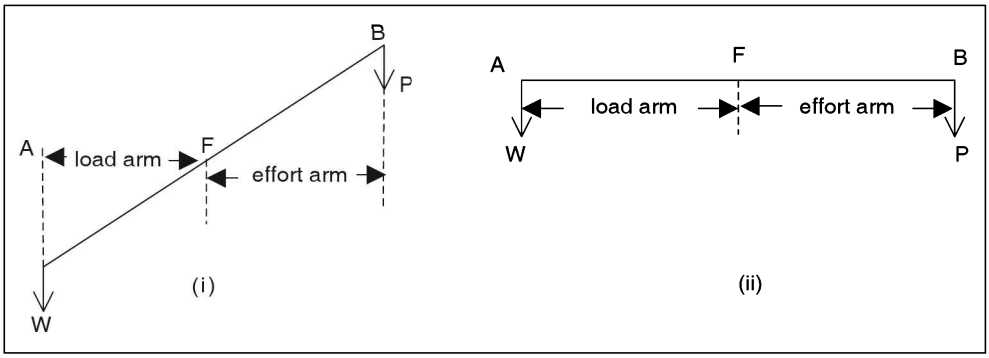


Figure 2.42 See-saw

In a see-saw, the load is applied at the end A and the effort P to lift the load is applied at B. The perpendicular distance between the load and the fulcrum is called the load arm and the perpendicular distance between the effort and the fulcrum is called effort arm. Thus in the above diagram (ii) load arm is AF, the effort arm is BF and F is the fulcrum.

In the above example, for the see-saw to be in equilibrium the algebraic sum of the moment of forces about the fulcrum should be equal to zero.

$\Rightarrow W \times AF - P \times BF = 0$ or $W \times AF = P \times BF$

Thus, load × load arm = effort × effort arm

This is called the principle of lever. From principle of lever load × load arm = effort × effort arm

$$\Rightarrow \frac{\text{load}}{\text{effort}} = \frac{\text{effort arm}}{\text{load arm}}$$

But load/effort = mechanical advantage (M.A.)

$$\therefore \text{ M.A.} = \frac{\text{effort arm}}{\text{load arm}}$$

Thus mechanical advantage of a lever increases with increase in the length of its effort arm. In the case of cutting pliers or a nut cracker the length of the effort arm is greater than that of the load arm so less effort is used to cut a wire or crack a nut.

Depending upon the relative positions of load, effort and fulcrum, levers are classified into first order, second order and third order levers.

First order lever

In the case of a see-saw, cutters, crowbar, pliers, scissors etc., fulcrum lies between effort and the load. This type of lever is known as a first order lever.

Usually in a first order lever the length of the effort arm is greater than that of the load arm. Thus, the mechanical advantage of a first order lever, in general, is greater than or equal to one. But in scissors the effort arm is shorter than the load arm, and the mechanical advantage is less than one.

Second order levers

In nutcracker, squeezer, hinges, etc load lies between effort and fulcrum. These type of levers are called second order levers. In a second order lever the effort arm is greater than the load arm. Hence the mechanical advantage of second order lever is always greater than one.

Third order lever

In forceps, tongs, knife or when a person swims the effort lies between load and the fulcrum. This type of lever is called third order lever. In third order lever the load arm is longer than the effort arm. Thus the mechanical advantage of the third order lever is always less than one. First and second order levers are force multipliers and the third order levers are speed multipliers.

Pulley

It is a wheel or a flat disc which has a grooved edge or a rim. A pulley rotates about a fixed point which passes through a centre called the axle. The axle is fixed to a frame or a block. If the block is clamped, the pulley is called fixed pulley and if the block is not clamped, the pulley is called moveable pulley.

Single fixed pulley

It is a single pulley which is usually used to draw water from well. The frame of the pulley is clamped.

A rope is passed through the groove. To one end of the rope a load is tied and effort is applied at the other end. The effort is applied in the downward direction. The tension 'T' of the rope acts on both the sides of the rope upward.

If the pulley is not rotating and the load due to friction and moveable parts is neglected then both the load and the effort are equal to the tension in the string. Thus W = P = T

$$\text{Mechanical advantage} = \frac{\text{load}}{\text{effort}} = \frac{W}{P} = \frac{T}{T} = 1$$

In practice, the effort applied is greater than load due to friction in the moveable parts of the pulley.

If the effort moves through a distance 'x' load also is lifted through the same distance. Hence velocity ratio (V.R) $= \frac{\text{distance moved by the effort}}{\text{distance moved by the load}} = 1$

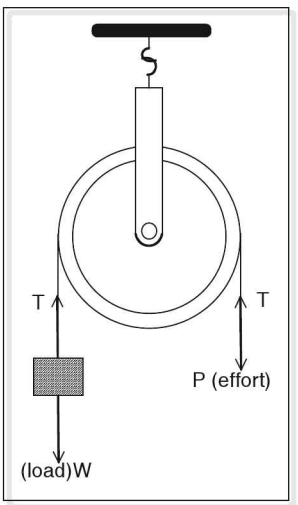


Figure 2.43 Single fixed pulley

Thus, there is no gain in mechanical advantage or in velocity ratio of the pulley. It is used only to change the direction of effort. It is easier to exert effort in the downward direction i.e., in the direction of gravity which makes lifting of loads convenient.

Efficiency of an ideal pulley

$$\eta = \frac{\text{mechanical advantage (I.M.A.)}}{\text{velocity ratio}}$$

Since M.A = 1 and V.R = 1

η = 1 or 100%

Thus the efficiency of an ideal pulley is 100%.

In practice the effort applied is used to lift useful loads, to overcome loads due to friction and load due to moveable parts. But the total load is equal to effort. If the useful load is W_1 and R is the load due to friction and moveable parts, total load = W_1 + R.

$\Rightarrow W_1 + R = P$

$$\Rightarrow \frac{W_1 + R}{P} = 1 \Rightarrow \frac{W_1}{P} + \frac{R}{P} = 1$$

$$\Rightarrow \frac{W_1}{P} = 1 - \frac{R}{P}.$$

But $\frac{W_1}{P} = \frac{\text{useful load}}{\text{effort}}$ = Actual Mechanical Advantage (A.M.A)

$$\therefore \text{A.M.A.} = 1 - \frac{R}{P}$$

Single moveable pulley

If the block of the pulley is not clamped, it is called moveable pulley.

In this case, the pulley moves as the effort pulls the rope. One end of a rope is fixed to a hook. The rope passes through the rim of the pulley and the effort is applied in the upward direction at the other end of rope. The moveable pulley divides the rope into two segments. Tension 'T' acts in an upward direction along the two segments of the rope. If P is the effort applied to overcome total load, then P = T and it is obvious from the figure that W = T + T = 2T where W is the total load.

Thus mechanical advantage = $\frac{\text{Load}}{\text{Effort}} = \frac{W}{P} = \frac{2T}{T} = 2$

Thus the effort applied is equal to half the load. Hence a single moveable pulley acts as a force multiplier. If the effort moves through a distance 'x', the load is lifted through a distance $\frac{x}{2}$. Thus velocity ratio of the single moveable pulley

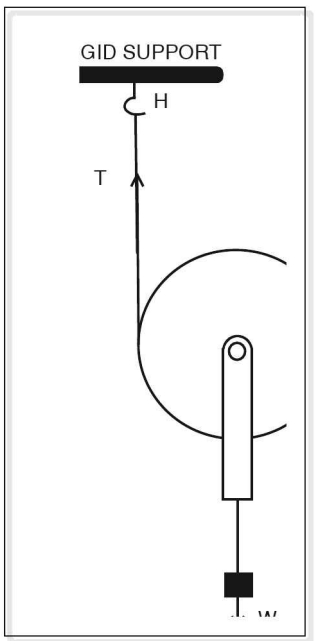


Figure 2.44 Single movable pulley

$$\text{V.R.} = \frac{\text{displacement of the effort}}{\text{displacement of the load}}$$

$$\text{V.R.} = \frac{x}{\frac{x}{2}} = \frac{2x}{x} = 2$$ The efficiency of a single moveable pulley

$$\eta = \frac{\text{M.A.}}{\text{V.R.}} = \frac{2}{2} = 1 \text{ or } 100\%$$

If W_1 is the useful load and R is load due to friction and moveable parts of the machine,

$$\text{Mechanical advantage} = \frac{\text{total load}}{\text{effort}} \Rightarrow \text{M.A.} = \frac{W_1 + R}{P}$$

$$\text{Since M.A.} = \frac{\text{Total load}}{\text{effort}} = \frac{W}{P} = 2, \text{ we get } \frac{W_1}{P} + \frac{R}{P} = 2 \text{ or } \frac{W_1}{P} = 2 - \frac{R}{P}$$

where $\frac{W_1}{P}$ is the actual mechanical advantage (A.M.A.). Hence $\text{A.M.A.} = 2 - \frac{R}{P}$

Since the ratio of R to P is a positive quantity, A.M.A. of a single moveable pulley is always less than 2.

Since it is not convenient to apply effort in the upward direction, a single moveable pulley is combined with a fixed pulley to give the same M.A. and V.R. and it is easier to apply the effort using the combination of the fixed pulley and moveable pulley because the effort is applied in the downward direction.

It consists of two pulleys. One pulley (B) is fixed and this pulley helps to change the direction of effort. The other pulley (A) moves along with the rope. The load is attached to the moveable pulley.

So one end of rope is fixed to a hook and the rope passes through the moveable pulley and the fixed pulley. The moveable pulley moves along the two segments of the rope. If the effort moves through a distance x, each length of the segment of rope is decreased by an amount equal to $\frac{x}{2}$.

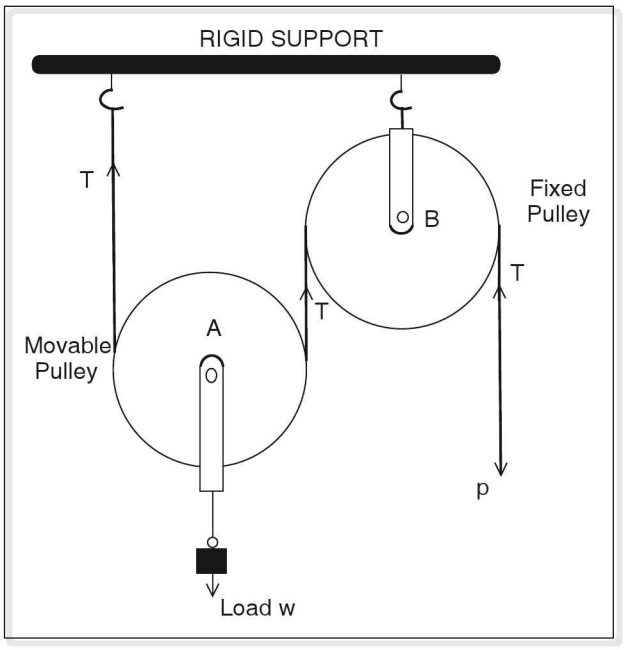


Figure 2.45 Single movable pulley with a fixed pulley

Sheath pulley system or block and tackle system

If consists of two sets of pulleys. One set of pulleys is attached to a rigid support at the top and is called the block. The lower set of pulleys carries a load and is called tackle.

One end of the string or rope is fixed to a hook of the tackle. The rope passes round the rim of all the pulleys and the effort is applied at the free end of the rope in the downward direction, as shown in the figure.

Tackle is supported by 'n' segments of the rope, where 'n' is the number of pulleys in the system. Tension on each segment of the rope is T and it acts vertically upwards.

If P is the effort used to overcome load W, then W = nT and P = T. Thus the ideal mechanical advantage of the pulley system I.M..A = $\frac{W}{P} = \frac{nT}{T} = n$

Thus the I.M.A. of a pulley system is equal to number of pulleys in the system. The effort required to lift the load is $\frac{W}{n}$. Thus the system of pulleys acts as a force multiplier and it also changes the direction of force.

Thus larger the number of pulleys, lesser is the effort. If the effort moves through a distance x then length of each segment of the rope is decreased by $\frac{x}{n}$.

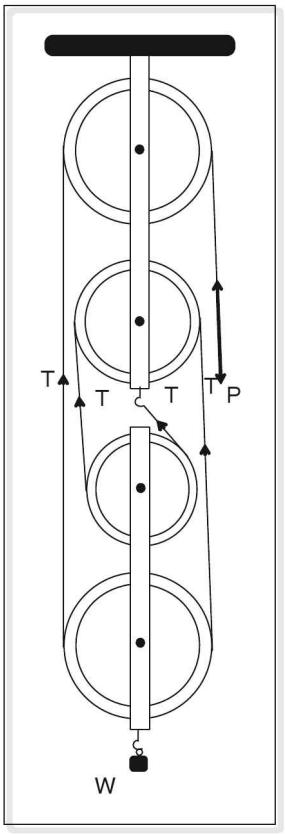


Figure 2.46 Block and tackle for four

Thus the velocity ratio of the pulley system (V.R.) = $\frac{\text{displacement of the effort}}{\text{displacement of the load}} \Rightarrow \text{V.R.} = \frac{x}{\left(\frac{x}{n}\right)} = n$

Note: Since M.A. = n there is no gain in energy.

The efficiency of system of pulleys

$$\eta = \frac{\text{M.A.}}{\text{V.R.}} = \frac{n}{n} = 1 \text{ or } 100\%$$

η is 100% only in an ideal system of pulleys. If W_1 is the useful load to overcome by the effort, then mechanical advantage M.A. = $\frac{W}{P}$ where W is the total load. But $W = W_1 + R$ where R is the load due to friction and moveable parts of the system.

Then M.A. = $\frac{W_1 + R}{P}$

Since M.A. = n, we get

$$\frac{W_1 + R}{P} = n$$

$$\frac{W_1}{P} + \frac{R}{P} = n \text{ But}$$

$\frac{W_1}{P}$ is the actual mechanical advantage of the pulley system A.M.A

$$\Rightarrow \text{A.M.A.} + \frac{R}{P} = n$$

$$\text{Hence A.M.A.} = n - \frac{R}{P}$$

Since $\frac{R}{P}$ is positive, A.M.A. of the pulley system is always less than n.

Effect of weight of pulleys in tackle

If W_2 is the weight of the pulleys in the tackle then the load and weight is balanced by nT, where n is the number of pulleys and T is the tension on each part of the rope. Thus,

$$W + W_2 = nT = nP \qquad (\because T = P)$$

where P is the effort

$$\Rightarrow W + W_2 = nP$$

$$\Rightarrow \frac{W}{P} + \frac{W_2}{P} = n \Rightarrow \frac{W}{P} = n - \frac{W_2}{P}$$

$$\text{Hence M.A} = n - \frac{W}{P} \qquad \left(\because \frac{W}{P} = \text{M.A}\right)$$

The efficiency of the system of pulleys,

$$\eta = \frac{\text{M.A.}}{\text{V.R.}}$$

$$\Rightarrow \eta = \frac{n - \frac{W_2}{P}}{n} \qquad (\because \text{V.R} = n)$$

$$\Rightarrow \eta = 1 - \frac{W_2}{nP}$$

Thus to achieve higher efficiency in the system of pulley, the weights of the pulleys in the tackle should be lesser.

Wheel and axle

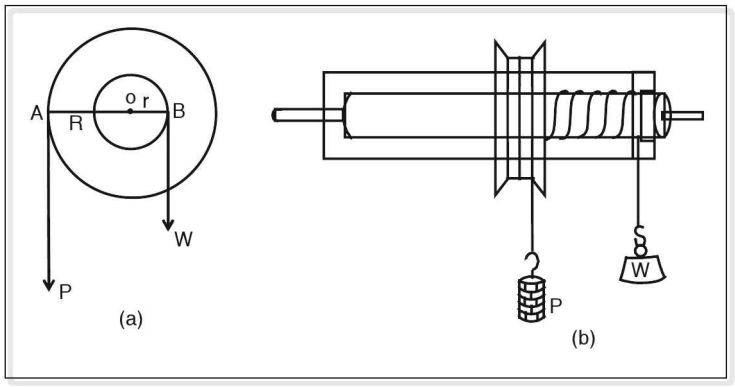


Figure 2.47 Wheel and axle

A wheel is fixed to a strong cylindrical axle such that they have a common axis. The radius of the wheel (R) is greater than the radius of the cylindrical axle (r). A string is wound around the cylinder. Load is attached to the free end of this string. Effort is applied to the free end of the string which is wound round the wheel. The windings of the rope on the wheel is in the opposite direction to that of the windings on the axle. The rotating effect of the wheel is equal to the rotating effect of the cylinder.

$\Rightarrow$ Moment of the wheel about the axis = Moment of the axle about the axis

$\Rightarrow P \times R = W \times r$ where P is the effort and W is the load.

$$\Rightarrow \frac{W}{P} = \frac{R}{r}$$

But $\frac{W}{P}$ is the mechanical advantage (M.A.)

$$M.A. = \frac{R}{r}$$

$$M.A. = \frac{\text{R adius of the wheel}}{\text{radius of the axle}}$$

Screw jack

When a tyre of a car or a heavy vehicle is punctured the vehicle should be lifted up so as to remove or replace the tyres. These vehicles are lifted up easily using a device known as screw jack.

A screw jack consists of a screw which works through a base which is at rest or stationary. The screw jack is placed under the vehicle or the load to be lifted and the effort is applied through a horizontal bar. The effort applied gives a turning effect to the bar and the bar in turn

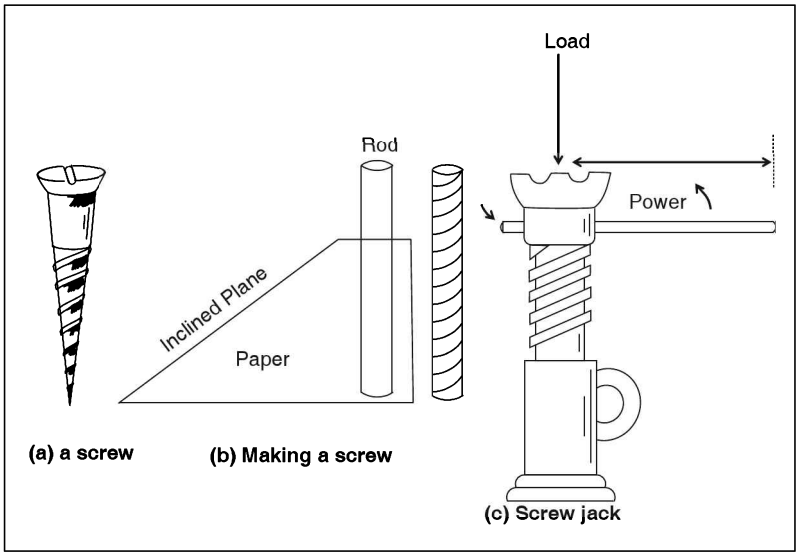


Figure 2.48 (a), (b), (c) The screw jack

rotates the screw. When the bar undergoes one complete rotation, the screw is raised by a height equal to the pitch. Pitch is the distance between two consecutive threads of the screw.

For an ideal screw jack, work done by effort is equal to work done on the load. If P is the effort applied to the horizontal bar and W is the load which is lifted through height h, then work done by the effort is $2\pi\ell \times P$ where ℓ is the length of the bar from the axis of the screw.

$2\pi\ \ell$ is the distance moved by the effort of the horizontal bar.

Work done on the load = $W \times h$, where 'h' is the pitch of the screw.

Work done by effort = work done on the load. But ratio of load to effort W/P is the mechanical advantage (M.A.) of the screw jack.

$$\Rightarrow 2\pi\ \ell \times P = W \times h$$

$$\Rightarrow \frac{W}{P} = \frac{2\pi\ell}{h}$$

Thus M·A. of the screw jack = $\frac{2\pi\ell}{h}$

Inclined plane

In order to climb up a hill or top of a building, we make use of slopes or a ladder or a staircase. We make use of slopes to take vehicles in and out of the house, garage or parking slots at higher level. Flyovers are constructed with slopes. Thus by using the slopes, we can easily move the load against gravity to a higher location.

Inclined plane is a smooth rigid plane inclined to the horizontal.

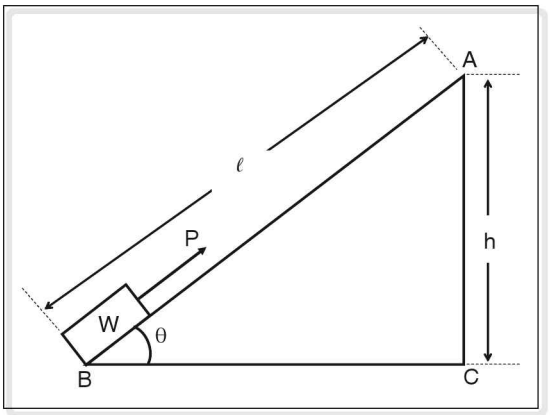


Figure 2.49 Inclined plane

Mechanical advantage of an inclined plane when efforts acts parallel to the plane.

Consider an inclined plane AB inclined to the horizontal at an angle θ. Let W be the load to be moved along the plane AB by applying effort P.

If the surface of the plane is considered to be frictionless then the output work is equal to the input work. i.e., work done by the effort is equal to work done on the load.

$$\Rightarrow P \times AB = W \times AC$$

$$\frac{W}{P} = \frac{AB}{AC}$$

$$\frac{W}{P} = \frac{\text{Length of the plane } (\ell)}{\text{height of the plane (h)}}$$

From figure in the previous page $\frac{AC}{AB} = \sin\theta$

$\therefore \frac{W}{P} = \frac{1}{\sin\theta} = \frac{\ell}{h}$.

But $\frac{W}{P}$ is the mechanical advantage (M.A.) of the inclined plane.

$\therefore$ M.A. $= \frac{W}{P} = \frac{\ell}{H} = \frac{1}{\sin\theta}$

From the above expression, it is obvious that lesser the inclination of the slope, lesser is the effort to raise a given load along the inclined plane. Thus greater effort is required to climb a steep hill.

Experiment to determine mechanical advantage of an inclined plane.

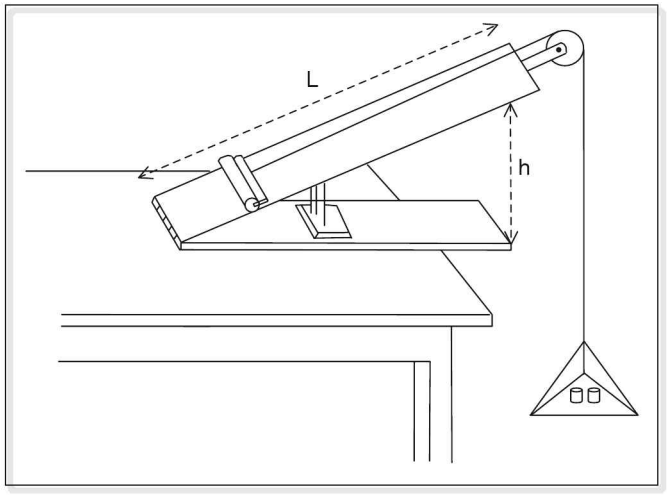


Figure 2.50 Inclined plane

Apparatus required: Inclined plane, meter scale, physical balance, scale pan, roller and weights.

Procedure

1. Inclined plane is inclined at a suitable angle.
2. Weight of the roller 'W' is found using a physical balance.
3. The roller is attached to one end of the string and the string is passed over a pulley. To the free end of the string is attached the scale pan whose mass (S) is known.
4. The length 'L' of the plane and the height 'H' is measured using a metre scale.
5. Suitable weights are added in the scale pan, so that the roller just ascends the plane. The weight W_1, in the scale pan is noted down.
6. The weights in the scale pan are gradually decreased until the roller just descends. The weight W_2 in the scale pan is noted down.

7. Experiment is repeated for different heights of the plane and the results are tabulated.

Observation

Weight of the roller W = _____ kg_{wt}

Weight of the scale pan S = _____ kg_{wt}

Length of the plane, L = _____ m

1					
2					
3					
4					
5					

P is the effort = $S + \left(\frac{W_1 + W_2}{2}\right) kg_{wt}$.

Result: It is found that $\frac{W}{P} \approx \frac{L}{H}$ and M.A. = $\frac{W}{P}$.

☛ *Example*

1. A machine is used to displace a load of 400 N through 50 cm by an effort of 100 N. If the efficiency of the machine is 80%,
   (i) find its mechanical advantage and velocity ratio.
   (ii) find the distance moved by the effort and
   (iii) find the resistance due to the moveable parts of the machine and friction.

Solution

(i) In the given problem

load = W = 400 N.

Distance moved by the load = y = 50 cm = 0·5 m.

Effort = P = 100 N.

Efficiency = η = 80%.

$$\text{M.A.} = \frac{\text{Load}}{\text{Effort}} = \frac{W}{P} = \frac{400}{100} = 4$$

Thus the M.A. of the machine is 4.

To find V.R, we can use the relation between, η, V.R. and M.A.

$$\eta = \frac{\text{M.A.}}{\text{V.R.}}$$

$$\Rightarrow \frac{80}{100} = \frac{4}{\text{V.R.}}$$

$$\therefore \text{V.R.} = \frac{4 \times 100}{80}$$

V.R. = 5.

V.R. of the machine is 5.

(ii) To find the distance moved by the effort we use the relation

$$\text{V.R.} = \frac{\text{distance moved by the effort}}{\text{distance moved by the load}} = \frac{x}{y}$$

where x is the distance moved by the effort, and y is the distance moved by the load

$$\Rightarrow 5 = \frac{x}{0.5}$$

$\therefore x = 0{\cdot}5 \times 5 = 2{\cdot}5$ m

Hence distance moved by the effort = 2·5 m.

(iii) Resistance due to moveable parts of the machine and friction.

R = P (V.R – A.M.A)

where R is the resistance due to moveable parts and friction.

A.M.A. is the actual mechanical advantage

$$\text{A.M.A.} = \frac{\text{useful load}}{\text{effort}}$$

Useful load = 400 N

$$\therefore \text{A.M.A.} = \frac{400}{100} = 4$$

$\therefore$ R = 100 (5 – 4) = 100 N.

Thus the load due to moveable parts and friction is 100 N.

☛ *Example*

Calculate the resistance offered by a lemon when an effort of 0·06 N is applied. The effort arm of the squeezer is 10 cm and the load arm is 3 cm. In length assume that the effort is applied at the end of the effort arm.

Solution

In the given problem
Effort = P = 0·06 N
Effort arm = 10 cm = 0·1 m
Load arm = 3 cm = 0·03 m
According to the principle of a simple machine
Load × load arm = effort × effort arm

$$\Rightarrow W = \frac{0.06 \times 0.1}{0.03} = 0.2\text{N}$$

Hence resistance offered by the lemon = 0.2 N

☛ *Example*

A person draws 10 kg water from a well at a depth of 5 m using a single fixed pulley by applying an effort of 120 N. Find the mechanical advantage, velocity ratio and efficiency of the machine. ($g = 10\ m\ s^{-2}$)

Solution

In the given problem
Load = W = $10 \times 10 = 100$ N
Effort = P = 120 N

$$\therefore \text{M.A.} = \frac{W}{P} = \frac{100}{120} = 0.833$$

$$\text{V.R.} = \frac{\text{distance moved by the effort}}{\text{distance moved by the load}}$$

In a single fixed pulley the distance moved by the effort = distance moved by the load.

$\therefore$ V.R. = 1

$$\text{efficiency} = \eta = \frac{\text{M.A}}{\text{V.R}}$$

$$\eta \frac{0.833}{1} = 0.83 \text{ or } 83\%$$

☛ *Example*

A block and tackle system of pulleys consists of 8 pulleys. A load of 1000 kg_f is lifted by an effort of 250 kg_f. Find the mechanical advantage, velocity ratio and efficiency.

Solution

In the given problem number of pulleys n = 8
W = load = 1000 kg_f
Effort = P = 250 kg_f

$$\therefore \text{M.A.} = \frac{\text{load}}{\text{effort}} = \frac{1000}{250} = 4$$

If the effort moves through a distance 'x' then the load moves through a distance x/n

$$\text{V.R.} = \frac{\text{distance moved by the effort}}{\text{distance moved by the load}} = \frac{x}{x/n} = n = 8$$

$$\text{Efficiency } \eta = \frac{\text{M.A.}}{\text{V.R.}} = \frac{4}{8} = 0.5 = 50\%$$

Thus the efficiency of the given system of pulleys is 50%.

☛ *Example*

A screw jack is used to lift a vehicle of weight 10000 N through a height of 0·25 m from the ground. If the length of the rod is 0·3 m, find the effort and mechanical advantage used to lift the load. (Assume the jack to be ideal)

Solution

In the given problem load = W = 10000 N
Height through which the load is lifted is h = 0·25 m

Length of the rod = 0·3 m
For an ideal machine work done on the load is equal to work done by the effort
$\Rightarrow W \times h = 2\pi r \times P$ where P is the effort

$$\therefore P = \frac{W \times T}{2\pi r}$$

$$= \frac{10,000 \times 0.25}{2\pi r \times 0.3} = \frac{25000}{6\pi} = 1326.29\,N$$

$$M.A. = \frac{load}{effort} = \frac{10000}{1326.29} = 7.539$$

Thus the M.A. = 7·539

Circular Motion

When we open or close a door, the door rotates about a fixed axis (hinges). Each particle or point on the door describes an arc with its centre at the foot of the perpendicular line drawn from the particle to the axis of rotation. Thus the particles of the door undergo a circular motion.

Hence a body is said to be in circular motion if every particle in it moves along a circular path and the centre of all these circles lies on the axis of rotation.

If a particle moves along a circle with uniform speed, then the motion of the particle is called as uniform circular motion.

☛ *Examples*

1. Motion of a person in a giant wheel or a merry-go-round.
2. Motion of the moon round the earth.

Consider a particle in a rigid body undergoing uniform circular motion along a circle of radius r. In uniform circular motion, though the speed of the particle is constant, its linear velocity is not constant. The magnitude of linear velocity is constant but its direction is along the tangent to the circumference of the circular path. Thus the direction of the linear velocity changes continuously. According to Newton's first law, the particle should move along the tangent due to inertia of direction. Instead the particle changes its direction and moves along a circular path due to a net external force acting on it. Thus the particle undergoes displacement along the curve.

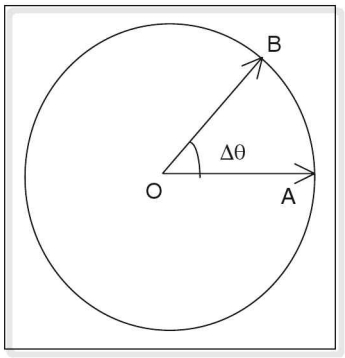


Figure 2.51

Thus the radius vector $\vec{OA}$ drawn from the centre of circular path to the particle rotates. The angle through which the radius vector rotates is known as angular displacement. The angular displacement, $\Delta\theta = \angle AOB$.

Let the particle be at point 'A' at time t_1. Due to the net external force acting on it, the particle undergoes uniform circular motion and let the particle after time t_2 to be at B. Let $\Delta\theta$ be the angular displacement of the radius vector, (angular displacement = $\Delta\theta = \frac{s}{r}$ where s is linear displacement and r is the radius of circle).

The rate of angular displacement is called angular velocity represented by ω.

$\omega = \frac{\Delta\theta}{t_2 - t_1} = \frac{\Delta\theta}{\Delta t}$ where Δt is small interval of time. Hence

$$\omega = \frac{\text{change in angular position}}{\text{time}} = \frac{\text{angular displacement}}{\text{time}}$$

Angular displacement is measured in radians and time in second. Hence the S.I. unit of angular displacement is radians per second. When the time interval Δt tends to zero, then the ratio of $\Delta\theta$ to Δt is called instantaneous angular velocity.
Instantaneous angular velocity

$$\omega = \underset{\Delta t \to o}{Lt} \frac{\Delta\theta}{\Delta t}$$

Relation between v, r and ω

Consider a particle moving with constant speed v along a circular path of radius r. Let θ be angular displacement in time interval t, then the angular velocity,

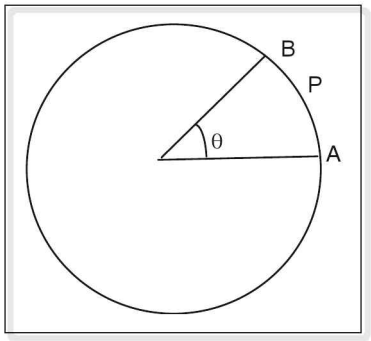


Figure 2.52

$\omega = \frac{\theta}{t}$, s = APB = length of the minor arc i.e., linear displacement of the particle.

But $\theta = \frac{s}{r}$

$\therefore \omega = \frac{s}{rt}$

But displacement/time = linear velocity, v

$\Rightarrow \omega = \frac{v}{r}$ or $v = \omega r$

Alternative method

Consider a particle moving in a circular path of radius r with uniform speed 'v'. Let $\Delta\theta$ be an extremely small angular displacement of the radius vector $\overrightarrow{OA}$ in time interval Δt.

Length of the arc AB = $s = r\Delta\theta$

Linear velocity of the particle = $\frac{\text{arc length}}{\Delta t} = \frac{r\Delta\theta}{\Delta t}$

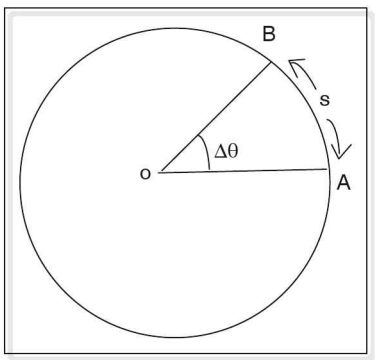


Figure 2.53

But $\frac{\Delta\theta}{\Delta t}$ = angular velocity = ω

$\therefore v = r\omega$

Thus the above equation gives the relation between linear velocity and angular velocity.

Angular momentum

Rotating bodies possess angular momentum. The angular momentum is defined as $\bar{L} = \bar{r} \times \bar{p}$; when $\bar{p}$ is the linear momentum. If $\bar{r}$ and $\bar{p}$ are perpendicular to each other,

$L = mvr$. Thus, angular momentum = mvr

$\because v = \omega r$.

$L = mr^2\omega$

Period

The time taken to complete one revolution or 2π radians is called period. It is denoted by T.

$$\text{Period} = \frac{2\pi}{\text{angular velocity}}$$

$$T = \frac{2\pi}{\omega} \text{ or } \omega = \frac{2\pi}{T}$$

Note: 2π radians = 360°

From the earlier discussion we know that when a particle undergoes a circular motion, the magnitude of the velocity remains the same but the direction of the velocity changes at every point. The direction of velocity is along the tangent to the circumference of the circle . Thus the particle is continuously accelerated. This acceleration acts in the direction of the force. This force is always perpendicular to the tangent to the circumference of the circle which makes the body move with constant speed. Thus the force is directed towards the centre of the circle. Hence the acceleration is directed towards the centre of the circular path and this acceleration is called centripetal acceleration or normal acceleration or radial acceleration.

Centripetal acceleration 'a' is given by $a = \frac{v^2}{r}$ where v is the magnitude of linear velocity and r is the radius of the circular path.

Also, since $v = r\omega$, $a = \omega^2 r$

Centripetal force

It is seen that when a particle undergoes uniform circular motion, a force is required to make the particle change its direction of motion continuously to move along the circular path with a constant speed. This force is always directed towards the centre of the circular path. This force is called as centripetal force or radial force (inward force).

Thus centripetal force is the force which is always directed towards the centre of the circle (i.e., along the radius vector) when the body undergoes circular motion.

Centripetal force F = mass of body × centripetal acceleration

$\therefore \quad F = m\frac{v^2}{r} = m\omega^2 r \qquad (\because \ v = \omega r)$

Centripetal force is provided by some other external force.

☛ *Examples*

1. The centripetal force required for a planet to go round the sun is provided by the gravitational force between the sun and the respective planet.
2. The centripetal force required for electrons to move round the nucleus is provided by electrostatic force of attraction between the nucleus and the electron.

Centrifugal force

In kinematics we have studied that rest or motion is defined depending on a reference. For example, a bird flying is said to be in motion when the reference frame is attached to the ground. But for another bird flying alongside, the first bird appears to be at rest.

Thus we find that Newton's first law can be applied depending on the reference frame. Hence, whenever no net force acts on a body producing a change of rest or of uniform motion, then there always exists a reference frame in which a body has no acceleration. Such a reference frame is called inertial reference frame or inertial frame.

For example, a car moving with uniform velocity can be considered as an inertial frame, since no force acts on any body placed in the car.

However, traveling in a train accelerating forwards, you would observe objects like a box placed on the floor of the coach sliding backward as if a force is acting on them. Thus a reference frame attached to an accelerating train is a non-inertial reference frame, and the force which appears to be acting on the box is referred to as a pseudo force or fictious force.

For example, standing passengers in a bus are thrown forward when the driver suddenly applies brakes.

Similarly, standing on the rotating platform of a merry-go-round, the hanging seats along the circumference appear to be pushed outward as if a radially outward force acts on them. This pseudo force arising from a circular motion or rotatory motion is referred to as centrifugal force.

Another example:

Consider a man sitting in a circular chamber rotating about a central vertical axis at a uniform angular velocity, ω rad s^{-1}.

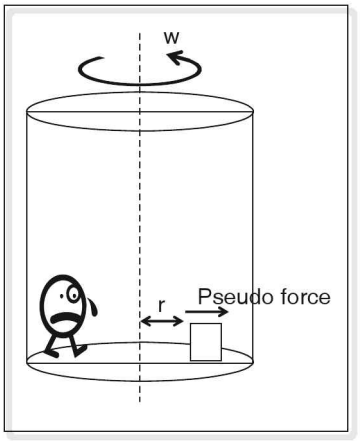


Figure 2.54

Let a box be resting on the floor at a distance of r from the centre.

As the chamber continues to rotate at uniform angular velocity, the man would observe that the box would slide and move towards the wall of the chamber. The box would remain at rest only after it comes into contact with the wall.

Similarly, every particle of the chamber also experiences such a force directed radically outwards.

The centripetal force and centrifugal force are equal in magnitude but opposite in direction.

Thus, the magnitude of the centrifugal force is given the expression.

$$\text{Centrifugal force} = \frac{mv^2}{r} = m\omega^2 r$$

It is always directed towards the centre.	It is directed away from the centre.
It is a real force.	It is a pseudo force.
It is always associated with an external agent, e.g., tension in the string etc.	Being a pseudo force, it is not associated with any external agent.
It is the cause of a circular motion.	It is the effect of circular motion.
It is the pull on the body towards the centre of the circle.	It is a tendency to move away from the centre of the circle.

Similarities between centripetal force and centripetal acceleration

1. Their magnitude of both forces is given by the expression $\frac{mv^2}{r}$ or $m\omega^2 r$
2. In uniform circular motion, their magnitude is constant.
3. The magnitude of each force depends on the mass and the speed of rotation.

Note: Centripetal force and centrifugal force are not action-reaction pair since both are acting on the same body.

Application of centrifugal force

Centrifuge

Centrifuge is a device used to separate minute particles of varying densities which cannot be separated by sedimentation. Its working is similar to the butter churn used to churn curds. It consists of a number of vials suspended to a horizontal frame which is supported by a vertical stand. The vials are filled with solid liquid mixture from which solid particles are to be separated and the vials are rotated in a horizontal plane at a very high speed by an electric motor. Higher density particles move towards the bottom of the vials (small lubes) because centrifugal force for heavier particles are greater than the lighter particles. Thus the lighter particles move towards the axis and can be easily skimmed off. Thus particles of different densities are separated.

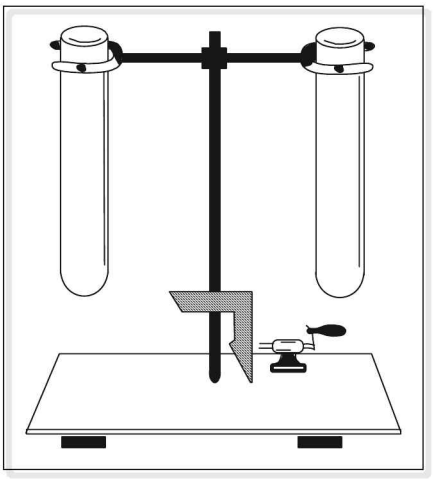


Figure 2.55 Centrifuge

Centrifuge is used to separate cream from milk, to separate hormones, viuses proteins, etc from blood.

Laundry drier or washing machine dryer

It consists of a perforated cylindrical vessel which can rotate about its axis at a very high speed. The clothes to be dried are placed in the cylinder and are rotated with high speed. The adhesive force between water and the cloth is less compared to the centrifugal force acting on the water molecules.

This centrifugal force, forces the water molecules to escape through the holes of the cylindrical vessel. Thus the wet clothes get dried.

Centrifugal governor

It works on the principle of centrifugal force. This device was applied to steam engine by James Watt. It is a special governor that controls the speed of an engine by regulating the fuel intake into it irrespective of load or fuel supply conditions and thus maintains a near constant speed.

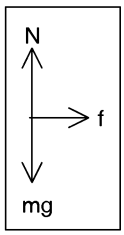
Centrifugal pumps

This device is used to transfer liquids or gases from low pressure regions. This device works on the principle of centrifugal force. Exhaust fans and blowers are centrifugal pumps.

Banking

The centripetal force that helps a vehicle to negotiate a curve is the frictional force between the tyres and the road. Since the vertical displacement of the vehicle moving along a circular path is zero, the frictional force involved is only static frictional force, which is limited by $\mu_s N$.

Since N = mg,



$f \leq \mu_s mg$

Equating f to centripetal force, $\frac{mv^2}{R}$, we get $\frac{mv^2}{R}, \leq \mu_s mg$

$\frac{v^2}{R} \leq \mu_s g$

Thus both the speed of the vehicle and the radius of curvature are limited to the extent that the ratio $\frac{v^2}{R}$ must be less than $\mu_s g$.

If either the speed increases or the curve to be negotiated is sharper, then frictional force alone would not be sufficient to produce the corresponding centripetal acceleration and a component of the normal force is utilized.

This is the reason a cyclist negotiating a curve would tend to bend in the direction of the curve; and roads are provided with banking.

Banking of roads

Banking is the inclination given to roads and rail tracks at turnings or along curved paths, by raising the outer edge of the path slightly above the inner edge towards the centre of the curvature.

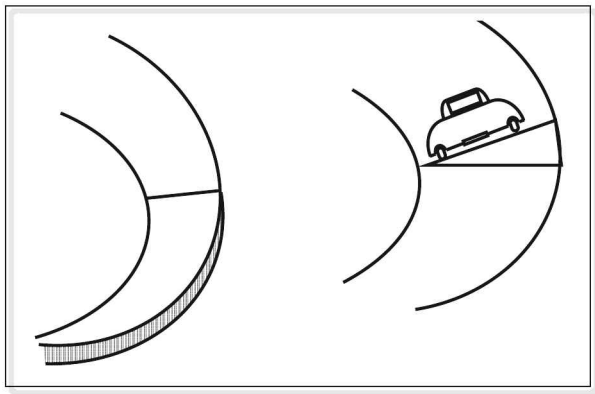


Figure 2.56

Consider a body moving along a curved path, where no banking is provided.

A car moving on a road or a train moving on the rails requires a centripetal force while taking a turn. To some extent, this centripetal force is provided by the friction between the road and tyre. When the vehicles are heavy, the required centripetal force may not be provided by the frictional force, as it is limited.

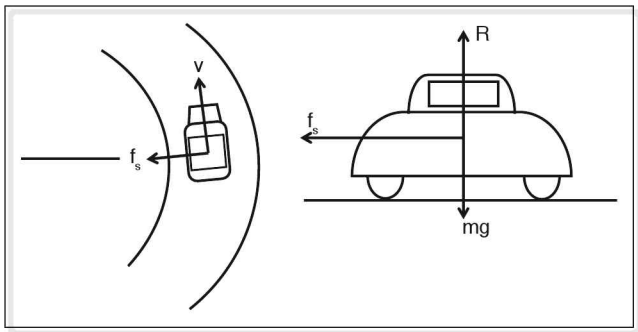


Figure 2.57

As the magnitude of f_s cannot exceed $\mu_s R$, where μ_s is the coefficient of static friction,

$\therefore f_s \leq \mu_s R$

$f_s \leq \mu_s mg$ i.e., $\frac{mv^2}{r} \leq \mu_s mg$

$\mu_s \geq \frac{v^2}{rg}$

If this condition is not satisfied, the car or train would skid outwards.

Now, consider a curved road with banking. The angle made by the line joining the outer edge to the inner edge with the horizontal line is called the 'angle of banking'.

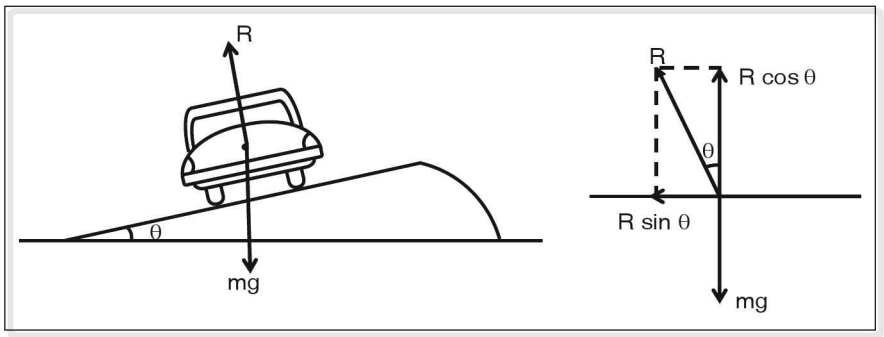


Figure 2.58

Let the angle of banking be θ. On resolving the normal reaction into components we can say that

$R \cos \theta = mg$

$R \sin \theta = \frac{mv^2}{r}$

Thus, we have $\tan \theta = \frac{v^2}{rg}$

The roads are banked, i.e., the angle of banking is provided for the average expected speed of the vehicles. If the speed of the vehicle v is less than the average expected speed (v_0) for a given banking, the vehicle would skid towards the centre.

If $v > v_0$ the car would skid outwards.

Centrifugal reaction

Take a stone and tie it to a string and whirl the stone with the other end of the string tied to the forefinger so that it describes a horizontal circle. The string exerts a force on the stone towards the centre. This force is centripetal force. At the same time, the string exerts an outward force on the finger and this force is known as centrifugal reaction.

Thus centrifugal reaction is a reaction force which acts on an agent, which makes the body to move along a curve or a circular path.

It is important to note that centrifugal reaction is not centrifugal force. Centrifugal reaction is a reaction force according to Newton's third law of motion, whereas centrifugal force is a pseudo force in a non-inertial frame.

☛ *Example*

A car takes a turn along a curved road of radius 40 m and the angle of banking of the road is 45°. At what maximum speed can the car be driven so that it may not skid. ($g = 10\ m\ s^{-2}$)

Solution

The maximum speed for the car is given by v.

$\tan\theta = \frac{v^2}{rg}$ where r is the radius of the road, g is the acceleration due to gravity and θ is the angle of banking = 45°

r = 40 m

g = 10 m s^{-2}

$$\tan\theta = \frac{v^2}{rg}$$

$$\tan 45° = \frac{v^2}{40\times 10} \Rightarrow \qquad (\because \tan 45 = 1)$$

$v^2 = 400$

$v = \sqrt{400} = 20$

The maximum speed of the vehicle = 20 m s^{-1}

☛ *Example*

A record player makes 60 rpm (revolution per minute). Find its angular velocity and its period.

Solution

In the above problem, the number of revolutions per minute = n = 60. Hence, the number of revolutions for one second, $n = \frac{60}{60}$ rps = 1

One revolution corresponds to 2π radiaus.

$\therefore$ Angular velocity = $\omega = 2\pi \times n$

$= 2\pi \times n$

$= 2\pi$ radians s^{-1}

☛ *Example*

A figure skater spins on her toes with her arms stretched out. If the length of each arm is 60 cm and if she spins at a uniform linear speed of 4 m s^{-1} , measured at the tip of the arm find the

(a) angular velocity

(b) period

(Consider length of arm = distance from tip of the arm to the axis of rotation)

Solution

a) In the above problem, since the ballet dancer spins on her toe, the length of the arm

= radius = r

r = 60 cm = 0·6 m

Linear speed = v = 4 m s^{-1}

Angular velocity $\omega = \frac{v}{r}$ rad s^{-1}

$$W = \frac{4}{0.6} = \frac{40}{6} = 6.66 \text{ rad } s^{-1}$$

(b) period, that is time taken for one revolution,

$$T = \frac{2\pi}{\omega} = \frac{2\pi}{6.66} = 0.9425 \text{ s}$$

☛ *Example*

A block of mass 150 g moves with a uniform speed in a circular groove of radius 30 cm. Find its centripetal acceleration and centripetal force acting on the block if the linear speed of the block is 0·8 m s^{-1}.

Solution

In the given problem
Mass, m = 150 g = 0·150 kg
Radius, r = 30 cm = 0·3 m
Linear speed v = 0·8 m s^{-1}
Hence centripetal acceleration, $a = \frac{v^2}{r}$

$$\therefore a = \frac{(0 \cdot 8)^2}{(0 \cdot 3)} = \frac{0 \cdot 64}{0 \cdot 3} = \frac{6 \cdot 4}{3} = 2 \cdot 13 \text{ m } s^{-2}$$

centripetal force $F = \frac{mv^2}{r} = m\omega^2 r$ or
F = m·a
= 0·15 × 2·13 = 0·195 N

☛ *Example*

A bucket containing water is fastened to a rope and is whirled in a vertical circle of radius 50 cm. Find the minimum speed at the highest point, so that the water does not spill.

Solution

The bucket should be rotated such that the weight of the bucket and the water in it are balanced by the centrifugal force.
Let the mass of the bucket filled with water be = m
∴ weight = mg
weight = centrifugal reaction acting on the bucket filled with water

$$\Rightarrow mg = \frac{mv^2}{r}$$

$(\because \text{centrifugal reaction} = \frac{mv^2}{r})$
where r is radius or the length of the rope = 500 cm = 0·5 m

$\therefore v^2 = rg$

$v = \sqrt{rg}$

$v = \sqrt{0.5 \times 9.8} = \sqrt{4.90}$

$v = 2{\cdot}214 \text{ m s}^{-1}$

Planetary motion

Motion of the planets, moon, sun, etc puzzled many scientists. It was Ptolemy in the second century A.D. who put forward a theory known as geo-centric theory. According to his theory, Sun along with the other planets revolves round the stationary earth, which is the centre of the universe.

Geo-centric theory is also known as Ptolemic theory. Geo-centric theory was accepted until Copernicus, a Polish monk proposed a theory known as 'helio centric theory'. According to helio-centric theory, earth along with other planets revolves round the sun in a circular orbit. Helio-centric theory was accepted for many years. It was Kepler, assistant of Tyco Brahe who analysed the observations of Brahe and supported the helio-centric theory and put forward his results in the form of laws called as Kepler's laws.

First law of kepler

First law of Kepler states that all planets revolve round the sun in an elliptical orbit of different radii with the sun at one of their foci points.

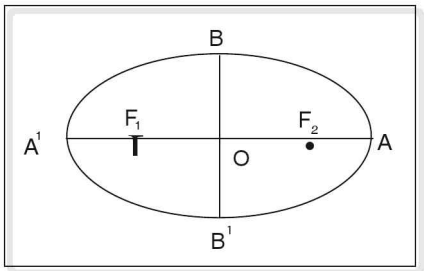


Figure 2.59 Ellipse

AA^1 is the major axis, BB^1 is the Minor axis.

Ellipse consists of two axes, longer axis ($A A^1$) is called the major axis and the shorter axis ($B B^1$) is called the minor axis. Two points F_1 and F_2 which are at fixed positions on the major axis are called as focii.

Kepler's first law is also called law of orbits.

Second law of kepler

According to the second law of Kepler an imaginary line drawn from sun to the planet, also called radius vector, sweeps out equal areas in equal intervals of time.

Consider a planet orbiting around the sun as shown in the figure. Let P_1 and P_2 be the positions of the planet at the beginning and end of a given time interval, say a month. After a certain duration, let P_3 and P_4 be the positions of the planet at the beginning and end of a month duration. Then according to Kepler's second law, the area SP_1P_2 is equal to the area SP_3P_4. This law is called law of areas.

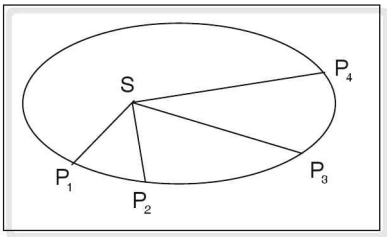


Figure 2.60

Third law of kepler

Statement: The cube of the average distance of the planet from the sun varies as square of period of its revolution. The length of the line drawn from the sun to the planet is not the same in an elliptical orbit. Hence average or mean distance of the planet from the sun is taken.

If T is period of revolution of the planet and d is the average distance of the planet from the sun, then $T^2 \propto d^3$.

If T_1 and T_2 are periods of revolutions of two planets, and d_1 and d_2 are the mean distances of the planets from the sun respectively then according to Kepler's third law,

$$\frac{T_1^2}{T_2^2} = \frac{d_1^3}{d_2^3}$$

Newton's Universal Law of Gravitation

Newton was investigating the nature of force that makes the moon go round the earth. His observation of an apple falling towards the earth led him to conclude that earth attracts all bodies towards it and he found that the force of attraction is not limited only to terrestrial bodies, but to all bodies in the universe. Force between the bodies is known as gravitational force. Newton made use of Kepler's second law in stating the universal law of gravitation.

Statement

Every body in the universe attracts every other body with a force which is directly proportional to the product of their masses and inversely proportional to the square of the distance between them.

If m_1 and m_2 are masses of two bodies separated by a distance d, then according to universal law of gravitation, the gravitational force between them, F is given by

$F \propto m_1 m_2$ --- (1)

and $F \propto \frac{1}{d^2}$ --- (2)

Combining (1) and (2), we get

$$F \propto \frac{m_1 m_2}{d^2}$$

$\Rightarrow F = G\frac{m_1 m_2}{d^2}$ where G is the universal gravitational constant.

S.I. unit of G:

$$G = \frac{F \times d^2}{m_1 m_2}$$

$$\text{Unit of G} = \frac{\text{unit of F} \times (\text{unit of distance})^2}{(\text{unit of mass}) \times (\text{unit of mass})} = \frac{N \times m^2}{kg \times kg}$$

Thus the unit of G in S.I. system is N m^2 kg^{-2} and the value of G = $6{\cdot}67 \times 10^{-11}$ N $m^2 kg^{-2}$.

Dimensional formula of G

Dimensional formula of G

$$= \frac{[\text{Force}] \times [\text{distance}]^2}{[\text{mass}]^2} = \frac{[M^1L^1T^{-2}] \times [L^2]}{[M^1] \times [M^1]}$$

$$= [M^{-1}L^3T^{-2}]$$

Hence the dimensional formula of G is $[M^{-1} L^3T^{-2}]$

G =F when $m_1 = m_2 = 1$ unit and the distance between the two bodies is one.

$$G = \frac{F \times 1^2}{1 \times 1} = F$$

Thus universal gravitational constant is equal to the gravitational force acting between two bodies of unit mass separated by a unit distance.

When a ball or any object is thrown up, it is pulled towards the earth due to earth's gravitational force. As a result the velocity of the body increases by $9{\cdot}8$ m s^{-1} for every second at sea level near the equator of the earth. Thus the body is uniformly accelerated.

This uniform acceleration of a freely falling body due to earth's gravitational force is known as acceleration due to gravity. Acceleration due to gravity is denoted by 'g'.

Note: Acceleration due to gravity is the same for all bodies irrespective of their mass, shape and size.

Relation between G and g

Consider an apple of mass m falling from a tree. The apple is acted upon by gravitational force due to earth and hence it is attracted towards the earth. From Newton's law of gravitation, the gravitational force acting on the apple is given by $F = \frac{GMm}{R^2}$ --- (1)

Where M is the mass of the Earth and R is the radius of the earth. From Newton's second law, the apple is accelerated towards centre of the earth due to gravitational force.

$\therefore$ F = mg --- (2)

From equations (1) and (2), we get

$$mg = \frac{GMm}{R^2} \text{ or } g = \frac{GM}{R^2}$$

This equation shows the relation between g and G and it shows that g is independent of mass of the body i.e., all bodies irrelevant of their masses are accelerated to the same extent. Thus when we drop a wooden ball and a solid iron ball of different masses simultaneously from the same height, both reach the earth at the same time.

From the relation $g = \frac{GM}{R^2}$, the value of g can be found out.

Mass of the earth = 6×10^{24} kg

Radius of the earth = $6{\cdot}4 \times 10^3$ km = 6.4×10^6 m.

$$\therefore\ g = \frac{GM}{R^2} = \frac{6 \cdot 67 \times 10^{-11} \times 6 \times 10^{24}}{(6 \cdot 4 \times 10^6)^2} = 9.77 \because 9{\cdot}8 \text{ m s}^{-2}$$

The value of 'g' calculated above is near the equator at sea level. The value of 'g' varies with height, depth and latitude.

Variation of 'g' with height

Consider a body of mass 'm' at a height 'h' above the surface of the earth. If M and R are mass and radius of the earth respectively then acceleration due to gravity at height 'h' is given by

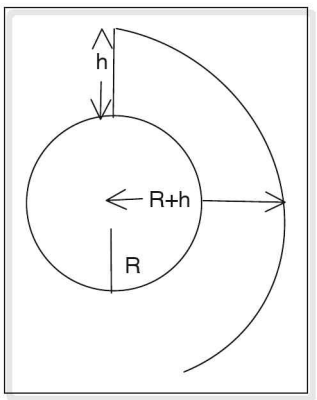


Figure 2.61

$$g_h = \frac{GM}{(R+h)^2} \quad \text{--- (1)}$$

On the surface of the earth

$$g = \frac{GM}{R^2} \quad \text{--- (2)}$$

From equation (1) and (2), we get

$$\frac{g_h}{g} = \frac{R^2}{(R+h)^2}$$

$$\frac{g_h}{g} = \frac{R^2}{R^2\left(1+\frac{h}{R}\right)^2}$$

$g_h = \frac{g}{\left(1+\frac{h}{R}\right)^2}$. Expanding $\left(1+\frac{h}{R}\right)^{-2}$ using binomial theorem and neglecting the higher powers. When h < < R, i.e., close to the earth's surface.

we get $\left(1+\frac{h}{R}\right)^{-2} \because 1-\frac{2h}{R}$

$\Rightarrow g_h = g\left(1-\frac{2h}{R}\right)$

Thus acceleration due to gravity decreases with height.

Variation of 'g' with depth

As we go deep into the earth, the effective mass of the earth decreases. Effective mass of the earth at a depth 'h' is given by $M_d = \frac{4}{3}\pi (R-h)^3 \rho$, where ρ is the average density of the earth.

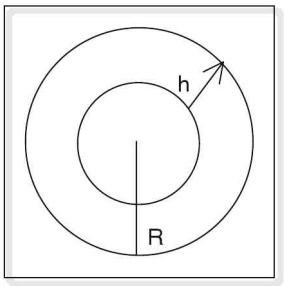


Figure 2.62

Thus acceleration due to gravity decreases with depth and the acceleration at depth 'h' is given by

$g_d = g\left(1-\frac{h}{R}\right)$

At the centre of the earth, depth, h = R

$\therefore g_d = 0.$

Variation of 'g' with latitude

Earth is not exactly spherical in shape, but is ellipsoidal. Its polar diameter is lesser by 40 km approximately than the equatorial diameter. Since $g \propto \frac{1}{R^2}$, acceleration due to gravity is greater or maximum at the poles and minimum at the equator.

Earth is not uniformly dense. There are a variety of mineral deposits, water, oil etc, inside the earth and on the surface there are massive land formations like mountains etc. Hence the value of g is affected by all these factors. Small variations in g is measured by a device known as gravity meter.

MASS AND WEIGHT

The force exerted by the earth on a body is called weight. Thus the weight of a body (w) is equal to mg where 'm' is the mass of the body and 'g' is acceleration due to gravity. From earlier discussion, it is seen that g varies from place to place, hence the weight of the body also varies from place to place and is different on different planets since 'g' is different on different planets, whereas the mass is the amount of matter contained in the body and thus mass of a body remains constant everywhere.

Suppose we measure the mass and the weight of a body using a common balance and a spring balance respectively, at two places, say Kanyakumari and Mount Everest, we observe that mass of the body at the two places is same but its weight on Mount Everest is less than that at Kanyakumari.

It is the amount of matter contained in the body.	It is the force exerted by the earth on the body i.e., W = mg.
Mass remains same anywhere in the universe.	Weight changes or varies from place to place.
S.I. unit of mass is kilogram.	S.I. unit of weight is newton and gravitational unit is kg_f or kg_{wt}.
Mass is measured by a common balance or physical balance.	Weight is measured by spring balance.

Measurement of weight – spring balance

The weight of a body (mg), as discussed earlier, can be determined using a spring balance. Before we proceed to know about how the measurement of weight is done by using a spring balance, we need to know about the balance and the principle behind it.

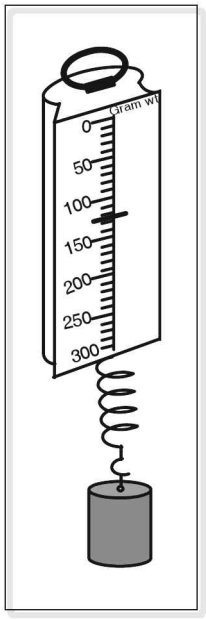


Figure 2.63
Spring balance

A spring balance consists of a spring and a pointer attached to the spring sliding over a scale as shown in the following figure. There is a hook attached to the lower end of the spring so as to enable us to attach the given load to the balance.

Working principle of a spring balance

If a load is attached to the hook of a spring balance, there is an extension in the length of the spring due to the attached load. Thus there is a downward force acting on the spring all along its length. The force acting per unit area of cross-section of the spring is called stress. If 'F' is the downward force acting on the spring and 'A' is the area of cross section of the spring, then stress = $\frac{F}{A}$. Due to this stress, there is an extension ($\Delta\ell$) in the length of the spring. This extension or elongation of the spring takes place throughout its length (ℓ). The ratio of elongation to its original length or elongation per unit length is called 'strain'. Thus $\text{strain} = \frac{\Delta\ell}{\ell}$.

The amount of strain produced in the spring is directly proportional to the stress applied on it.

This proportionality is applicable to the spring within the elastic limit (the condition in which the spring tends to loose its elastic property) only. Thus, within the elastic limit, for an elastic body, stress is proportional to the strain. This is known as Hooke's law.

Thus according to Hooke's law

$$\text{stress} \propto \text{strain} \Rightarrow \frac{\text{stress}}{\text{strain}} = \text{constant}$$

$$\therefore \frac{F/A}{\Delta\ell/\ell} = \text{constant.}$$

$$\Rightarrow \frac{F}{A} \times \frac{\ell}{\Delta\ell} = \text{constant}$$

$$\Rightarrow \frac{F}{\Delta\ell} \times \frac{\ell}{A} = \text{constant}$$

As 'ℓ' the initial length of the spring and 'A' the area of cross-section of the spring are constant, we have $\frac{F}{\Delta\ell}$ = constant. The downward force 'F' acting on the spring is the weight of the load 'w' and so $\frac{w}{\Delta\ell}$ = constant .

Experiment to determine weight of a body using a spring

Consider a spring of considerable length suspended from a rigid support as shown in the figure. Let 'ℓ' be its length, when no load is attached to the free lower end of the spring. When a load is attached at the lower end of the spring, it has an extension $\Delta\ell$. According to Hooke's law, within the elastic limit, the weight of the load attached at the bottom of the spring (w) is directly proportional to the extension ($\Delta\ell$). Thus $w \propto \Delta\ell$ and so $\frac{w}{\Delta\ell}$ = constant . This is the principle used in measuring the weight of a substance.

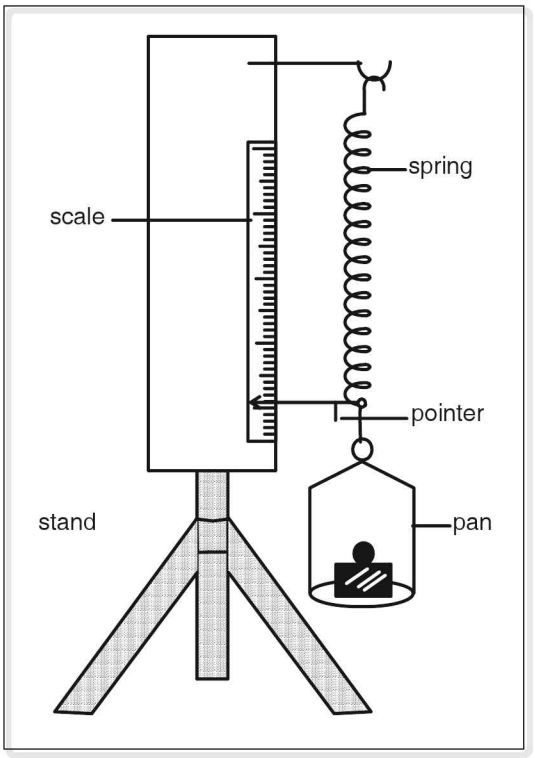


Figure 2.64 Spring balance

Attach a pan of negligible weight at the bottom of the spring. Let 'ℓ_o' be the length of the spring shown by the pointer on the scale. Now place 10 g weight in the pan, and note the pointer reading. Let it be 'ℓ_1'.

Then the extension in the spring $\Delta\ell = \ell_1 - \ell_o$. Now increase the weight in the pan in steps of 10 g and in each case note the pointer reading 'ℓ_1', so that the extension $\Delta\ell$ in each case can be calculated. The procedure is repeated till the weight in the pan is around 100 g. Now gradually decrease the weight in the pan in steps of 10 g and again note the pointer reading 'ℓ_2' in each case. Calculate the mean of the pointer readings while loading and unloading the pan so that extension $\Delta\ell$ is calculated accurately. In each case calculate $\frac{D\ell}{\text{weight}}$.

The observations are tabulated as below.

Sl. No.	Weight in the pan (w) (g)	Pointer positions		Mean position	Extension	$\frac{D\ell}{w}$
		While loading ℓ_1 (cm)	While unloading ℓ_2 (cm)	$\ell = \frac{\ell_1 + \ell_2}{2}$ (cm)	$\Delta\ell = \ell - \ell_0$ (cm)	(cm g^{-1})
1	10					
2	20					
3	30					
4	40					
5	50					
6	60					
7	70					
8	80					
9	90					
10	100					

A graph is plotted between w (in g_{wt}) along the X-axis and $\Delta\ell$ (in cm) along the Y-axis. The graph obtained is a straight line making a positive angle with the X-axis and passing through the origin, as shown in the figure.

To determine the weight of a body, the body is kept in the pan; then the extension in the spring is measured. From the graph of weight versus extension, the weight of the body is determined.

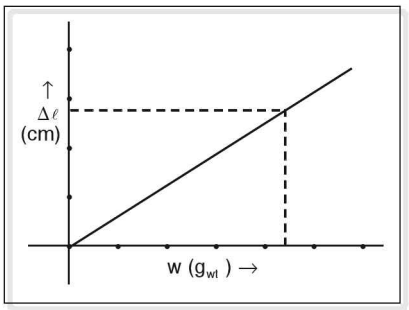


Figure 2.65 w versus $\Delta\ell$ graph

Rockets

Rockets are vehicles which are used to carry objects having heavy mass like space craft, satellite etc. The mass of rocket plays an important role in accelerating the rocket. Rocket works on the principle of conservation of linear momentum which states that when net external force acting on the system is zero, the total momentum of the system remains constant.

During the launching of a rocket fuels are burnt and the gases escape with a high speed. These gases are ejected through a nozzle present at the rear end of the rocket. The rocket acquires the same momentum as that of the ejected or exhaust gases but in the opposite direction.

Unlike other vehicles, rockets carry their own oxidant which provides oxygen for combustion of the fuel. Hence rockets can operate in space. Rocket carries payload or space-craft, which is usually placed in the front portion of the rocket.

The mass of the rocket not only includes the mass of the space craft but also mass of the fuels and oxidants called rocket propellant. The mass of the rocket has an effect in its acceleration.

The product of rate of the fuel consumption and its escape velocity gives the force or thrust exerted on the rocket. If R is the rate of fuel combustion and V_{ex} is the exhaust velocity of the fuel then

RV_{ex} = force exerted on the rocket.

RV_{ex} = mass of rocket × acceleration of the rocket.

Thus, greater the mass of the rocket, lesser is its acceleration. But as the rocket consumes fuel, the mass of the rocket keeps on decreasing. Further, acceleration due to gravity decreases as the rocket goes higher and the resistance of air also decreases as it ascends. The ratio of mass of pay load (M) to mass of propellant (m) is called 'pay load ratio'. Thus pay load ratio = M/m.

The propellants i.e. fuel and oxidant are kept in different chambers and they are supplied in controlled quantities to the engine and the combustion of the fuel takes place in the engine. The exhausted gases come out through the nozzle present at the rear end.

Multi-stage rockets

In order to carry heavy load, multi-stage rockets are used. A multi-stage rocket consists of a number of parts of a rocket which are placed (joined) one above the other. As the rocket ascends to a height say about 100 km, it discards the chambers of propellants which have been used. Thus the mass of the rocket is reduced. This is the first stage. The second stage carries the rocket to a still greater height and at a certain height the chambers of the propellant get detached from the rocket. The next stage takes over and the process continues. Thus the fuel consumed is reduced. This increases the efficiency of the rocket.

Rockets which are used to launch space vehicles or scientific instruments are multistage rockets and generally have massive first booster stage and two smaller stages.

Rockets are used to launch space-craft or satellites and missiles. Rockets are equipped with radio transmitters which help in transmission of data collected back to the earth.

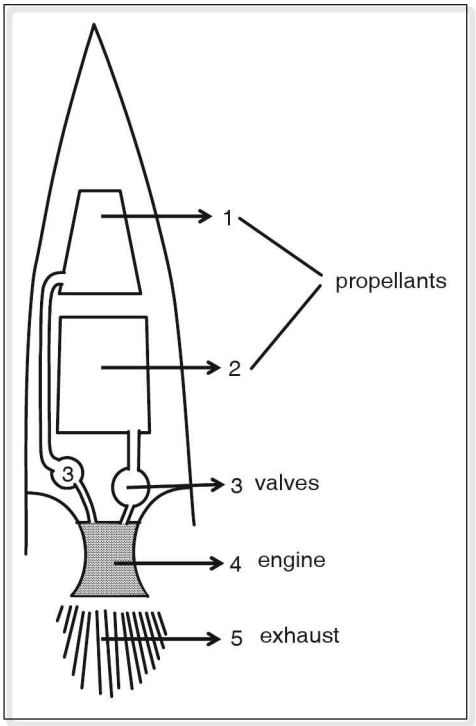


Figure 2.66 Rocket

Satellite

Moon is the natural satellite of the earth. INSAT, Aryabhatta, Sputnik etc are artificial satellites. Thus satellites are celestial or artificial objects which revolve round a body of larger mass.

Artificial satellites are man-made satellites and these are launched using multi stage rockets. A rocket takes a satellite to the desired height and places the satellite in an orbit by giving sufficient horizontal velocity.

Principle of Orbiting Satellite

When an object is dropped, it reaches the earth along the vertical line in the absence of air resistance (free fall).

If the object is thrown horizontally with the same initial velocity, it takes a curved path, but still it reaches the earth.

If the initial horizontal velocity is further increased, the object takes a less curved path and the object would fall on the earth at a farther point or the range increases. Thus as the initial velocity increases, greater is the radius of the curved path.

If the object is given an initial horizontal velocity, such that the radius of the curved path described by the object is equal or greater than the radius of the earth, the object will never return to earth. Instead it orbits round the earth and the object becomes an artificial satellite.

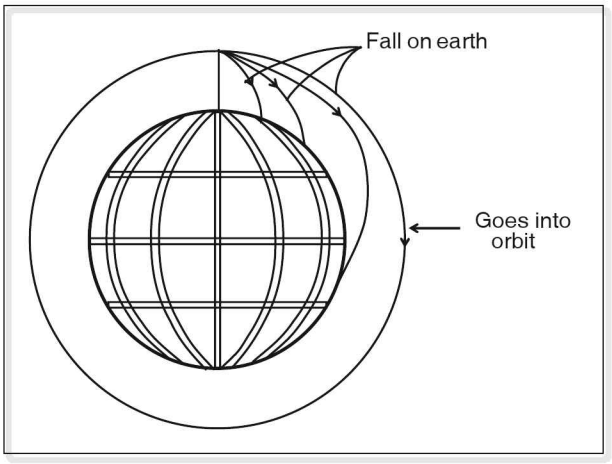


Figure 2.67

If the value of initial velocity of the object is less than the required value the object falls to earth.

Orbital velocity (v_0)

Satellites are taken to a required height and are given required horizontal velocity such that it revolves round the planet. This horizontal velocity with which the satellite revolves round the planet is called as orbital velocity.

For any satellite to go around the planet, it requires a centripetal force. The centripetal force required for the satellite is provided by gravitational force of attraction between the satellite and the planet.

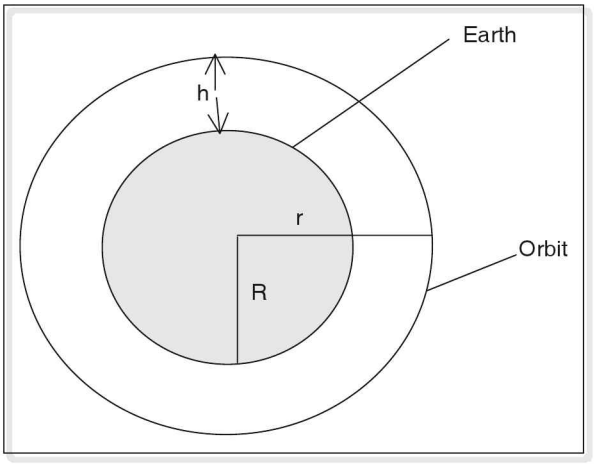


Figure 2.68

$\therefore$ Centripetal force = gravitational force between satellite and the planet.

$$\frac{mv_0^2}{r} = \frac{GMm}{r^2} \quad \text{--- (1)}$$

where M is the mass of the planet,

r is the radius of the orbit,

m is the mass of the satellite,

G is universal gravitational constant and

v_0 is the orbital velocity of the satellite.

$$\Rightarrow v_0^2 = \frac{GM}{r}$$

But r = R + h, where R is the radius of the planet and h is the height of the satellite from the surface of the planet.

$$\therefore\ v_0 = \sqrt{\frac{GM}{R+h}}$$

If h << R, then

$$v_0 = \sqrt{\frac{GM}{R}}$$

$$\text{or } v_0 = \sqrt{gR}\left(\because g = \frac{GM}{R^2}\right)$$

Thus the orbital velocity is independent of the mass of the satellite.

Escape velocity (v_e)

It is the minimum velocity required for a satellite or any body to escape from the earth's gravitational pull.

For any satellite to escape from earth's gravitational pull, work has to be done against gravitational force. This work done appears as kinetic energy of the satellite.

Thus kinetic energy of the satellite is equal to work done against gravitational force.

$$\therefore\ \frac{1}{2}mv_e^2 = \frac{GMm}{R^2} \times R$$

(work = force × displacement)

where v_e is the escape velocity M is the mass of the planet and R is the radius of the planet.

$$\therefore\ v_e = \sqrt{\frac{2GM}{R}} \text{ or}$$

$$v_e = \sqrt{2gR}\left(\because g = \frac{GM}{R^2}\right)$$

where g is acceleration due to gravity on the surface of the planet. Thus the escape velocity is same for all bodies because it is independent of mass of the body.

Substituting the value of g = 9·8 m s^{-2} and radius of the earth, R = 6400 km, it is found that the escape velocity of a body on the earth is 11.2 km per second. i.e., In order to escape from Earth's gravitational pull, an object should cover a distance of 11.2 km in one second.

Relation between orbital velocity and escape velocity

Orbital velocity of a satellite orbiting close to earth's surface is given by

$v_0 = \sqrt{gR}$ --- (1)

Escape velocity of a body on the surface of the earth is given by $v_e = \sqrt{2gR}$ -- (2),

where R is the radius of the earth.

From equation (1) and (2)

$$v_e = \sqrt{2}v_0 \text{ or } v_0 = \frac{1}{\sqrt{2}}v_e$$

Uses of artificial satellites

1. Communication satellite:

   Satellites are used for global communication, telephone communication, fax messages, TV transmission and internet.

   Communication satellites are geo-stationary satellites.

   A geo-stationary satellite appears to be at rest in space at a certain height above the equator. This satellite has the same period of revolution as that of the earth, i.e., revolvution per day, hence this satellite appears to be at rest with respect to earth.

   By placing three satellites in proper positions, the whole globe can be linked.

   INSAT series of satellites are geo-stationary satellites.

   Latest geo-stationary satellite launched by India is INSAT-3E. This was launched on 28th September 2003.

2. Weather forecasting:

   These satellites are used to give information about weather changes, about cyclones, hurricanes, floods etc.

3. Remote sensing satellites

   These satellites are used to identify underground water resources, mineral resources, etc, to find the areas of ocean, forest, deserts etc., to monitor air, water, sound and thermal pollution.

   IRS (Indian Remote sensing Satellite) is a remote sensing satellite.

   Bhaskara was the first IRS. It was launched in 1979. After Bhaskara many sophisticated remote sensing satellites are launched. These satellites are monitored at master control station at Hassan.

4. Satellites and space–stations are used to obtain the information of the planets, stars, galaxies, comets etc.

   Experiments conducted in such space shuttles and space station helps in growth of science and technology. Satellites are also launched by many countries for espionage.

ISRO

Indian space research organisation was started in 1969, but Indian space programme was started in 1963 at Thumba in Kerala with the launching of an American research rocket.

Space technology started in India in the 1970. Dr Homi J. Bhaba and Dr Vikram A. Sarabhai took India to the forefront in space technology. Space technology was further developed by Prof. Satish Dhavan, U.R. Rao and Dr Kasturi Rangan by launching various satellites.

The first Indian rocket was launched in November, 1971 from Thumba in Kerala. It was a sounding rocket which was used to study atmosphere.

Rohini RH–75 was a single stage rocket with only one scientific instrument. Now multi stage rockets are used.

The first Indian satellite Aryabhatta was launched in 1975 by a launch vehicle (rocket) called CI Inter Cosmos of USSR. This satellite weighed about 360 kg and was placed in an orbit at a height of 550 km. This satellite worked efficiently for six months.

Indian technology has been improved and the advancement of the technology has made it possible to build satellite launch vehicles. Four launch vehicles have been built and tested successfully. The launch vehicles are

1. ASLV - Augmented satellite Launch Vehicle
2. PSLV - Polar Satellite Launch Vehicle
3. GSLV - Geo-synchronous Satellite Launch Vehicle
4. SLV - Satellite Launch Vehicle

Space Vehicle Arians launched geo-stationary satellite APPLE in 1981 from French Guiana.

The latest Indian Geo-stationary satellite INSAT–3A and INSAT-3E were launched from French Guiana. INSAT-3A was launched in 19th April 2003. INSAT-3E weighs around 27 tonnes and has a life time of 15 years. This satellite is used in TV services and to provide telecommunication for Indian subcontinent.

Various centres of ISRO are located at Bangalore, Ahmedabad, Lucknow, Thumba and Thiruvananthapuram.

☛ *Example*

The time taken by Mars is 1.88 years to go round the Sun. Find the ratio of average distance between the Mars and the Sun to that between the Earth and the Sun.

Solution

From Kepler's third law:

$$\frac{T_1^2}{T_2^2} = \frac{d_1^3}{d_2^3}$$

where T_1 and T_2 are the periods of Mars and Earth respectively.

$T_1 = 1{\cdot}88$ years

$T_2 = 1$ year

$$\therefore \frac{T_1^2}{T_2^2} = \frac{d_1^3}{d_2^3}$$

$$\frac{(1.88)^2}{1^2} = \frac{d_1^3}{d_2^3} \quad \frac{3.534}{1} = \left(\frac{d_1}{d_2}\right)^3$$

Taking cube root $\frac{d_1}{d_2} = 1.52$

☛ *Example*

Two equal masses are separated by a certain distance and the force between them is 8 N. When the distance of separation is increased by 1m, the force between them reduces to 2 N. Find the initial distance of separation and their masses.

Solution

Let the mass of each body be = m kg

Given $m_1 = m_2 = m$ kg

Let the initial distance of separation be = d m

From universal law of gravitation the force between them is given by $F = \frac{G\, m_1\, m_2}{d^2}$

where G is the universal gravitational constant.

Let F_1 be the force when the bodies are separated by a distance d.

$$F_1 = \frac{G\, m_1\, m_2}{d^2} = \frac{Gm^2}{d^2}$$

Let F_2 be the force when the distance is increased by 1 m

$$F_2 = \frac{G\, m_1\, m_2}{(d+1)^2} = \frac{Gm^2}{(d+1)^2}$$

$$\frac{F_1}{F_2} = \frac{\frac{Gm^2}{d^2}}{\frac{Gm^2}{(d+1)^2}}$$

$$\Rightarrow \frac{8}{2} = \frac{(d+1)^2}{d^2} \Rightarrow 4 = \frac{(d+1)^2}{d^2}$$

Taking square root on both sides.

$$2 = \frac{d+1}{d}$$

$2d = d + 1$

or $d = 1$ m

Since $m_1 = m_2 = m$ and d= 1 m and G = $6{\cdot}67 \times 10^{-11}$ N m^2 kg^{-2}, using

$F = \frac{G m^2}{d^2}$ we get

$$8 = \frac{6.67 \times 10^{-11} \times m^2}{1^2}$$

$$\Rightarrow m^2 = \frac{8}{6 \cdot 67 \times 10^{-11}} \Rightarrow m^2 = \frac{8 \times 10^{11}}{6 \cdot 67}$$

$$\Rightarrow m^2 = 1 \cdot 199 \times 10^{11} = 0.1199 \times 10^{12}$$

$$m = \sqrt{0.1199 \times 10^{12}} = 0.346 \times 10^6 \text{ kg}$$

$$m_1 = m_2 = 0 \cdot 346 \times 10^6 \text{ kg}$$

☛ *Example*

What is the acceleration due to gravity on the top of Mount Everest, whose height is 8848 m, if acceleration due to gravity on the surface of the earth is 9·8 m s^{-2}?
Radius of the earth = 6400 km

Solution

In the given problem height h = 8848 = 8.848 km

Acceleration due to gravity $g^1 = g\left(1 - \frac{2h}{R}\right)$

$$g^1 = 9 \cdot 8\left(1 - \frac{2 \times 8 \cdot 848 \times 10^3}{6400 \times 10^3}\right)$$

$g^1 = 9 \cdot 8\ (1 - 0 \cdot 0028) = 9 \cdot 8 \times 0 \cdot 9972$
$g^1 = 9 \cdot 77 \text{ m s}^{-2}$

☛ *Example*

A satellite is to revolve round the earth in an orbit of height 8000 km. With what speed should the satellite be projected into the orbit? What will be its time period?

Solution

Radius of the orbit in which the satellite revolves r = R + h = 6400 km + 8000 km = 14400 km
where 'R' is the radius of the earth and h is the height of orbit of the satellite.

The orbital velocity of the satellite is given by $v_0 = \sqrt{\frac{GM}{R+h}} = \sqrt{\frac{R^2 g}{R+h}}$

Where 'g' is acceleration due to gravity on the surface of earth.

$$\therefore v_{0} = 6.4 \times 10^6 \sqrt{\frac{9.8}{14.4 \times 10^6}} \text{ m s}^{-1}$$

$$= 6.4 \times 10^6 \sqrt{0.68 \times 10^{-6}}$$

$= 6.4 \times 10^3 \times 0.825$
$= 5.28 \times 10^3 \text{ m s}^{-1}$
$= 5.28 \text{ km s}^{-1}$

Time period of the satellite, $T = \frac{2\pi r}{v_0}$

$\therefore T = \frac{2 \times \pi \times 14400 \times 10^3}{5280}$ s

$= 17127.3$
$= 285.5$ minutes
$= 4.76$ hours

test your concepts

Very short answer type questions

1. Define IMA and AMA.
2. Define angular displacement and angular velocity.
3. The functioning of a common balance is based on the principle of _____.
4. The area under the force–displacement graph gives _____.
5. Define linear momentum.
6. Why is the combination of fixed pulley and a moveable pulley preferred to one moveable pulley?
7. State Hooke's law.
8. A metre scale is suspended at 50 cm mark to a rigid support with a weight of 150 g attached at 5 cm mark. To balance the scale, a weight of 450 g should be attached at _____ cm mark.
9. Define work and power (write their dot product).
10. Define centripetal force.
11. A particle moves in a circle whose radius is 20 cm with a linear speed of 10 m s^{-1}. Then its angular velocity is _____ rad s^{-1}.
12. The weight of a body is minimum at ____ on the earth.
13. Define couple. Give an example of a couple.
14. What is the mechanical advantage of a sheath of pulleys?
15. Can Newton's law hold good in all frames of reference?
16. A body of mass 'm' is dropped from a height 'h'. The kinetic energy of the body on reaching the ground is _____.
17. What are parallel forces?
18. Define orbital velocity and escape velocity.

19. What is moment of a force or a torque?

20. Wheel and axle is a lever of _____ order.

21. What is centre of gravity?

22. What is the effect of weight of pulleys in the tackle?

23. Define centripetal acceleration.

24. What is meant by stable equilibrium?

25. Define uniform circular motion.

26. The mechanical advantage of an inclined plane 2 m long and 1 m high is equal to _____.

27. Give an expression which relates orbital velocity and escape velocity.

28. What is simple machine?

29. The escape velocity on the surface of the earth is _____ km s^{-1}.

30. Give the expression for mechanical advantage of a screw jack.

Short answer type questions

31. Obtain the relation between angular velocity and linear velocity.

32. Why is the efficiency of a multi-stage rocket greater than that of a single stage rocket?

33. Define angular momentum and give its expression.

34. Explain the working of centrifuge.

35. What is banking of roads and rails?

36. What is inertial frame of reference and non-inertial frame of reference?

37. A rocket with a mass of 20,000 kg is blasted upwards with an initial acceleration of 5 m s^{-2}. Calculate the initial force of the blast.

38. Mention the uses of artificial satellites.

39. State Newton's law of gravitation.

40. A swing is suspended from two parallel ropes at its ends. Two person weighing 50 kg_{wt} and 30 kg_{wt} sit at a distance 20 cm and 80 cm from one end. The weight of the swing is 20 kg_{wt}. Find the tensions on the rope. Length of the plank of the swing is 100 cm.

41. Why is centrifugal force is called pseudo force?

42. What is the relation between g and G?

43. A ball is tied to a string of length 1 m and is whirled in a horizontal circle with a linear speed of 5 m s^{-1}. If the mass of the ball is 1000 g. Find its

(i) angular velocity,

(ii) centripetal force acting on the ball and

(iii) centrifugal force acting on the ball.

44. What is the effect of mass on the acceleration of a rocket?

45. Obtain the relation between M.A, V.R. and η.

Essay type questions

46. Describe an experiment to determine the weight of a body.

47. Obtain an expression for banking angle.

48. Explain the working of wheel and axle.

49. Explain the principle of orbiting satellite.

50. Explain the working of a screw-jack.

CONCEPT APPLICATION

Concept Application Level—1

Direction for questions 1 to 7: State whether the following statements are true or false.

1. The work done by the earth on a satellite revolving around it is zero.

2. The escape velocity of a body on a planet is always greater than the orbital velocity of the body around the planet in an orbit closer to it.

3. The gravitational potential energy of a body is directly proportional to its height above the surface of the earth.

4. The mechanical advantage of an inclined plane is always greater than unity.

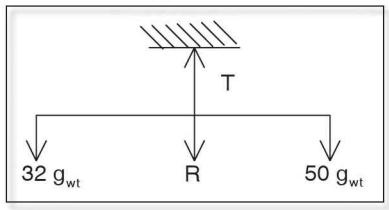
5. The resultant of all internal forces acting on a body at rest is zero.

6. Centrifugal force is a reaction to centripetal force.

7. A 10 N force is applied at a distance of 1 m from the hinge of door such that the line of action of force is parallel to the plane of the door. The moment of force produced on the door is 10 N m.

Direction for questions 8 to 14: Fill in the blanks.

8. Two forces act on either side of a rigid body of negligible mass suspended by string as shown in the figure. If R is the resultant force the tension of the string T = _____ g_{wt}.



9. The ratio of the weight of a body to its mass gives the _____ at the given place.
10. Two planets have average radii of their orbits of revolution in the ratio 64 : 9. Then the ratio of their period of revolution is _____.
11. Sugar crystals can be separated from molasses using a _____.
12. A particle moving along a circular path with the uniform angular velocity 'ω' is under the action of a force 'F' directed towards the centre of the circular path. If the radius of the circular path is doubled without changing the linear velocity, the centripetal force acting on the body would be _____.
13. The ratio of escape velocity of a tennis ball to that of a basket ball on the surface of the earth is _____.
14. If IMA and V.R. are the ideal mechanical advantage and velocity ratio of a pulley then its efficiency is _____.

Direction for question 15: Match the entries in column A with appropriate ones from column B.

15.

A.	Escape velocity	( )	a.	impulse
B.	Orbital velocity	( )	b.	$\frac{1}{2}mr^2\omega^2$
C.	Angular momentum	( )	c.	$\sqrt{gR}$
D.	Linear momentum	( )	d.	$\sqrt{2gR}$
E.	Centrifugal force	( )	e.	inertial frame of reference
F.	Kinetic energy	( )	f.	non-inertial frame of reference
G.	Change in momentum	( )	g.	mvr
H.	The frame of reference inside a moving satellite	( )	h.	mv
I.	Newton's laws are valid	( )	i.	M.A. = 1
J.	Single fixed pulley	( )	j.	$\frac{mv^2}{r}$

Direction for questions 16 to 30: For each of the questions, four choices have been provided. Select the correct alternative.

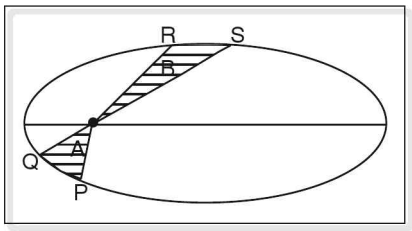
16. The escape velocity of a body on a planet depends on
(1) acceleration due to gravity on the planet (2) the radius of the planet
(3) the mass of the planet (4) All the above

17. The stability of a flexible body depends on
(1) height of the centre of gravity from the ground (2) base area of the body
(3) shape of the body (4) all the above

18. The relation between momentum (P), mass (m), kinetic energy (KE) and velocity (V) of two bodies is

(1) $P_1V_1 = P_2V_2$ where KE is constant
(2) $KE_1m_1 = KE_2m_2$ where P is constant
(3) $KE_1P_2^2 = KE_2P_1^2$ when m is constant
(4) All the above

19. In the above diagram the shaded regions A and B are the areas covered by a planet around the sun d_A and d_B, t_A and t_B are the distances traveled by the planet and the time taken by it to cover the paths PQ and RS respectively. Choose the correct statement.



(1) $d_A = d_B$ if $t_A = t_B$
(2) $d_A < d_B$ if $t_A > t_B$
(3) $d_A = d_B$ if $t_A < t_B$
(4) $d_A > d_B$ if $t_A = t_B$

20. The gravitational force of attraction between two bodies at a certain distance is 10 N. If the distance between them is doubled the force of attraction _____.

(1) decreases by 50%
(2) decreases by 75%
(3) increases by 50%
(4) increases by 75%

21. Action and reaction

(1) always act on two different bodies
(2) are equal in magnitude
(3) act in opposite directions
(4) All the above

22. Two particles of equal mass are revolving with the same linear speed on circular paths of radii r_1 and r_2. The ratio of the centripetal force acting on them will be _____.

(1) $\frac{r_2}{r_1}$
(2) $\sqrt{\frac{r_2}{r_1}}$
(3) $\left(\frac{r_1}{r_2}\right)^2$
(4) $\left(\frac{r_2}{r_1}\right)^2$

23. Choose the correct statement.

(1) In a simple machine, the input energy is always equal to output energy.
(2) Actual mechanical advantage > Ideal mechanical advantage.
(3) Velocity ratio = $\frac{\text{displacement of the effort}}{\text{displacement of the load}}$
(4) The mechanical advantage of a single fixed pulley is greater than one.

24. While opening a tap with two fingers, the forces applied are

(1) equal in magnitude
(2) parallel to each other
(3) opposite in direction
(4) All the above

25. A rubber ball of mass 250 g hits a wall normally with a velocity of 10 m s^{-1} and bounces back with a velocity of 8 m s^{-1}. The impulse is _____ N s.

(1) –0.5
(2) +0.5
(3) –4.5
(4) +4.5

26. The false statement about universal gravitational constant (G) among the following is _____.

(1) Its dimensional formula is $[M^{-1} L^3 T^{-2}]$
(2) It is a vector quantity
(3) Its value is 6.67×10^{-11}
(4) Its unit is N m^2 kg^{-2}

27. The weight of an object is zero at _____.

(1) a height equal to half of the radius of earth
(2) the centre of the earth
(3) the equator
(4) both (1) and (2)

28. A force acts on an object while in the state of motion. If the force acting on it is equal to the sum of all the resistance forces, then the object _____.

(1) stops after covering some distance
(2) moves with uniform acceleration
(3) moves with uniform velocity
(4) moves with uniform retardation

29. In the case of an oscillating simple pendulum if the air resistance is taken into consideration, then _____.

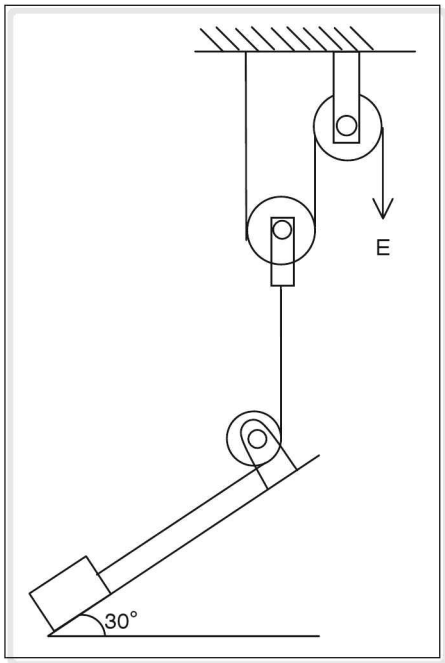
(1) amplitude of oscillation remains the same
(2) frequency remains the same
(3) the resultant acceleration remains the same
(4) All the above

30. The magnitude of acceleration due to gravity decreases

(1) as the height from the surface of the earth increases
(2) as the depth from the surface of the earth increases
(3) as one moves from the pole of the earth to its equator
(4) All the above

Concept Application Level—2

31. An object of mass 50 kg is lifted by using an arrangement as shown in the figure. If the load due to the weight of the pulley and the resistance due to the friction of the pulley is 2.5 kg_{wt}, calculate the effort required to lift the object. If $M.A_1$ and $M.A_2$ are the mechanical advantages of the inclined plane and the pulleys respectively, derive an expression for the resultant mechanical advantage if the friction force between the object and the inclined plane is negligible.



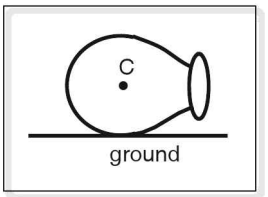
32. A load of 60 kg is lifted using a wheel and axle, the sum of whose diameters is 1.2 m, by a man and his son who can share their efforts in the ratio 2 : 1. If the radius of the axle is 10 cm, find the effort contributed by the son. ($g = 10\ m\ s^{-2}$).

33. An object of certain mass is dropped from the top of a building of height 10 m. At a certain point on its path the ratio of its potential energy to its kinetic energy is found to be 1:3. The work done by the gravitational force of the earth in moving the body up to this point is 225 J. Determine the mass of the object and the kinetic energy when it reaches the ground. (Take $g = 10\ m\ s^{-2}$)

34. A stone is tied to a massless and inextensible string and rotated in a vertical plane. Discuss the variation in centripetal force of the stone from the top to the bottom positions of the circular path if minimum speed is imparted to it so as to continue its circular motion.

35. A scooterist takes a turn through the outer edge of a road of width 10 m. The safe velocity and the radius of curvature of the road are 4.9 m s^{-1} and 24.5 m respectively. Calculate the speed with which he should drive the scooter to avoid skidding.

36. A metallic vessel having a shape that is a combination of a hemisphere and a truncated cone is placed on the ground in a position as shown in the figure. 'C' is the centre of curvature of the hemispherical part of the vessel. The height of the truncated conical part is less than the radius of curvature of the hemispherical part of the vessel.



(i) Locate the centre of gravity of the vessel.
(ii) What is the state of equilibrium of the vessel in the position shown?
(iii) When the vessel is left in the position shown, what is the effect of the forces acting on the vessel? Discuss.

37. The frame of reference attached to a bus can be either an inertial or a non-inertial frame of reference. Explain with suitable example when it can be inertial and when non-inertial frame of references.

38. A spring balance whose maximum extension of its spring is 20 cm can sustain a maximum load of 20 kg_{wt} within its elastic limit on the surface of the earth. Now the same balance is taken to the moon. Given that $g_{earth} = 6\ g_{moon}$, what is the maximum mass of the load that can be attached to the balance to have half its maximum extension?

39. A body is projected vertically upwards from ground with a velocity 'u' and it gains a maximum height 'H'm. At a point 'x'm height ($x < H$), the ratio of its potential to Kinetic energy is 1 : 2. Now if the body is projected upwards with a velocity '3u' from the ground, find the ratio of its potential to kinetic energy at a point whose height is '3x'm.

40. The magnitude of acceleration due to gravity at an altitude 'h' from earth is equal to its magnitude at a depth 'd'. Find the relation between 'h' and 'd'. If the 'h' and 'd' both increase by 50%, are the magnitudes of acceleration due to gravity at the new altitude and the new depth equal?

41. A ball of mass 500 g hits a wall with a velocity 10 m s^{-1} such that it makes an angle 60° with the normal. It bounces back with the same speed making an angle of 60° with the normal. Determine the change in momentum.

42. Why cannot an artificial satellite be parked in its orbit using a single stage rocket instead of using a multistage rocket for the purpose?

43. Find the power of a machine gun that can fire 240 bullets per minute, each weighing 50 g, with a velocity of 100 m s^{-1}.

44. A spirit level cannot be used to find out whether the top of the table placed on a rotating platform is horizontal. Explain.

45. An object thrown in the downward direction from a height 10 m with an initial velocity 15 m s^{-1} strikes the ground and bounces back to a height of 12 m. Determine the percentage loss of its energy during collision with the ground. (Take g = 10 m s^{-2})

Concept Application Level—3

46. When 'n' number of springs of spring constants $k_1, k_2, \ldots\ldots k_n$ are connected in parallel show that the effective spring constant $k = k_1 + k_2 + \ldots.. k_n$.

47. A stone of mass 'm' falls from a height 'h' into sand pit and moves down into the sand through a distance 's' before it comes to the state of rest. Calculate the resistance force offered by the sand on the stone.

48. A bullet of mass 50 g is fired into a wooden block of mass 1.95 kg placed on a horizontal wooden surface. On striking, the bullet is embedded into the block and both together cover a distance of 200 m in a straight path before coming to rest. Given the coefficient of friction between the block and the surface is 0.2 and the mass of the rifle from which the bullet is triggered is 14 kg, find the magnitude of the velocity of the recoil of the rifle.(Neglect the resistance of air on the bullet).

49. An object is suspended from the ceiling of a lift by means of a string. When the lift starts moving in the upward direction the string breaks and the object falls into the lift. Explain giving reasons.

50. A particle loosely held at the circumference of a platform rotating with a uniform velocity flies off during the rotatory motion. Explain the kind of motion it undergoes,

(i) as seen by an observer on the ground.

(ii) as seen by an observer on the platform.

key points for selected questions

Very short answer type questions

1. $IMA = \frac{\text{Total load}}{\text{effort}}, AMA = \frac{\text{useful load}}{\text{effort}}$

2. Angular displacement – change in angular position.

Angular velocity – rate of change angular displacement.

3. moments

4. work

5. Product of mass and velocity

6. To apply effort conveniently

7. With in elastic limit, stress is directly proportional to strain.

8. 65

9. Work is product of force and displacement. Power is rate of doing work.

$W = \vec{F}.\vec{d}$, Power $= \vec{F}.\vec{v}$

10. The force directed radially inward

11. 50

12. equator

13. Two equal, parallel, unlike forces acting on the same body at different points.

14 MA = n (n = number of pulleys in sheath)

15. No

16. mgh

17. Two forces acting along parallel lines.

18. Orbital velocity – Minimum velocity to make a satellite revolve around the planet

Escape velocity – Minimum velocity to make a body escape from earth's surface.

19. Product of force and perpendicular distance.

20. first

21. The point where the total weight of a body acts.

22. They decrease the efficiency

23. The acceleration directed towards the centre of circular path.

24. A body regains its original position after application of slight force.

25. Angular velocity is constant

26. 2 and cosec 30°

27. $V_e = \sqrt{2}\, V_o$

28. A device which allows application of force conveniently to overcome force at on other point.

29. 11.2

30 M.A. $= \dfrac{2\pi\ell}{h}$

Short answer type questions

31. $V = r\omega$

32. Multistage rockets discard the chambers of propellants which have been used.

33. $\vec{L} = \vec{r} \times \vec{p} = mvr = mr^2\omega$

34. Centrifugal force $= \dfrac{mv^2}{r} = m\omega^2 r$

35. The angle made by the line joining the outer edge to the inner edge with the horizontal line is called the 'angle of banking'

$\tan\theta = \dfrac{v^2}{rg}$

36. The reference frames in which a body has no acceleration and possesses acceleration

37. (i) Force required

(ii) Gravitational pull

(iii) Upward force

(iv) Resultant force

Ans: net force $= 2{\cdot}96 \times 10^5$ N

38. Communication links, remote sensing etc.

39. $F = G\dfrac{m_1 m_2}{r^2}$

40. (i) Theorem or principle of moment

(ii) Clockwise moment

(iii) Equilibrium

Ans: $T_2 = 44\ kg_{wt}$

$T_1 = 56\ kg_{wt}$

41. It is invisible force.

42. $g = \dfrac{GM}{R^2}$

43. Ans: $\omega = 5\ rad\ s^{-1}$

Centripetal force = 25 N

Centrifugal force = 25 N

44. As fuel burns the mass of the rocket decreases and its acceleration increases.

45. Efficiency $\eta = \dfrac{M.A.}{V.R.}$

"Efficiency is defined as the ratio of M.A. to V.R".

Essay type questions

46. $W = mg$

Spring balance

Weights

Pointer

Extension

47. $\tan\theta = \dfrac{v^2}{rg}$

48. Force multipliers

49. Horizontal speed

Curved path

50. Pitch

KEY

Concept Application Level—1

True or false

1. True
2. True
3. False
4. True
5. False
6. False
7. False

Fill in the blanks

8. 82
9. acceleration due to gravity.
10. 512 : 27
11. centrifuge
12. $\frac{F}{2}$
13. 1 : 1
14. $\frac{\text{I.M.A.}}{\text{V.R}}$

Match the following

15. A : d
    B : c
    C : g
    D : h
    E : j
    F : b
    G : a
    H : f
    I : e
    J : i

Multiple choice questions

16. Choice (4)
17. Choice (4)
18. Choice (4)
19. Choice (4)
20. Choice (2)
21. Choice (4)
22. Choice (1)
23. Choice (3)
24. Choice (4)
25. Choice (3)
26. Choice (2)
27. Choice (2)
28. Choice (3)
29. Choice (2)
30. Choice (4)

Concept Application Level—2, 3

Key points

31. (i) For inclined plane MA = $\frac{1}{\sin\theta}$

    (ii) For single movable pulley, MA = 2

    (iii) 13.75 kg_{wt}, $MA_1 \times MA_2$

32. (i) From given information, find the load in newtons.

    Find the radii of wheel and axle.

    The mechanical advantage (M.A) of wheel and axle is, M.A = $\frac{R}{r} = \frac{\text{Load}}{\text{Effort}}$.

    Find the value of effort.

    Here, the son contributes one-third of the total effort.

    Then find the value of effort contributed by the son.

    (ii) 40 N

33. (i) Work done by net force = change in kinetic energy.

    (ii) For a freely falling body, loss of potential energy = gain of kinetic energy

    (iii) 9 kg, 900 J

34. At the top position of the circular path, is centrifugal force, $F = \frac{mv^2}{r}$ equal to the weight of the stone and tension (T) in the string.

At the bottom position of the circular path, the tension (T) in the string is equal to the weight of the stone and centrifugal force (F).

35. (i) $\tan\theta = \frac{v^2}{rg}$

When he moves along the outer edge, radius of curvature

$$r_1 = r + \frac{\text{width of the road}}{2}$$

(ii) $\sqrt{28.91}\ m\ s^{-1}$

36. Which portion of the metallic vessel (i.e.,) a hemisphere or a truncated cone, contains more mass?

Hemispherical portion has more metal. So, centre of gravity (G) of the vessel lies near the bottom.

Find whether the height of the centre of gravity remains constant in any position where the hemisphere part touches the ground.

Then find the state of equilibrium of the vessel.

Here, centre of gravity 'G' lies near to the base of the vessel i.e. near the hemisphere. Then find the direction of moment of forces acting on the metallic vessel with respect to C.

37. (i) Frame of reference that is at rest or moving with uniform velocity is inertial frame of reference.

(ii) Frame of reference that moves with acceleration is non - inertial frame of reference.

38. (i) Find the maximum load (i.e.),, 20 kg_{wt}, in terms of newtons.

Convert 20 cm length in terms of S.I.

Find the value of force constant (k) from the formula, $F = kx$ ----- (1)

Now, find the maximum load that is possible on the surface of the moon if $F = mg_m = kx$ (g_m acceleration due to gravity on the moon).

Substitute the value of $\frac{g_E}{6}$ for $g_{m.}$ ----- (2)

Now, find the mass of the load that is attached to have its maximum extension ---- (3)

Divide the value obtained in (3) by (2).

Then this gives maximum mass of the load that can be attached to the balance to have half its maximum extension.

(ii) 60 kg

39. (i) Find the ratio of potential to kinetic energy at a height 'x' from the ground for the body projected vertically upwards from the ground with a velocity 'u'.

Take the mass of the body as 'm'.

Find the value 'x' in terms of 'u' and "g".

For the same body, when projected up with velocity 3u, the ratio of potential to kinetic energy is at a height '3x' from the ground.

Find the velocity of the body at height '3x' in terms of u, g and x using the formula $v^2 = u^2 - 2gh$, and then simplify.

(ii) 1 : 8

40. (i) Here, acceleration due to gravity at height 'H', from earth is $g_h = g_o\left(1 - \frac{2h}{R}\right)$.

Then, acceleration due to gravity at a depth 'd' from earth is $g_d = g_o\left(1 - \frac{d}{R}\right)$.

If $g_h = g_d$, then find the relation between 'h' and 'd' – (1)

If the 'g' and 'd' both increase by 50%, then

$h = h + \frac{h}{2}$ and $d^1 = d + d/2$.

Substitute the value obtained from (1) and find whether $g_{h'} = g_{d.}$

(ii) $d = 2h$

41. (i) Resolve momentum into horizontal and vertical components.

(ii) Change in momentum is due to change in vertical component of the momentum.

(iii) -5 kg m s^{-1}

42. If the mass of the satellite is more, is it easily possible to lift it to higher altitudes?

Find the difference between the single stage rocket and multistage rocket.

43. (i) How many bullets are fired from the machine gun in one second?

What is the kinetic energy of the bullet?

The work done by the gun in one second is equal to the total amount of kinetic energy of the bullets fired in one second.

Find the value of the power.

(ii) 1 kW

44. What is centripetal force?

Will heavier particle move towards the centre of curvature?

Will air bubble move away from the centre of curvature?

45. (i) Total energy $= \frac{1}{2} mv^2 + mgh_1$

Change in energy $= \left(\frac{1}{2}mv^2 + mgh\right) -$

(mgh_2)

% change in energy

$$= \frac{\left(\frac{1}{2}mv^2 + mgh_1\right) - (mgh_2)}{\frac{1}{2}mv^2 + mgh_1} \times 100$$

(ii) 43.5%

46. (i) When springs are connected in parallel.

$\vec{F} = \vec{F}_1 + \vec{F}_2 + \vec{F}_3$

(ii) $k = k_1 + k_2 + k_3 \ldots . k_n$

47. (i) The net force acting on the stone as it moves in sand

$ma = mg - R.F$

$R.F. = mg - ma$

$v^2 - u^2 = 2as$

$\Rightarrow v^2 = 2gh$. This is the initial velocity in the sand. Using this find 'n' in sand.

(ii) $mg\left(1 + \frac{h}{s}\right)$.

48. (i) Find the masses of bullet (m_b) and block (m_w) in C.G.S. system.

According to the law of conservation of momentum:

$m_1 u_1 + m_2 u_2 = m_1 v_1 + m_2 v_2$.

Then, take initial velocity of bullet and block as 'u' and 'zero'.

When the bullet is embedded into the block, both have a common velocity.

Let 'v' be the common velocity.

Find the relation between 'u' and 'v'.

By law of conservation of momentum, when both bullet and block together cover 200 m before coming to rest, then, the frictional force, $F = \mu (m_b + m_w) g$ ----(1)

Find the work done to overcome the friction.

The kinetic energy of the system
$= \frac{1}{2} (m_b + m_w) v^2$ = work done

Then, $(m_b + m_w) v^2 = \mu (m_b + m_w)$
$g s$ --- (2)

Find the value of "v" from (2).

Find the mass of the gun m_g from given information.

Apply the law of conservation of momentum.

$u_g = u_b = o$

Then, $m_g u_g + m_b u_b = m_g v_g + m_b u_b$ ---- (3)

Find the value 'v_g' from (3)

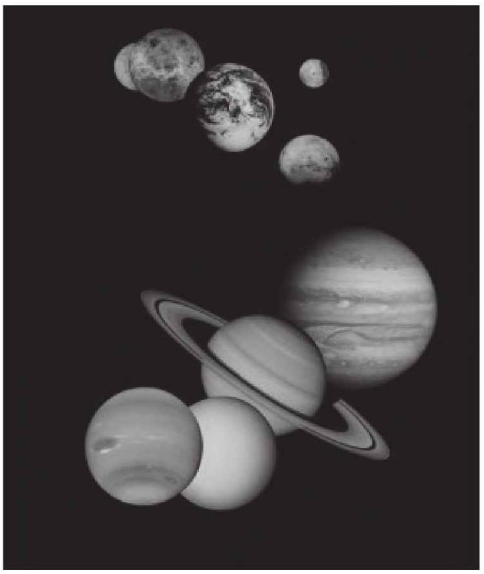
(ii) 4 m s^{-1}

49. When the lift start moving in the upward direction with acceleration 'a' then T = m (g + a).

50. (i) How the motion appears from inertial and non inertial frames of reference?

(ii) Direction of velocity is along the tangent.

(iii) How the centrigal force acts on the body?



3

Heat

INTRODUCTION

In the earlier classes, we have learnt that heat is a form of energy, and that some of the other forms of energy can be converted to heat while heat itself can be converted to some of the other forms of energy. Similarly, by the work-energy theorem we understand that work done on a body can appear in the form of heat.

For example, when we rub our palms the biochemical energy is expended to do the 'work' which results in the warmth in the palms.

Thus we see that when some form of energy is converted to heat energy, it makes a substance or a body 'hotter'. This degree of hotness (or coldness) is called temperature.

Furthermore, we have also seen that when hot lead shots are dropped into cold water, the temperature of the lead shots decreases while the temperature of the water increases, indicating that heat energy always flows from hot bodies to cold bodies, and this flow of heat stops when the temperatures of the two bodies are equal—a condition known as thermal equilibrium.

To elaborate on this let us take another example. If an ice-cube at 0°C is dropped into a vessel containing water at 0°C, the ice-cube does not melt, since the two substances are at the same temperature and heat does not flow from water to ice even though water has more heat content (336 J g^{-1}) than ice.

Thus, while heat is a form of energy, rise (or fall) of temperature is one of its effects. Another important effect of heat is that, on supply of heat, matter changes its state—ice melts to become water, a liquid boils to vaporise. Similarly when heat is taken away from a hot substance, it can also result in change of state – water freezing to form ice, steam condensing into water.

We have also studied that the thermal expansion is another effect of heat and solids as well as liquids and gases expand on heat.

We shall now study calorimetry – measurement of heat energy in greater detail. We will also direct our attention to working principle of heat engines, an important application of thermal energy.

Heat

Every body contains some amount of heat. This is equal to sum total of average kinetic energy and potential energy of molecules present in the body.

When heat is supplied to a body, average kinetic energy of its molecules increases, thus increasing its temperature. We say that the body has become hot. The opposite happens if heat is removed from a body. Then we say that the body has cooled. Thus heat is a form of energy which gives us sensation of hotness or coldness.

Units of heat

Since heat is a form of energy, its units are same as that of other forms of energy.

S.I. unit of heat is joule (J).

Calorie (cal) is another unit which is specially used for heat.

One calorie is the amount of heat required to raise the temperature of 1 g of water by one degree celsius.

The two units of heat are related as below

1 cal = 4.1863 J ≅ 4.2 J

1 kcal = 4186.3 J ≅ 4200 J

where 1 kcal = 1000 cal is a higher unit of heat.

Heat capacity

To increase the temperature of a body, we have to supply heat to it. But the rise in temperature is not the same for all bodies for the same amount of heat supplied.

If we supply the same amount of heat to 400 g and 700 g of water we find that rise in temperature is more in the former than in the latter case.

If equal masses of water and milk are supplied with the same amount of heat the rise in temperature of milk is more than that of water.

This rise in temperature and heat supplied to a body are related by a term called heat capacity or thermal capacity of body. It is denoted by 'C'.

It is defined as an amount of heat needed to raise the temperature of a given mass of a body by 1 K or 1°C.

It is equal to the ratio of total heat supplied to the total rise in temperature of a body.

Let Q joule of heat be supplied to a body so that its temperature rises by $\Delta\theta$ K. Then

heat capacity (C) of the body $= \frac{Q}{\Delta\theta}$

$\therefore C = \frac{Q}{\Delta\theta}$

Thus units of heat capacity $= \frac{\text{Unit of heat}}{\text{Unit of temperature}}$

Hence S.I. unit of heat capacity is $\frac{\text{joule}}{\text{kelvin}}$ ($J\ K^{-1}$)

Other common units of heat capacity (C) are

$J\ °C^{-1}$, $cal\ °C^{-1}$ or $kcal\ °C^{-1}$

$1\ cal\ °C^{-1} = 4.2\ J\ K^{-1} = 4.2\ J\ °C^{-1}$

$1\ kcal\ °C^{-1} = 4200\ J\ K^{-1} = 4200\ J\ °C^{-1}$

It can be seen that here °C and K are interchangeable since we consider only the change in temperature.

Factors affecting heat absorbed or given out by a body

Consider two similar containers A and B containing half a litre and one litre of water respectively. The initial temperatures of water in both the containers (which is equal to room temperature, say θ_1) is noted with the help of a thermometer. Both these containers are placed on two similar burners that supply heat at a constant rate; and heated. To attain an equal higher temperature, say θ_2, it is observed that the container B takes more time compared to the time taken by the water in the container A. Thus heat absorbed by B is greater than that absorbed by A. This implies that larger the mass of a body, greater is the heat absorbed or given out by it.

Now consider two similar containers P and Q having equal volumes of water. Thus the mass of water in both the containers is equal. Now both the containers are placed on burners that supply heat at a constant rate and heated. If the required rise in temperature of P is more than that of Q, it is observed that P has to be heated for more time compared to Q. Thus the heat absorbed by P is greater than that absorbed by Q. This implies that the amount of heat absorbed or given out by a body depends on the change in its temperature.

The experimental results shows that

1. Heat absorbed or given out by a body is directly proportional to the mass of the body
   $Q \propto m$ --- (1)
2. Heat absorbed or given out by a body is directly proportional to the rise or fall in temperature of the body
   $Q \propto \Delta\theta$ --- (2)

Combining equation (1) and (2), we get $Q \propto m\ \Delta\theta \Rightarrow Q = m\ s\ \Delta\theta$ where s is a constant of proportionality which we shall discuss later.

☛ *Example*

1 g of water needs 4.2 J and 100 g of water needs 420 J of heat energy respectively so as to raise their temperature by 1°C. Calculate the heat capacity of water in both cases.

Solution

Heat capacity (C) = $\frac{\text{heat}}{\text{rise in temperature}}$

1st case

For 1 g of water

$C = \frac{4.2\ J}{1°C} = 4.2\ J\ °C^{-1}$

2nd case

For 100 g of water

$C = \frac{420\ J}{1°C} = 420\ J\ °C^{-1}$

Specific heat capacity

From the above example it is seen that water has different heat capacities for different masses. This is observed in all the substances. As the mass changes, heat capacity also changes. But for a given mass of a given substance, it is constant. Therefore heat capacity of a unit mass of substance is taken as reference and is called specific heat capacity. It is denoted by 'c' or 's'

Specific heat capacity is defined as the amount of heat required to raise the temperature of unit mass of substance (1 kg or 1 g) by one kelvin or one degree celsius. Heat absorbed or given out by a body is given by

$Q = ms\Delta\theta$

Thus the constant of proportionality in the above equation is the specific heat capacity of the given body.

Units of specific heat capacity

From the above equation we get $s = \frac{Q}{m\Delta\theta}$. Thus unit of $s = \frac{\text{Unit of Q}}{(\text{Unit of m})\ (\text{Unit of } \Delta\theta)}$

S.I. unit of specific heat capacity is $\frac{\text{joule}}{\text{kilogram} \times \text{kelvin}} = J\ kg^{-1}\ K^{-1}$.

As change in temperature in kelvin and celsius scales is equal, it can be written as $J\ kg^{-1}\ C^{-1}$.

Other common units of specific heat are $cal\ g^{-1}\ °C^{-1}$ or $kcal\ kg^{-1}\ °C^{-1}$.

$1cal\ g^{-1}\ °C^{-1} = 4.2\ J\ g^{-1}\ °C^{-1}$ and $1kcal\ kg^{-1}\ °C^{-1} = 4200\ J\ kg^{-1}\ °C^{-1}$

The specific heat capacity of a substance is independent of its mass. It is a characteristic property of a substance. Its value is different for different substances.

Water	1.00	4.20	4200
Alcohol	0.52	2.18	2180
Kerosene Oil	0.50	2.10	2100
Turpentine Oil	0.42	1.76	1760
Aluminium	0.21	0.882	882
Flint glass	0.12	0.504	504
Iron	0.115	0.483	483
Copper	0.095	0.399	399
Zinc	0.093	0.391	391
Brass	0.092	0.386	386
Silver	0.056	0.235	235
Tin	0.055	0.231	231
Mercury	0.033	0.139	139
Gold	0.032	0.134	134
Lead	0.031	0.130	130

$Q = mc\Delta\theta \Rightarrow c = \frac{Q}{m\Delta\theta}$

$\Rightarrow \frac{Q}{\Delta\theta} = mc = C$ (heat capacity)

$\Rightarrow$ Heat capacity = mass $\times$ specific heat capacity

☛ *Example*

10 g of kerosene oil absorbs 210 J of heat to raise its temperature by 10°C. Find its specific heat capacity.

Solution

Given mass(m) of the kerosene oil = 10 g

Heat absorbed (Q) = 210 J

Rise in temperature ($\Delta\theta$) = 10°C

Specific heat capacity (s) = ?

$Q = ms\Delta\theta$

$\Rightarrow 210\ J = 10\ g \times s \times 10°C$

$\therefore s = \frac{210\ J}{10g\ \ 10°C} = 2.1\ J\ g^{-1}\ °C^{-1}$

☛ *Example*

The specific heat capacity of water is 4.2 J g^{-1} $°C^{-1}$. Calculate the heat capacity of 20 g of water. How much heat has to be supplied to it to raise the temperature by 10°C?

Solution

Given

Specific heat capacity (s) = 4.2 J g^{-1} $°C^{-1}$

mass (m) = 20 g

Rise in temperature ($\Delta\theta$) = 10 °C

Heat capacity for 20 g(C) = ?

Heat supplied (Q) = ?

C = ms = 20 g × 4. 2 J g^{-1} $°C^{-1}$

C = 84 J $°C^{-1}$

$$C = \frac{Q}{\Delta\theta} \Rightarrow Q = C\,\Delta\,\theta$$

= 84 J g^{-1} $°C^{-1}$ × 10°C

$\Rightarrow$ Q = 840 J.

Water has highest specific heat capacity

It is 1 cal g^{-1} $°C^{-1}$ (4.2 J g^{-1} $°C^{-1}$) or 1 kcal kg^{-1} $°C^{-1}$ (4200 J kg^{-1} $°C^{-1}$)

Thus one gram of water needs 1 cal or 4.2 J of heat to raise its temperature by 1°C i.e., the temperature of water doesn't rise or fall quickly as compared to that of other substances. Hence we find water bodies to be comparatively cool even in sweltering heat of summer.

The high specific heat capacity of water gives rise to some natural phenomenon and has many advantages in our day to day life. Let us discuss some consequences of it.

1. Land breeze and sea breeze: The specific heat capacity of water is nearly five times that of land. This means water needs nearly five times more heat than land to raise its temperature by 1°C. Similarly water has to lose five times more heat for 1°C fall in temperature. Thus land heats up and cools five times quicker than water. In coastal areas this gives rise to cool sea breeze from sea to land during day time and land breeze from land to sea during night. This keeps the temperature in coastal areas moderate.
2. We apply hot water bag to the swollen parts of our body when sprained to get relief from pain. This is called fomentation as water contains very high amount of heat energy at comparatively low temperature.
3. Water is used as coolant in number of applications. It can absorb large amount of heat without its temperature increasing appreciably. Hence it is used as a coolant in car radiators, nuclear plants and in industries.
4. In cold countries during winter, farmers water their fields in the evening. Owing to its high specific heat capacity, temperature of water doesn't fall quickly. In the absence of external water, due to drop in atmospheric temperature, water in fine capillaries and veins of plants freezes and capillaries burst due to increase in volume, destroying the crop.

☛ *Example*

A 250 W heater is used to heat 500 g of mercury. The heater is on for 1 minute and the initial temperature of mercury is 20°C. Calculate the final temperature of mercury. [Specific heat capacity of mercury = 0.14 J g^{-1} $^{\circ}C^{-1}$]

Solution

Given,

Power of heater (P) = 250 W

Heat energy given by heater = P × t = 250 W × 1 minute = 250 W × 60 s

Heat absorbed by mercury = $ms\Delta\theta$

m = 500 g

$\Delta\theta$ = 20°C

Let x = final temperature of mercury.

Heat given by heater = heat absorbed by mercury.

$\therefore$ 250 W × 60 s = 500 g × 0.14 J g^{-1} $^{\circ}C^{-1}$ × (x − 20) °C

$$\frac{250 \times 60}{500 \times 0.14} = x - 20$$

$\therefore x$ = 234.28°C.

Principle of Mixtures

Take few lead shots whose temperature is 70°C and add them to water at 30°C contained in an insulating container and stir well. The temperature of lead shots decreases and that of water rises. This means lead shots lose heat and water gains heat. If no heat is lost to surroundings then heat lost by lead shots is equal to the heat gained by water. This is the basis of principle of mixtures which states that when a hot body is mixed with a cold body, the hot body loses heat and the cold body gains heat till they are in thermal equilibrium and the heat lost by the hot body is equal to the heat gained by the cold body, provided heat is not lost to the surroundings. This principle is in accordance with the principle of conservation of energy and is used to determine specific heat capacity of solids and liquids.

Calorimeter

Heat is a form of energy. Like other forms of energy, it can be measured. The branch of physics which deals with the measurement of heat is called calorimetry.

The device which is used to measure quantity of heat is called calorimeter. It uses the principle of mixtures to find the specific heat capacity of different bodies. Hence the method is also called the method of mixtures.

Calorimeter consists of a thin cylindrical copper vessel. Its outer and inner surfaces are highly polished to prevent heat loss by radiation. The copper vessel is kept in rectangular wooden box and space between them is filled with materials like asbestos or cotton wool which are good insulators of heat. This prevents heat loss by conduction.

At its top, the box has a wooden lid with two holes, one for inserting a thermometer and the other for a stirrer which is also made of copper. The lid is coated with tin or aluminum foil.

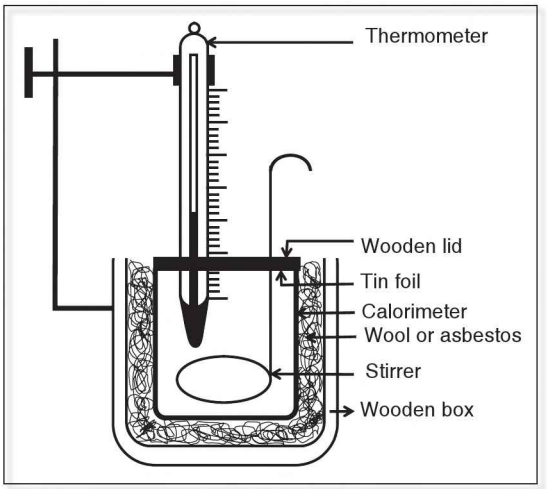


Figure 3.1

According to the principle of calorimetry or method of mixtures, heat lost by the hot body = heat gained by the cold body.

Let

m_1 = mass of hot body.

m_2 = mass of cold body.

θ_1 = temperature of hot body.

θ_2 = temperature of cold body.

θ = temperature of both bodies under thermal equilibrium.

s_1 = specific heat capacity of hot body.

s_2 = specific heat capacity of cold body.

Fall in temperature of hot body = $\theta_1 - \theta$

Rise in temperature of cold body = $\theta - \theta_2$

Heat lost by hot body = $m_1 s_1(\theta_1 - \theta)$

Heat gained by cold body = $m_2 s_2 (\theta - \theta_2)$

$\therefore m_1 s_1(\theta_1 - \theta) = m_2 s_2(\theta - \theta_2)$

To determine specific heat capacity of a given solid by method of mixtures

A solid whose specific heat capacity is to be measured is taken in the form of small pieces. It is first heated separately and then added to calorimeter which contained cold water. Devices like Sinclair's heater and hypsometer are suitable for this purpose.

Sinclair's heater

It consists of a cylindrical double walled copper vessel fitted with an outlet tube as shown in figure.

Water is heated at the lower portion and steam is passed out through steam outlet. Through an opening at the top, a tube is introduced in the outlet tube. This tube is open at both ends. The bottom open end of the tube is tappered so that it fits well at the slope of the outlet tube. The solid which is to be

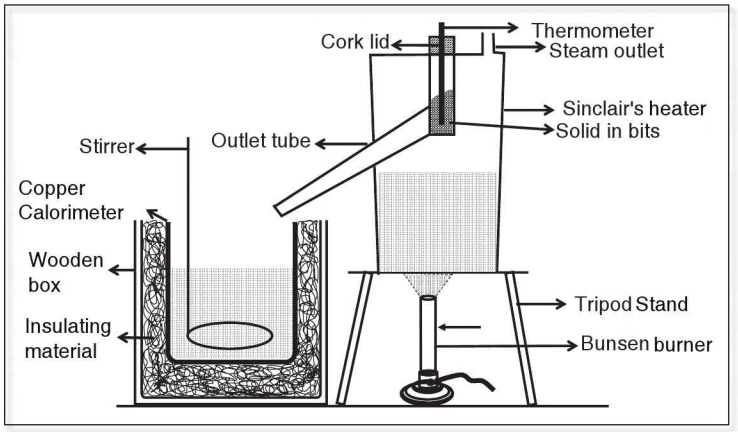


Figure 3.2 Sinclair's heater

heated is introduced in the tube in the form of small pieces. To measure its temperature, a thermometer is inserted in the tube through a rubber cork. Care should be taken that the bulb of the thermometer is completely immersed in the solid pieces. The cylindrical vessel is now heated by bunsen burner and temperature of solid pieces is noted. The solid gets heated as it is surrounded by steam formed by boiling water. This ensures uniform heating. A wooden partition (not shown in the figure) is kept on the side of this arrangement to prevent heat radiations from reaching the calorimeter into which the hot solid pieces have to be transferred. Once the temperature of solid is sufficiently high and steady, heating is stopped and calorimeter is brought below the outlet tube. The open ended tube along with thermometer is lifted and heated solid pieces are allowed to fall into calorimeter and calorimeter is then closed quickly.

While the solid whose specific heat capacity is to be measured is being heated in the sinclair's apparatus, the following procedures can be completed.

1. Take dry, clean and highly polished calorimeter and weigh it along with its stirrer using a physical balance. Let it be m_1 g.
2. Now pour water into it at room temperature till it is half full and weigh it again. Let it be m_2 g.
3. Put calorimeter in wooden box and note the temperature of water. Let it be θ_1°C.
4. Note the temperature of hot solid pieces let it be θ_2°C. Introduce them into the water in the calorimeter and close the wooden lid quickly to prevent heat loss.
5. Stir the mixture well with stirrer and note the steady temperature of mixture. Let it be θ°C.
6. Remove calorimeter from the wooden box and weigh it along with solid pieces. Let it be m_3 g.

Calculations

When the hot solid pieces are added to water it loses heat which is absorbed by water and calorimeter till they are in thermal equilibrium.

Fall in temperature of solid = $\theta_2 - \theta$

Rise in temperature of water = $\theta - \theta_1$

Rise in temperature of calorimeter = $\theta - \theta_1$

Mass of solid = $(m_3 - m_2)$g

Mass of water = $(m_2 - m_1)$g

Heat lost by solid = $(m_3 - m_2)\, s\, (\theta_2 - \theta)$

Where s = specific heat capacity of the solid which is to be determined.

Heat gained by water = $(m_2 - m_1)s_w\, (\theta - \theta_1)$

Heat gained by calorimeter = $m_1 s_C (\theta - \theta_1)$

Heat lost by the solid = heat gained by the water + heat gained by the calorimeter.

$$(m_3 - m_2)s\, (\theta_2 - \theta) = (m_2 - m_1)s_w\, (\theta - \theta_1) + m_1 s_C\, (\theta - \theta_1)$$

$$\therefore s = \frac{[m_1 s_C + (m_2 - m_1)s_w\,](\theta - \theta_1)}{(m_3 - m_2)\ (\theta_2 - \ \theta)}$$

While determining specific heat capacity by the above method, certain precautions need to be taken to ensure the accuracy of measurements. They are as follows.

1. The calorimeter should be clean, dry and highly polished.
2. Good heat insulators should be used in the wooden box of the calorimeter.
3. The thermometer should be highly sensitive.
4. The final temperature of the mixture should be noted one minute after the temperature has stabilized.
5. Calorimeter should not be kept near a source of heat.

☛ *Example*

A solid of mass 70 g at 80°C is dropped into 100 g of water at 20°C contained in a copper calorimeter of mass 60 g and specific heat capacity 0.40 J g^{-1} $°C^{-1}$. The final steady state temperature of the mixture is 30°C. Find the specific heat of solid. [Specific heat of water = 4.2 J g^{-1} $°C^{-1}$].

Solution

Compile the given information in tabular form as shown below:

Substance	Mass (gram)	Specific heat capacity J g^{-1} $°C^{-1}$	Initial temperature	Final temperature = 30°C	
				rise in temperature (°C)	Fall in temperature
Solid	70	X	80°C	-	80°C - 30°C $\Delta\theta_F$ = 50°C
Water	100	4.2	20°C	$\Delta\theta_R$ = 30°C - 20°C = 10°C	-
Calorimeter	60	0.40	20°C	$\Delta\theta_R$ = 30°C - 20°C = 10°C	-

Heat lost by solid= m s $\Delta\theta F$ = 70 g × x × 50 °C

Heat gained by water = ms $\Delta\theta_R$ = 100 g × 4.2 J g^{-1} $°C^{-1}$ × 10 °C = 4200 J

Heat gained by calorimeter = ms $\Delta\theta_R$ = 60 g × 0.4 J g^{-1} $°C^{-1}$ × 10 °C = 240 J

Applying principle of calorimetry we get,

3500x = 4200 J + 240 J

$$\therefore x = \frac{4440}{3500} = 1.268 \text{ J } g^{-1} °C^{-1}$$

Measurement of specific heat capacity of a liquid by method of mixtures

The procedure for measurement of specific heat capacity of a liquid by method of mixtures is same as that for a solid. Liquid whose specific heat capacity is to be determined is taken in calorimeter instead of water. A solid whose specific heat capacity is known is heated and introduced into it.

According to the principle of calorimetry, heat lost by the solid = heat gained by the liquid + heat gained by the calorimeter.

$$(m_3 - m_2)\, s\, (\theta_2 - \theta) = (m_2 - m_1)\, s_\ell\, (\theta - \theta_1) + m_1 s_C\, (\theta - \theta_1)$$

$$\therefore (m_2 - m_1)\, s_\ell\, (\theta - \theta_1) = (m_3 - m_2)\, s\, (\theta_2 - \theta) - m_1 s_C\, (\theta - \theta_1)$$

$$s_\ell = \frac{(m_3 - m_2)s(\theta_2 - \theta) - m_1 s_C(\theta - \theta_1)}{(m_2 - m_1)\,(\theta - \theta_1)}$$

where m_1, m_2 and m_3 are masses of empty calorimeter, calorimeter with liquid, calorimeter with liquid and solid respectively; θ_1, θ_2 and θ are initial temperature of calorimeter and liquid, initial temperature of the hot solid and final steady temperature of the mixture respectively and s_ℓ, s_C and s are the specific heats of liquid, calorimeter and solid respectively. In the above procedure, the solid is selected such that it does not dissolve or react with the given liquid.

Change of state

Matter can exist in three states or phases namely solids, liquids and gases. The state of matter can be changed by changing the temperature or pressure. For example, when ice is heated, it converts to water which converts to vapour or steam on further heating. When ice and water are heated their temperatures rise. But, when ice changes into water or water changes into steam, there is no rise in temperature although heat is supplied continuously. Reverse of this is also true, i.e., when steam on cooling converts to ice, there is no fall in temperature during change of state. Thus change of state always takes place at constant temperature. When a solid on heating changes to liquid at a constant temperature, the change of state is called melting or fusion. The constant temperature at which the solid melts is called its melting point. Melting point of ice is 0°C. The conversion of liquid into solid on cooling at constant temperature is called freezing. The constant temperature at which liquid freezes to solid is called freezing point. Freezing point of water is 0°C. The melting point of a substance when it is converted from solid to liquid state has the same value as its freezing point when it is converted from liquid to solid state. Change of a liquid to gas on heating at a constant temperature is called its boiling or vaporization. The constant temperature at which a liquid changes to gas is called boiling point or ebullition point. The change from gaseous to liquid state on cooling at a constant temperature is called condensation. The constant temperature at which condensation occurs is called liquefaction point.

Let us perform a simple experiment to determine the melting point of a solid. This method is suitable for solids which have their melting point between 30°C and 100°C. Take some powdered naphthalene in a thin glass test tube fitted with a thermometer. Arrange the apparatus as shown in the following figure.

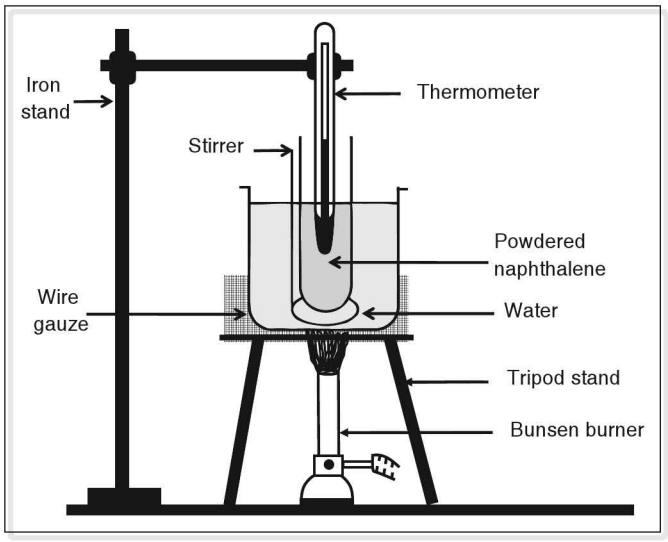


Figure 3.3

Turn on the bunsen burner and take thermometer reading after every 1 minute. Plot a graph of temperature versus time, taking temperature on Y-axis and time on X-axis.

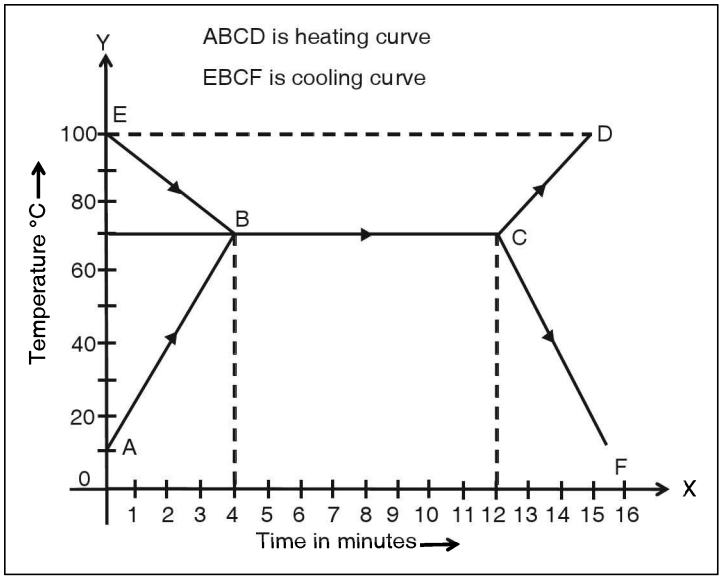


Figure 3.4

Initially we get a straight line inclined to X-axis (AB of graph) indicating that as naphthalene is heated its temperature rises proportionately. After some time we get a line parallel to X-axis (BC of graph) indicating that there is no rise in temperature though naphthalene is being heated continuously. By definition, we know that a change of phase from solid to liquid is taking place during BC. This represents melting of naphthalene. After some time, we again get a straight line graph (CD) inclined to X-axis. This indicates that the temperature of the liquid naphthalene rises on heating. ABCD represents the heating curve of naphthalene.

AB → rise in temperature of solid naphthalene on heating.
BC → melting of naphthalene on heating at constant temperature.
CD → rise in temperature of liquid naphthalene on heating.

The heating is stopped and naphthalene is allowed to cool and again temperature versus time graph is plotted giving us cooling curve EBCF.

EB → fall in temperature of liquid naphthalene on cooling.
BC → freezing of naphthalene at constant temperature on cooling.
CF → fall in temperature of solid naphthalene on cooling.

The temperature on Y-axis corresponding to BC is the melting point of the solid naphthalene or freezing point of the liquid naphthalene.

Latent heat of fusion

We have seen that during melting or vaporization the temperature of a substance remains constant although heat is continuously being supplied to it. Why doesn't its temperature rise? What happens to heat supplied during change of state?

The heat supplied to a body during its change of state does not result in rise of temperature as it is used for changing its state. The intermolecular space increases when a solid melts overcoming the intermolecular force of attraction.

Thus the heat supplied to a solid during its change of state is stored as potential energy of liquid molecules. The heat energy that is used to change its state without any rise in temperature is called latent heat or hidden heat. The heat required to change a solid into liquid at its melting point without any rise in temperature is called latent heat of fusion. Thus heat needed to melt ice into water at 0°C is called latent heat of fusion of ice. The latent heat of fusion depends on the mass of substance. The higher the mass of a solid, the larger is its latent heat of fusion. The latent heat required to convert a unit mass of a solid to liquid at constant temperature is called its specific latent heat of fusion.

Units of specific latent heat of fusion

$Q = mL$ where Q, m and L are heat supplied to a solid, its mass and its specific latent heat of fusion respectively.

$$\text{Thus unit of } L = \frac{\text{unit of } Q}{\text{unit of } m}$$

$\Rightarrow$ S.I unit of 'L' is J kg^{-1}. Other common units are cal g^{-1}, kcal kg^{-1} and J g^{-1}.

The specific latent heat of fusion of ice is amount of heat needed to convert unit mass of ice (1 kg or 1g) at 0°C to water at 0°C. The specific latent heat of fusion of ice is highest among all substances. Its value is 336×10^3 J kg^{-1} or 336 J g^{-1}. In common units it is 80 cal g^{-1}.

Specific latent heat of fusion of some substances

J g^{-1}	336	180	146	88	37.7	27	12.5
J kg^{-1}	336000	180000	146000	88000	37700	27000	12500
cal g^{-1}	80	43	35	21	9	6.42	3

Application and consequences of high specific latent heat of fusion of ice

The specific latent heat of fusion of ice is 336000 J kg^{-1}. Thus 1kg of ice absorbs 336×10^3 J of heat when it melts without any rise in its temperature. Conversely 1kg of water releases 336×10^3 J of heat when it freezes.

Few natural consequences of this are discussed below:

1. Water bodies such as lakes, ponds and rivers do not freeze suddenly in cold countries with the onset of winter. Also when they do freeze, they give 336 J g^{-1} of heat to the surroundings making the weather moderate.
2. During snow storm, snow is formed due to freezing of water vapour in the atmosphere. They give out large amount of heat to the surroundings which makes the surrounding weather moderate.
3. During summer, snow melts slowly on mountains as it needs large amount of heat. Hence there are no flash floods in rivers and they contain water for the whole year.

General applications

1. Soft drinks and fruit juices are cooled with ice rather than with cold water as ice can absorb 336 J g^{-1} of heat without increasing the temperature.
2. Ice cream at 0°C is felt cooler than water at 0°C as ice can absorb 336 J g^{-1}, while water absorbs 4.2 J g^{-1}.

Determination of latent heat of fusion of ice

To determine latent heat of fusion of ice by the method of mixtures, the following procedure is followed.

1. Take a clean, dry and highly polished calorimeter with stirrer and measure its mass using an accurate physical balance. Let it be m_1 g.
2. Fill in the calorimeter with water at room temperature till it is half filled. Measure the mass of the calorimeter with water. Let it be m_2 g.
3. Place the calorimeter in the wooden box having good heat insulating material like cotton wool or asbestos. With the help of a highly sensitive thermometer, note the temperature of water. Let it be θ_1°C.
4. Now take ice pieces made of pure distilled water and crush them into small pieces. Wipe them dry with the help of filter paper and drop them in calorimeter. Allow the ice pieces to melt by gently stirring the water. Repeat the process till the temperature of mixture falls by about 15°C. Note the stabilized steady temperature of the mixture. Let it be θ °C. The temperature of ice before dropping them in water is 0°C.
5. Remove the calorimeter from the wooden box and record its mass with the mixture. Let it be m_3 g.

Calculations

1. Mass of water at room temperature = $(m_2 - m_1)$ g.
2. Mass of ice = $(m_3 - m_2)$ g.
3. Rise in temperature of water formed from ice = $\theta - 0° = \theta$°C
4. Fall in temperature of water and calorimeter at room temperature = $(\theta_1 - \theta)$°C

   Heat gained by ice to change its state to water at 0°C

   = mass of ice × specific latent heat of fusion of ice

   = $(m_3 - m_2) L_{ice}$

   Heat gained by water formed from ice = (mass of ice) (specific heat capacity of water) (rise in temperature of water formed from ice)

   = $(m_3 - m_2)$ g × 4.2 J g^{-1} $°C^{-1}$ × θ°C

   = $4.2(m_3 - m_2) \times \theta$.

   Heat lost by calorimeter = $m_1 \times 0.40$ J g^{-1} $°C^{-1}$ × $(\theta_1 - \theta)$°C = $0.4m_1(\theta_1 - \theta)$ J.

   Heat lost by water = $(m_2 - m_1) \times 4.2$ J g^{-1} $°C^{-1}$ × $(\theta_1 - \theta)$°C = $4.2 (m_2 - m_1) (\theta_1 - \theta)$

According to principle of mixture,
Heat gained by ice = Heat lost by calorimeter + Heat lost by water at room temperature.

$$(m_3 - m_2)\, L_{ice} + 4.2(m_3 - m_2)\theta = 0.4m_1(\theta_1 - \theta) + 4.2(m_2 - m_1)\,(\theta_1 - \theta)$$

$$\therefore (m_3 - m_2)L_{ice} = [\{0.4m_1 + 4.2(m_2 - m_1)\}\,(\theta_1 - \theta) - 4.2(m_3 - m_2)\theta]$$

$$\therefore L_{ice} = \left[\frac{\{0.4m_1 + 4.2(m_2 - m_1)\}\,(\theta_1 - \theta) - 4.2(m_3 - m_2)\theta}{(m_3 - m_2)}\right]$$

Precautions

In addition to precautions already discussed, following precautions need to be taken while finding specific latent heat of fusion of ice.

1. Ice should be prepared using distilled water and should be dried thoroughly before adding to water in the calorimeter.
2. Water in the calorimeter should be stirred gently.
3. Final temperature should be recorded only when all the ice has melted and one minute after the temperature has stabilized.

☛ *Example*

A calorimeter of mass 50 g and specific heat capacity 0.40 J g^{-1} $°C^{-1}$ contains 200 g of water at 40°C. What amount of ice at 0°C should be added to it so that the final temperature is 10°C [specific heat capacity of water = 4.2 J g^{-1} $°C^{-1}$ and specific latent heat of fusion of ice = 336 J g^{-1} $°C^{-1}$]?

Solution

Ice	m	336 J g^{-1} $°C^{-1}$	0	$\Delta\theta = 10 - 0 = 10$	-
Water	200	4.2 J g^{-1} $°C^{-1}$	40		$\Delta\theta = 40 - 10 = 30$
Calorimeter	50	0.40 J g^{-1} $°C^{-1}$	40		$\Delta\theta = 40 - 10 = 30$

Heat gained by ice to form water at 0°C = mL_{ice} = m × 336 J g^{-1} $°C^{-1}$ = 336m J

Heat gained by water formed from ice = ms $\Delta\theta$ = m × 4.2 J g^{-1} $°C^{-1}$ × 10°C = 42m J

Heat lost by water at 40°C = ms $\Delta\theta$ = 200 g × 4.2 J g^{-1} $°C^{-1}$ × 30°C = 25200 J

Heat lost by calorimeter at 40°C = ms $\Delta\theta$ = 50 g × 0.40 J g^{-1} $°C^{-1}$ × 30°C = 600 J

Heat gained by ice to form water + Heat gained by water formed from ice = Heat lost by water + Heat lost by calorimeter

$\Rightarrow 336\ m + 42m = 25200 + 600$

$\Rightarrow 378\ m = 25800 \Rightarrow m = \frac{25800}{378} = 68.25\ g.$

Factors affecting melting point

Melting point of a solid is affected by pressure and impurities present in it.

Effect of pressure

In general, melting point increases with increase in pressure. But for substances which expand on freezing like water, iron, etc., the melting point decreases with increase in pressure. Thus pure ice which melts at 0°C under normal pressure will melt at temperature less than 0°C if pressure is increased.

Regelation

It is a phenomenon in which melting point of ice is lowered with increasing pressure and it is restored to normal value when entire pressure is removed. This can be illustrated as under. Take a string and attach weight, to its two ends and place the string over an ice block. You will find that string passes through the ice block without breaking it and falls down.

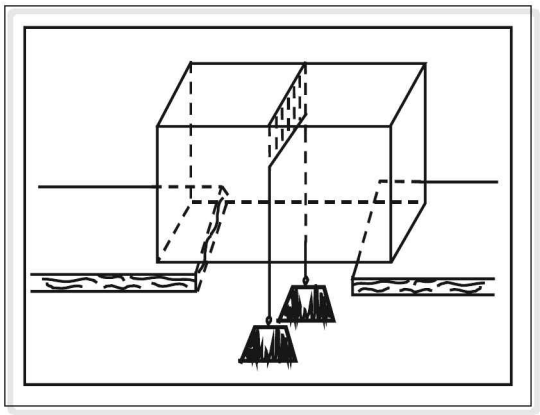


Figure 3.5

This is due to regelation. When the string is kept on the top surface of the ice block it exerts pressure on the block due to which it melts and the string goes down. The water above string again freezes as the external pressure is removed. In this way the string slips down the whole of the ice block. It is also due to regelation that an ice skater is able to slide over ice. The ice below the skates melts due to increase in pressure forming a thin layer of water which allows the skater to move ahead smoothly. As soon as skater goes ahead, the water freezes again.

Effect of impurities on melting point

Addition of impurities decreases the melting point of ice.

Freezing mixture

A mixture of ice and an impurity like common salt which is soluble in water in proper ratio is called freezing mixture. The melting point of freezing mixture is lower than 0°C. For example, the melting point of the mixture of ice and common salt mixed in the ratio 3 : 1 is −22°C. The melting point of mixture of ice and calcium chloride mixed in the ratio 3 : 4 respectively is −55°C. Freezing mixtures are used in preparing ice creams and in laboratories for preserving biological specimens.

Latent heat of vaporization

Temperature of liquids increases when heated till it reaches the boiling point. At boiling point, while liquid vaporizes, its temperature remains constant. During vaporization, the heat supplied is used to overcome intermolecular forces of attraction to increase intermolecular space and it is stored in vapour molecules in the form of potential energy. This heat is called latent heat of vaporization. The latent heat of vaporization of a given liquid increases with mass. In order to have a unique or specific value, we define latent heat of vaporization for unit mass of a given liquid. The latent heat of vaporization of a unit mass of a liquid is called its specific latent heat of vaporization.

Units of specific latent heat of vaporization

$Q = m L$; where Q, m and L are heat supplied to a liquid, mass of a liquid and its specific latent heat of vaporization respectively.

$$\text{Thus, unit of } L = \frac{\text{unit of } Q}{\text{unit of } m}$$

Hence, S.I unit of 'L' is joule per kilogram (J kg^{-1}) other common units are cal g^{-1}, kcal g^{-1} and J g^{-1}

Specific latent heat of vaporization of steam

Amount of heat required to convert unit mass of water (1 kg or 1 g) at 100 °C to steam at 100 °C is called specific latent heat of vapourization of steam. Steam has highest specific latent heat of vapourization. This value is 2268 J g^{-1} or 226.8×10^4 J kg^{-1} or 540 cal g^{-1}

Methyl alcohol	64 °C	1120	112×10^4
Acetone	57 °C	520	52×10^4
Benzene	80 °C	400	40×10^4
Carbon disulphide	46 °C	360	36×10^4
Carbon tetrachloride	77 °C	190	19×10^4
Ethyl alcohol	79 °C	850	85×10^4
Mercury	357 °C	290	29×10^4
Turpentine oil	156 °C	290	29×10^4
Water	100 °C	2268	226.8×10^4

Advantages and natural consequences of high specific latent heat of vaporization of steam

1. For water in soil to evaporate, 226.8×10^4 J kg^{-1} heat is needed which slows down vaporization. Thus plants do not wilt easily even in summer.
2. One gram of water at 100°C contains 100 cal of heat. If the same water is to be converted into steam at the same temperature it needs 540 cal of additional heat. Thus 1 g of steam at 100°C contains 640 cal of heat. Therefore burns caused by steam at 100°C are more severe than that caused by water at 100°C.

3. Steam can carry large amount of heat. Therefore it is used in thermal power stations where heat energy from coals is carried by steam and converted to electric energy.

Determination of specific latent heat of vaporization by method of mixtures

Follow 1 to 3 steps mentioned in determination of latent heat of fusion of ice by calorimeter.

4. In this case the wooden lid of the calorimeter box has a third hole for introducing steam into the calorimeter. Steam is generated in a boiling flask and passed through steam trap which removes condensed water vapours. The steam coming out of steam trap is introduced in the calorimeter till the temperature of the mixture rises by about 20°C. The mixture is stirred gently for uniform distribution of temperature. The final steady temperature of mixture is recorded. Let it be θ°C. The temperature of steam is 100°C.
5. Remove the calorimeter and record its mass. Let it be m_3.

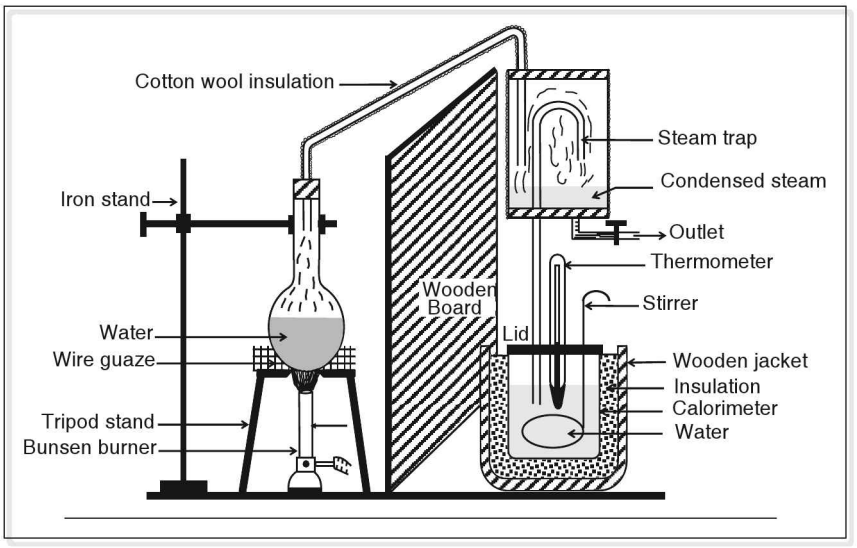


Figure 3.6

Calculations

Mass of water at room temperature = $(m_2 - m_1)$g, Mass of steam condensed and mixed with water = $(m_3 - m_2)$ g

Fall in temperature of water formed from steam $\Delta\theta = (100 - \theta)°C$

Rise in temperature of calorimeter and water at room temperature $\Delta\theta = (\theta - \theta_1)°C$

Heat lost by steam to form water at 100°C

$= (m_3 - m_2) \times L_{steam}$

$= (m_3 - m_2)x$, where $L_{steam} = x$, the latent heat of vaporisation of water.

Heat lost by water at 100°C formed from steam

$= (m_3 - m_2)\ g \times 4.2\ J\ g^{-1}\ °C^{-1} \times (100 - \theta)°C$

$= 4.2(m_3 - m_2)\ (100 - \theta)J$

Heat gained by calorimeter = $m_1 \times 0.40\ J\ g^{-1}\ °C^{-1} \times (\theta - \theta_1)°C$

Heat gained by water at room temperature = $(m_2 - m_1) \times 4.2\ J\ g^{-1}\ °C^{-1} \times (\theta - \theta_1)°C$

$= 4.2(m_2 - m_1)\ (\theta - \theta_1)\ J$

By principle of mixtures

Heat lost by steam = heat gained by calorimeter + heat gained by water at room temperature.

$= (m_3 - m_2)x + 4.2(m_3 - m_2)\ (100 - \theta) = 0.40m_1\ (\theta - \theta_1) + 4.2(m_2 - m_1)\ (\theta - \theta_1)$

$$\therefore x = \left[\frac{\{0.4m_1 + 4.2(m_2 - m_1)\}\ (\theta - \theta_1) - 4.2(m_3 - m_2)(100 - \theta)}{m_3 - m_2}\right]$$

Precautions

Other than customary precautions with calorimeter, observe the following precautions for steam.

1. Steam should be perfectly dry.
2. Steam should be passed only till temperature of mixtures rises by about 20°C.

☛ *Example*

100 g of water at 10°C is taken in a calorimeter of mass 50 g and specific heat capacity 0.40 $Jg^{-1}\ °C^{-1}$. Calculate the mass of steam which should be passed into the water in the calorimeter so that the temperature of mixture becomes 40°C. [Specific heat capacity of water = 4.2 J $g^{-1}\ °C^{-1}$ and specific latent heat of vaporization of steam = 2260 J g^{-1}.]

Solution

Steam	x	2260 J $g^{-1}°C^{-1}$	100 °C	-	$\Delta\theta$ = 100 °C – 40 °C = 60°C
Water	100	4.2 J $g^{-1}\ °C^{-1}$	10 °C	$\Delta\theta$ = 40°C – 10 °C = 30°C	-
Calorimeter	50	0.4 J $g^{-1}\ °C^{-1}$	10 °C	30 °C	-

Heat given by steam to form water at 100°C = m L = 2260x J

Heat lost by water formed from steam at 100°C = m s $\Delta\theta$

$= m \times 4.2\ J\ g^{-1}\ °C^{-1} \times 30°C = 126x\ J$

Total heat lost by steam and the water formed from steam = 2260x + 126x = 2386x J.

Heat gained by water at 10°C = $100\ g \times 4.2\ J\ g^{-1}\ °C^{-1} \times 30°C = 12600\ J$

Heat gained by calorimeter at 10°C = $50\ g \times 0.4\ J\ g^{-1}\ °C^{-1} \times 30°C = 600\ J$

By principle of calorimetry

2386x J = 12600 J + 600 J

$\therefore x = 5.53$ g (approx).

To determine boiling point of a liquid

Take water in a closed vessel fitted with a thermometer. Heat it with bunsen burner and note the rise in temperature for every 1 minute. Plot temperature versus time graph, taking time on X-axis and temperature on Y-axis.

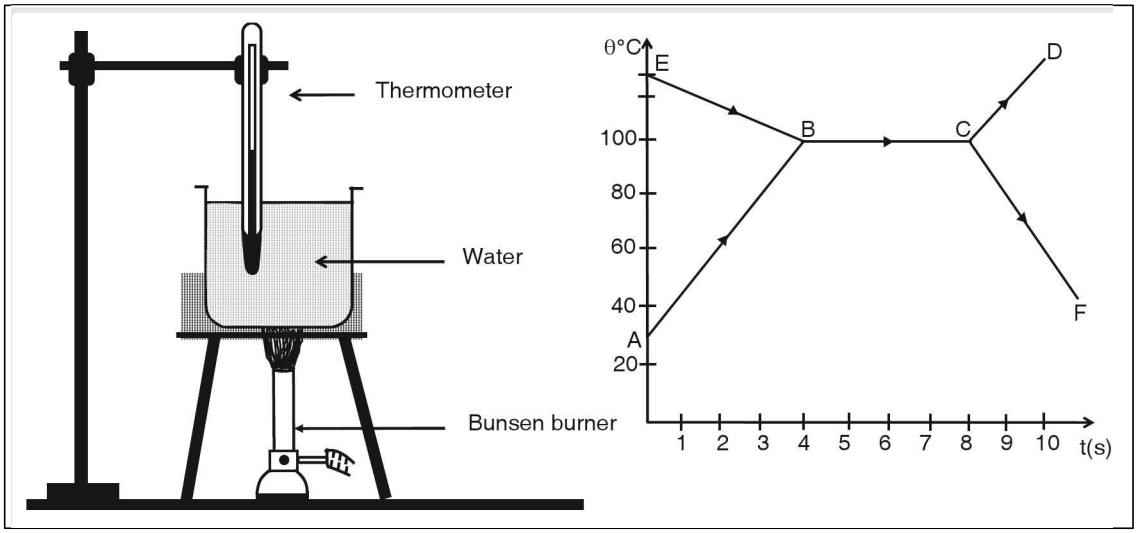


Figure 3.7

ABCD represents heating curve of water.

AB represents rise in temperature of water on heating.

BC represents change of state of water to steam at constant temperature.

CD represents rise in temperature of steam on heating.

Heating is stopped and vessel is allowed to cool and cooling curve EBCF is plotted.

EB represents fall in temperature of steam on cooling.

BC represents change in state of steam to water at a constant temperature.

CF represents fall in temperature of water on cooling.

Effects of pressure and impurities on boiling point

Pressure

Increase in pressure increases the boiling point of water and decrease in pressure decreases its boiling point.

Let us discuss an activity to understand the above fact.

Take some water in a flask. Close the mouth of flask with two holed rubber cork fitted with a delivery tube and a thermometer. Heat the flask and note the temperature of water. It begins to boil at 100°C and steam begins to form. Now partly close the delivery tube so that pressure due to steam increases inside. Check the flask for the temperature at which water boils. You will find that it has increased. Stop heating

and allow the water to cool. Close the flask completely. On cooling the steam inside the flask condenses thereby decreasing the pressure. You will find water boils at lower temperature. Some examples of effect of pressure on boiling point are

1. Pressure cooker is used to cook food quickly.
2. Water boils at a temperature much lower than 100°C at higher altitudes.

Impurities

Addition of impurities increases the boiling point. When 100 g of common salt is added to 100 g of water, the boiling point increases by 1°C.

Evaporation

Wet clothes when hung on rope dry. The water in the wet clothes vapourizes without boiling. If you drop some spirit on your hand, it vapourizes leaving cooling sensation on your hand. The process by which water vapourizes without boiling is called evaporation. The molecules on the surface of water are attracted to water molecules below the surface by cohesive forces and by the surrounding air molecules by adhesive force. If the energy of few molecules on the surface is high, they escape to the surrounding air. Since this process takes continuously at all temperatures we find that wet clothes dry after some time. This is due to continuous evaporation. Since the energetic molecules leave water, the average kinetic energy of molecules decreases causing fall in temperature and gives cooling effect. Evaporation is purely a surface phenomenon. Therefore water stored in earthern porous container cools faster than that in a metal container. Since evaporation involves escaping of surface water molecules to surroundings, the rate of evaporation depends on the following factors

1. surrounding temperature
2. wind velocity
3. surface area of water

Boiling is different from evaporation, as boiling takes place rapidly at a fixed temperature only and throughout the liquid.

Refrigeration

When a liquid evaporates, it extracts heat energy required for phase change from surrounding. Therefore evaporation has cooling effect. Refrigerator is a device which is used to store food, vegetables and other perishable items at low temperature. It works on the principle that evaporation of a liquid gives cooling effect. Basically it consists of the following main parts connected to the main frame.

1. pump
2. cooling chamber
3. evaporation tube
4. condensation tube
5. copper fins

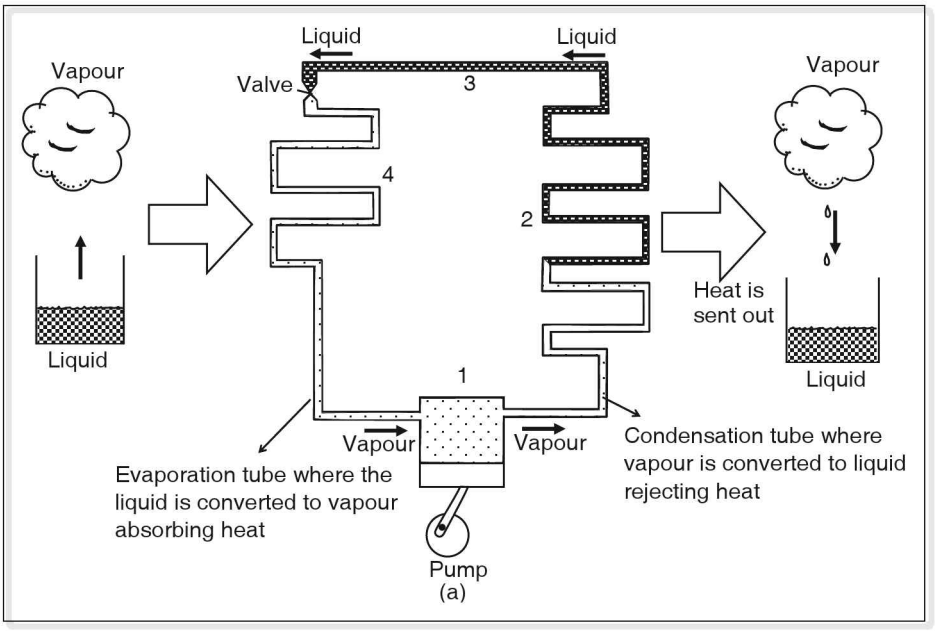


Figure 3.8

Pump situated behind the main structure of refrigerator at the bottom, circulates volatile liquid like liquid freon through the evaporation tube. Evaporation tube passes through cooling chamber. When freon reaches cooling chamber it evaporates extracting the required heat from cooling chamber. The temperature of cooling chamber decreases considerably. The vapors then reach condensation tube and pump exerts pressure which condenses them to liquid which is again circulated through evaporation tube. The condensation tube is connected to copper fins which radiate the heat released during condensation. This process is repeated continuously and temperature in cooling chamber is brought down to desired level.

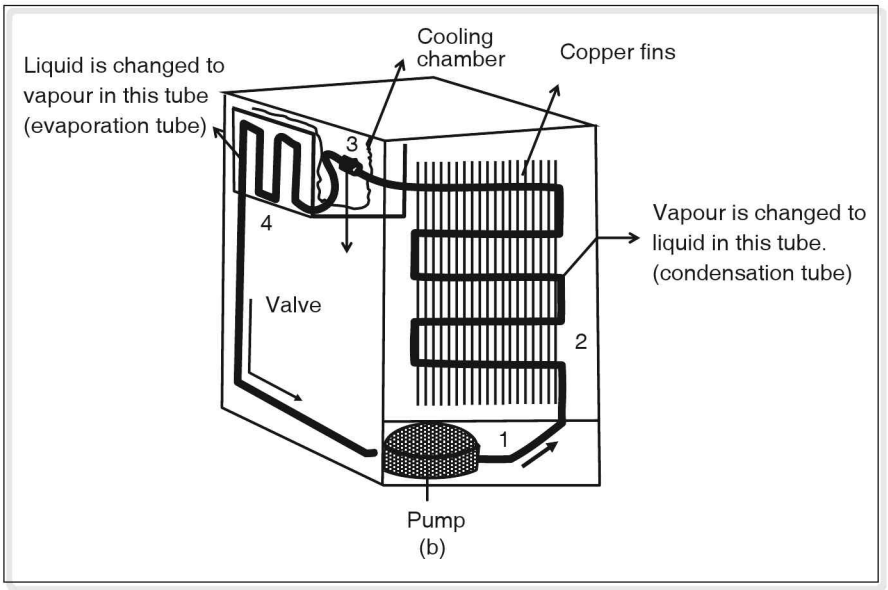


Figure 3.9

Sublimation

A solid when heated converts to liquid at melting point and liquid on further heating changes to gas at boiling point. There are few substances which convert to gas from solid state without changing to liquid state. This process is called sublimation.

☛ *Example* naphthalene balls, camphor, etc., are some solids that undergo sublimation.

Humidity and dew point

Almost 70% of earth's surface is covered by water in the form of lakes, rivers, oceans, etc. Thus water evaporates continuously and forms water vapour in the atmosphere. To prove that the atmosphere always contains certain amount of water vapuor, let us discuss a simple experiment. Take some water in a beaker and add a few pieces of ice to it. Melting of ice causes temperature of the mixture to fall. After some time as the temperature has fallen to certain value you will find that the outer surface of the beaker has become blurred due to the formation of condensed water film. This water film develops due to condensation of water vapour present in air due to the fall in temperature.

The presence of vapour makes the air moist or damp. This moisture or dampness of air due to water vapour is called humidity. The higher the vapour content present in air, the higher is the humidity of air. Humidity is high in coastal areas. The amount of water vapour present in one cubic metre of air is called its absolute humidity. There is a limit to the mass of water vapour which can be present in a given volume of air. When the mass of water vapour present in air is equal to this limit, i.e. maximum possible value, the air is said to be saturated with water vapour and if it is less than air, it is said to be unsaturated. Sometimes we feel air to be dry and at other times, damp. This feeling not only depends on the exact amount of vapour present in air but also on the difference between this amount and the quantity of water vapour required to saturate the air. If this difference is less, air is felt more damp and if this difference is more, air is felt dry. It is because of this factor that even though the actual amount of water vapour in air is more on a dry summer day than on a moist winter day, the air still is felt drier on summer day. This means that the quantity of water vapour required to saturate given volume of air increases with an increase in temperature. The extent of moistness of air felt by us is determined by a quantity called relative humidity.

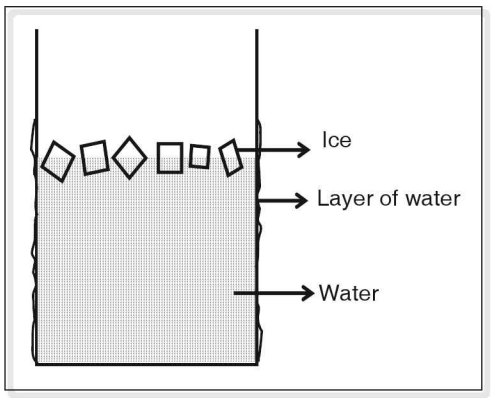


Figure 3.10

Relative humidity

It is defined as ratio of amount of water vapour actually present in a given volume of air to the amount of water vapour required to saturate the same volume of air at the same temperature.

$$\text{Relative humidity} = \frac{\text{amount of vapour in a given volume of air}}{\text{amount of vapour needed to saturate same volume of air at same temperature}} \times 100$$

Thus relative humidity of completely saturated air is 100%. The air is felt damp if relative humidity is greater than 60%, otherwise it is felt dry. As already discussed the amount of vapour required to saturate a given mass of air depends on temperature. Keeping the amount of unsaturated air and the vapour content in it constant, if its temperature is continuously decreased, the air will get saturated. Now it cannot hold any more amount of vapour for the given temperature. If the temperature is decreased slightly the vapour in the given volume of air condenses. This is how a thin film of water was formed on the outer surface of the beaker discussed earlier. The water droplets so formed are also called dew. The temperature at which the given volume of air becomes saturated with water vapour without actually increasing the vapour content in it is called dew point. At this temperature, the water vapour in air just begins to condense. During day time you rarely find dew, but during night when the temperature falls below dew point, dew is formed on cold bodies. Sometimes dew is formed on dust particles in the atmosphere forming fog or mist.

Anomalous behaviour of water

Liquids expand on heating and contract on cooling. Water shows exception to this rule below 4°C. If water at room temperature is cooled, it contracts till its temperature reaches 4°C. On further cooling, as its temperature falls below 4°C, it begins to expand instead of contracting. Conversely if water is heated from 0°C to 4°C, it contracts instead of expanding. This behaviour of water is known as anomalous behaviour of water and the expansion of water on cooling below 4°C is called anomalous expansion of water. The anomalous behaviour of water can be demonstrated with the help of Hope's apparatus.

Hope's apparatus

It consists of a cylindrical vessel surrounded by a cylindrical trough at its centre. Two thermometers can be inserted in the vessel at the top and bottom as shown in figure.

The vessel is filled with water at 10°C and the trough with a freezing mixture of ice and salt. The changes in temperature of water are recorded by the thermometers.

Following observations are made.

1. Initially, the temperature shown by both thermometers is same. As the time elapses the lower thermometer records lower temperatures while temperature shown by the upper thermometer is almost constant. The temperature shown by the lower thermometer decreases gradually till 4°C and then remains constant at 4°C.
2. The temperature shown by the upper thermometer begins to fall till it is 0°C. But the temperature recorded by the lower thermometer remains constant at 4°C.
3. After the upper thermometer shows 0°C, a thin layer of ice is formed at the top surface of water.

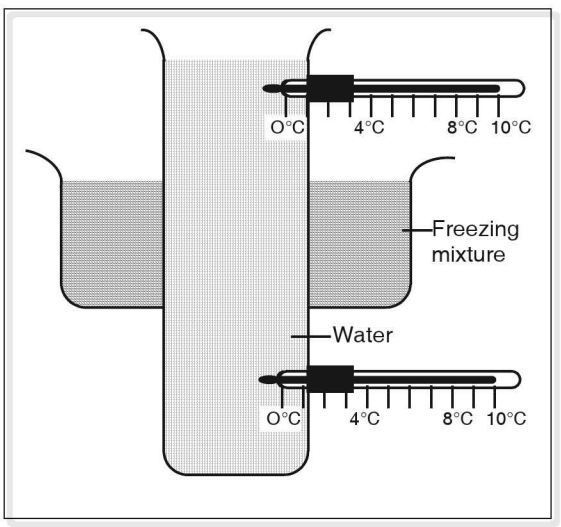


Figure 3.11 Hope's Apparatus

The above observation can be explained as follows:

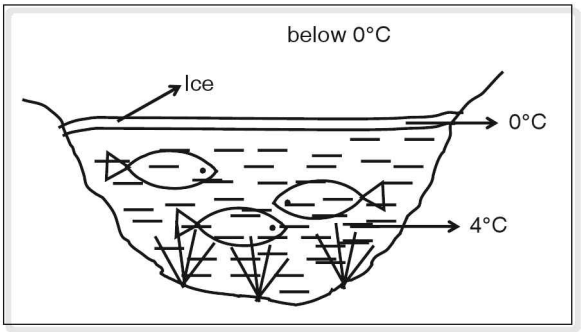


Figure 3.12

Initially, the temperature of water is 10°C. Due to the freezing mixture, the temperature of water at central portion begins to fall and it contracts. As its density increases, the cold water goes to the bottom. Therefore the lower thermometer records fall in temperatures. This continues till temperature of water at the central portion reaches 4°C. Now as temperature falls further, water expands instead of contracting and its density decreases. Therefore cold water below 4°C goes up instead of going down. Therefore temperature of water at the top of the vessel begins to fall till it reaches 0°C as shown by the upper thermometer. At 0°C water freezes into ice. As density of ice is less than water, layer of ice is formed on the top surface. The temperature of water at the bottom of the vessel remains constant at 4°C as ice being a bad conductor of heat, prevents loss of heat from water below it to the surrounding.

Consequences of anomalous behaviour of water.

1. The anomalous behaviour of water plays very significant role in the survival of aquatic flora and fauna in cold countries. As explained in Hope's apparatus, the water surface of lakes, ponds or rivers may get frozen in winter, but the temperature of water at bottom remains constant at 4°C.
2. Water pipes in cold countries burst when temperature falls below 4°C as water expands.
3. Fruits and vegetables also get damaged if temperature falls below 4°C as water contained in their capillaries expand.

Heat Engine

Heat is a form of energy. It can be converted to other forms of energy and vice versa. Energy is ability to do work. Thus heat can also be used to do useful work.

Some examples of work done by heat are:

1. The weight of a pressure cooker is raised when sufficient amount of steam is accumulated in it.
2. James Watt, a Scotish scientist, observed that while making tea, the lid of the tea kettle flew up as steam was generated due to heating of water. James Watt later developed modern steam engine which can transport marine loads over large distances rapidly. Heat energy cannot be used directly to do useful work. The device which converts heat into useful mechanical work is called heat engine. Heat engines are classified into two types namely,external combustion engine and internal combustion engine.

External combustion engine

It consists of a cylinder fitted with an air tight piston. The piston is connected to a crank shaft. The cylinder is connected to a boiler through an inlet valve and to a condensor through an outlet valve. The boiler and condensor are connected to each other through a pump.

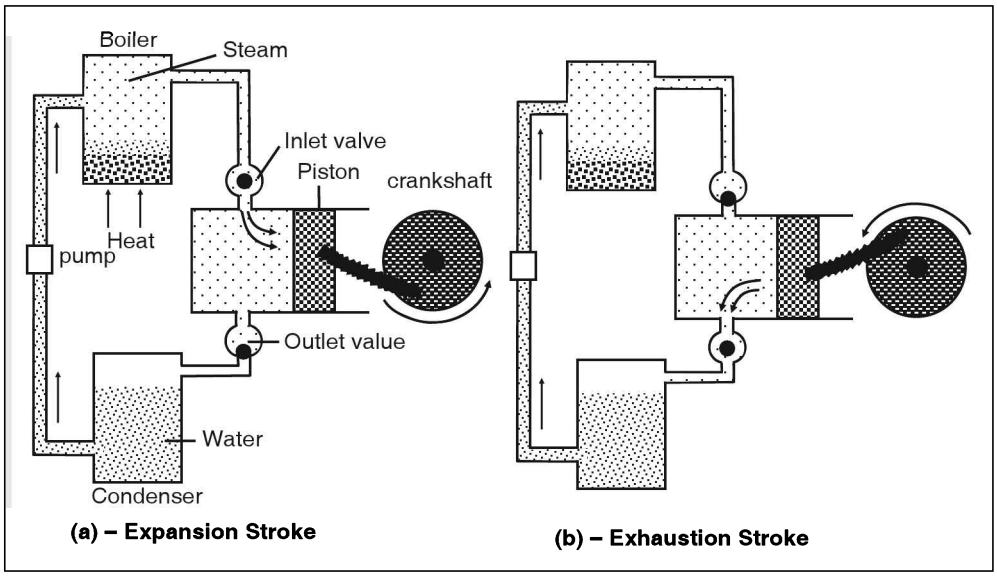


Figure 3.13

Water is heated in the boiler and steam is generated. The pressure of accumulated steam opens the inlet valve and forces the piston outwards. This is called an expansion stroke. The outward or inward movement of the piston is called a 'stroke'. At the end of expansion stroke, the inlet valve is closed and the piston begins to move inwards.

The outlet valve is opened and used steam is passed to condensor where it condenses. The water is then pumped to boiler.

This action is repeated. In external combustion engine, steam is used to do useful work therefore it is also called steam engine.

Steam engine is not used extensively as it has certain disadvantages.

1. Steam engines are bulky and cannot be used to run smaller machines.
2. Steam engines can be dangerous as due to high pressure of steam, the boiler may burst.
3. Steam engine cannot be started instantly. To obtain steam at high temperature some time is needed.

Internal combustion engine

Due to the drawbacks, steam engine cannot be used in many applications. To overcome the drawbacks of a steam engine, a new heat engine was devised. In this heat engine, combustion of fuel and conversion of heat energy to mechanical energy is done inside the cylinder itself. Therefore it is called internal combustion engine. Depending on the fuel used, internal combustion engines are classified into petrol engines and diesel engine.

Petrol engine

It consists of a cylinder and an airtight piston which is connected to a crankshaft. Inlet and outlet valves and a spark plug are provided at the head of the cylinder. Petrol engine is connected to carburettor. It mixes petrol and air in proper proportion and supplies it to the cylinder through the inlet valve.

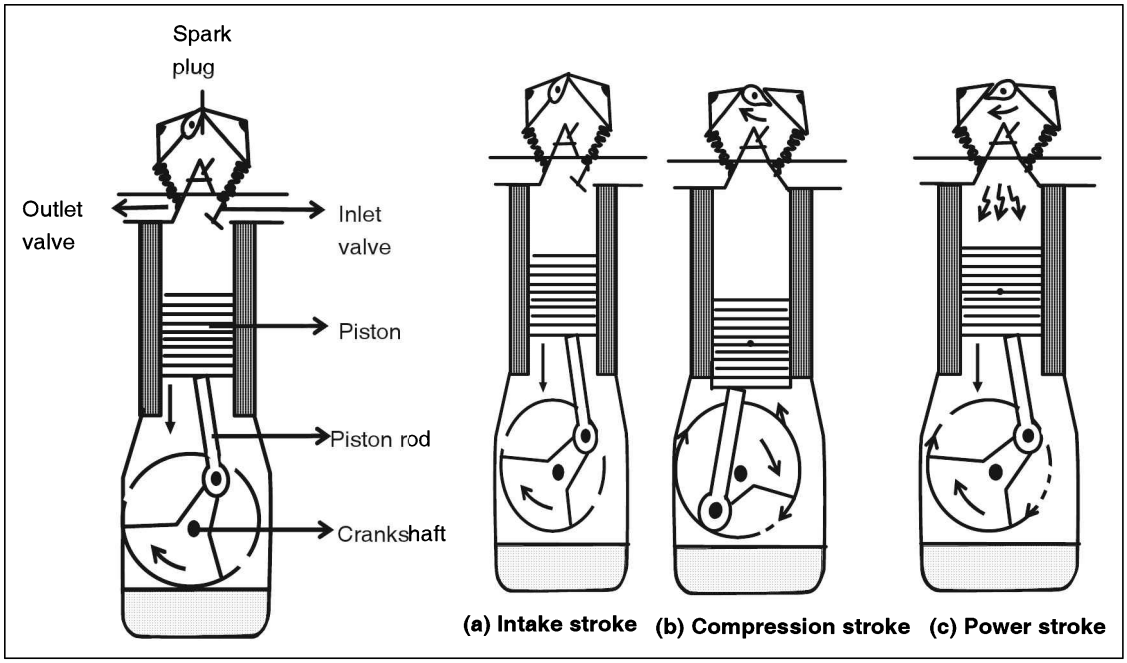


Figure 3.14

One complete operation or cycle of engine involves four stages or four strokes. Each stage is described below:

1. **Intake stroke:** The piston is moved outward. This decreases the pressure in cylinder which opens the inlet valve. The fuel mixture enters the cylinder from the carburettor.
2. **Compression stroke:** The piston is moved inwards towards the cylinder head which compresses the fuel mixture. Both the inlet and the outlet valves are closed.
3. **Power stroke:** The compressed fuel is burnt by the sparks from spark plug. The by-products of combustion like water vapour and carbon dioxide expand rapidly pushing the piston outwards with great force.
4. **Exhaust stroke:** The gases released during combustion are expelled through outlet valve as the piston is moved back and the outlet valve is opened. These four strokes constitute one work cycle. As the cycles are repeated piston moves to and fro rapidly which rotates the crankshaft. The internal combustion engine has to be started with the help of self starter, kick starter or push starter. The internal combustion engine is used in almost all vehicles like scooters, motorcycles, cars, etc.

Diesel engine

Diesel engine is similar to a petrol engine with few changes in construction. The spark plug is replaced by fuel injection pump. Carburettor is not used in a diesel engine. One work cycle of diesel engine consists of four strokes as in petrol engine.

Intake stroke: Pistons move out and filtered air is taken into the cylinder.

Compression stroke: The air is compressed to $\frac{1}{16}$th of its volume as piston moves in. The compressed air becomes hot enough to ignite the fuel. Hence there is no need of spark plug.

Power stroke: At the end of compression stroke, required amount of diesel is injected into the cylinder by fuel injection pump. Diesel bursts into flame instantly. Due to the high pressure generated by combustion and expansion of gases, the piston is pushed out with large force.

Exhaust stroke: The spent gases are expelled in the exhaust stroke. Although internal combustion engine has many advantages over steam engine, its biggest drawback is that it is a single largest source of air pollution. Some of the pollutants released like carbon monoxide, sulphur dioxide and lead compounds are highly toxic and harmful.

Efficiency of heat engine

It is defined as ratio of useful work done by heat engine to the total amount of heat supplied to it. It is generally expressed as a percentage.

$$\text{Efficiency} = \frac{\text{work done (useful)}}{\text{total heat supplied}} \times 100$$

No heat engine is 100% efficient. Efficiency of few types of heat engines are given below:

Table below lists efficiencies of some commonly used engines.

1	Steam engine	15
2	Jet engine	15
3	Petrol engine	30
4	Steam turbine	35
5	Diesel engine	40

Advantages of an internal combustion engine over a steam engine.

1. An internal combustion engine is more efficient and safer than a steam engine.
2. It can be started instantaneously.
3. Internal combustion engines are less bulky, and can be manufactured in a range of different sizes. They are suitable for small machines also.

test your concepts

Very short answer type questions

1. What is freezing mixture?
2. Define humidity, absolute humidity and relative humidity.
3. In diesel as well as in petrol engines, one complete cycle of operation involves _____, _____, _____ and _____ strokes.
4. State two factors which affect the boiling point.
5. Density of water is maximum at _____ °C.
6. In diesel engines, carburetor is replaced by _____.
7. The expansion of water on cooling below 4°C is called _____.
8. Calorimeter works on the principle of _____.
9. Define heat capacity and specific heat capacity.
10. What is fog or mist?
11. The specific latent heat of fusion of ice is _____ $J\ kg^{-1}\ K^{-1}$ or _____ $cal\ g^{-1}$.
12. State the principle of mixtures.
13. Define evaporation.
14. The latent heat of vaporisation of a unit mass of a liquid is called its _____.
15. Define the following:
    (i) melting or fusion (ii) freezing (iii) boiling (iv) condensation
16. What is an internal combustion engine?
17. How much heat is absorbed when 10 g of ice at 0°C is completely converted to water at 0°C?
18. Define latent heat.
19. State the principle on which a refrigerator works.
20. 1 g of steam at 100°C cools to form water at 80°C. The amount of heat liberated is _____. (Total heat released = Q + H)
21. Define latent heat of fusion and specific latent heat of fusion of ice.
22. The amount of heat required to raise the temperature of unit mass of a substance by one kelvin is called _____.
23. What is a heat engine and name two types of heat engines?
24. Define latent heat of vapourization and specific latent heat of vapourization of water.
25. What is sublimation?
26. The direction of heat flow between two bodies is determined by_____.
27. State two factors which affect the melting point.
28. Define the efficiency of heat engine.

29. The temperature at which the given volume of air becomes saturated with water vapour without actually increasing the vapour content in it is called _____.

30. Define regelation.

Short answer type questions

31. Give a few examples and some practical applications of high specific heat of vapourization of water.

32. A body of mass 1 kg absorbs 4 kJ of heat when its temperature rises from 30°C to 40°C. Calculate its specific heat capacity.

33. Explain sublimation.

34. State a few disadvantages of steam engine.

35. Give some natural consequences and a few practical applications of high specific heat capacity of water.

36. 1kg of ice at 0°C is converted into steam at 100°C. Calculate the heat required
Specific heat of fusion of ice = 336×10^3 J kg^{-1}
Specific heat of vaporization of water = 2260×10^3 J kg^{-1}
Specific heat capacity of water = 4200 J kg^{-1} $°C^{-1}$

37. Differentiate between evaporation and boiling.

38. Explain the working of a refrigerator.

39. Describe a calorimeter.

40. An electric heater of power 500 W is used to raise the temperature of 2 kg of a liquid from 30°C to 40°C in 2 minutes. Find the heat capacity if the specific heat capacity of the liquid is 3 kJ kg^{-1} $°C^{-1}$

41. Explain the formation of dew.

42. 10000 J heat is required to raise the temperature of a metal ball by 20 K. Calculate its heat capacity.

43. Mention few natural consequences, and some practical applications of high specific latent heat of fusion of ice.

44. State a few advantages of an internal combustion engine.

45. Mention a few consequences of anomalous behaviour of water.

Essay type questions

46. Describe an experiment to determine specific latent heat of fusion of ice by method of mixtures.

47. Explain the experiment to find the boiling point of water. Also explain its heating and cooling curves.

48. Describe an experiment to determine the specific heat capacity of a given solid by method of mixtures.

49. Describe the method to determine specific latent heat of vaporization of water using a calorimeter.

50. Describe an experiment to measure the melting point of naphthalene. Draw heating and cooling curves and explain them.

CONCEPT APPLICATION

Concept Application Level—1

Direction for question 1 to 7: State whether the following statements are true or false.

1. Due to anomalous behaviour of water, volume of ice is more than that of the same mass of water.
2. When common salt is added to water, the boiling point of water increases.
3. Heat required to change the state of substance at constant temperature is called latent heat.
4. Refrigerator is a device which works on the principle that evaporation of a liquid gives cooling effect.
5. The specific heat capacity of water is 4200 J $kg^{-1} K^{-1}$.
6. The rate of evaporation of a given liquid depends only on the surface area of the liquid exposed to air.
7. The SI units of heat and work are the same.

Direction for questions 8 to 14: Fill in the blanks.

8. _____ is used as a coolant in car radiators, nuclear plants and in industries.
9. The process that is involved when a substance changes to gas directly from solid state without changing to liquid state is called _____.
10. The dimensional formula of heat is _____.
11. Heat absorbed or lost by a substance is proportional to _____ and _____.
12. Boiling is a _____ process where as evaporation is a _____ process in respect with the speed of the process.
13. The device which is used to measure specific heat capacity of solids and liquids is _____.
14. Calorimeter is highly polished from inside and outside to prevent heat loss by _____.

Direction for question 15: Match the entries in column A with appropriate ones from column B.

15.

Column A			Column B
A. Water equivalent	( )	a.	Conversion from solid state to gaseous state directly
B. High specific heat capacity	( )	b.	Heat lost by hot body = heat gained by the cold body.
C. Bomb calorimeter	( )	c.	$\frac{\text{work output (or) heat utilized}}{\text{heat produced}}$
D. Principle of calorimeter	( )	d.	Liquid converts to gas
E. Evaporation	( )	e.	226×10^4 J kg^{-1}.

(Continued on following page)

F.	Latent heat of vaporization of pure water	( )	f.	Burns caused by steam.
G.	Sublimation	( )	g.	Petrol engine
H.	Thermal efficiency	( )	h.	Water
I.	Scalding	( )	i.	Calorific values
J.	Carburettor	( )	j.	gram (or) kilogram

Direction for questions 16 to 30: For each of the questions, four choices have been provided. Select the correct alternative.

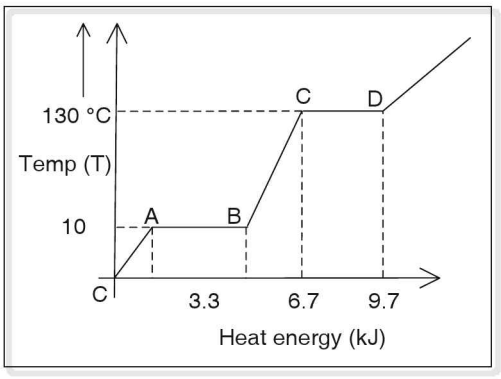
16. In a device, the complete conversion of heat energy into mechanical work _____.
(1) is not possible at any cost
(2) depends on type of device in use
(3) depends on the atmospheric temperature, pressure etc
(4) depends on the type of fuel used

17. If water is converted into ice and maintained at 0°C, which of the following statements is true for the molecules of the substance?
(1) Only kinetic energy decreases
(2) Only potential energy decreases
(3) Both the energies increase
(4) Potential energy increases and kinetic energy decreases

18. Steam engine is not extensively used because
(1) it is bulky
(2) it takes long time to generate the steam
(3) high pressure of steam may burst the boiler which is dangerous
(4) All the above

19. The specific heat capacity of a substance depends on its
(1) nature (2) mass (3) rise in temperature (4) Both (2) and (3)

20. Study the following graph drawn between the amount of heat energy supplied to 500 g of a substance and its rise in temperature. Identify the correct option.
(1) The melting point of the given substance is 10°C.
(2) The latent heat of fusion of the given substance is 2300 J.
(3) The boiling point of the substance 130°C.
(4) All the above.



21. Sea water and river water are heated to their boiling points. Then
(1) river water boils at lower temperature.
(2) sea water boils at lower temperature.
(3) both boil at the same temperature.
(4) Cannot be said

22. The food cooks faster in pressure cooker since
(1) the pressure exerted by water vapour enclosed in the cooker increases the boiling point of water.
(2) the pressure exerted by the water vapour enclosed in the cooker decreases the boiling point of water.
(3) the water vapour pressure does not influence the boiling point of the water.
(4) there is no loss of heat to the surrounding as the pressure cooker is airtight.

23. During change of state which of the following is true for molecules of a substance?
(1) Kinetic energy increases (2) Potential energy increases
(3) Both increase (4) Both remains constant

24. At 40°C, the actual amount of water vapour present in 1 m^3 of air is 45 g. If the temperature decreases to 30° C, then amount of water vpaour that condenses is _____ g. [Take the mass of water vapour present in 1 m^3 of saturated air at 30°C as 30 g].
(1) 75 (2) 10 (3) 25 (4) 15

25. In thermal power stations, steam is preferred to convert heat to electric energy because
(1) steam is good conductor of heat
(2) steam is good conductor of electricity
(3) steam can flow quickly
(4) every gram of steam can give 2260 J of heat.

26. A hot body at temperature θ_1 is mixed with a cold body at θ_2 both having the same heat capacity, such that the rise in temperature of the cold body is equal to the fall in temperature of the hot body. The resultant temperature θ is
(1) $\frac{\theta_1 + \theta_2}{2}$ (2) $\theta_1 + \theta_2$ (3) $\frac{\theta_1 - \theta_2}{2}$ (4) $\theta_1 - \theta_2$

27. Snow balls are formed due to
(1) melting of ice (2) freezing of atmospheric moisture
(3) regelation (4) sublimation

28. A substance of mass m_1 with specific heat capacity s_1 and initial temperature θ_1 is mixed with another substance of mass m_2, specific heat capacity s_2 and initial temperature θ_2 respectively. Then, their equilibrium temperature is _____ (take $\theta_1 > \theta_2$)
(1) $\frac{m_1s_1\theta_1 - m_2s_2\theta_2}{m_1s_1 - m_2s_2}$ (2) $\frac{m_1s_1\theta_1 - m_2s_2\theta_2}{m_1s_1 + m_2s_2}$
(3) $\frac{m_2s_2\theta_2 - m_1s_1\theta_1}{m_1 + m_2}$ (4) $\frac{m_1s_1\theta_1 + m_2s_2\theta_2}{m_1s_1 + m_2s_2}$

29. Heat gained by 1 g of water when heated from 0°C to 100°C is (Hint : used $H = m s\Delta\theta$)
(1) 4200 cal (2) 4.2 cal (3) 100 cal (4) 1 cal

30. Two bodies of masses 50 g and 100 g are taken if their water equivalents are 10 g and 20 g respectively. Then the values of their specific heat capacities are _____ cal g^{-1} °C^{-1} and _____cal g^{-1} °C^{-1}.
(1) $\frac{1}{5}$ and $\frac{1}{5}$ (2) $\frac{1}{3}$ and $\frac{1}{3}$ (3) $\frac{11}{5}$ and $\frac{11}{5}$ (4) $\frac{1}{2}$ and $\frac{1}{2}$

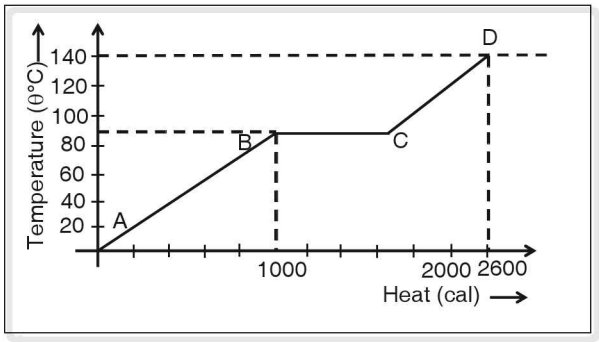
Concept Application Level—2

31. A metallic calorimeter of mass 100 g contains 200 g of ice and the initial temperature of the calorimeter and the ice is −10°C. Heat is supplied to the system containing calorimeter and ice at a constant rate of 50 °cal s^{-1}. Find the time required to raise the temperature of the system to 50°C. [Neglect the loss of heat to the surroundings. Take specific heat capacity of calorimeter and ice as 0.2 °cal g^{-1} °C^{-1} and $\frac{1}{2}$ cal g^{-1} °C^{-1}. Take latent heat of fusion of ice as 80 cal g^{-1} and specific heat capacity of water as 1 cal g^{-1} °C^{-1}.

32. The atmospheric temperature varies over a wide range but human body temperature is more or less constant at 37°C. Explain.

33. What happens when a thermos flask containing liquid is vigorously shaken? Will there be any change in the temperature of the liquid?

34. The graph below shows the rise in temperature of a solid of specific heat capacity 0.5 cal g^{-1} °C^{-1} and the heat supplied to it. Find its mass and specific latent heat of fusion.
Also find the specific heat capacity of the body in its liquid state.



35. A bullet of 50 g is fired from a revolver with a velocity of 100 m s^{-1} into an ice block at 0 °C. The bullet stops after it gets embedded into ice. Assuming that all the kinetic energy of the bullet is converted into heat energy which is absorbed by the melting ice, find the quantity of the ice that just melts.

36. A diesel engine does not contain a spark plug. Then how does ignition takes place in it? Explain.

37. How much energy in watt-hour is required for an electric heater to convert 2 kg of ice at 0 °C to water at 25°C. Assume the water equivalent of the given heater to be 10 g. (Assume that initially the heater and the ice are in thermal equilibrium).

38. Ice of mass 200 g at 0°C is converted to water at 0°C in 8 minutes supplying heat to it at constant rate. Calculate the time taken to raise the temperature of water to 100°C. (take latent heat of fusion of ice = 336 J g^{-1} °C^{-1} and specific heat capacity of water = 4.2 J g^{-1} °C^{-1}.

39. The initial temperature of half litre of water is 20°C. On supplying heat to it its temperature rises to 72 °F. For every 1°C rise in temperature per gram of water, 4.2 J of heat is needed. Calculate the heat capacity of water.

40. When 40 g of ice at 0°C is added to 160 g of water at 40°C, find the final temperature?

41. Two identical calorimeters 'X' and 'Y', of water equivalent 10 g each, contain equal quantities of same liquid. The mass, initial temperature and specific heat capacity of the liquid present in both calorimeters are 50 g, 25°C and $\frac{1}{2}$ cal g^{-1} °C^{-1} respectively. A 10 g metal piece (say A) of specific

heat capacity 0.4 cal $g^{-1}\,°C^{-1}$ is dropped in calorimeter "X". and 25 g of metal piece (say B) of specific heat capacity "S_B" is dropped into calorimeter "Y". Due to this the equilibrium temperature in 'X' and 'Y' rises to 30°C and 40°C respectively. If the initial temperature of 'B' is twice that of 'A', then find the value of "S_B".

42. Calculate the mass of steam at 100°C needed to just melt 320 g of ice at −10°C. Take specific heat of ice = 0.5 cal $g^{-1}\,°C^{-1}$, specific heat of vaporization of steam = 540 cal g^{-1}, specific heat of fusion of ice = 80 cal g^{-1}.

43. A thermally insulated container has 500 g of mercury in liquid state at its freezing point. The container is connected to a vacuum pump and air above the liquid (mercury) is pumped out of the container. Due to this some of the liquid evaporates and the remaining liquid solidifies. Then find the mass of the solid that is found by solidification of liquid. Take latent heat of fusion, L_F = 12500 J kg^{-1} and Latent heat of vaporization L_V = 290000 J kg^{-1}.

44. 100 g of water at 70°C is added to 120 g of water at 30°C contained in a vessel. The final temperature of mixture is 40°C. Calculate the thermal capacity of the vessel.

45. A mercury thermometer whose markings are in centimeter scale, is used to measure the temperature. It shows 3 cm and 15 cm when kept in melting ice and boiling water respectively. Find the equilibrium temperature, as recorded by the above thermometer when 10 g of iron ball at 229° C is dropped into 50 g of water at 25°C. Here, the water equivalent of the container is negligible and heat lost to the surroundings is almost zero. (Take specific heat capacity of iron and water as 0.1 cal $g^{-1}\,°C^{-1}$ and 1 cal $g^{-1}\,°C^{-1}$ respectively).

Concept Application Level—3

46. A closed vessel is partly filled with water and connected to a vacuum pump. With help of this vacuum pump, if the pressure inside the vessel is maintained at low level, discuss the effect on water considering various conditions.

47. Three liquids A, B and C of densities in the ratio of 1 : 2 : 3 respectively are taken. Their initial temperatures are 60°C, 50 °C and 30°C respectively. When equal parts (by volumes) of 'A' and 'B' are mixed, the resultant temperature of the mixture is 55°C. When equal parts (by weights) of A and C are mixed, then the equilibrium temperature of the mixture is 40°C. Find the ratio of specific heat capacity of the liquids.

48. The intensity of sunlight incident on a block of ice is 800 w m^{-2}. If ice absorbs 60% of the light, calculate the thickness of ice that would melt in 1 minute (density of ice = 917 kg m^{-3}).

49. Why is that the hot water is preferred to cold water to put off the fire?

50. Prove that when 'x' grams of steam at 100°C is mixed with 'y' grams of ice at 0°C and allowed to attain thermal equilibrium, the final temperature of mixture is $\theta = \dfrac{80(8x - y)}{x + y}$.

Very short answer type questions

1. Mixture of ice and salt taken in proper proportion.
2. Moisture of air due to water vapour, water vapour present in m^3 of air, extent of moistness of air felt.
3. intake, compression, power and exhaust
4. Pressure, soluble impurities
5. at 4°C
6. Fuel injection pump
7. Anamolous expansion of water
8. Mixtures or calorimetry
9. $C = \frac{Q}{\Delta\theta}, S = \frac{Q}{M\Delta\theta}$
10. Dew formed on dust particles
11. 336×10^3 or 80
12. Heat lost by the hotter body = Heat gained by the colder body.
13. Liquid vapourizing without boiling
14. specific latent heat of vaporization
15. (i) Conversion from solid to liquid
    (ii) Conversion from liquid to solid
    (iii) Conversion from liquid to gas
    (iv) Conversion from gas to liquid
16. Conversion of heat energy to mechanical energy inside the cylinder.
17. 3360 J
18. The heat supplied during phase transition.
19. Evaporation leads to cooling
20. 2344 Jg
21. The heat supplied to convert unit mass of a solid to liquid – specific latent heat while latent heat is conversion from solid to liquid.
22. specific heat capacity of the body
23. A device in which heat is utilised to do some useful work, is called heat engine. Internal and external combustion engine.
24. Conversion from liquid to gas, conversion of unit mass of liquid to gas.
25. Solid converted to vapour directly.
26. temperature
27. Pressure, soluble impurities
28. η = Ratio of output to input.
29. dew point
30. When pressure is applied, the body melts and on release of pressure it resolidifies.

Short answer type questions

31. Thermal power stations, plants do not wilt easily even in summer because water in soil evaporate slowly etc.
32. $Q = ms\Delta\theta$
    Ans: 400 J kg^{-1} $°C^{-1}$
33. Solid converts into gaseous state.
34. (i) Size
    (ii) Safety
    (iii) Starting time
    (iv) Efficiency
35. Land and sea breezes
36. $Q = Q_1 + Q_2 + Q_3$
    $Q = m L_{ice}, Q_2 = ms\Delta\theta, Q_3 = m L_{steam}$
    Ans: 3016×10^3 J
37. (i) Surface phenomenon
    (ii) Rate of process
    (iii) Temperature
    (iv) Effect
    (v) Factors affecting
38. (i) Refrigeration
    (ii) Working principle
    (iii) Construction
    (iv) List of major parts

(v) Working of each part
(vi) Use

39. (i) Definition
(ii) Construction
(iii) Copper vessel and stove
(iv) Wooden box
(v) Working principle
(vi) Use

40. $Q = ms\,\Delta\theta$,
$Q = k\,t$
Ans: 6 kJ $°C^{-1}$

41. (i) Saturation
(ii) Effect of temperature on saturation
(iii) Dew point
(iv) Temperature below dew point

42. $C = Q/\Delta\theta$
Ans: 500 J K^{-1}

43. No flash floods, surrounding weather will become moderate, the cream preparation etc.

44. (i) Efficiency
(ii) Starting time
(iii) Size
(iv) Safety

45. Survival of aquatic flora and fauna. water pipes burst and fruits, vegetables get damaged when temperature falls below 4°C.

Essay type questions

46. Principle of calorimetry.
Heat lost by hot body = Heat gained by cold body.

47. $Q = ms\Delta t$; $Q = mL_V$
Heating curve: Solid converts to liquid then to gaseous state.
Cooling curve: Gas converts to liquid and then to the solid state.

48. $Q_{lost} = Q_{gained}$

50. Latent heat of fusion



Concept Application Level—1

True or false

1. True
2. True
3. True
4. True
5. True
6. False
7. True

Fill in the blanks

8. water
9. sublimation
10. $[ML^2T^{-2}]$
11. Mass, Change in temperature
12. quick; slow
13. calorimeter
14. Radiation

Match the following

15. A : j
B : h
C : i
D : b
E : d
F : e
G : a
H : c
I : f
J : g

Multiple choice questions

16. Choice (1)
17. Choice (2)
18. Choice (4)
19. Choice (1)
20. Choice (4)
21. Choice (1)
22. Choice (1)
23. Choice (2)
24. Choice (4)
25. Choice (4)
26. Choice (1)
27. Choice (3)
28. Choice (4)
29. Choice (3)
30. Choice (1)

Concept Application Level—2,3

Key points

31. (i) The heat energy required by the system (containing both calorimeter and ice) to raise its temperature from $-10°C$ to $0°C$ is

$Q_1 = m_c s_c (10) + m_i s_i (10)$

Substitute, mass of calorimeter, m_c mass of ice, m_i and specific heat capacities of calorimeter and ice and find Q_1

Find the heat energy absorbed by ice to convert into water at $0°C = Q_2 = mL_F$

Substitute for latent heat of fusion, L_F and mass m of ice

Let the heat energy absorbed by the water and calorimeter to increase its temperature from $0°C$ to $50°C = Q_3$.

$Q_3 = m_w s_w (50) + m_c s_c (50)$

The total heat supplied $= Q = Q_1 + Q_2 + Q_3$

The rate of heat supplied to the system $= R = 50$ cal s^{-1}

Time (t) taken by the system, $t = \frac{Q}{R}$

(ii) 9.4 minute

32. (i) Define specific heat capacity.

(ii) What is the major constituent of human body and what is its specific heat capacity?

(iii) What is the consequence of this specific heat capacity on the temperature of human body?

33. When the thermos flask is shaken, work is done on the molecules of the liquid in the flask. By the work-energy theorem, work done is equal to change in energy.

34. (i) Observe the graph carefully and note the quantities represented on X and Y axes.

(ii) From the graph and using $Q = ms\Delta\theta_R$, find mass (m) of the solid.

(iii) Find L using the graph and $Q = mL$.

(iv) From the graph find the specific heat capacity of liquid using $Q = ms\,\Delta\theta_R$.

(v) 0.75 cal g^{-1} $°C^{-1}$

35. (i) The loss of kinetic energy of the bullet

$= \frac{1}{2}mv^2 - \frac{1}{2}mu^2$

Here, $v = 0$ m s^{-1} and $u = 100$ m s^{-1}

m = mass of the bullet $= \frac{50}{1000}$ kg

We know, $W = JQ$

Mechanical equivalent of work

Here, $J = 4.2$ J cal^{-1}

$$Q = \frac{\frac{1}{2}mu^2}{4.2} \text{ cal ------- (1)}$$

Now, this heat energy (Q) is used to melt the ice at $0°C \Rightarrow Q = mL_F$ (2)

Find the value of 'm' from (2)

$m = \frac{Q}{L_F}$

Take the value of $L_F = 80$ cal g^{-1}

(ii) 0.744 g

36. (i) What is the use of spark plug?

(ii) How is the temperature of air affected when it is highly compressed?

(iii) What would happen if diesel is brought in contact of this air?

37. (i) Heat Q supplied to melt the ice at 0°C to water at 0°C = mL_F

Here L_F = latent heat of fusion of ice.

Heat supplied to increase the temperature from 0°C to 25°C = $ms\Delta t$

Here, $\Delta t = 25 - 0 = 25°C$

s = specific heat capacity of water.

m = mass of the substance.

Take 1 watt hour = 3600 J

(ii) 245.3 watt hour

38. (i) Use Q = mL to find the heat absorbed by ice when it melts.

(ii) This heat is supplied over 8 minutes. Find heat supplied per minute.

(iii) Use $Q = ms\Delta\theta_R$ to find the heat required to raise the temperature of water from 0°C to 100°C.

(iv) Use heat supplied per minute × time = heat gained by water, to find time (t).

(v) 10 minutes

39. (i) Use:

heat capacity = mass of substance × heat required per gram to raise its temperature by 1°C.

(ii) 2100 J $°C^{-1}$

40. (i) Let θ be the resultant temperature of the mixture.

(ii) Use Q = mL to find heat absorbed by ice at 0°C to convert into water at 0°C.

(iii) Use $Q = ms\Delta\theta_R$ to find heat absorbed by melted ice to attain the temperature θ°C.

(iv) Use $Q = ms\Delta\theta_F$ to find heat lost by water to attain the temperature θ°C.

(v) Add 2 and 3 and equate with 4 to find θ.

(vi) 16°C

41. (i) In calorimeter X:

Find the initial temperature of metal A by using the principle,

heat lost by metal A = heat gained by liquid calorimeter X

$\Rightarrow m_A . s_A . (\Delta t)_A = \{50\ g \times 0.5 \times (30 - 25) + (5°C \times 10\ g)]\}$

$\Rightarrow 10 \times 0.4 \times (T - 30°C) = (50 \times 0.5 \times 5 + 50\ cal)$ ------- (1)

In calorimeter Y

Find the initial temperature of metal B (2T).

The heat lost by metal B = heat gained by (liquid + calorimeter).

$\Rightarrow m_B . S_B\ (2T - 40) = 50\ g \times 0.5 \times (40 - 25) + (15\ °C \times 25\ g)$ ------- (2)

(ii) $S_B = 0.2\ cal\ g^{-1}\ °C^{-1}$

42. (i) Use $Q = ms\Delta\theta_R$ to find the heat absorbed by ice to raise its temperature from - 10°C to 0°C.

(ii) Use $Q = mL_{ice}$ to find heat absorbed by ice at 0°C to form water at 0°C.

(iii) Add 1 and 2 to find total heat absorbed.

(iv) Use $Q = mL_{vap}$ to find the heat lost by steam at 100°C when it condenses to water at 100°C.

(v) Use $Q = ms\Delta\theta_F$ to find the heat given when water at 100°C cools to 0°C.

(vi) Add 4 and 5 to find total heat given.

(vii) Equate 3 and 6 to find the mass of steam.

(viii) 42.5 g

43. (i) Let the total mass of the liquid = $M = \frac{1}{2}$ kg

Let, the mass of the liquid that is converted to solid = m

If L_v is the latent heat of vaporization and L_F latent heat of fusion then

$mL_F = (M - m)L_V$

Substitute the appropriate values and find the value of m.

(ii) 0.479 kg

44. (i) Use $Q_l = ms\Delta\theta_F$ to find heat lost by hot water.

(ii) Use $Q_g = ms\Delta\theta_R$ to find heat gained by cold water.

(iii) If H_v is the heat capacity of the vessel, use $H_v\Delta\theta_R$ to find heat gained by it.

(H_v = mass of the vessel × specific heat capacity of the vessel)

(iv) Add 2 and 3 and equate with 1 to find the heat capacity of the vessel.

(v) 756 J $°C^{-1}$

45. (i) Let the equilibrium temperature of the ball and water after the ball is dropped in to water be T.

Apply the principle of mixture (or) calorimetry:

Heat lost by the hot body = heat gained by the cold body.

$m_b\, s_b(\Delta t)_b = m_w\, s_w(\Delta t)_w$

$\Rightarrow 10 \times 0.1 \times (22\text{ g} - T) = 50 \times 1.$

x (T – 25) ------ (1)

Find the value of T from (1).

To find the temperature in new scale, use the formula

$$\left(\frac{T - LFP}{UFP - LFP}\right)_{\text{scale 1}}$$

$$= \left(\frac{\text{Length(x)} - LFP}{UFP - LFP}\right)_{\text{scale 2}} \text{-------- (2)}$$

$$= \left(\frac{T - 0}{100 - 0}\right) = \left(\frac{x - 3}{15 - 3}\right)$$

(ii) 6.48 cm

46. Consider the following conditions:

1. temperature of water is above room temperature
2. temperature below room temperature
3. boiling point of water at low pressures.
4. heat losses/exchange to the surrounding.
5. insulation of the vessels.

47. (i) Find the ratio of the densities of A, B and C from given information.

Here, density (d) $= \dfrac{\text{mass (m)}}{\text{volume (V)}}$ ____ (1)

When equal volumes of A and B liquids are taken, then find the ratio of their densities. Use principle of mixtures (or) calorimetry and find the ratio of specific heats of A and B. Get the equilibrium temperature of A and C from the given data.

Again, heat lost by "A" liquid = heat gained by "C" liquid. $\Rightarrow m_A\, s_A\, (\Delta t)_A = m_C\, s_C\, (\Delta t)_C$

Given that $m_A = m_C$

Find the ratio of s_A and s_C

(ii) 2 : 1 : 4

48. (i) From the given data find the heat absorbed (Q) by the ice block per unit area in 1 minute

(60 seconds).

(ii) If A is the area of the block, total heat absorbed is QA.

(iii) If t is the thickness of the block, its volume = At and mass = 917At

($\because$ density of ice = 917 kg m^{-3} (given))

(iv) Heat absorbed during melting of ice

$= mL_{ice}$

$= 917\text{ At} \times 336000\text{ J}$

($\because L_{ice} = 336000\text{ J g}^{-1}$)

(v) Equate 4 and 2 to find 't'.

(vi) 0.09 m

49. (i) Which one gets converted into water vapour hot water or cold water?

(ii) Engulfed water vapour reduces the supply of oxygen.

50. (i) Let the resultant temperature be θ.

(ii) Use $Q = mL_{vap}$ to find heat lost when steam at 100°C condenses to water at 100°C.

(iii) Use $Q = ms\Delta\theta_F$ to find heat lost when water at 100°C cools to θ°C.

(iv) Add 2 and 3 to find total heat lost.

(v) Use $Q = mL_{ice}$ to find heat gained when ice at 0°C melts to form water at 0°C.

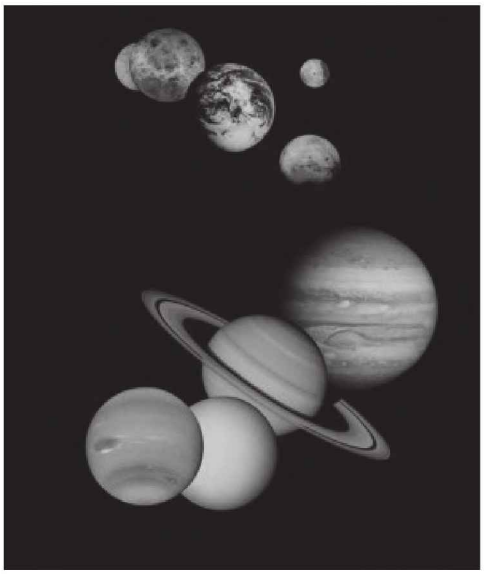
(vi) Use $Q = ms\,\Delta\theta_R$ to find heat gained when water at 0°C heats to θ°C.

(vii) Add 5 and 6 to find total heat gained.

(viii) Equate 4 and 7 to find θ in the given form.

4

Light



INTRODUCTION

Nature around us is colourful and diverse. It is impossible to enjoy nature without light. Light causes sensation of sight. Investigation of nature of light has opened a treasure of knowledge and development.

Nature of light

Light is a form of electromagnetic radiation and a non-mechanical wave. Non-mechanical waves do not require a material as a medium for propagation. Light exhibits dual nature, i.e., both particle nature as well as wave nature, and some phenomena of light can be explained by considering light as a wave (interference, diffraction and polarization) while other phenomena like photo-electric effect, Compton effect, etc., can be explained by treating light as a particle.

Light travels at a speed of nearly 3×10^8 m s^{-1} in air or vacuum. Wavelength of visible light is 400 nm to 800 nm (4000 A° - 8000 A°) and the corresponding frequency is 7.5×10^{14} Hz to 3.75×10^{14} Hz.

Light travels along a straight line and this property is called the rectilinear propagation. The straight line is called a ray, and a bundle of rays is called a beam of light.

When light is incident on a surface, a part of it is reflected, a part of it is refracted and the remaining part is absorbed. Light is reflected by highly polished surfaces (mirrors) and it is refracted partially by transparent media.

Light rays can be focused using mirrors and lenses. If light rays meet at a point after reflection or refraction by a mirror or through a lens respectively, an image is obtained on a screen. These images are called real images.

Thus real images are formed when the rays after reflection or refraction meets at a point. If the rays appear to meet when produced backwards, then the image formed is virtual. The real image can be caught on a screen but the virtual image cannot be formed on a screen.

Reflection of Light

When light is incident on a surface, separating two media, a part of the light is bounced back to the first medium. This is known as reflection.

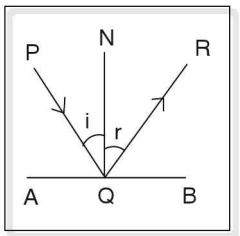


Figure 4.1

AB is the surface, PQ is the incident ray and QR is the reflected ray.

$\angle i$ = angle of incidence

$\angle r$ = angle of reflection

Laws of reflection

1. The incident ray, the reflected ray and the normal to the reflecting surface at the point of incidence lie in the same plane.
2. The angle of incidence is equal to the angle of reflection ($\angle i = \angle r$).

Reflection by a plane mirror

An object AB is placed at a distance 'u' on one side of the plane mirror. Each point of AB acts as a source. To find the distance and nature of the image, consider a ray AC incident on the normal and another ray AO incident on the mirror at an angle 'i'. The ray AC retraces its path since it is incident along the normal. But the ray AO is reflected on OD at an angle 'r'. The two reflected rays are produced backwards and they intersect at A^1. Similarly B^1 will be the virtual image of point B. Every point of the object will form the corresponding image between A^1 and B^1. Thus A^1B^1 is the virtual, erect image of the object AB.

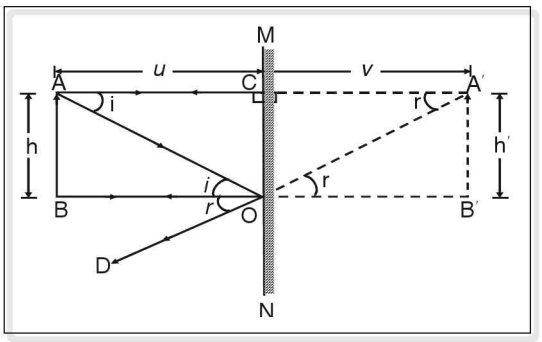


Figure 4.2 An extended object AB and its virtual image A′B′ formed in a plane mirror MN.

Image formed by plane mirror

1. The image is virtual and erect.
2. It is of the same size as that of the object.
3. The distance between the object and the mirror is the same as that between the image and the mirror.
4. It is laterally inverted.

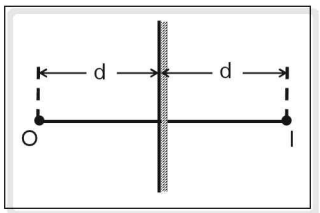


Figure 4.3

Spherical mirror

Spherical mirrors are a part of a sphere. Consider a hollow sphere. Now consider slicing a portion out of a sphere and the portions of the sphere are silvered on the inner or outer surface.

Generally, the smaller piece is used as a mirror.

If a mirror is obtained from a sphere silvered on the outer side, it is called a concave mirror (shown in A).

If a mirror is obtained from a sphere silvered inside, it is called a convex mirror (shown in B).

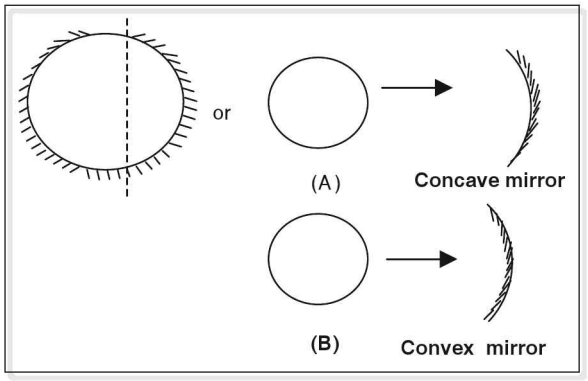


Figure 4.4

Terms relating to spherical mirrors

Length AB is the measure of aperture, C is the centre of curvature, P is the pole, and PC is the principal axis.

Aperture

The portion available for reflection is called aperture; APB is the aperture.

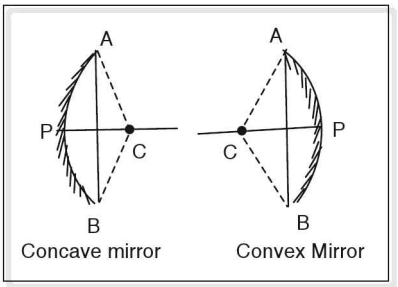


Figure 4.5

Pole

It is the geometric centre of the reflecting surface. P is the pole.

Centre of curvature

It is the centre of the sphere of which the mirror forms a part. C is the centre of curvature.

Principal axis

It is the straight line passing through the centre of curvature and the pole. The line passing through P and C in the figure is the principal axis.

Radius of curvature (R)

It is the radius of the sphere of which the mirror forms a part. PC is the radius of curvature.

Principal focus

Consider a parallel beam of rays incident on a spherical mirror. In case of the concave mirror, the parallel beam after reflection converges at a point F called the principal focus.

In case of a convex mirror, it appears to diverge from a point F called the focus.

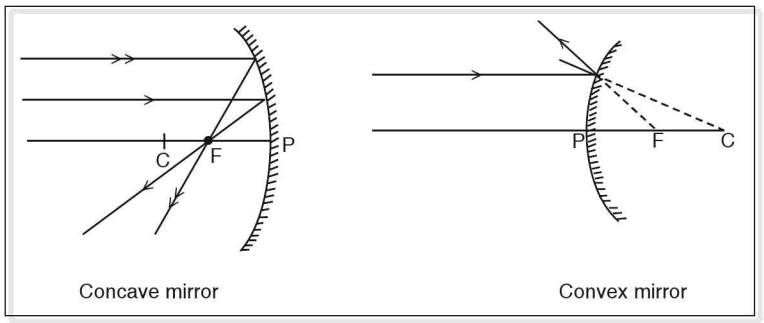


Figure 4.6

Thus a concave mirror is called a converging mirror and a convex mirror is called a diverging mirror.

Focal length

It is the distance between the pole and the principal focus. PF is the focal length. Focal length is measured in m or cm.

Relation between focal length and radius of curvature

Consider a concave mirror of small aperture. P is the pole of the mirror. PC is the radius of curvature = R

CM is the normal to the spherical surface.

PF is the focal length = f

OM is the ray parallel to principal axis and is incident on the mirror at an angle 'i'.

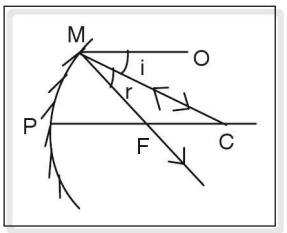


Figure 4.7

The ray OM after reflection passes through principal focus (F). A ray CM, incident normally, retraces its path after reflection.

From laws of reflection

$\angle i = \angle r$ --- (1) ($\angle r = \angle CMF$)

where angle r is the angle of reflection and i is the angle of incidence.

$\angle MCF = \angle i$ (alternate angles) --- (2)

From (1) and (2)

$\angle MCF = \angle r$ --- (3)

From geometry

FC = FM --- (4) ($\because$ base angles of Δ^{le} CMF are equal)

Since we are considering a small aperture,

FM = PF --- (5)

PF + FC = PC

PF + PF = PC --- (6) ($\because$ PF = FC) [from (4) and (5)]

PF is the focal length = f

PC is the radius of curvature = R

$\therefore$ f + f = R

$\Rightarrow$ 2f = R

or $f = \frac{R}{2}$

Thus for a spherical mirror, the radius of curvature is twice its focal length.

Rules to determine the path of reflected rays

1. In case of a concave mirror, incident rays parallel to the principal axis on reflection pass through principal focus whereas in a convex mirror the reflected rays only appear to pass through the principal focus.

2. An incident ray passing through the centre of curvature gets reflected along the same path in case of the concave mirror. In case of the convex mirror, the reflected ray only appears to pass through the centre of curvature

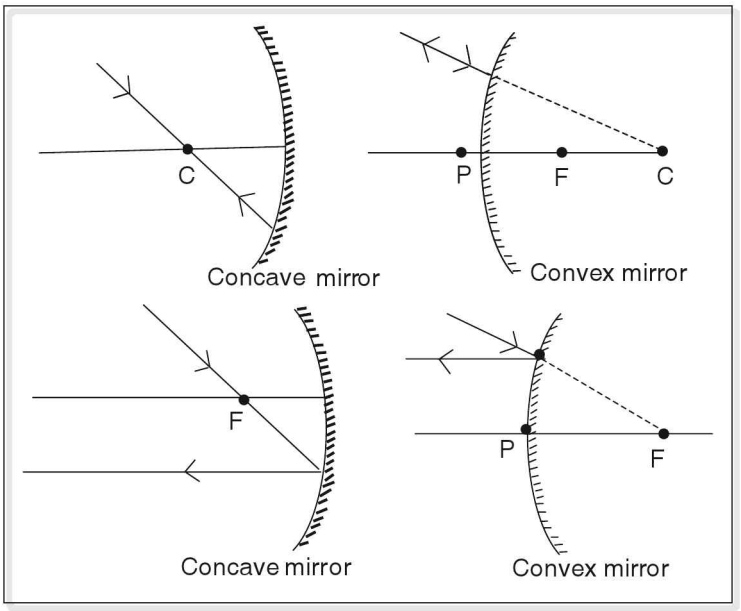


Figure 4.8

3. In case of the concave mirror an incident ray passing through the principal focus reflects parallel to the principal axis. In case of the convex mirror, an incident ray which appears to pass through the focus reflects parallel to the principal axis.

Formation of image by convex mirror

OJ is an extended object and IG is an image formed by a convex mirror of small aperture.

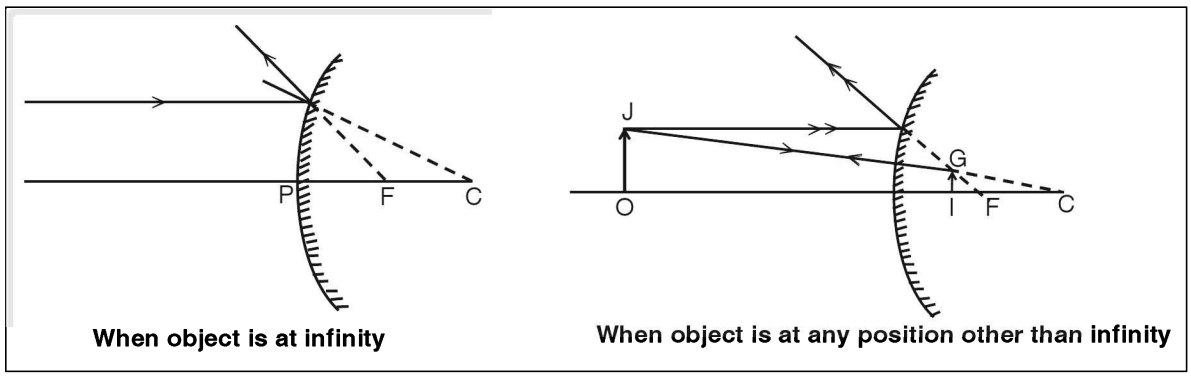


Figure 4.9

Formation of image by concave mirror

AB is an extended object and A_1B_1 is an image formed by the concave mirror of small aperture.

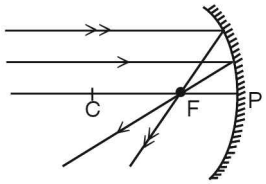
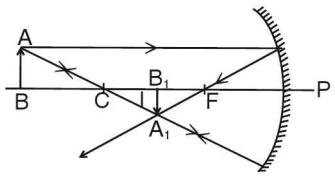
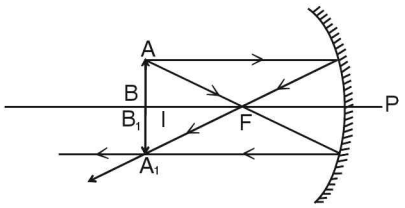
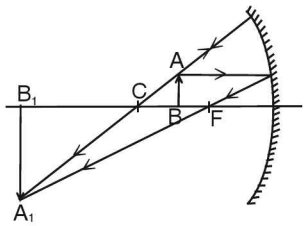
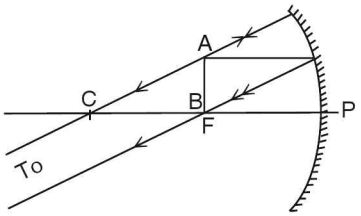
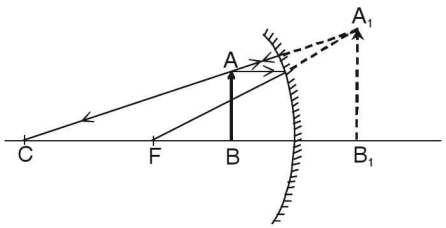
$u = \infty$		Real, inverted and highly diminished, $v = f$
$R < u < \infty$		Real, inverted and diminished, $f < v < R$
$u = R$		Real, inverted and of equal size as that of object, $v = R$
$f < u < R$		Real, inverted and magnified, $v > R$
$u = f$		Real, inverted and highly magnified, $v = \infty$
$u < f$		Virtual, erect and magnified and formed on the opposite side of the mirror as that of the object.

Image formed by mirror of large aperture is not distinct and the image suffers defects and distortions.

New Cartesian sign convention of reflection of spherical mirror

1. All distances are measured from the pole.
2. Distances measured in the same direction of the incident ray are taken as positive and the distances measured in the direction opposite to the incident ray are taken as negative.
3. Perpendicular distances (height) above the principal axis are positive and perpendicular distance below the principal axis are negative thus focal length and radius of curvature is negative for concave mirror and positive for convex mirror.

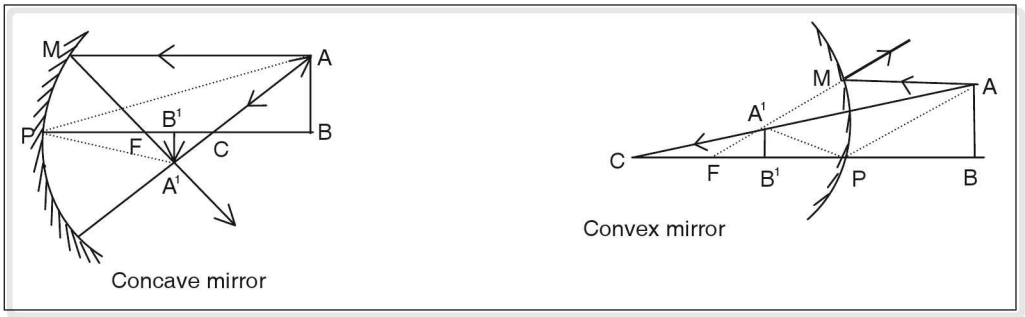


Figure 4.10

Consider a spherical mirror (concave or convex mirror) of radius of curvature 'R'. Let 'f' be the focal length of the spherical mirror. AB is an object placed at a distance 'u' from the pole P.

A^1B^1 is an image of an object at a distance 'v' from the pole P. PC is the radius of curvature.

Triangle ABP and triangle A^1B^1P are similar.

$$\therefore \frac{A^1B^1}{AB} = \frac{PB^1}{PB}$$

Using Cartesian sign convention, we get

$$\frac{PB^1}{PB} = \frac{-v}{-u}$$

$$\frac{A^1B^1}{AB} = \frac{-v}{-u} = \frac{v}{u} \text{ --- (1)}$$

From similar triangles, we have ABC and A^1B^1C

$$\frac{A^1B^1}{AB} = \frac{B^1C}{BC} = \frac{PC - PB^1}{PB - PC}$$

$$\frac{A^1B^1}{AB} = \frac{PB^1}{PB}$$

Using Cartesian sign convention, we get

$$\frac{PB^1}{PB} = \frac{v}{-u}$$

$$\frac{A^1B^1}{AB} = \frac{v}{-u} \quad \text{--- (1)}$$

From similar triangles ABC and A^1B^1C, we have

$$\frac{A^1B^1}{AB} = \frac{B^1C}{BC} = \frac{PC - PB^1}{PC + PB}$$

(Continued on following page)

Using Cartesian sign convention and (1)	Using Cartesian sign convention and (1) we get
$\frac{v}{u}=\frac{-R-(-v)}{-u-(-R)}$	$\frac{v}{-u}=\frac{R-v}{R+(-u)}$
$\frac{v}{u}=\frac{-R+v}{-u+R}$	$\frac{v}{-u}=\frac{R-v}{R-u}$
$-uv+vR=-uR+vu$ $\Rightarrow vR+uR=uv+uv \Rightarrow vR+uR=2uv$ dividing throughout by uvR	$\Rightarrow vR-uv=-uR+uV$ $\Rightarrow uR+vR=uv+uv \Rightarrow uR+vR=2uv$ dividing throughout by uvR
$\frac{1}{u}+\frac{1}{v}=\frac{2}{R}$	$\frac{1}{v}+\frac{1}{u}=\frac{2}{R}$
$\frac{1}{u}+\frac{1}{v}=\frac{1}{f}$	$\frac{1}{u}+\frac{1}{v}=\frac{1}{f}$

$\because \frac{R}{2}=f$

$\therefore$ The relation between object distance, image distance and the focal length for a mirror is $\frac{1}{f}=\frac{1}{u}+\frac{1}{v}$

Magnification (m)

It is the ratio of image height to the object height

$$m=\frac{\text{height of the image } (h_i)}{\text{height of the object } (h_o)}=\frac{A^1B^1}{AB}$$

$$m=\frac{h_i}{h_o}=\frac{-v}{u} \qquad \left(\because \frac{A^1B^1}{AB}=\frac{v}{u}\right)$$

(using Cartesian sign convention)

Magnification is positive if the image is erect and virtual and it is negative, if the image is real and inverted.

☛ *Example*

Find the radius of curvature of a concave mirror of focal length 15 cm.

Solution

Focal length = $\frac{1}{2}$ radius of curvature (R)

Given

F = 15 cm

R = 2f = 2 × 15 = 30 cm

Since it is a concave mirror, 'R' is negative (from Cartesian sign).

☛ *Example*

A rear view mirror of an automobile has an focal length of 2 m. If it locates a person standing at a distance 4 m from it, find the nature and distance of the image.

Solution

Since the mirror is a rear view mirror, the mirror is a convex mirror

Given

Focal length = f = 2 m (focal length is positive)

Object distance = u = – 4 m

∵ From Cartesian sign convention object distance is in the opposite direction to that of the incident ray

Using mirror formula, we get

$$\Rightarrow \frac{1}{2} = \frac{1}{-4} + \frac{1}{v} \Rightarrow \frac{1}{v} = \frac{1}{2} + \frac{1}{4}$$

$$\frac{1}{v} = \frac{4+2}{4 \times 2} \Rightarrow \frac{1}{v} = \frac{6}{8}$$

$$v = \frac{8}{6} = 1.33 \text{ m}$$

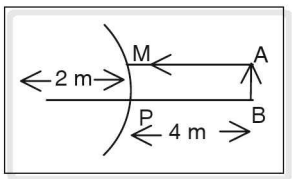


Figure 4.11

Since v is positive, image is virtual (it is formed at the back of the convex mirror)

Magnification (m)

$$m = \frac{\text{height of the image}}{\text{height of the object}} = -\frac{v}{u}$$

$$m = \frac{-1 \times 3}{-4} = 0 \times 3325$$

Since magnification is positive, the image is erect and virtual.

☛ *Example*

An object of length 2 cm is placed at distance 2.5 f from a concave mirror where 'f' is its focal length. Find the length of the image.

Solution

Height of the object = 2 cm

Focal length = –f (for concave mirror f is negative)

Object distance u = –2.5 f (from Cartesian sign convention)

Using the mirror formula, we get

$$\frac{1}{f} = \frac{1}{u} + \frac{1}{v}$$

$$\frac{1}{-f}=\frac{1}{-2.5f}+\frac{1}{v}\Rightarrow\frac{1}{v}=\frac{1}{-f}+\frac{1}{2.5f}\Rightarrow\frac{1}{v}=\frac{-1}{f}+\frac{2}{5f}$$

$$\frac{1}{v}=\frac{-5f+2f}{5f^2}\text{ or }v=\frac{-5f}{3}\text{ magnification }m=\frac{h_i}{h_0}=\frac{-v}{u}$$

$$\frac{h_i}{2}=-\frac{\left(\frac{-5f}{3}\right)}{\left(\frac{-5}{2}f\right)}=\frac{-2}{3}\Rightarrow h_i=\frac{-2}{3}\times 2=\frac{-4}{3}$$

Image is 1.33 cm long. Minus sign shows it is inverted.

☛ *Example*

A rod of length 10 cm lies along the principal axis of a concave mirror of focal length 10 cm in such a way that the end close to the pole is 20 cm away from it. Find the length of the image of the rod.

Solution

Focal length f = –10 cm

∴ Radius of curvature R = 2f = –20 cm

Length of the rod AB = 10 cm

Distance of the end A = 20 cm = R

Since the distance of object 'A' is on the centre of curvature the image A^1 will be at 'C'.

Distance of the end B = 20 + 10 = 30 cm

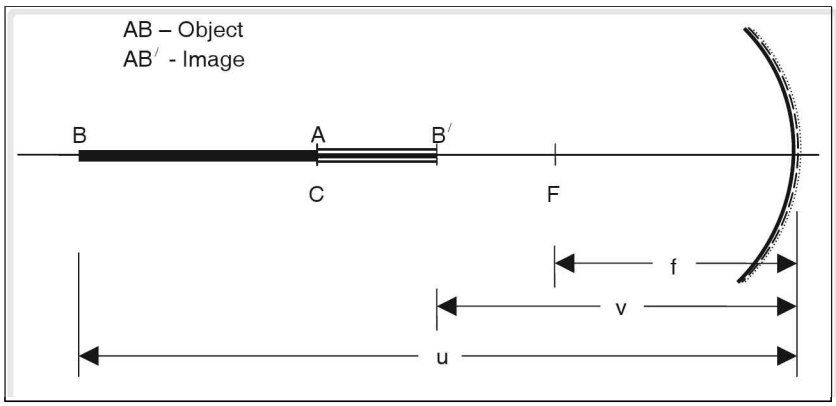


Figure 4.12

∴ u = –30 cm (from Cartesian sign convention)

f = –10 cm

v = ?

From mirror formula

$$\frac{1}{f}=\frac{1}{u}+\frac{1}{v}$$

$$\frac{1}{-10}=\frac{1}{-30}+\frac{1}{v}$$

$$\frac{1}{v}=\frac{-1}{10}+\frac{1}{30}\Rightarrow\frac{1}{v}=\frac{-30+10}{30\times 10}$$

$$\frac{1}{v} = \frac{-20}{300}$$

$$v = \frac{-30}{2} = -15 \text{ cm}$$

Length of the image = A^1B^1 = 20 – 15 = 5 cm

Refraction

When light passes from one medium to another medium, there is a change in the path of light in the second medium. This is known as refraction.

Deviation in the path of light is due to the change in the speed of light when it passes from one medium to another.

Rectilinear propagation of light is obeyed only if the light propagates in the same medium, i.e., if the optical density of the medium remains constant.

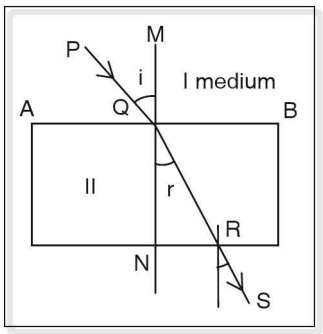


Figure 4.13

PQ is the incident ray

QR is the refracted ray

RS is the emergent ray

MN is the normal

∠i is the angle of incidence

∠r is the angle of refraction

When a light ray is incident from rarer medium to the denser medium, the refracted ray bends towards the normal i.e., ∠i > ∠r

When a light ray travels from denser medium to a rarer medium, the refracted ray bends away from the normal i.e., ∠i < ∠r

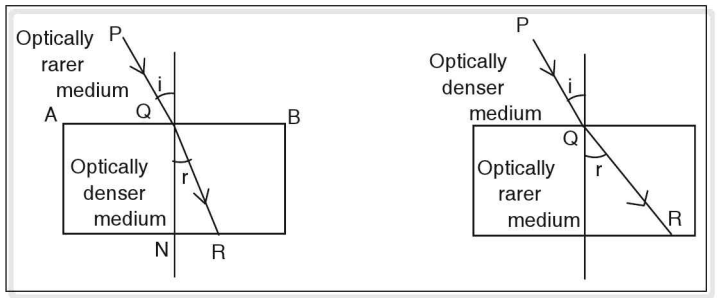


Figure 4.14

Thus the deviation of the ray of light depends on the optical density of the two media.

Laws of refraction

I Law

The incident ray, the refracted ray and the normal to the surface separating the two media at the point of incidence, all lie on the same plane.

II Law

The ratio of sine of angle of incidence to the sine of angle of refraction is a constant for a given pair of media and for a given colour of light. This law is also known as Snell's law.

Mathematically $\frac{\sin i}{\sin r} = \mu$ where μ is a constant called refractive index.

Snell's law is not applicable when angle of incidence is equal to zero, i.e., when the ray is incident along the normal.

Refractive index of a medium

Refractive index of a medium is the ratio of velocity of light in air or vacuum to its velocity in a given medium. This is known as the absolute refractive index.

$$\therefore \mu = \frac{\text{velocity in air or vacuum (c)}}{\text{velocity in given medium (v)}} \Rightarrow \mu = \frac{c}{v}$$

where μ is absolute refractive index. If μ_1 and μ_2 are the refractive indices of two media and v_1 and v_2 are the velocity of light in the medium one and medium two respectively,

then

$$\mu_1 = \frac{c}{v_1} \qquad \text{--- (1) and}$$

$$\mu_2 = \frac{c}{v_2} \qquad \text{--- (2)}$$

Dividing (2) by (1), we get

$$\frac{\mu_2}{\mu_1} = \frac{\frac{c}{v_2}}{\frac{c}{v_1}}$$

$$\frac{\mu_2}{\mu_1} = \frac{v_1}{v_2} = {}^1\mu_2$$

Where ${}^1\mu_2$ is the relative refractive index of medium 2 with respect to medium 1.

$$\therefore {}^1\mu_2 \text{ or } {}^1\mu_2 = \frac{\mu_2}{\mu_1} = \frac{v_1}{v_2}$$

Air	1.00
Water	1.33
Glass	1.5
Diamond	2.42

Experiment to verify second law of refraction or to find absolute refractive index of a glass

Take a glass slab and place it on a white paper which is fixed to a drawing board.

1. Trace the boundary ABCD of the glass slab and remove it.
2. Draw a MN normal to the glass surface AB. Make a suitable angle (angle of incidence) with the normal. This is the line of incidence.
3. Fix two pins P_1 and P_2 on the line of incidence.
4. Replace the glass slab on the boundary traced. i.e., on ABCD.
5. Locate the image of the pins P_1 and P_2 looking through the opposite face of the glass slab through CD.
6. The image of the pins P_1^1 and P_2^1 appears to be in line with P_1 and P_2. Fix two pins at these points, i.e. P_1^1 and P_2^1.
7. Mark the points of the pins P_1, P_2, P_1^1 and P_2^1.
8. Remove the glass slab.
9. Join P_1^1 and P_2^1 to meet the surface CD, and join QR.
10. PQ is the incident ray, and QR is the refracted ray.
11. Measure the angle of refraction, $\angle r = \angle NQR$.
12. Find the refractive index of the glass

$$\mu = \frac{\sin i}{\sin r}.$$

13. Repeat the experiment for different angles of incidence. Tabulate the results.
14. In each case it is found that the ratio of sin i to sin r is a constant.

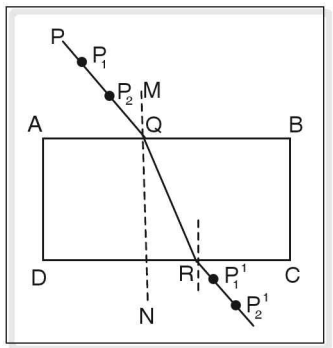


Figure 4.15

Principle of reversibility of light

If a ray of light travels from medium 1 to medium 2 along a certain path, it retraces the path, when it travels from medium 2 to medium 1. Thus the path of light is reversible.

AB is the refracting surface separating the two media. MN is the normal.

A ray of light PQ is incident at an angle 'i' and is refracted along QR at an angle 'r'.

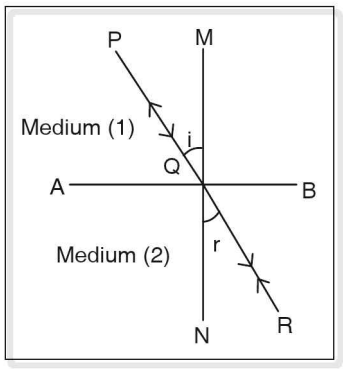


Figure 4.16

Refractive index of medium 2 with respect to medium 1 is

$$^{1}\mu_{2} \text{ or } {}_{1}\mu_{2} = \frac{\sin i}{\sin r} \quad \text{--- (1)}$$

When a ray of light travels from medium 2 to medium 1, according to the principle of reversibility, RQ is the incident ray and PQ is the refracted ray, refracted at an angle 'i'.

Refractive index of medium 1 with respect to medium 2 is

$$^{2}\mu_{1} = \frac{\sin r}{\sin i} \quad \text{--- (2)}$$

☛ *Example*

The velocity of light in air is 3×10^8 m s^{-1}. If the refractive index of water is 4/3. Find the velocity of light in water.

Solution

Given
Velocity of light in a C = 3×10^8 m s^{-1}

Refractive index of water $\mu_w = \frac{4}{3}$

Refractive index of water = $\frac{c}{v_w}$

Where v_w is the velocity of light in water

$$\frac{4}{3} = \frac{3 \times 10^8}{v_w}$$

$$v_w = \frac{3 \times 3 \times 10^8}{4}$$

$v_w = 2.25 \times 10^8$ m s^{-1}

Velocity of light in water is 2.25×10^8 m s^{-1}.

☛ *Example*

The angle of incidence of a ray of the light is 45°. If ray travels from air to glass, find the angle of refraction. Refractive index of glass, $\mu_g = 1.5$.

Solution

Given i = 45° = angle of incidence

Refractive index of glass = $\mu_g = 1 \cdot 5$

Angle of refraction r = ?

$\mu_g = \frac{\sin i}{\sin r}$ (from Snell's law)

$1.5 = \frac{\sin 45}{\sin r}$

$1.5 = \frac{\frac{1}{\sqrt{2}}}{\sin r}$

$\sin r = \frac{1}{1.5 \times \sqrt{2}}$

sin r = 0.4728. From sine trigonometric table, we get r ≈ 28°.

☛ *Example*

Refractive index of diamond is 2.5 and that of glass is 1.5. Calculate the refractive index of diamond with respect to glass.

Solution

Refractive index of glass $\mu_g = 1.5$

Refractive index of diamond $\mu_d = 2.5$

From relative refractive index

$${}^{g}\mu_d = \frac{\mu_d}{\mu_g} = \frac{2 \cdot 5}{1 \cdot 5} = 1 \cdot 667$$

Refraction through glass slab

Consider a glass slab ABCD placed in air. A ray PQ incident from air to glass at an angle i_1 is refracted along QR at an angle i_2. The ray QR is incident along glass – air interface, hence it undergoes refraction along RS at an angle i_3 (RS is called as an emergent ray)

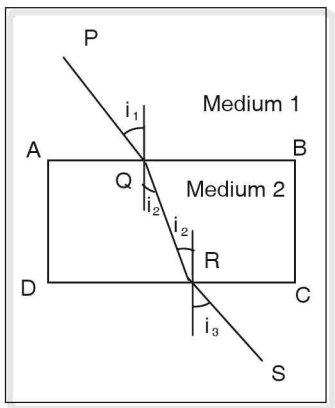


Figure 4.17

From Snell's law

$$\frac{\mu_2}{\mu_1} = \frac{\sin i_1}{\sin i_2} \quad \text{--- (1)}$$

where μ_2 is refractive index of glass and μ_1 is refractive index of air.

From equation (1)

$\mu_1 \sin i_1 = \mu_2 \sin i_2$ --- (2)

At glass air interface

$$\frac{\mu_1}{\mu_2} = \frac{\sin i_2}{\sin i_3}$$

$\mu_1 \sin i_3 = \mu_2 \sin i_2$ --- (3)

Comparing (2) and (3)

$\mu_1 \sin i_1 = \mu_1 \sin i_3$

Since $\mu_1 = 1$ (refraction index of air)

$\sin i_1 = \sin i_3$

Thus the emergent ray is parallel to the incident ray.

Lateral displacement

Consider an optical medium (ABCD) of thickness 't'. PQ is the incident ray, incident at an angle i, the ray PQ undergoes refraction along QR at an angle 'r'. In the absence of an optical medium the ray PQ would have travelled straight without any deviation along PQT. Due to the presence of a medium the incident ray undergoes shift. This is known as lateral displacement. Thus lateral displacement is the perpendicular distance between the emergent ray and the incident ray produced. RT measures lateral displacement.

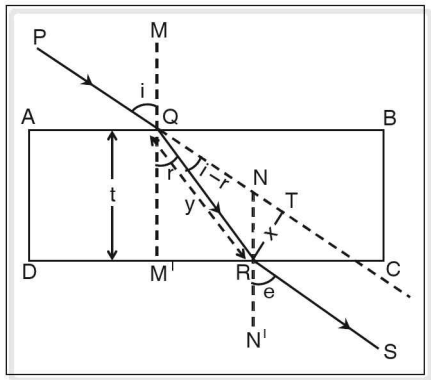


Figure 4.18

The lateral displacement depends on the following factors.

1. Thickness of the medium.
   Lateral displacement is directly proportional to the thickness.
2. Refractive index of the medium.
   Lateral displacement is directly proportional to the refractive index of the medium.
3. Angle of incidence
   Lateral displacement varies directly as the angle of incidence
4. Wavelength of light
   Lateral displacement varies inversely as the wavelength of light.

☛ *Example*

A stick immersed obliquely in water appears to be bent. This is due to lateral displacement. When a ray of light passes from water to air, it bends away from the normal, hence the stick appears to be bent and short.

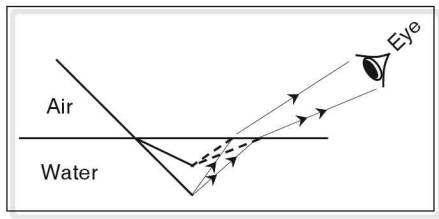


Figure 4.19 Apparent bent in a stick

Real and apparent depth

Consider an object 'O' placed in an optically denser medium. A ray OM passes without deviation, since it is incident along the normal. A ray OP incident at an angle i undergoes refraction at an angle 'r' along PR. When refracted ray PR is produced backwards, it meets the normal at I. The object appears to come from I and not from O. I is the virtual image of the object O. IM is the apparent depth and OM is the real depth.

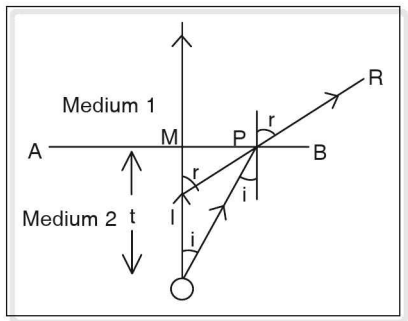


Figure 4.20

$$\text{Refractive index of the medium} = \frac{\text{real depth}}{\text{apparent depth}}$$

Real depth–it is the actual depth of the object.
Apparent depth – it is depth of the image of the object.

☛ *Example*

1. The floor of a swimming pool appears to be raised when seen from the diving board. This is due to refraction.
2. Sun appears before it actually rises. This is due to refraction of light.

☛ *Example*

A coin placed in a beaker containing water appears at a depth of 3 cm from the water surface. What is the real depth of the coin? Refractive index of water = $\frac{4}{3}$

Solution

Real depth = ?
Apparent depth = 3 cm
Refractive index of water = $\frac{4}{3}$

$$\text{Refractive index} = \frac{\text{real depth}}{\text{apparent depth}}$$

$$\Rightarrow \frac{4}{3} = \frac{\text{real depth}}{3}$$

$$\frac{4}{3} \times 3 = \text{real depth}$$

real depth = 4 cm

Total internal reflection

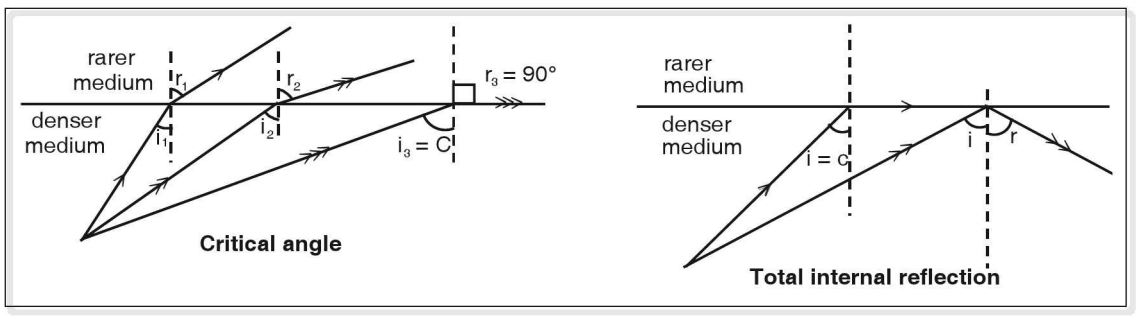


Figure 4.21

When a ray of light passes from denser to rarer medium, the refracted ray bends away from the normal. Hence the angle of refraction is greater than the angle of incidence. Gradually if the angle of incidence increases, angle of refraction also increases. At a particular angle of incidence, the refracted ray just grazes the refracting surface or the angle of refraction is equal to 90°. This angle of incidence in the denser medium is called critical angle. If the angle of incidence is further increased the rays do not undergo refraction, it gets reflected into the same optically denser medium. This is known as total internal reflection.

Conditions for total internal reflection

1. The light ray should travel from denser medium to rarer medium.
2. Angle of incidence in the denser medium should be greater than the critical angle.

☛ *Example*

1. Diamonds are cut into sharp angles. This gives rise to a number of refracting surfaces. Light rays entering diamond suffers total internal reflection. Light rays strike the diamond-air interface at greater than 24° (which is the critical angle of diamond). Thus the light rays get trapped. These trapped light rays make the diamond to sparkle.

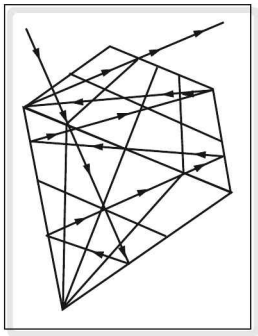


Figure 4.22 Diamond

Mirage

Inverted images of distant objects on the ground are usually seen during hot summer days. These images are seen as a result of an optical illusion called mirage. Such an illusion caused due to perceived presence of water though in reality water is non existent.

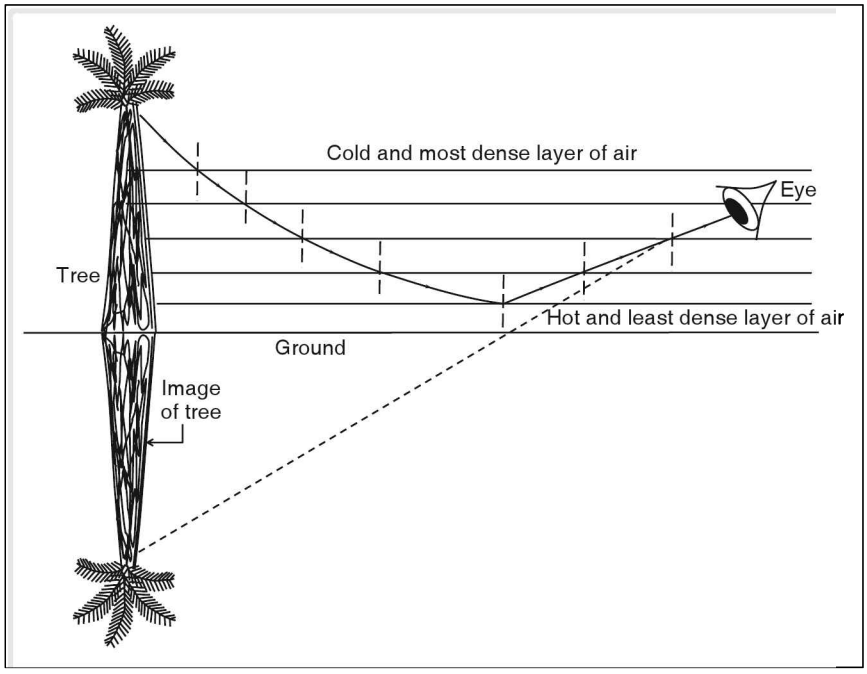


Figure 4.23 Formation of mirage

During peak summer days, the layer of air just above the earth is less dense than the air above it. Light travelling from distant or tall object passes from layer of high refractive index (denser medium) to layer of low refractive index (rarer medium). Thus the refracted ray bends more and more away from the normal.

This happens till the angle of incidence is equal to the critical angle. Once the angle of incidence is greater than the critical angle, light rays undergo total internal reflection and reach the eye and the person sees the image of the object, when there is no water around.

Twinkling of stars

Twinkling of stars is due to atmospheric refraction.

Light from stars undergoes continuous refraction as it enters the earth's atmosphere. Since the refracted ray bends towards the normal stars appear at different position and appear slightly higher than their actual positions. The apparent position of the star changes gradually due to change in the earth's atmosphere. Thus the apparent position of the star fluctuates. As a result, the path of light from the star varies. (flux entering the eye also varies).

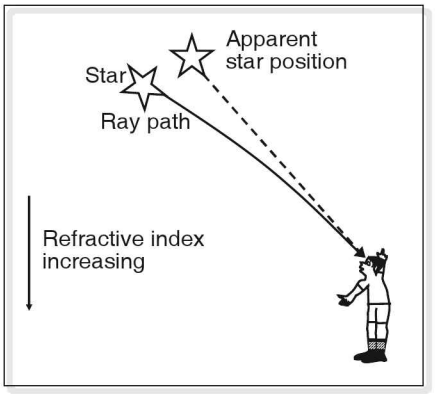


Figure 4.24 Apparent star position due to atmospheric refraction.

Due to change in the light flux, star appears brighter some times and fainter at the other time. Thus the twinkling of stars is due to fluctuation of apparent position of the star.

Planets do not twinkle because they appear bigger than stars. A planet can be considered as a collection of a large number of point sized objects which nullify the twinkling effect.

Relation between critical angle and refractive index

Consider a ray of light PQ travelling from denser to rarer medium. Let the ray be incident at the critical angle. The angle of refraction will be equal to 90°.

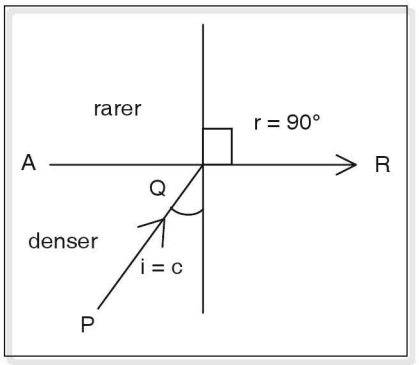


Figure 4.25

From Snell's Law

$$^{1}\mu_{2} = \frac{\sin i}{\sin r} \quad \text{--- (1)}$$

but i = c and r = 90°

$$^{1}\mu_{2} = \frac{\sin c}{\sin 90}$$

$$^{1}\mu_{2} = \frac{\sin c}{1} \quad (\because \sin 90 = 1) \Rightarrow \frac{\mu_2}{\mu_1} = \sin c$$

$$[^{1}\mu_{2} = \frac{\mu_2}{\mu_1} \text{ from relative refractive index}]$$

If the second medium is air, then $\mu_2 = 1$ and $\mu_1 = \mu$ $\frac{1}{\mu} = \sin c$ or $\mu = \frac{1}{\sin c}$

Differences between reflection and total internal reflection

For reflection to take place, smooth polished surface is required.	No smooth polished surface is required for total internal reflection to occur.
It takes place at all angles of incidence.	Total internal reflection can occur only if the angle of incidence is greater than critical angle.
Some percentage of light rays are absorbed by the reflecting material.	No rays of light are absorbed.

Total reflecting prisms

Total reflecting prisms are right angled isosceles prism. They are based on the principle of total internal reflection. These prisms are used to rotate light through 90° or 180° and are also used to obtain erect images of an inverted object.

Deviation through 90°

When a ray of light is incident normally on the surface of a right angled isosceles prism, it does not undergo refraction but strikes the surface AC at an angle 45°, which is greater than the critical angle of glass i.e., 42°. The ray undergoes total internal reflection and strikes the face BC normally and emerges out of it. Thus the ray is rotated through 90°. Surface AC acts like a mirror.

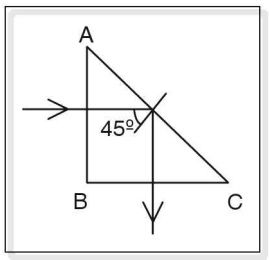


Figure 4.26

Deviation through 180°

When a ray of light is incident on the surface AC of the prism, it passes without deviation and it strikes the surface AB at angle 45°, greater than the critical angle for glass. Total internal reflection takes place and the ray further strikes the surface BC at an angle 45°. The ray undergoes another total internal reflection and then it emerges out through AC, because the ray strikes the surface normally and the ray of light is rotated through 180°.

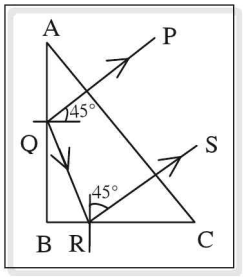


Figure 4.27

Erecting prism or non deviation prism

Rays after refraction from AB strikes the surface AC at an angle 45°, which is greater than the critical angle for glass and it undergoes total internal reflection. At surface BC, the rays undergo refraction and emerges out from the surface BC and the image MN appears as inverted.

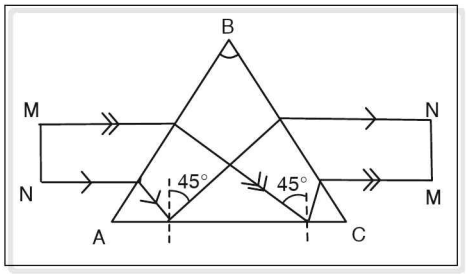


Figure 4.28

Refracting periscope or prismatic periscope

Refracting periscope consists of two right angled isosceles prisms. The light rays entering the periscope are rotated through 180° by the totally reflecting prisms which make the image of a distant object visibled.

This periscope is similar to that of reflecting periscope, but the reflecting periscope makes use of plane mirrors instead of reflecting prisms.

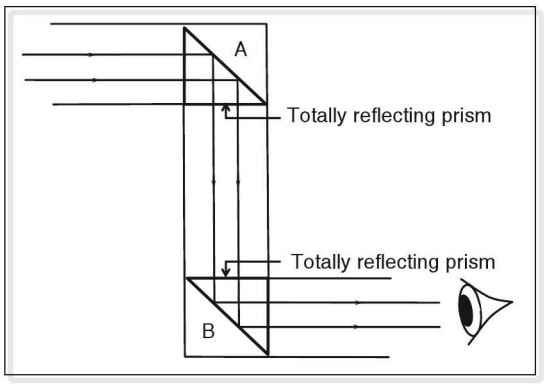


Figure 4.29 Prismatic periscope

Advantages of refracting periscope over reflecting periscope

1. Final image formed by the refracting periscope is bright and sharp, since no rays of light are absorbed.
2. The quality of the image is affected by the presence of dust and moisture in reflecting periscope. Moisture and dust does not have any appreciable effect on the image quality in refracting periscope.

Optical fibres or light pipes

Optical fibres are devices used to transmit light in any desired path. It works on the principle of total internal reflection.

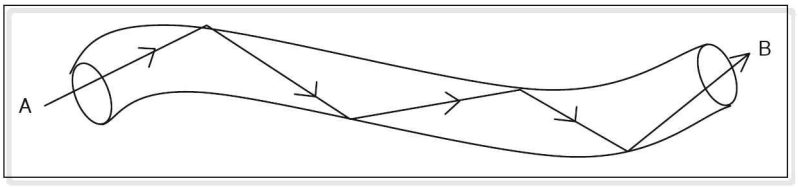


Figure 4.30

It consists of a thin fibre glass. When a ray of light enters a fibre, it suffers a series of total internal reflections and finally emerges out from the other end. Since the light energy is not absorbed, intensity of light is not lost.

Uses of optical fibres

1. **Used in communication:** Optical fibres are used in long distance telephony. Sound energy is converted into electrical energy. Electrical energy is converted into light energy. This light energy is fed to optical fibres and it is transmitted over a long distance through optical fibres. At the receiving end light energy is finally converted to sound energy.
2. Optical fibres are used as endoscopes or fibrescopes. Endoscopes is a device used to examine internal organs of a human body by physicians.

☛ *Example*

Refractive index of a glass is 1.5. Find its critical angle.

Solution

From the relation between critical angle and refractive index.

$$\mu_{glass} = \frac{1}{\sin c}$$

$$1.5 = \frac{1}{\sin c}$$

$$\text{Sin } c = \frac{1}{1.5} = 0.6666$$

$c = \sin^{-1} 0{\cdot}666 \sim 42°$.

Critical angle for glass is 42°.

Refraction through prisms

Prism is a transparent medium. It is bounded by two triangular surfaces and three rectangular surfaces.

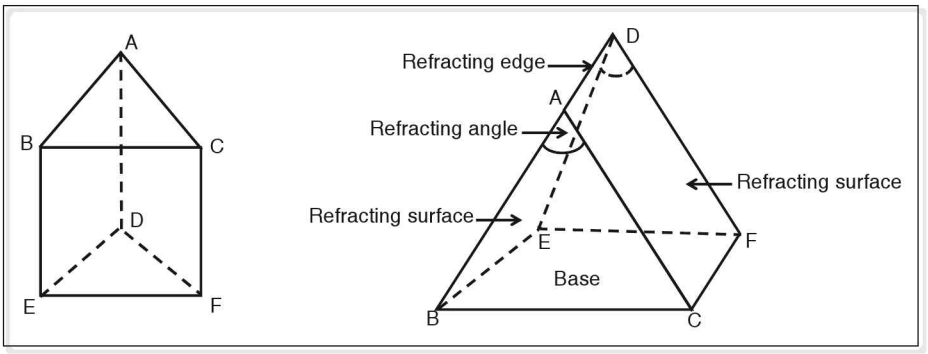


Figure 4.31 Prism

Rectangular surfaces are called refracting surfaces. The angle between the refracting surfaces is called the refracting angle or the angle of the prism. The line joining the refracting surfaces is called the refracting edge. Section of a prism perpendicular to the refracting edge is called the principal section.

Deviation of light by prism

ABC is the principal section of an equilateral prism. ∠A is the angle of the prism. PQ is the incident ray, incident at an angle i_1. Since the ray PQ is travelling from optically rarer medium to denser medium, the refracted ray QR bends towards the normal. r_1 is the angle of refraction on the face AB. The ray QR is incident at an angle r_2 on the surface AC of the prism and emerges out along RS at an angle i_2. i_2 is called the angle of emergence.

The angle between incident ray produced (PQ) and the emergent ray (RS) is called the angle of deviation. Angle of deviation = ∠STZ = d

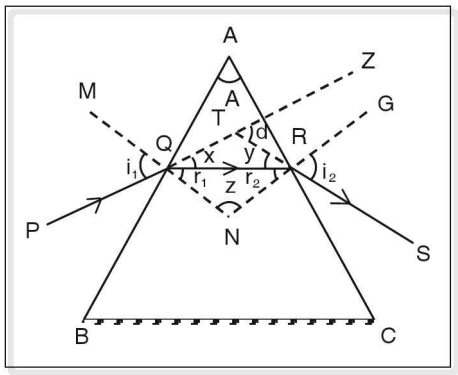


Figure 4.32

The angle of deviation depends on the following factors.

1. Angle of the prism
2. Angle of incidence
3. Refractive index of material
4. Colour of light

1. Dependence of angle of deviation on angle of incidence.



Experiment I

1. Place an equilateral prism on a white paper fixed to a drawing board. Trace the principal section ABC of the prism and remove the prism.
2. Draw normals MN and NK to the refracting surfaces AB and AC respectively.
3. Make a suitable angle (i) with the normal. This line is the incident ray (PQ).
4. Fix two pins P_1 and P_2 on the line of incidence as shown and bring the prism back to its position.
5. Look for the images of the pins P_1 and P_2 through the refracting surface AC. Fix two more pins P_1^1 and P_2^1 such that the images of the pins P_1 and P_2 appear in the same straight line.

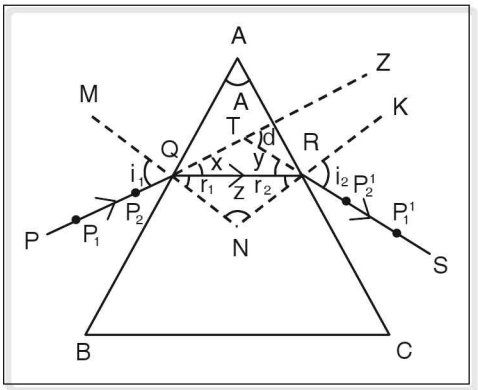


Figure 4.33

6. Encircle the points of the pins. Remove the prism and the pins.
7. Join the points P_2^1 and P_1^1 and produce it to meet the extended line PQ at T.
8. Join QR, QR is the refracted ray. RS is the emergent ray.

   $\angle SRK = i_2$ represents the angle of emergence.

   $\angle r_1$ is the angle of refraction on the surface AB.

   $\angle r_2$ represents the angle of incidence on the surface AC of the prism.
9. Produce the incident ray (PQ) and the emergent ray (RS) backward. The rays PQ and RS intersect at a point T.

   $\angle STZ$ is the angle of deviation = d

   It is found that angle of deviation, $d = \angle i_1 + \angle i_2 - \angle A$ ---- (1)

   Where $\angle A = \angle r_1 + \angle r_2$
10. Measure $\angle i_1, \angle i_2, \angle r_1$ and $\angle r_2$ and then find angle of deviation by using equation (1).

It is found that the angle of deviation decreases with increase in the angle of incidence upto a particular angle. This particular angle of incidence where the angle of deviation is minimum is called the minimum deviation position of the prism. Further as the angle of incidence is increased, the angle of deviation also increases. It is found that at the minimum deviation position,

(a) angle of incidence is equal to angle of emergence i.e., $\angle i_1 = \angle i_2$

(b) angle of refraction r_1 is equal to angle of refraction r_2. i.e. $\angle r_1 = \angle r_2$. Then $d = 2\angle i - \angle A$ and $\angle A = 2r$.

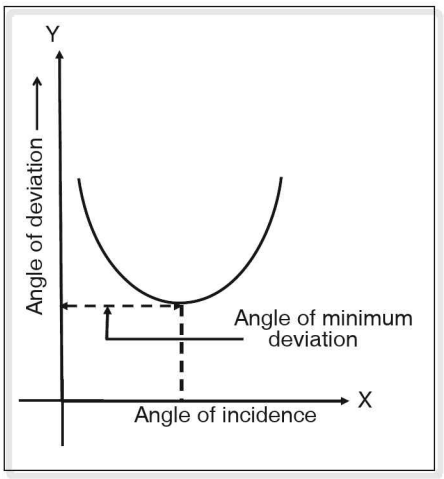


Figure 4.34

2. Dependence of angle of deviation on refractive index.

If the hollow glass prism is filled with transparent liquids like water, benzene etc. and the experiment is conducted as explained above, it is found that as the refractive index of the material increases the angle of deviation also increases.

3. Dependence of angle of deviation on the colour or the wave length of light.

The experiment is repeated as explained above, but the pins are first painted red and the angle of deviation is found out. In the next trial the pins are painted violet and the angle of deviation is found out for the same angle of incidence as that of the red pins.

It is found that angle of deviation for violet is greater than the angle of deviation of the red colour.

$\therefore$ Lesser the wavelength, greater is the angle of deviation.

4. Dependence on angle of the prism.

Repeat the experiment by taking different angles of the prism. (90°, 60°, 30° prism). It is found that the angle of deviation decreases with increase in the angle of the prism.

Refraction through lenses

Lens

A lens is a transparent material which is bounded by two spherical surfaces or one spherical surface and one plane surface.

Lenses are of two types namely,

1. convex lens or converging lens.
2. concave lens or diverging lens.

Convex lens

Convex lens has one or two spherical surfaces, such that it is thicker at the centre than at the edges. Convex lenses are of three types, namely,

1. Double convex
2. Plano convex
3. Concavo–convex lens

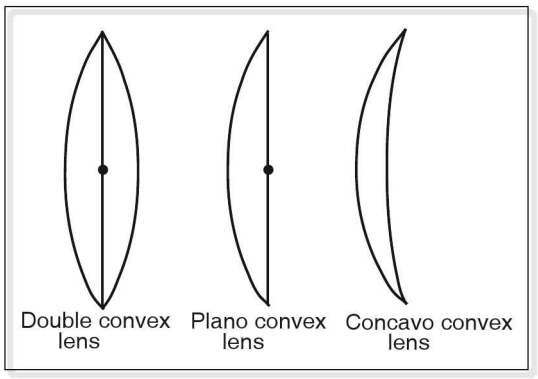


Figure 4.35

Concave lens

Concave lens has one or two spherical surfaces, such that it is thinner at the centre than at its edge.

Concave lens are also of three types, namely,

1. double concave lens
2. plano concave lens
3. convexo–concave lens

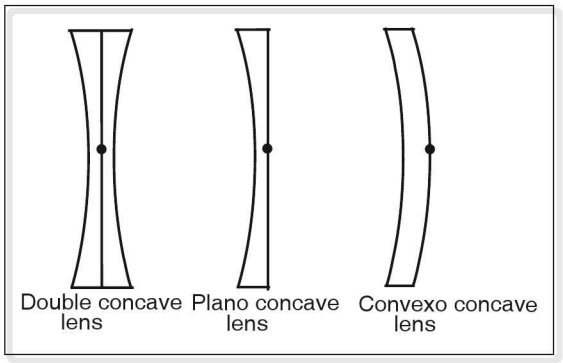


Figure 4.36

Convergent action of convex lens

A lens can be considered to be made up of prisms. The bases of the prisms are downwards in the upper half and prisms in the lower half have their bases upwards. Prisms at the edges have greater refracting angle.

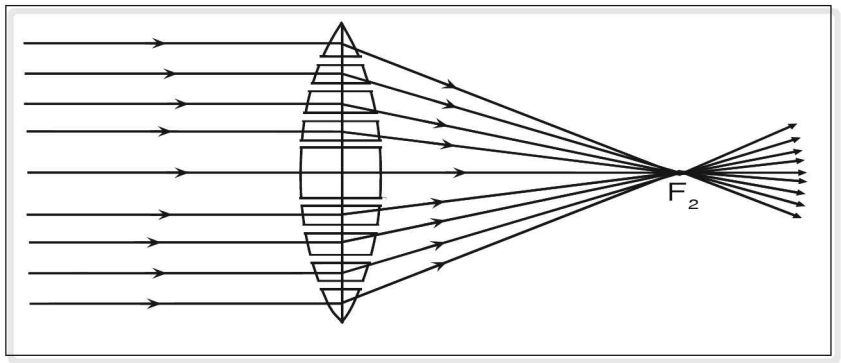


Figure 4.37

When a ray of light is passed through a prism, the emergent ray always bends towards the base (broader side) of the prism and greater the refracting angle, lesser will be the deviation. Thus the rays from the outer edges are deviated more towards the base line and the central line passes without deviation. All these rays converge at a point and this point is called the focus.

When a convex lens is held facing the sun and the screen is placed at the other side of the lens, by adjusting the distance between the screen and the lens, a sharp and bright spot is obtained on the screen. Thus a convex lens is a converging lens.

Diverging action of concave lens

Concave lens too can be considered to be made up of prisms. The prisms in the upper half have their bases upwards or their refracting angles downwards and the prisms in the lower half have their refracting angles upwards. The rays near the outer edges of the lens are deviated maximum towards the bases of prism and the central ray goes without deviation. These rays do not meet but when these rays are produced backwards, they meet at a point. This point is called focus.

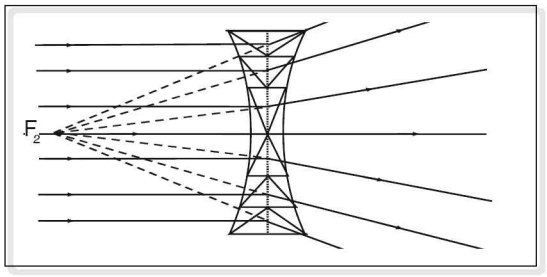


Figure 4.38

When smoke is produced around the concave lens, held in sunlight, the emergent rays are seen to move away from each other. Hence a concave lens is a diverging lens.

☛ *Some definitions*

Each surface of a lens is a part of a sphere. Surface 1 is a part of a sphere S_1 and surface 2 is a part of a sphere S_2.

1. **Centre of curvature of lens (C)**
   It is centres of the spheres of which the surfaces of a lens form a part.
2. **Radii of curvature (R)**
   It is the radius of the spheres from which the spherical surface was obtained. Since a lens generally has two curved surfaces, it has two radii of curvature.
3. **Principal axis**
   It is a straight line joining the centre of curvatures of the two spherical surfaces of the lens.

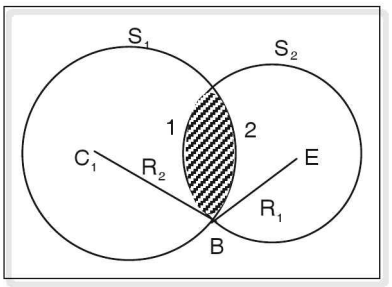


Figure 4.39

Optical centre

In case of thin lens it is a point on the principal axis such that a ray of light passing through it emerges without deviation or it is a point, where any ray of light directed towards it will have its emergent ray parallel to the direction of incidence. Optical centre is denoted by 'O'

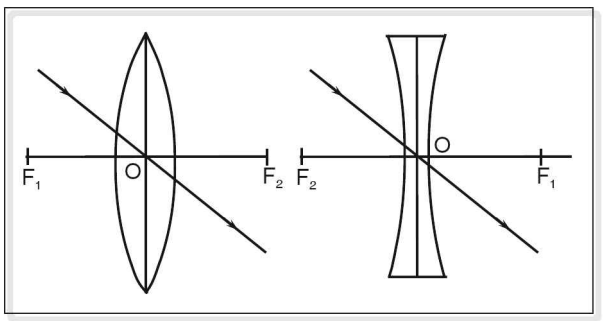


Figure 4.40

Principal focus

It is the point on the principal axis to which the incident rays close and parallel to the principal axis of the lens converge after refraction, in case of a convex lens or appear to diverge from in case of a concave lens.

Since the light ray can be incident in either direction of the lens, there are two principal foci, which are at equal distance from the optical centre, provided the medium on either side of lens is the same.

These points are referred to as first focal point and second focal point.

Focal length

Distance between optical centre and the principal focus is called focal length and is denoted by f.

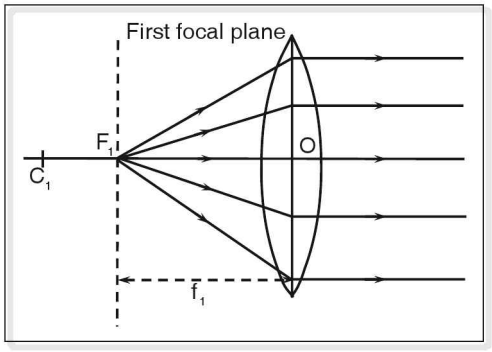


Figure 4.41

Focal plane

It is a vertical plane passing through the focal point.

Rules for construction of images formed by thin lenses

To find out the images formed by the lens by considering convenient rays.

1. The rays parallel and close to the principal axis after refraction pass through the principal focus in a convex lens or appear to diverge from the principal focus in case of a concave lens.
2. Any ray of light which is directed through the principal focus or appearing to meet at it (concave lens) after refraction always emerges parallel to the principal axis.
3. Any ray of light which passes through the optical centre of a lens emerges without deviation.

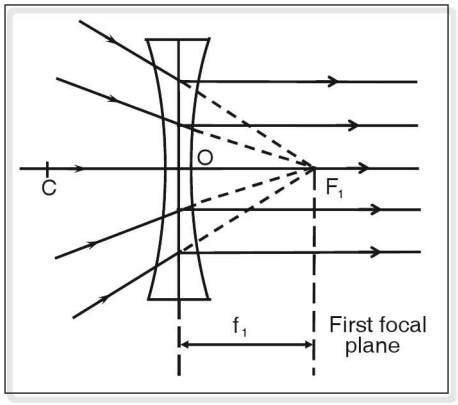
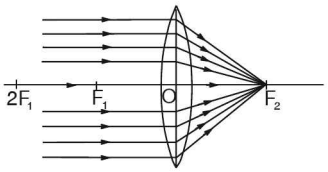
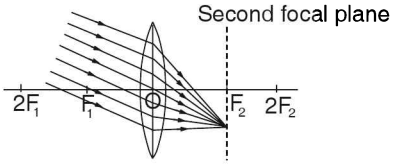
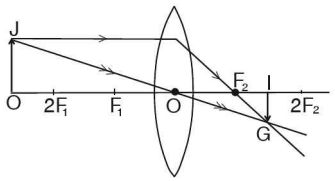
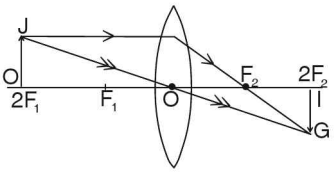
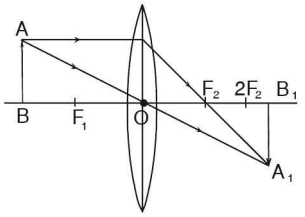
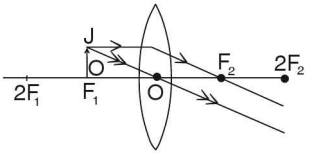
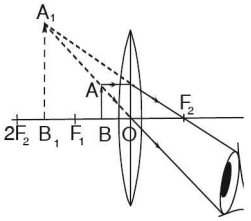


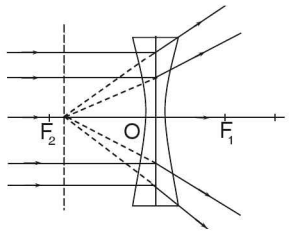
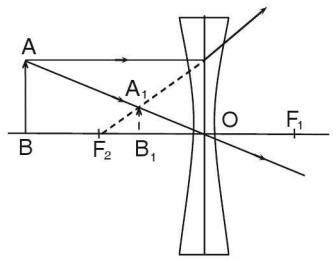
Figure 4.42

The above stated rules are used to obtain the position of the image. The point where the rays intersect gives the image position.

Formation of images by convex lens

1		Infinity, rays are parallel to the principal axis	Formed at principal focus on the other side	Real, inverted and diminished	Burning glass or magnifying glass
2		Infinity, rays are not parallel to the principal axis	At the focal plane on the opposite side of the lens as that of object	Real, inverted highly diminished	Eye lens of astronomical telescope
3		Beyond 2F	On the opposite side, between F and 2F	Real, inverted diminished	Photographic camera
4.		At 2F	At 2F on the opposite side	Real, inverted, same size	Terrestrial telescope
5		Between 2F and F	Beyond 2F on the opposite side	Real, inverted and enlarged	Cine projection
6		At F	At infinity	Real, inverted and enlarged	Search lights and spot lights
7		Between F and the optical centre	On the same side	Virtual, erect and enlarged	Simple microscope, reading finger prints, to observe biological specimen and weaving patterns

Formation of images by concave lens

1		Infinity	At F on the same side	Virtual, erect and diminished to a point	Eye lens in Galilean telescope
2		At any distance	On the same side between optical centre and F	Virtual, erect and diminished	Correcting short sight

Cartesian sign convention

Depending on the distance of the object from the lens, the image formed on the lens is on the same side or on the opposite side. Therefore it is necessary to use sign convention. The sign convention rules are as follows.

1. All distances are measured from optical centre.
2. Distance measured in the direction of incident light is positive and distance measured against incident light is negative.
3. Distances measured above the principal axis are positive and distances measured below the principal axis are negative.
4. According to sign convention, focal length of a convex lens is positive and that of the concave lens is negative.

Lens formula

AB is an extended object placed at a distance 'u' from the optical centre O. A^1B^1 is the image at a distance 'v' from O.

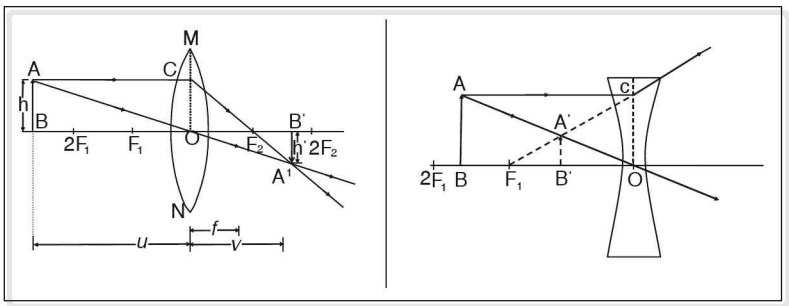


Figure 4.43

Using Cartesian sign
Object distance OB = −u
Image distance OB^1 = v
Focal length of a convex lens = f
From similar triangles AOB and A^1OB^1

$$\frac{A^1B^1}{AB} = \frac{OB^1}{OB} = \frac{v}{-u} \text{ --- (1)}$$

from similar triangles OCF_2 and $A^1B^1F_2$

$$\frac{A^1B^1}{OC} = \frac{B^1F_1}{OF_1} \quad \text{--- (2)}$$

$$\frac{B^1F_1}{OF_1} = \frac{OF_1 - OB^1}{OF_1}$$

$$\frac{B^1F_1}{OF_1} = \frac{OF_1 - OB^1}{OF_1} = \frac{v-f}{f}$$

∴ (2) can be written as

$$\frac{A^1B^1}{OC} = \frac{v-f}{f} \text{ but OC = AB}$$

$$\frac{A^1B^1}{OC} = \frac{v-f}{f} \text{ ----- (3)}$$

Comparing (1) and (3)

$$\frac{v}{-u} = \frac{v-f}{-f}$$

vf = −uv + uf --- (4)
dividing (4) by uvf

$$\frac{1}{u} = \frac{-1}{f} + \frac{1}{v}$$

or

$$\frac{1}{f} = \frac{1}{v} - \frac{1}{u}$$

Using Cartesian sign
Object distance = −u
Image distance = −v
Focal length of concave lens = −f
From similar triangles A^1B^1O and ABO

$$\frac{A^1B^1}{AB} = \frac{OB^1}{OB} = \frac{-v}{-u} = \frac{v}{u} \text{ --- (1)}$$

From similar triangles OCF_1 and $A^1B^1F_1$

$$\frac{A^1B^1}{OC} = \frac{B^1F_1}{OF_1} \quad \text{--- (2)}$$

$$\frac{B^1F_1}{OF_1} = \frac{OF_1 - OB^1}{OF_1} = \frac{-f-(-v)}{-f}$$

$$\therefore \frac{B^1F_1}{OF_1} = \frac{-f+v}{-f}$$

∴ eqn (2) can be written as

$$\frac{A^1B^1}{OC} = \frac{-f+v}{-f} \text{ but OC = AB}$$

$$\frac{A^1B^1}{OC} = \frac{-f+v}{-f} \text{ ------- (3)}$$

comparing (1) and (3)

$$\frac{v}{u} = \frac{-f+v}{-f}$$

−vf = −uf + uv --- (4)
dividing (4) by uvf

$$\frac{-1}{u} = \frac{-1}{v} + \frac{1}{f}$$

or

$$\frac{1}{f} = \frac{1}{v} - \frac{1}{u}$$

$$\therefore \frac{1}{f} = \frac{1}{v} - \frac{1}{u} \quad \text{--- (5)}$$

Equation (5) is called the lens formula. This equation is general and is valid in all situations.

Magnification

It is the ratio of size of the image to the size of the object. Magnification is denoted by 'm'.

$$m = \frac{\text{height of the image } (A^1B^1)}{\text{height of the object } (AB)}$$

$m = \frac{A^1B^1}{AB}$ From equation (1) of the previous page.

$$\frac{A^1B^1}{AB} = \frac{v}{u}$$

$$\therefore\ m = \frac{v}{u}$$

Thus magnification $m = \frac{h_i}{h_o} = \frac{v}{u}$

h_i is the height of the image and h_o is the height of the object.

Magnification is positive for a convex lens forming a virtual image and magnification is negative, if the image formed by the convex lens is real.
Image formed by concave lens is always virtual, hence magnification for concave lens is positive.

Power of the lens

Nature and size of the image depends on the focal length of the lens. A lens of short focal length bends the light to a large extent. Thus the image formed by lens of short focal length will be closer to the optical centre.

Power of the lens is the measure of degree of convergence or divergence of the lens.

Power of the lens is mathematically defined as reciprocal of focal length and power is denoted by P.

$$P = \frac{1}{f}$$

Unit of power is dioptres, denoted by 'D', when f is in metre.

$$\therefore\ 1\ D = \frac{1}{1\ m}$$

Thus 1 dioptre is the convergence or divergence produced by a lens of 1metre focal length.

Combination of thin lenses

If two or more thin lenses are kept in contact, then their equivalent focal length

$$\frac{1}{f} = \frac{1}{f_1} + \frac{1}{f_2} + \frac{1}{f_3} \ldots\ldots$$

where f_1, f_2, f_3 ..... are focal lengths of the lenses then the power of the combination of the lenses.

$P = P_1 + P_2 + P_3$.....

Where $P_1 = \frac{1}{f_1}$; $P_2 = \frac{1}{f_2}$.....

☛ *Example*

An object 5 cm high is placed at a distance of 10 cm from a convex lens of focal length 15 cm. Find the nature, position, size and magnification of the image.

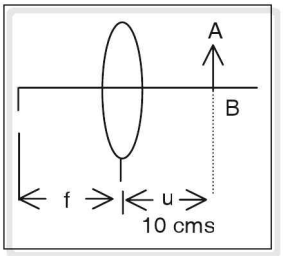


Figure 4.44

Solution

Using Cartesian sign convention

Object distance = −10 cm

Focal length of a convex lens = 15 cm

Height of the object = h_0 = 5 cm

Using lens formula

$$\frac{1}{f}=\frac{1}{v}-\frac{1}{u}$$

$$\frac{1}{15}=\frac{1}{v}-\frac{1}{-10}$$

$$\frac{1}{15}=\frac{1}{v}+\frac{1}{10}\Rightarrow\frac{1}{v}=\frac{10-15}{15\times10}$$

$$v=-\frac{150}{5}=-30\text{ cm}$$

negative sign indicates it is a virtual image

$$m=\frac{v}{u}$$

$$m=\frac{-30}{-10}=3$$

positive sign indicates the image is erect.

$$m=\frac{h_i}{ho}\Rightarrow 3=\frac{h_i}{5}$$

∴height of image = 5 × 3 = 15 cm.

☛ *Example*

A convex lens throws on a screen 12 m from the lens a magnified image of the object. If the magnification is 15, find the focal length of the lens.

Solution

Magnification m = 15
Image distance = v

$$\text{Magnification} = +\frac{v}{u} \qquad \text{--- (1)}$$

From lens formula

$$\frac{1}{f} = \frac{1}{v} - \frac{1}{u}$$

using Cartesian sign

$$\frac{1}{f} = \frac{1}{v} - \frac{1}{-u}$$

$$\frac{1}{f} = \frac{1}{v} + \frac{1}{u}$$

$$\therefore \frac{1}{u} = \frac{1}{f} - \frac{1}{v}$$

$$\frac{1}{u} = \frac{v-f}{vf}$$

$$u = \frac{vf}{v-f} \quad \text{--- (2)}$$

substituting (2) in (1)

$$m = \frac{v(v-f)}{vf}$$

$$m = \frac{v(v-f)}{vf}$$

$$\therefore 15 = \frac{12-f}{f}$$

$$15f = 12 - f$$

$$16f = 12$$

$$f = \frac{12}{16} = 0.75 \text{ m.}$$

☛ *Example*

Two convex lenses A and B of focal lengths 0.3 m and 0.2 m respectively are placed in contact. Calculate the effective focal length and the power of the combination. If B is a concave lens what is the effective focal length and the power?

Solution

Reciprocal of effective focal length of the combination of the lens

$$\frac{1}{f} = \frac{1}{f_1} + \frac{1}{f_2}$$

$$\frac{1}{f} = \frac{1}{0.3} + \frac{1}{0.2}$$

$$\frac{1}{f} = \frac{0.2 + 0.3}{0.06}$$

$$f = \frac{0.06}{0.5} = 0.12 \text{ m}$$

power of the combination

$$P = \frac{1}{f} = \frac{0.05}{0.06} = 8.33 \text{ D}$$

Thus the combination behaves like a convex lens of focal length 0·12 m

If lens B is a concave then the effective focal length is

$$\frac{1}{f} = \frac{1}{f_1} - \frac{1}{f_2}$$ (By Cartesian sign)

$$\frac{1}{f} = \frac{1}{0.3} - \frac{1}{0.2}$$

$$\frac{1}{f} = \frac{0.2 - 0.3}{0.06}$$

$$\frac{1}{f} = \frac{-0.1}{0.06}$$

$$f = \frac{-0.06}{0.1} = -0.6 \text{ m}$$

Power of the combination

$$P = \frac{1}{f} = \frac{1}{-0.6} = -1.67 \text{ D}$$

The combination behaves like an concave lens of focal length 0.6 m.

☛ *Example*

A convex lens of focal length 0.12 m produces an image which is thrice the size of the object. Find the distance between the object and the lens.

Solution

Focal length of the lens = 0.12 m

Magnification $m = \frac{h_i}{h_o} = \frac{v}{u}$

$h_i = 3h_o$

$\therefore m = 3$

$$m = \frac{v}{u}$$

$$3 = \frac{v}{u}$$

$v = 3u$

from lens formula

$$\frac{1}{f} = \frac{1}{v} - \frac{1}{u}$$

object distance for convex lens = −u (using Cartesian sign)

$$\frac{1}{0.12} = \frac{1}{3u} - \frac{1}{-u}$$

$$\frac{1}{0.12} = \frac{1}{3u} + \frac{1}{u}$$

$$\frac{1}{0.12} = \frac{4}{3u}$$

$$u = \frac{4 \times 0.12}{3} = 0.16 \text{ m}$$

To obtain real image u = 0·16 m
To obtain virtual image v is −ve

$\therefore v = -3u$

$$\frac{1}{0.12} = \frac{-1}{3u} + \frac{1}{u}$$

$$\frac{1}{0.12} = \frac{-u + 3u}{3u^2}$$

$$\frac{1}{0.12} = \frac{2u}{3u^2}$$

$$u = \frac{2 \times 0.12}{3} = 0.08 \text{ m}$$

To obtain virtual image, object distance = 0.08 m.

To determine focal length of a convex lens

1. By distant method

1. Mount the lens on a lens stand.
2. Direct the lens towards a distant object such as a tree etc.
3. Place the screen behind the lens.
4. Adjust the distance between the screen and the lens until a clear and sharp image of the object is obtained.
5. Measure the distance between the screen and the lens. This distance gives the approximate focal length.

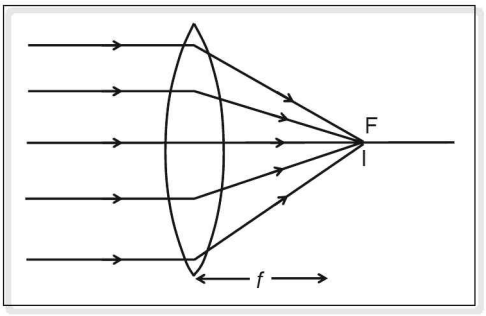


Figure 4.45 f-by Distant Object Method

2. By plane mirror method

1. Mount the convex lens on the lens stand.
2. Place a needle in front of the lens and a plane mirror on the other side of the lens, so that the plane of the mirror is vertical and perpendicular to the axis of the lens.
3. Adjust the distance of the lens or the object (needle) until a real and inverted image is formed just above the object.
4. If the tip of the image in the plane mirror coincides with the tip of the object, then the distance between the needle and the lens gives the focal length.

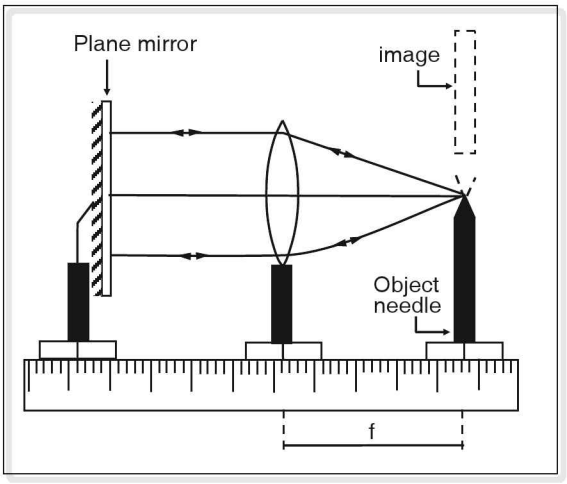


Figure 4.46

3. Focal length of u – v method

1. Mount a convex lens on a lens stand
2. Place an illuminated object in front of it and a screen behind the lens.
3. Adjust the distance between the screen and the lens until a clear and sharp image is obtained on the screen.
4. Measure the distance between object and the lens. This gives object distance (u).
5. Distance between the screen and the lens gives the image distance v.
6. Focal length is calculated using the formula $f = uv/u + v$

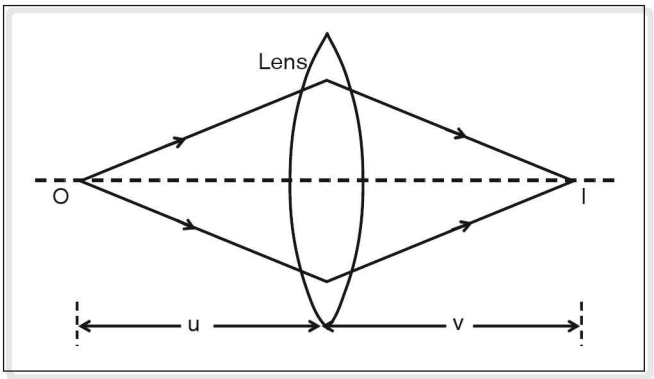


Figure 4.47 u-v method

Application of lens

1. Lenses are used in optical instruments like telescope, microscope, camera and sepctrometer.
2. Lens are used to rectify eye sight of a person having defective eye sight.

Dispersion of light

Early discovery of Newton

When Newton was observing stars using astronomical telescope, he found that the stars were coloured near the fringes. He concluded that the colours observed were due to the nature of light but not due to the defect in lens. To prove this, he placed an equilateral prism in a path of narrow beam of light entering through a small aperture of a window in a dark room. Emergent light from the prism was received in the screen. He found that emergent rays were spread out in bands of seven colours with red at one end merging gradually with orange, yellow, green, blue, indigo and violet.

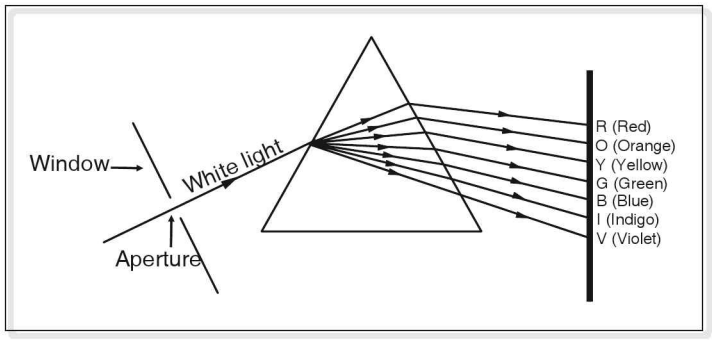


Figure 4.48 Dispersion by a prism

Thus the phenomenon of splitting of white light into its constituent colours is known as dispersion. The medium which brings about dispersion is called dispersive medium.

For example, prism is a dispersive medium.Vacuum or air is a non dispersive medium.

The band of colours (wavelength) obtained is known as spectrum. Spectra are classified as pure spectrum and impure spectrum.

Pure spectrum

It is the spectrum in which the constituent colours occupy their respective position (VIBGYOR) and are distinct. For example, spectrum obtained from prism and spectroscope arrangement.

Impure spectrum

It is spectrum in which the bands of colours overlap. There are no well defined boundaries between the bands of colours.

☛ *Example* rainbow

Cause for dispersion

Dispersion occurs because different colours (wavelengths) are refracted or deviated to different extents because speed of different colours of light is different in a given medium.

$$\mu = \frac{\text{velocity of light in air or vacuum}}{\text{velocity of light in a given medium}}$$

The refractive index of violet colour in glass is greater than the refractive index of red, or in other words, speed of violet is least in glass and the speed of red is the highest.

Refrangibility

It is the ability of light to get refracted when passing from one medium to another. It depends on the refractive index of the medium for a given wavelength of light. In the visible light the refrangibility for violet is maximum and that of red is least. ($\delta_{violet} > \delta_{red}$).



Experiment II

Experiment to show how white light splits into its constituent colours and it is not the prism which colours white light.

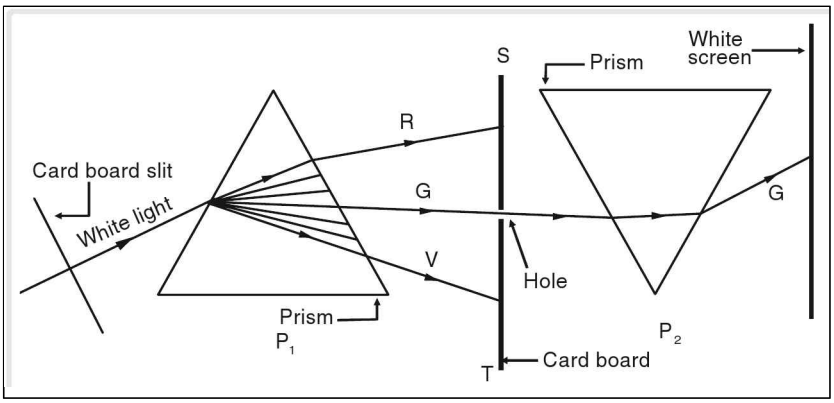


Figure 4.49

White light from a slit is made to pass through an equilateral prism placed in a dark room. Dispersed light from this prism is obtained on the white screen having a small hole. The coloured light passing through this hole is made to fall on a second prism P_2, which is placed in an inverted position. The light passing through the second prism gets refracted but does not undergo dispersion.



Experiment III

Newton's experiment on recombination of white light

Newton took two identical prisms P_1 and P_2. P_2 was placed inverted with respect to P_1. White light emerging out of the slit was made to fall on P_1. The spectrum produced by P_1 was interrupted by the prism P_2. He found that white light emerged out of P_2. Thus the spectrum produced by P_1 recombined in the prism P_2.

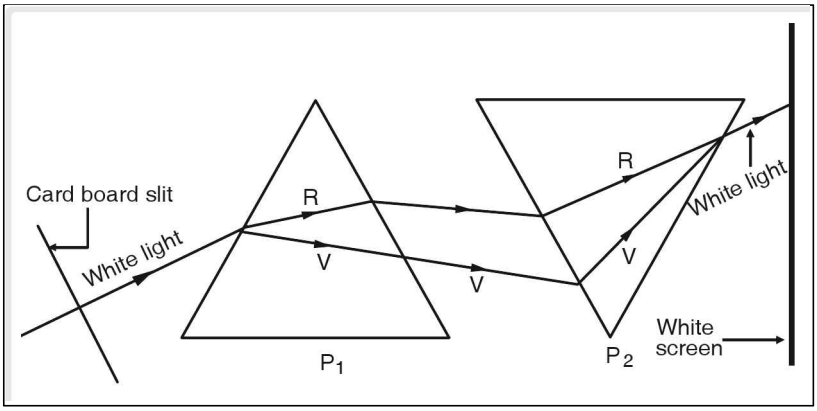


Figure 4.50 Recombination of dispensed light

Second prism P_2 bends the dispersed colours, towards its base through different angles such that the rays emerging out from the prism P_2 are incident at the same point. Thus P_1 is a dispersing prism and P_2 is a recombined prism.

Colour of the objects and pigments

From the above facts, it is seen that white light consists of seven colours.

Colours It is the sensation produced in the brain, when a certain wavelength in the visible region strikes the retina.

Pigments It is an optically active substance which absorbs most of the wavelengths in the visible regions and reflects selective wavelengths. For example, leaf appears green in colour due to the presence of chlorophyll pigment. Pigments are colourless and get activated when white light is incident on them.

When white light is incident on the object it absorbs some of the colours and reflects certain colours. The colour of an object is determined by the wavelength it reflects.

Sunflowers appear yellow because of the pigment present in them which absorbs all colours except yellow. An object appears white if it does not absorb any colour and it appears black if it absorbs all the colours. A colour filter is a transparent body which allows certain wavelengths and absorbs all other wavelengths.

Primary colours of light

Red, blue and green are called primary colours. Other colours are obtained by mixing primary colours in proper proportions. If the primary colours are mixed in proper proportion white light is obtained. Colours obtained by mixing two primary colours are called composite colours or secondary colours.

☛ *Example*

Red + Blue = Magenta
Red + Green = Yellow
Blue + Green = Cyan

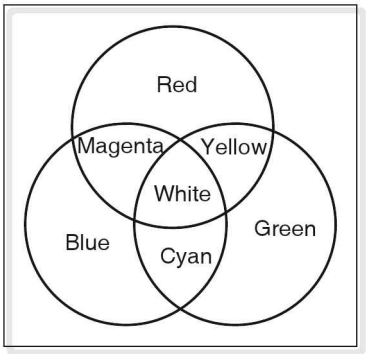


Figure 4.51

Complimentary pair of colours

A pair of colours, in which one is a primary colour and the other is a secondary colour, which on mixing in proper proportion gives white light.

☛ *Example*

Yellow + Blue = White

Colour on the vertex of the triangle represents a primary colour. Colours at the mid points represent composite colours. Colours at the line joining the vertex and midpoint of the side opposite to it represent complimentary colours.

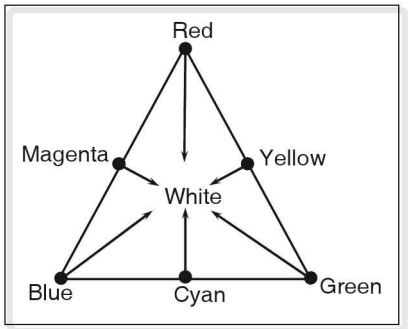


Figure 4.52 Colour triangle

Raman Effect

Tyndall and Rayleigh discovered that when a beam of light is passed through a transparent and homogeneous medium, a part of the incident light is deflected sideways and scattered. The frequency or wavelength of the scattered beam was found to be same as the incident beam.

Lord Rayleigh explained the blue colour of the sky. He said that the blue colour of the sky is due to the scattering of light by molecules present in the atmosphere. The intensity of the scattered light in a medium is inversely proportional to the fourth power of the wavelength.

$$I_{max} \propto \frac{1}{\lambda^4}$$

I_{max} is the maximum intensity, λ is the wavelength of light. Thus the wavelength of blue light is minimum and it is scattered most. Rayleigh scattering is also called coherent scattering. Coherent scattering is a scattering in which there is no change in the wavelength of light, i.e., scattered light has the same wavelength as that of the incident light.

At sunrise and sunset, sun appears red in colour. This is due to Rayleigh scattering, At sunset and sunrise, light from the sun travels maximum distance. As a result most of the blue light is scattered and the light reaching the earth's surface lacks blue colour. It is red colour which reaches the earth's surface, since it is least scattered.

When Sir C.V. Raman was experimenting the scattering of light by liquids, he discovered that when a monochromatic light is passed through organic liquid like benzene, toluene etc, it results in the formation of the scattered light which consists of wavelengths of frequencies higher and lower that of the incident light in addition to that of the incident light. This is known as Raman effect. Raman's scattering is an incoherent scattering. Incoherent scattering is a type of scattering in which the scattered light consists of frequencies other than those of the incident light. Raman effect is used to study the structure of molecules.

Spectroscope

Spectroscope is an instrument used to obtain pure spectrum. Spectroscope consists of collimator, prism table and a telescope.

Collimator consists of a slit from which light rays are rendered parallel by the lens. This light is made to fall on a prism placed on the prism table which is in line with the collimator. The rays get dispersed by the prism and the spectrum is observed through a telescope.

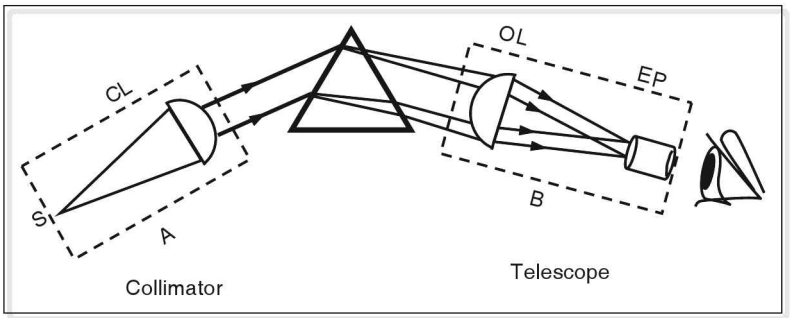


Figure 4.53 Spectroscope

Uses of a spectroscope

Spectrometer helps us to observe spectra. By studying and analysing spectra, information about the sources of light can be obtained. Special spectrometer is required to study Raman lines.

Emission spectra

Light from candle or an incandescent bulb consists of all wavelengths present in the visible region. This spectra is known as continuous emission spectra.

When gases and vapours are made to emit light, the spectrum consists of a number of bright lines against a dark back ground. This spectra is known as line emission spectra.

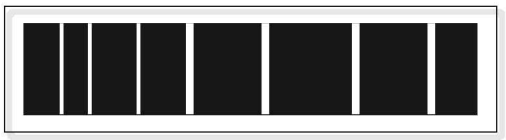


Figure 4.54

Line emission spectra is a characteristic of an element. i.e different elements give rise to different line spectra. By studying line spectra, elements present in the gas or molecules can be discovered.

By measuring intensities of the line spectra, an approximate amount of each element present can be found out. This technique is called spectro-chemical analysis. This is used in medicine and crime detection.

Absorption spectra

When white light is passed through gases or vapour at a low temperature some of the wavelengths are absorbed. This is known as absorption spectrum.

☛ *Example*

When a light from carbon arc is passed through sodium vapour, two lines are blotted out in the yellow region (D_1 and D_2) of the spectrum.

When a solar spectrum is observed by a powerful telescope, the spectrum consists of a number of dark lines. These lines are called Fraunhofer lines, named after Fraunhofer who first studied these lines.

By analysing Fraunhofer lines, elements present in the sun can be found out. Helium was discovered by studying Fraunhofer lines.

Electromagnetic radiations

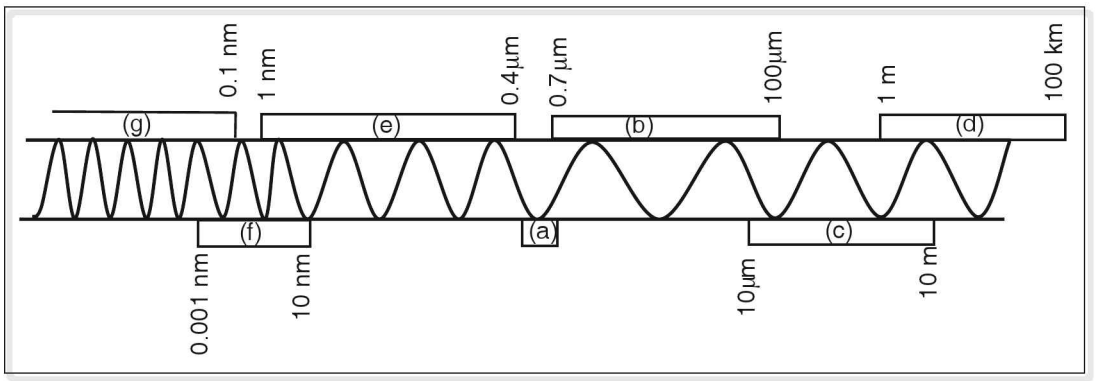


Figure 4.55 Various regions of electromagnetic spectrum and their wavelength ranges

(a) Visible spectrum (b) IR spectrum (c) Microwaves (Radio waves)
(e) UV spectrum (f) X-rays (g) γ-rays

Radiations, IR, visible, UV, radio waves, X-rays, gamma rays and micro waves constitute electro magnetic spectrum. These radiations can be described in terms of oscillating electric and magnetic fields which are at right angles to each other and are perpendicular to the direction of propagation. All radiations travel with the speed of light but differ in their wavelengths or frequencies.

$C = \nu\lambda$

C – where C is the speed of light and λ is the wavelength, ν is the frequency

Differences in the properties of different electromagentic radiations are due to the difference in their frequencies.

Visible spectrum

The wavelength of visible light ranges from 00 nm to 750 nm (4000 A° to 7500 A°). The spectrum within this range is called visible spectrum (VIBGYOR)

Invisible spectrum

It is the region of spectrum which does not strike the retina.

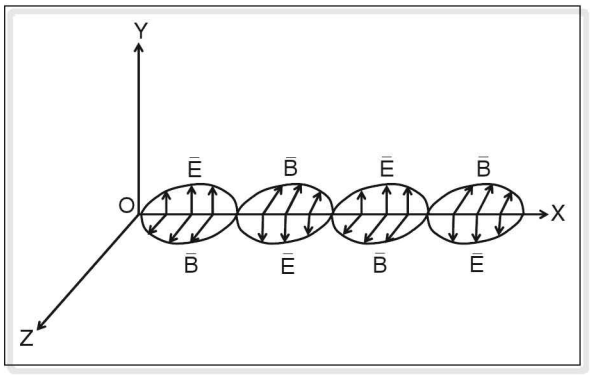


Figure 4.56 Electromagnetic wave

Radio waves

The wavelength of radio waves ranges from 1 m to 100 km. They are produced by electro magnetic oscillators of low frequencies. Radio waves are received from extra-terrestial sources known as radio astronomy. Radio waves of short wavelengths are used in communication which includes satellites, Radars and TV. Longer wavelengths are used in radio broadcasting.

Microwaves

The wavelength of microwave ranges from 10 μm to 10 m. These waves are produced by electromagnetic oscillators of high frequency in an electric circuit. They are used in radio detection and ranging, microwave ovens, telemetry and satellite communications.

Infrared spectrum

It is an electromagnetic radiation beyond the red region. Its wavelength varies from 8000×10^{-10} m to $4,00,000 \times 10^{-10}$ m

Infrared rays were discovered by William Herschel. He obtained a spectrum by passing sunlight through a glass prism. He observed a rise in temperature when he moved a thermometer with a blackened bulb from violet region to red region and there was a sudden increase in temperature beyond the red region. Infrared rays produce heating effect but not the visible effect.

Sources of infrared rays

Sun, lamp, burning gases. Any substance which radiates heat is a source of infrared rays. All living beings emit IR rays.

Properties of IR rays

1. They are electromagnetic radiations which are not affected by electric and magnetic field.
2. They travel with the speed of light.
3. They do not affect photographic plates.
4. They are least scattered by fog and mist because of their long wavelengths.
5. They can be detected by thermopile or thermometer with blackened bulb.
6. They obey the laws of reflection and refraction. Hence they can be focused by parabolic mirrors or lenses.

7. Infrared rays having wavelength beyond $20{,}000 \times 10^{-10}$ m can be absorbed by glass.
8. They can penetrate rock salt.

Uses

1. Since they produce heating effect, they are used to relieve pains from muscles and swollen joints.
2. Because of their long wavelength they can travel long distances in fog without much scattering and are used in infrared photography.
3. They are used in surface to air missiles.

   A missile is fired in the direction of an enemy plane. Sensor of infrared radiation present in the missile detects the heat trail left by the enemy plane and sends the message to computer. The computer guides the missile towards the target.
4. They are used to detect diseases in crops. Any change in the heat radiation of crops is due to some kind of disease. Heat radiation is measured through satellites.
5. Night vision devices are used to see objects in dark. These devices can detect infrared radiation emitted by hot objects.

Ultraviolet spectrum

The region of spectrum beyond violet region is called ultraviolet spectrum.

This spectrum was discovered by J. Ritter. Ritter discovered that when a silver chloride paper was placed beyond violet region, it first turned to violet, then brown and black. Thus the ray which produced strong chemical effect in silver chloride paper was named as chemical ray or actinic ray. Later on the rays were named as ultraviolet rays.

Glass absorbs UV rays but quartz does not absorb these rays.

Sources

1. Any source of light at temperature greater than 2500°C is a good source of UV rays.
2. Sunlight, mercury vapour lamps.

Properties

1. They are electromagnetic waves having wavelength between 4000 A° to 100 A° (1 A° = 10^{-10} m)
2. They are not affected by electric or magnetic fields.
3. They obey the laws of reflection and refraction.
4. They produce fluorescence when incident on substances like zinc sulphide, quinine sulphate solution.

   Fluorescence is a phenomenon in which a substance absorbs light of shorter wavelength and reflects a light of longer wavelength.
5. They produce chemical effects.
6. They travel with the velocity of light.
7. They are absorbed by stratosphere and convert ozone to oxygen.

Uses

1. Ultraviolet rays when absorbed by a human body produces vitamin D which helps in growth of strong bones and teeth but causes blindness when seen directly.
2. Ultraviolet rays are used to detect original gems from fake ones.
3. Ultraviolet rays are used to detect forgery. The document to be checked is placed under ultraviolet rays and any alterations can be easily detected.
4. The UV rays kill germs and microorganisms. Hence they are used to sterilize surgical instrument and drinking water.

X-rays

They have wavelength ranging from 0.01 A° to 10 nm (100 A°). X-rays are produced when a high speed electron is stopped by a target. X-rays of shorter wavelength are used to study structure of materials because the spacing between the atoms in a solid ranges from 0·1 A° to 10 A°. These X-rays are called hard X-rays.

X-rays of longer wavelength called soft X-rays can penetrate flesh but they are stopped by bones and are used in radiography.

Gamma rays

Gamma rays have wavelength ranging from 0·0001 nm to 0·1 m. Gamma rays are produced by radioactive elements. Gamma rays have the highest penetrating power.

Gamma rays are used in treatment of cancer. They are used as γ ray microscope and are used to produce photo-electric effect.

Photoelectric effect

Max Planck proposed quantum theory in order to explain black body radiation. According to quantum theory, radiation consists of tiny packets of energy called quanta or photon. The energy of each quanta is given by $E = h\nu$, where h is Planck's constant whose value is 6.63×10^{-34} J s and ν is the frequency of radiation. These quanta travel with the velocity of light. According to quantum theory, radiation cannot be emitted or absorbed continuously. It can be done in integral multiples of $h\nu$.

When light of suitable frequency falls on a substance or metals, electrons are emitted out from the metal surface. This phenomenon is known as photoelectric effect. Electrons emitted are known as photoelectrons. UV rays, X-rays and γ-rays can cause photoelectric effect on certain metals.

The minimum frequency required for the incident light to eject electrons from the metal surface is known as threshold frequency. If the frequency of incident light is less than its threshold frequency, electrons are not emitted from the metal surface. If the frequency of incident light is greater than the threshold frequency electrons are ejected with kinetic energy or in other words velocity of photoelectrons increases with frequency. Threshold frequency is different for different metals.

Number of photoelectrons depends only on intensity of incident light and the photoelectrons increase with increase in intensity of incident light.

Einstein explanation of photoelectric effect

Einstein made use of Planck's quantum theory in order to explain photoelectric effect. Einstein assumed that tiny packets of energy called photons propagate through space and interact with matter. When a photon of energy $h\nu$ is incident on a metal, it interacts with an electron and transfers the whole of its energy to a single electron. The electrons are bound to the material and hence a certain amount of work has to be done to release the electron from the metal surface. If the frequency of incident radiation is lesser than the threshold frequency, electrons cannot be emitted and there is no photoelectric effect. If the frequency of the incident radiation is greater than the threshold frequency, part of the energy is used to release the electron and the remaining part is given as kinetic energy to an electron.

Application of photoelectric effect

1. Used to reproduce sound in cinematography.
2. Used in burglar alarms.
3. Used in automatic opening and closing of doors
4. Used in automatic control of traffic signals.
5. Used in exposure meter.
6. Used in automatic switching on and off street lights.

Theories of light (a historical review)

First century A.D.	The Greek scientists Euclid and Hero suggested that light travelled in straight lines and that light originated at the eye and travelled towards the object of vision.
~ 1000 A.D.	The Arab scientist Alhazen proposed that light radiating from a luminous source gets reflected from the objects to the eyes allowing them to be seen.
16th century	The English thinker Francis Bacon interpreted light as a disturbance that could be detected by the eye and that this disturbance gets 'stained' when they get reflected from an object causing the colours.
Early 17th century	Galileo and Kepler succeeded in the construction of a refracting telescope proving the laws of refraction, and Snell refined these laws.
17th century	Sir Isaac Newton and his contemporary Christian Huygens, a Dutch scientist, came up with contradictory theories on light stirring a debate over the truth in their proposals. While Huygen proposed his wave theory Newton came up with the corpuscular theory. Newton's theory was more readily accepted and prevailed for almost a century.

Newton's Corpuscular Theory

Based on his experiments in 1666 and the laws of mechanics, Newton proposed his corpuscular theory on the following assumptions.

1. Light is made up of point-like particles called corpuscles having negligible mass.
2. These particles are perfectly elastic and travel in straight lines.

3. Based on the laws of motion interfaces of different optical media exert forces of attraction and repulsion leading to the phenomena of reflection and refraction.
4. Speed of light is greater in a denser medium.
5. The different sizes of the corpuscles account for the different colours.

He went on to explain that just as a perfectly elastic collision takes place when a ball hits or strikes a rigid surface, reflection of light takes place when corpuscles strike a reflecting surface.

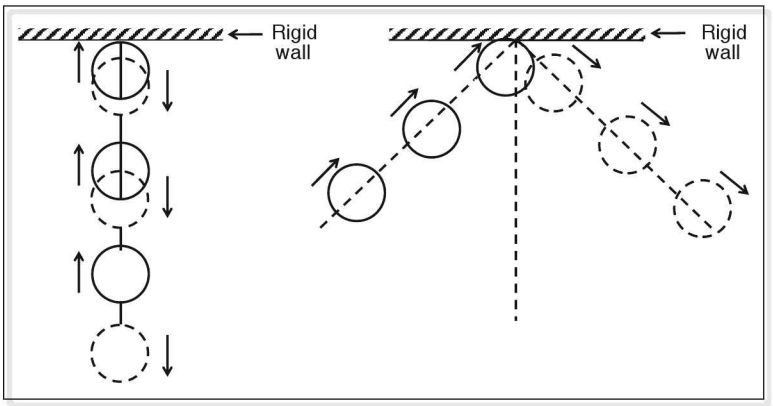


Figure 4.57 Elastic collision of a rubber ball and a rigid wall

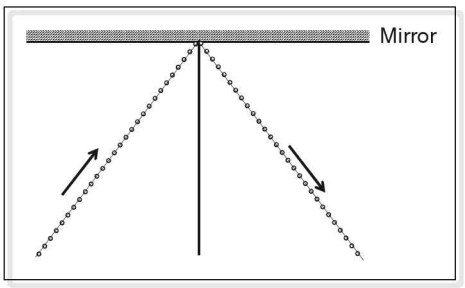


Figure 4.58 Reflection of light corpuscles

To explain the phenomenon of refraction, Newton suggested that as the light particles enter a denser medium they are attracted by particles of the medium and so travel faster resulting in the bending of the path traced.

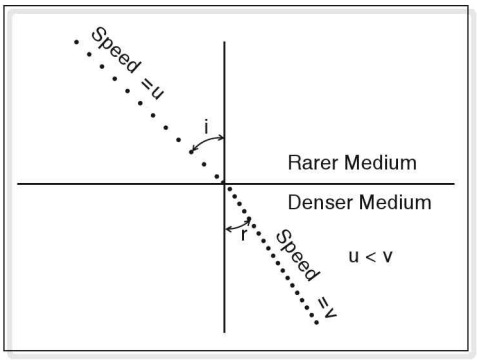


Figure 4.59 Motion of particles during refraction

Drawbacks of Newton's corpuscular theory

1. Newton proposed that velocity of light in a denser medium is greater than its velocity in a rarer medium but it was found that velocity of light in the rarer medium is greater than its velocity in the denser medium.
2. There was no justification for the assumption that different colours were due to different sizes of the corpuscles.
3. Newton's theory could not explain phenomenon like interference, diffraction and polarization.

Huygen's Wave Theory

A group of scientists led by the Dutch physicist Christian Huygens challenged Newton's corpuscular theory by claiming that the wave theory gave a better explanation of reflection and refraction. They put forth the following proposals to explain the behaviour of light.

1. All space is filled with an invisible substance called ether and light travels in the form of longitudinal mechanical waves.
2. Every luminous body which is a source of light, either point or extended (a line) would be at the geometrical centre of the propagation of wave and is called the primary source.
3. Any point in the medium which would act as a source of light for further propagation is referred to as a secondary source.
4. All points in the medium which receive the light from the source, either primary or secondary, at the same time would lie along a curved surface called wavelet.
5. The envelope of all wavelets in the same phase–receiving light at the same time–would form a wave front.
6. The 'ray of light' would be along the normal, drawn at a point.
7. At large distances from the primary source, (luminous body) the radius of curvature of a wave front would be large such that small protions of the wave front may be approximated to a plane.

On the basis of the above hypotheses, spherical, cylindrical or plane wave fronts may be constructed.

Newton–huygen debate

1. Huygen reiterated that reflection and refraction as well as dispersion of light could be convincingly explained with the help of wave theory.
2. Newton attacked the wave theory. He pointed out that if the wave theory was true then light should bend round the corners, just as sound and water waves do.

   (Interference and diffraction were not known until Fransesco Grimaldi's experiments later indicated that light does bend at corners).

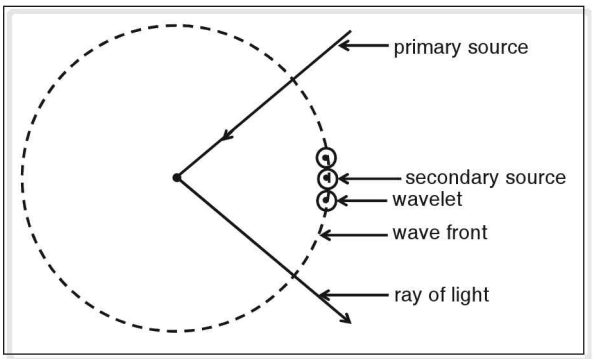


Figure 4.60 Spherical wave front

3. Newton's supposition that speed of light increased in denser medium was disproved when Foucault and Fizean's experiments on speed of light pointed out that light travelled faster in a rarer medium.
4. The assumption made by Huygens that space was filled with matter called ether could not be proved and as such did not gain much support.
5. Newton's theory suffered a major setback when first Thomas Young (1801) and later Fresnel, Malus and Fraunhofer discovered interference and diffraction, which could not be explained by the corpuscular theory, thus giving strength to the wave theory of Huygens.

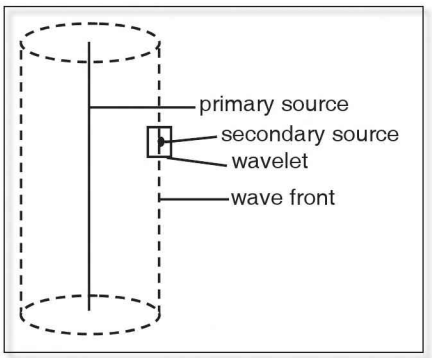


Figure 4.61 Cylindrical wave front

6. Huygen's postulate that light travelled as mechanical waves was disproved when in 1872 James Maxwell discovered the existence of electromagnetic waves; and Heinrich Hertz showed that these electromagnetic waves displayed all the phenomena of light, suggesting that light is also an electromagnetic wave, which is fundamentally a transverse wave.

Light was first believed to be possessing a particle nature giving rise to the particle theory of light. But then it was found that, while some observations could be explained using this theory there were observations which could not be explained. This led to the profounding of new theory put forward by Huygen called the 'wave theory of light'. But then certain observations explained by the particle theory of light could not be explained about the wave theory- to resolve this problem certain scientists came up with the idea that light behaves both as a particle and a wave leading to a new consolidated idea of 'dual nature of light'.

Ripple tank experiment

Certain phenomena like polarization of light confirmed the transverse wave nature of light. Even waves produced on the surface of water, also called ripples are also transverse waves. Hence certain phenomena of light which can be explained by its wave nature can be easily understood by taking an anology between water waves i.e., ripples and light waves. For this purpose, we consider a water tank made of glass or perspex (a transparent material) mounted on four legs kept as supports, at a height. The side view of the tank is shown in the following figure.

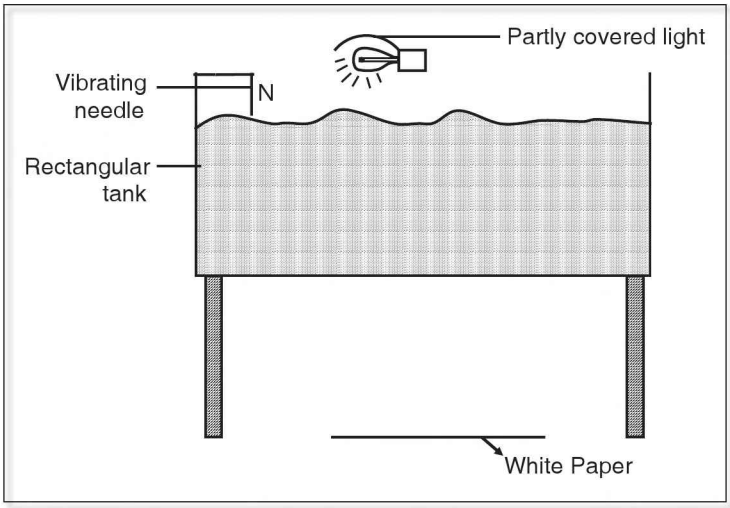


Figure 4.62 Ripple Tank (side view)

A small needle, which is fixed to one end of a metallic strip is connected to a vibrator and held above the water surface so that when the vibrator is switched on, the needle vibrates to and fro and produces ripples in the water. These ripples consist of crests and troughs and are circular in shape. The circular ripples produced at the needle expand while the wave moves in all directions.

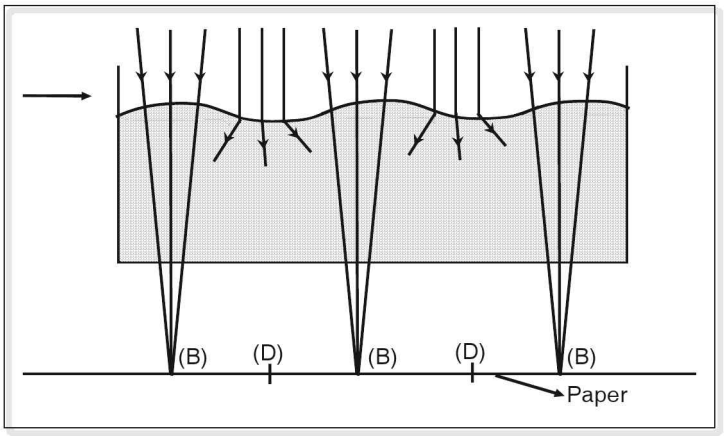


Figure 4.63 Formation of bright (B) and dark (D) bands on paper placed below the ripple tank

A white paper is placed below the tank on the ground surface. A bulb is held horizontally above the water surface. The light from the bulb passes through the water towards the paper kept below the ripple tank. The crest portion of the waves acts as a convex lens and thus forms a bright band of light on the paper placed below the tank. Similarly, the trough portion of the waves acts as a concave lens and thus forms a dark band of light on the paper placed below the tank. Thus the motion of ripples in the tank can be understood by the motion of dark and bright bands of light on the paper.

Reflection of waves in a ripple tank

Consider a ripple tank as described earlier, with a pin arrangement which vibrates and produces ripples in the tank. The circular ripples produced will be reflected by a hard surface like the wall of the tank. The top view of the ripple tank where the circular ripples are produced at source 'O' and reflected at the wall of the tank 'AB' is shown in the following figure.

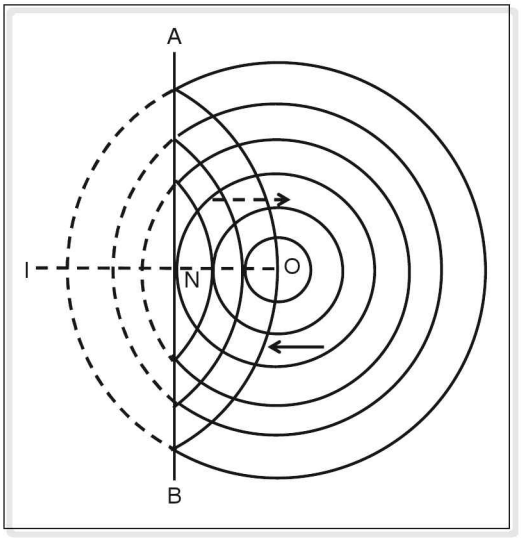


Figure 4.64

The ripples would reflect at AB such that the centre of the reflected ripples would be at 'I'. Thus 'I' is the image of 'O' and the distance ON = IN, which implies that the object distance is equal to image distance, considering AB as the reflecting surface.

The phenomenon of reflection can be studied by observing the movement of bright and dark bands of light on the white paper placed below the ripple tank.

Refraction of waves in a ripple tank

For refraction to take place we require two media with different densities such that waves from one medium travel to another medium. In a ripple tank this is possible by introducing a pile of plates at the bottom portion covering half the area. The side view of such an arrangement in a ripple tank is shown in the following figure.

The ripple tank is considered to be of two portions 'A' and 'B'. In portion 'B' at the bottom of water a pile of plates is placed such that the depth of water in the region 'B' is decreased and becomes shallow. So region 'A' is a deep water region. The velocity of water waves depends on the depth of water. The waves produced in the region 'A' while travelling toward region 'B' change their velocity; and hence their wavelength. In this case, a small straight strip like scale is used to produce waves instead of a small needle so as to produce plane waves in place of circular waves. The horizontal portion of the scale is in contact with water and vibrated to produce plane waves.

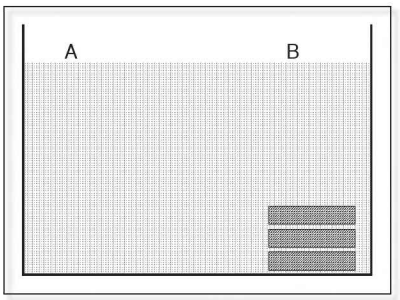


Figure 4.65

In the top view of the ripple tank shown in the following figure, XY is an imaginary line which acts as a boundary between shallow region 'B' and deeper region 'A'. The waves produced in region 'A' travel towards 'B' and in this process their velocity is decreased. Hence the wavelength of the waves is also decreased. This can be understood by observing the distance between two bright bands of light formed on the white paper below the tank in both the regions 'B' and 'A'. The bold lines in the above figure indicate the plane wave fronts and the lines with arrows which are normal to the waves indicate the direction of rays. It can be understood that the rays while travelling from rarer to denser medium bend towards the normal. A similar effect takes place in light.

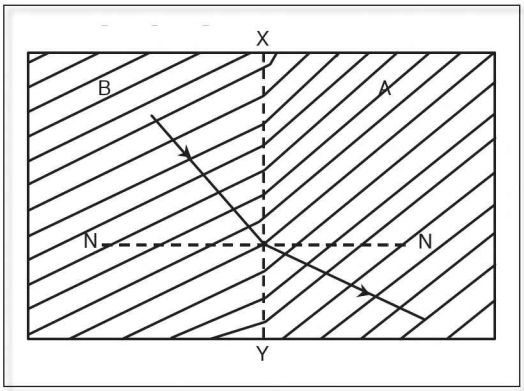


Figure 4.66

Superposition of waves

Consider two vibrating pins placed at different locations in a ripple tank. The waves produced from these two pins propagate in all possible directions and they meet in a common region. When the waves meet, the net effect (displacement) of the vibrating particles produced at a point is the sum of the individual effects (displacements) produced at the point by the two waves. The total displacement at a given point is equal to the vector sum of displacements at that point due to individual waves. This is caused due to the overlapping or combination of the two waves at the point. This phenomenon is known as superposition and according to the principle of superposition of waves, when two or more waves travel through the same medium and meet at a point, the resultant displacement of the vibrating particle at a point is the vector sum of the displacements due to the individual waves at that point.

If a crest or trough of one wave combines with the crest or the trough of another wave respectively at a given point, the net displacement at the point is maximum and it is called 'constructive superposition'. Similarly when the trough or crest of one wave combines with the crest or the trough of another wave respectively at a given point, a net displacement at the point is minimum and it is called as 'destructive superposition'.

Interference of waves

During the superposition of waves, if the waves produced from two sources have the same frequency and amplitude and travel in the same direction the superposition of such waves produce well defined constructive and destructive superposition in space and this phenomenon is called 'Interference'.

In interference, when constructive superposition takes place it is called 'constructive interference' and similarly when destructive superposition takes place it is called 'destructive interference'. To obtain a sustained interference pattern i.e., a regular arrangement of maxima and minima, the following conditions are essential.

(a) The waves that take part in interference must be in phase or have a constant phase difference between them. Such waves which have a constant phase difference between them are called coherent waves and the corresponding sources are called coherent sources.

(b) The frequency or wave length of the two waves must be equal.

(c) To get a maximum at a point in interference, the phase difference between the two waves must be $2n\pi$, where n = 0, 1, 2, etc.

(d) To get a minimum at a point in interference, the phase difference between the two waves must be $(2n-1)\pi$ where n = 1, 2, 3 etc.

Interference of water waves in a ripple tank

Consider two pins, N_1 and N_2 arranged in a ripple tank instead of a single pin. These are used to produce ripples in the tank. Let these two pins vibrate simultaneously with the same frequency such that circular ripples are produced at the two locations 'N_1' and 'N_2' and travel in all possible directions.

The top view of the ripple tank with the above described arrangement is shown in the following figure.

When the distance between the two sources (N_1 and N_2) of the waves is lesser compared to their distance from screen PQ an interference pattern is observed. The positions where two crests or two troughs of waves from two sources 'N_1' and 'N_2' combine and produce constructive interference are marked with dots and the positions where a crest from one wave combines with a trough of another wave and produces destructive interference are shown by circles. Thus on the screen 'PQ' we observe the interference pattern, with maximum displacement of the water particles at some points and minimum displacement of the water particles at some points. The points of maximum displacement fall on the antinodal line and the points of minimum displacement fall on the nodal line. We also observe that these nodal and antinodal lines are distributed uniformly on the screen.

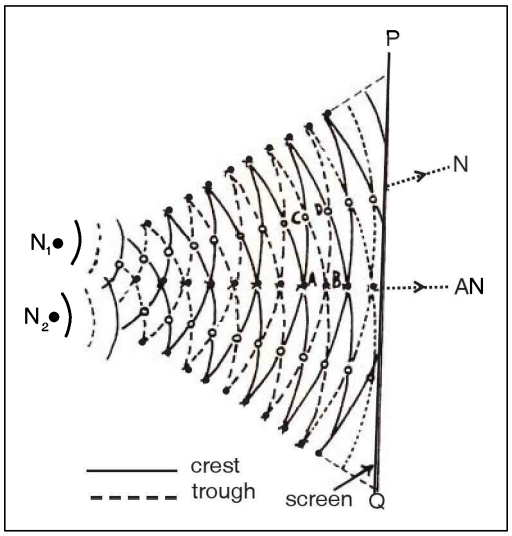


Figure 4.67 Interference of water waves in a ripple tank

Similar to the interference of water waves taking place in a ripple tank, interference also takes place in light and on the screen where we observe interference pattern, the brightness of the light varies from maximum to minimum at equidistant locations forming bands of light called 'fringes'. The band of light where brightness is maximum is called a bright band and is caused by constructive interference of light. Similarly, the band of light where brightness is minimum is called a dark band and is caused by destructive interference of light.

Diffraction of waves

Consider a ripple tank where a single pin is placed for producing ripples. The ripples produced are circular in shape. As the wave formed moves away from the source, the circular wavefronts are formed.

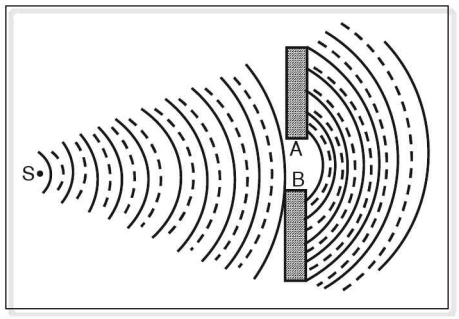


Figure 4.68 Bending of waves at the opening 'AB'

In such a situation, when an obstacle is placed such as a wooden plank with a wide hole 'AB' called an 'aperture' (only cross section of the plank is shown in the figure) as shown in the figure, the part of the wave striking the opening 'AB' will pass through it. But the hole AB itself becomes a 'secondary source' and waves travel in all directions beyond the hole on the other side of the plank. In the following figure, the top view of the ripple tank with waves that are passing through the opening 'AB' is shown.

This bending and travelling of the waves as they pass through the edge of an obstacle is known as diffraction. The bold lines in the above figure indicate the crests of wave fronts while the dotted lines indicate the troughs. If the diameter of the opening, i.e., the hole decreases one can observe that the ripples which pass through the opening are likely to be more circular and this opening will act as another source that produces waves.

The hole itself acts as an obstacle and when the size (i.e., diameter) of the hole decreases to such an extent that the diameter of the hole is comparable with the wave length of waves, the diffraction pattern is more prominent.

Diffraction of light

The wavelength of light is of the order of angstroms. So, when light waves encounter obstacles of very small sizes, the light waves bend around the edges of the obstacle and travel. This is known as diffraction of light. It is due to the diffraction of light that a sharp shadow of an object cannot be obtained on any screen placed behind it. Due to diffraction, light bends and travels into the region of geometrical shadow of the object and thus complete darkness is not found in the shadow. The brightness of light gradually decreases towards complete darkness as we go gradually into the geometrical shadow of the obstacle.

Diffraction of waves at a straight edge

If a lengthy strip like object, for example a half meter scale (AB), is held perpendicular to the motion of waves in a ripple tank such that the width of the scale is perpendicular to the surface of water as shown in the figure below, the bottom portion of the waves is obstructed by the obstacle 'AB' and the upper portion of the waves advances further.

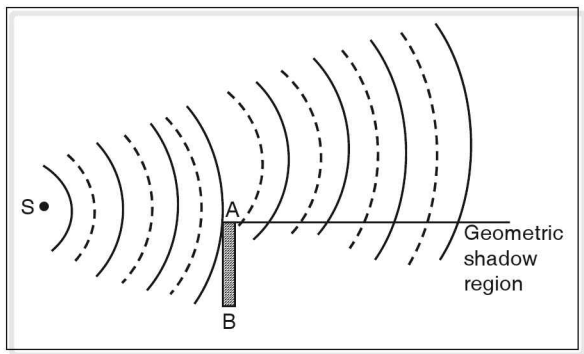


Figure 4.69

But the waves moving beyond 'AB' bend and travel into the geometric shadow region of the

obstacle. This is the diffraction of waves observed due to a straight edge. Similar diffraction is observed when light is diffracted at a straight edge and the light after diffraction encroaches into the geometric shadow of the straight edge where one can observe a gradual decrease in brightness of light in the shadow.

Visual photometry

When we tend to study in the light of a candle during night, we place ourselves near the candle. When we move away from the candle it becomes difficult for us to read; even though the amount of light emitted by the candle does not change. Hence for clear vision, we need a light of sufficient brightness. Our eyes cannot bear too much of brightness of light also. So there is a need to measure the brightness of light. The branch of physics which deals with the measurement of brightness of light is known as 'visual photometry'. Brightness of light from different sources cannot be compared just like that by observing and if it is done that way, it is not authentic. So certain standards and techniques are used to measure the brightness of light from a source. The branch of physics which deals with measurement of brightness of light emitted by different sources using certain standards and some special techniques is known as visual photometry. The following are some of the basic terms that are used in visual photometry.

(a) Luminous flux or light flux (ϕ)

A source of light emits radiations depending on its temperature and the surface of the body. But eye is sensitive only to the amount of visible energy emitted by it. Luminous flux is the amount of radiant visible (luminous) energy emitted per second from the source. It is denoted by ϕ and is measured in Lumen.

(b) Solid angle

Consider a small portion of an area on the surface of a sphere as shown in the figure. From the centre of the sphere, draw the normal lines on to the boundary of the area considered on the surface of the sphere. These lines form a cone.

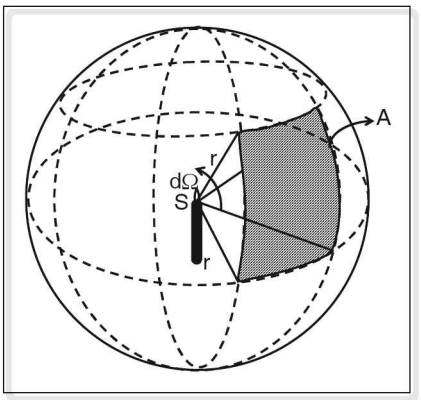


Figure 4.70 Concept of solid angle

The angle enclosed by the conical surface is known as solid angle. It is equal to the ratio of the area of the portion on the surface of the sphere to the square of the radius of the sphere.

If the area of the portion on the spherical surface is 'A' and the radius of the sphere is 'r' then the solid angle is given by $\Omega = \frac{A}{r^2}$; and is measured in steradians (sr)

Note: The total surface area of a sphere is $4\pi r^2$ where 'r' is its radius.

$\therefore$ Solid angle in a sphere $= \frac{4\pi r^2}{r^2} = 4\pi$ sr.

(c) Luminous intensity

A light source emits light in all directions. The luminous flux emitted by a light source per unit solid angle in a particular direction is called luminous intensity. It is denoted by I.

So $I = \frac{\varphi}{W}$.

Luminous intensity is measured in units of candela (cd)

(d) Candela (cd)

Luminous intensity of a point source is said to be 1 candela if one lumen of luminous flux is emitted per unit solid angle.

$1 \text{ cd} = 1 \text{ lm sr}^{-1}$

(e) Lumen (lm)

Lumen is the unit of luminous flux. One lumen is defined as the amount of luminous energy emitted per second by a uniform source of light of luminous intensity one candela within a cone of unit solid angle.

1 lm = 1 cd sr

1 lm = $\frac{1}{683}$ W at 555 nm of wavelength.

(f) Candle power

Candles emit different intensities of light in different directions. The luminous intensity of the light source in a particular direction is known as candle power. It is expressed in candela (cd).

Laser

It is the acronym of light amplification by stimulated emission of radiation.

Amplification is the process of increasing the energy and that is done in the light by stimulated emission of radiation; and so it is called laser. Lasers can be identified by their characteristic properties and because of them they differ from ordinary light.

Let us have a comparative study of ordinary light and a laser.

1. Phase difference	The light waves emitted from an ordinary source of light do not have a constant phase difference between them. The phase difference may change with time. Thus ordinary light is incoherent.	The light waves emitted from a laser source have a constant phase relationship between them which does not change with time. This is called temporal coherence.
2. Directionality	An ordinary light source emits light in all directions due to divergence.	Laser emits light only in one particular direction and hence has high directionality.

(Continued on following page)

3. Monochromacity	In case of an ordinary light, intensity is maximum at a particular wavelength. But the intensity of light is not zero in the neighbourhood wavelengths of that particular wavelength. So it has a band width and it is of the order of 1000 A°.	For a laser, the intensity of light is maximum at a particular wavelength and in the neighbourhood of that wavelength, the intensity of light is zero. So the bandwidth of laser is very less and it is of the order of 10 A°
4. Intensity	The light energy received per second per unit area is intensity. It is less in the case of ordinary light as an ordinary source spreads light uniformly in all directions.	The light from a laser source is limited to a small region, and it is not spread out like an ordinary light. So intensity of light is high in laser light.

The above differences between ordinary light and laser light help in the study of characteristics of laser light which makes it different from an ordinary light.

Characteristics of laser light

1. Generally light is emitted from a source by transition of electrons in atoms from an excited state to ground state. When transitions take place in a random manner, there is no phase relationship between light waves emitted during the transitions of electrons. But in the case of a laser source, this transition of electrons from an excited state to ground state takes place in a regular order. Due to this, there exists a definite phase relationship between light waves emitted during the transition, which does not change with time. Thus, we say that these waves are coherent and the phenomenon relating to the constancy of this phase relationship is known as 'temporal coherence'. This is an important property of laser light.
2. We know that the light from a normal source is divergent and moves in all possible directions, distributing the light energy over a large area. But the light from a laser source is spread over very small area when compared to ordinary light. Thus we can say that laser light is more limited to a particular direction and this property is known as 'directionality'.
3. The energy received per unit time per unit area is known as intensity. Since the laser light is not spread over a large area and is limited to a smaller one, all the light is incident on a small portion. So, the intensity of laser light is many times more than that of an ordinary light.
4. It is observed that the intensity of light varies with its wave length. As the wave length of light increases, its intensity increases and at a particular value of wavelength, it is maximum. Further, as the value of the wavelength becomes higher the intensity gradually decreases. So, the wavelength of a light is specified as the wavelength at which its intensity is maximum. The intensity of light does not abruptly become zero for certain wavelength range on either sides of the wavelength corresponding to maximum intensity. This range of wavelength, about a wavelength corresponding to maximum intensity is called 'bandwidth' and this band width is very less for laser light when compared to that of ordinary light. Hence we can specifically assign a single value of wavelength to a laser light than to mention the band width of wavelength. This property of laser light is known as 'monochromacity', as 'chromacity' refers to colour which is a characteristic of wavelength of light.

Working of lasers—The process

The process involved in working of lasers are

1. absorption
2. spontaneous emission
3. pumping and population inversion
4. stimulated emission of electro magnetic radiation.

Light is emitted from a source when electrons have a transition from higher energy level to a lower energy level. So, for a transition to take place, electrons are initially excited from the ground state to a higher state (energy level).

Absorption

Consider two energy levels E_1 and E_2. When a photon of energy $h\nu$ (ν is the frequency of radiation) is incident on an atom, the electron absorbs this energy, and if this energy $h\nu = E_2 - E_1$, the electrons jumps from the ground state E_1 to the excited state E_2. In cases of gases or solid a large number of atoms participate in the absorption of electromagnetic energy from an external source.

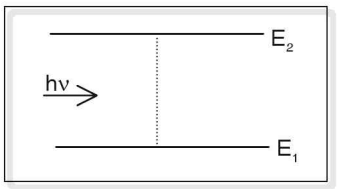


Figure 4.71

Spontaneous emission

Electrons absorb energy from an external source like ordinary light. Once energy is absorbed from an external source, the electrons move from ground state to an excited state. But the electrons in the excited state cannot be maintained there for a long time and tend to come back to the ground state, on their own. Thus electrons from excited state are transmitted to the ground state spontaneously and the resultant emission of light due to that spontaneous transition is known as spontaneous emission. But these spontaneous transitions of electrons take place at random and hence the light energy in spontaneous emission lacks coherence.

Population inversion

To have a coherent radiation, a special system called 'active system' is introduced whose energy level lies between the ground state and the excited state of electrons, and the electrons transited to this active level (active state) from excited state remain in the state for longer period than their stay in the excited state. The radiation which is emitted when electron transitions take place from excited state to the active state, is very less and is completely absorbed by the active medium as thermal energy. Hence this radiation is 'invisible radiation'. The stability of the electrons in the active state is more and is called the 'meta stable state'. As the transition of electrons from excited state to active state continues, a stage may be reached where the number of electrons in the active state may be more than that in the ground state. Thus the number of electrons in the active medium in the meta stable state is more than that in the ground state and this stage of increasing electrons in the meta stable state(i.e., active medium) is called 'population inversion'.

This achievement of population inversion can be done by inducing radiation to the active state from an external source. The process of achieving population inversion by inducing radiation from an external source is called pumping.

Consider three states of energy level E_1, E_2 and E_3.

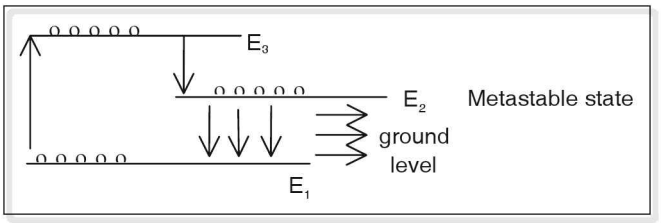


Figure 4.72

When a photon of energy $h\nu = E_3 - E_1$ is incident, the electrons from the ground state are excited to energy level E_3. But the lifetime of the electrons in the excited state is very short (10^{-8} s). As a result electrons return to the ground state by spontaneous emission but some electrons jump to energy level E_2 emitting a small amount of energy equal to $E_3 - E_2$. This energy is absorbed as thermal energy by the medium. This is known as non-radiative transition.

Electrons in the energy level E_2 have comparatively longer lifetime (milli seconds). This energy level is called meta stable state. At some instant the number of electrons in the meta stable state is greater than electrons in the ground state. This is known as population inversion.

Stimulated or induced emission

The induced radiation (whose energy is equal to the energy difference of active and ground levels) stimulates the electrons in the active state and so is also called stimulated radiation. When a large number of electrons in the active state are stimulated, they are transited from the active state to the ground state simultaneously which gives rise to a temporal coherent beam of light. This process is called 'stimulated emission' which increases the intensity of light.

The stimulated emission can be made dominant by placing the active medium discussed above in a cavity so that multiple reflections of light take place before the beam comes out, thereby amplifying the light energy.

Thus, the incident light is amplified by stimulated emission of radiation and hence LASER.

Actual laser and its working

Actual laser consists of an active medium in a resonator. Two mirrors, one perfectly reflecting and the other partially reflecting are fixed, one at each ends of the resonator. The mirrors are parallel to each other. Energy is supplied by a flash of light from an external source for optical pumping which results in population inversion.

During stimulated emission, light or photon travels to and fro (perpendicular to mirror surface) after being reflected by the mirrors. Intensity of light increases due to multiple reflection. This amplified light escapes through a partially reflecting mirror as a narrow beam of monochromatic light.

Types of lasers

Different types of lasers that are available are, solid state lasers like ruby laser, liquid and dye lasers and gaseous lasers like helium neon (He-Ne) lasers.

In lasers, solid, liquid or gas is used. Thus laser can be broadly classified as

1. solid laser,
2. liquid and dye laser
3. gaseous lasers

Ruby laser is a solid laser and the active system used is ruby (Al_2O_3, Cr^{3+}). Xenon discharge tube provides flash light required for optical pumping. Wavelength of ruby laser is 6943 A°. Helium – Neon laser is a gas laser. Neon acts as the active system. Radio frequency is used for optical pumping. Wavelength of He-Ne laser is 6328 A°.

Uses of lasers

Lasers have many applications in the fields of medicine, industry and in science and technology. Some of them are mentioned below.

1. These are used in bloodless surgeries, where surgeries are done without cutting the body parts with knife.
2. Lasers are used for welding and cutting heavy metallic parts in industries.
3. They are used to store and retrieve a large amount of information in compact discs.
4. They are used to destroy inter continental ballistic missiles in air itself without the help of any other missile.
5. They are used in three dimensional photography called holography.
6. They are used in the field of communication through fibre optics.
7. Lasers are used to study Raman scattering to analyse the molecular structure of a material.
8. Lasers are used to separate isotopes containing mixture of isotopes of an element.

Optical instruments

Instruments like mirrors, lenses etc which deal with light are called optical instruments.

The human eye is a sense organ which causes the sensation of vision, and this helps us to "see" the objects and also perceive their colours. But the human eye has certain limitations. Some of the limitations are listed below.

1. Defects of vision, like short sight, long sight, colour blindness, etc.
2. Range of visibility.
3. Inability to see very small objects like cells, etc.
4. Make visual recording for posterity.

To overcome these limitations, man has invented optical instruments like spectacles, microscopes, telescopes, etc. To record visual observations for further reference the photographic camera, video camera, etc. were invented.

There has been a continuous effort to make these instruments more sophisticated and effective with the advances made in science and technology helping us in this.

Let us first understand certain aspects of the functioning of the human eye.

The Human Eye

The eye is a spherical ball having a diameter of about 2.5 cm, with a slight bulge in the front portion. The posterior part of the eye-ball is filled with a fluid called vitreous humour which helps in preventing the eye-ball from collapsing. The exterior part of the eye ball is covered by a thick fibrous tissue called sclerotic which gives a definite shape to the eye-ball and protects the interior parts.

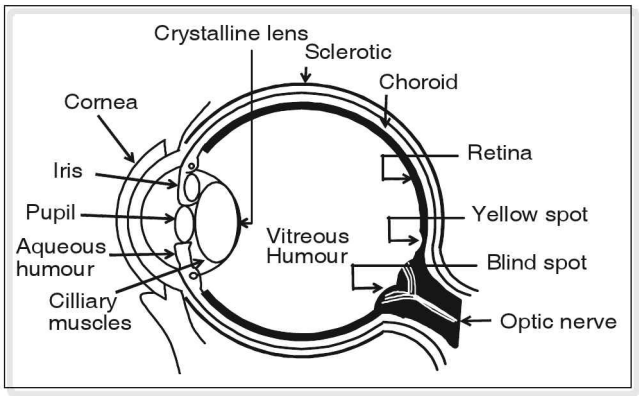


Figure 4.73

A crystalline lens is located in the front part of the eye. The focal length of the lens is varied with the help of ciliary muscles along the periphery such that the image of the object viewed falls on the retina, which forms the screen. This process of altering the focal length of the lens is known as accommodation of the eye. Several thousand nerves coming from the brain enter the eye from the rear and terminate along the retina in the form of light sensitive cells. These cells are of two types—rods and cones, and they help us to perceive the brightness and colour of the objects viewed.

The choroid between the sclerotic and the retina is a dull grey membrane preventing internal reflections and thus keeping the interior of the eye-ball dark.

The slight bulge in the front is due to the presence of a fluid called aqueous humour in front of the crystalline lens. This fluid forms a primary lens of almost a fixed focal length. It prevents collapsing of eye due to changes in atmospheric pressure.

Light enters the aqueous humour through the cornea which is a thin membrane holding the aqueous humour in place.

Behind the cornea is the iris which is an opaque muscular diaphragm having a small aperture or an adjustable opening called the pupil. This controls the light entering the retina. When the muscles contracts, the size of the pupil increases allowing more light to enter and vice versa.

The light passing through the pupil falls on the eye lens which is held in position by ciliary muscles. These rays after passing through the eye lens are focused on retina. Real inverted image is informed on the retina. Any image formed on the retina is converted into optical impulses which are sent to the brain through the optic nerve.

Yellow spot is situated at the centre of the retina and is the most sensitive part of the retina that helps us to view objects closely. The blind spot refers to the portion of the retina where there are no light sensitive cells.

Two eye-balls are present in the front part of the head to estimate the distances of objects and their relative visual depth.

The human brain cannot perceive more than two images formed on the retina within an interval of 1/16th second. This is called persistence of vision.

Formation of images on retina

The images formed on the retina are always real, inverted and highly diminished. While viewing distant objects (at infinity) the ciliary muscles relax and this leads to an increase in the focal length of the lens. In this case, the image is focused on the retina. While viewing nearer objects the ciliary muscles contract. This leads to a decrease in the focal length of the lens and the focus is formed in front of the retina such that the image would form on the retina i.e., between f and 2f.

Least distance of distinct vision: It is the minimum distance at which objects can be seen more distinctly without strain. It is also called the near point of the eye. It is denoted by D. For a normal person least distance of distinct vision is 25 cm. Infinity is referred to as far point.

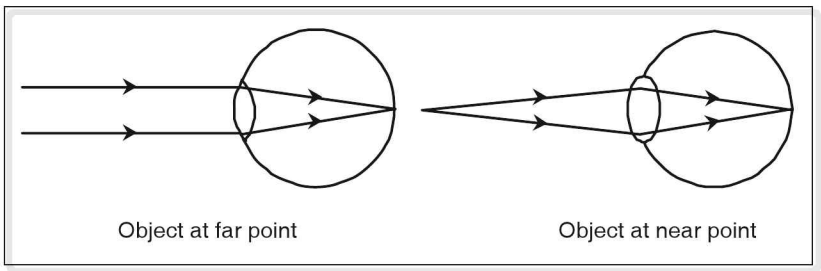


Figure 4.74

Perception of colours

The rod shaped cells on the retina respond only to the intensity of the incident light, but cannot perceive colours. The cone shaped cells are sensitive to colours. There are three types of cones responding to each of the three primary colours—red, green and blue, helping us perceiving and distinguishing between different colours. Absence of certain cone cells sensitive to specific colours would lead to a defect known as colour blindness in some people.

Defects of vision and rectification

1. Myopia or Short sightedness

It is the ability to see near objects and inability to see distant objects. This defect occurs due to slight elongation of the eye ball. As a result the distance between eye lens and retina is greater (focal length of lens decreases) and the image is formed in front of the retina and not on the retina.

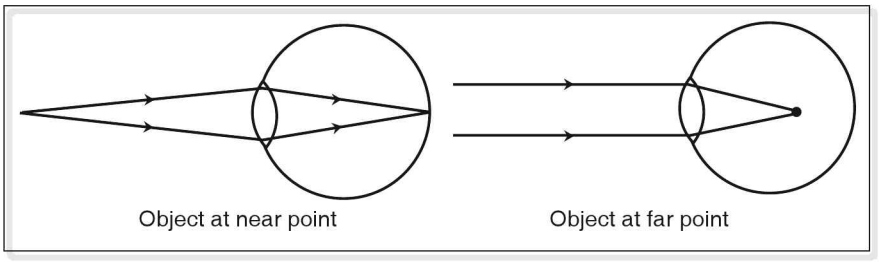


Figure 4.75

Correction

Myopia can be corrected by using a concave lens of suitable focal length to have the apparent position of the far point at near point.

Image formation with the corrective lens

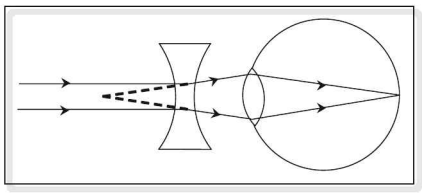


Figure 4.76

2. Hypermetropia or long-sightedness

Ability to clearly and comfortably see objects at a distance and inability to see clearly objects that are nearby is referred to as hypermetropia more commonly known as long-sight.

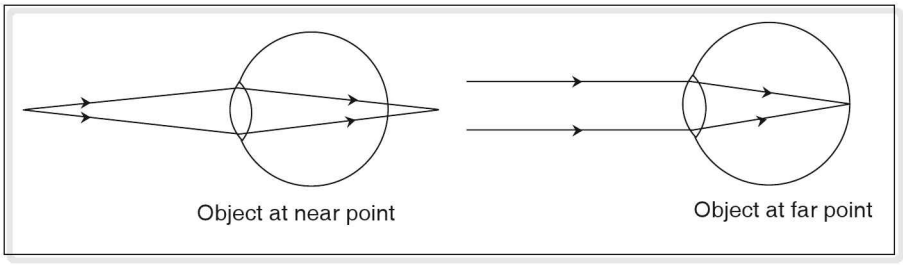


Figure 4.77

This defect is caused by slight flattering of the eye ball. The distance between the eye lens and the retina is less than in case of a normal eye. As a result the image formed by nearby objects is behind the retina.

Correction

Hypermetropia can be corrected by using a convex lens of suitable focal length to make the image form on the retina, i.e., by decreasing the combined focal length or by increasing the power of the combination.

Image formation with corrective lens

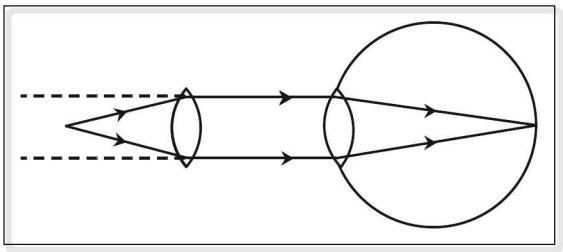


Figure 4.78

3. Presbyopia

Inability to see clearly objects that are near due to the ageing of the ciliary muscles is referred to as presbyopia. Thus, presbyopia is the condition of hypermetropia due to ageing.

Presbyopia can be corrected by using a 'reading glass', i.e., a suitable convex lens—a magnifying glass.

A bi-focal lens is often used when the defects of myopia and hypermetropia exist together, i.e., the person is unable to see both far distant objects as well as near objects.

4. Astigmatism

When the cornea is not spherical and is curved differently in different directions, then focusing simultaneously on horizontal and vertical lines becomes difficult. Such a condition is referred to as astigmatism.

Astigmatism can be corrected by using a cylindrical lens.

5. Cataract

Due to ageing, a thin membrane forms over the crystalline lens making it hazy. This leads to decrease in the brightness of objects and may also lead to blindness. This condition is called cataract and can be corrected by suitable surgery.

Bees have some retinal cones which are sensitive to ultraviolet light. Humans retina are not sensitive to ultraviolet light, and hence we cannot see objects in the ultraviolet light. Cones of chicks are sensitive to bright light only and hence chicks wake up with sunrise and sleep in the dusk.

Camera

Camera is used to record image on a film. When an object is beyond 2F a real, diminished, inverted image is formed between F and 2F on the other side of convex lens.

A camera consists of a convex lens of small focal length fixed to one end of a box. To the other end is a light sensitive film. The inner sides of the box are blackened to prevent internal reflection. Distance

between the screen and the lens can be adjusted by the knob. A circular plate having an aperture in it is placed behind the lens and it is called diaphragm. The diameter of the aperture can be varied and the aperture controls the amount of light falling on the film. The size of the aperture is expressed as f-numbers.

f-number is defined as the ratio of the focal length of the lens to the diameter of the aperture.

$$\therefore \text{f-number} = \frac{\text{Focal length of the lens}}{\text{Diameter of the aperture}}$$

According to the intensity of light, the f-number may be adjusted. In dim-light lower f-number is used and higher f-number is used in bright light.

To control the exposure time of the film, a shutter is provided, whose speed can be controlled.

To take a photograph of an object the distance between the lens and the film is adjusted, so as to get a clear image in the view finder. Then f-number and shutter speed are adjusted depending upon the intensity of light. When the button is pressed, the film gets exposed to the light entering through the lens. The film is then treated chemically, a process known as 'development' and then prints of the film are taken on a photographic paper.

Comparison between a camera and the eye

A camera in many ways is similar to the human eye. Perhaps the human eye was the source of inspiration for the design of a camera.

The important differences between the two are as follows:

1. In a camera, the distance between the lens and the screen can be adjusted but not the focal length of the lens, whereas, in the case of the eye, the ciliary muscles adjust the focal length keeping the distance between the lens and the retina constant.
2. The interior of the eye-ball is filled with vitreous humour but in a camera, the chamber is hollow and empty.
3. The image formed on the retina is temporary and its impression is recorded in the brain as memory. The image formed on the film is a permanent record.
4. Distance between crystalline lens and retina remains the same but in camera distance between the lens and the film can be changed.
5. Retina converts light signals to electrical impulses which are transmitted to brain whereas in camera film gets exposed when light is incident on it, due to chemical reactions.
6. Eye is subjected to a number of defects whereas in camera there are no such defects.

The simple microscope

A convex lens of short focal length which is used to observe small objects clearly is known as a simple microscope. It is also called a magnifying glass. To construct a simple microscope, a convex lens of short focal length is fixed in a circular metallic frame provided with a handle to hold.

Working

The size of the image of an object formed on our retina increases if the distance of the object from our eyes decreases. This happens due to the fact that as the object comes closer to our eyes, the angle subtended by the object at our eyes increases and so the size of the image of the object on the retina also increases. This is known as 'angular magnification'.

If an object of size 'h' is placed at the least distance of distinct vision (D) from the eye, the angle which it subtends at the eye can be given by $\alpha_0 = \frac{h}{D}$.

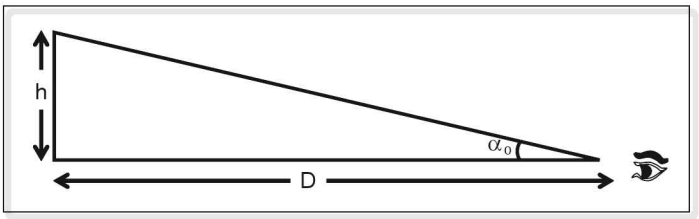


Figure 4.79

If this angle is increased, the size of the image of the object increases and this can be done by placing a convex lens of short focal length between the eye and the object such that the distance between the lens and the object is equal to the focal length of the convex lens. Then the image of the object is formed at an infinite distance. In this situation the image can be observed with the least strain to eye. This adjustment is called normal adjustment.

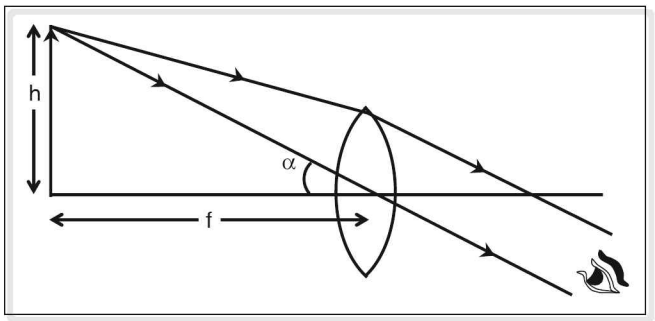


Figure 4.80

Here, the angle subtended by the object at the eye is equal to the angle subtended by the object at the optic centre of the lens (α) and it is greater than the angle subtended by the object, when it is placed at 'D' from the eye. So the image of the object is magnified. Here $\alpha = \frac{h}{f}$.

Magnification in normal adjustment

The angular magnification of the image of the object is the ratio of the angle subtended by the object at the eye in the normal adjustment situation to the angle subtended by it at the eye when it was at the least distance of distinct vision (D).

$$\therefore \text{Magnification } m = \frac{\alpha}{\alpha_0} = \frac{h/f}{h/D} = \frac{D}{f}$$

Magnification when the object is placed between the focus and optical centre of the lens

If the object is placed before the lens at a distance less than its focal length, a virtual, erect and a magnified image is formed on the same side of the lens to which the object is placed. The distance of the object from the lens can be adjusted so that the image is formed at a distance from the lens which is equal to 'D' from the lens.

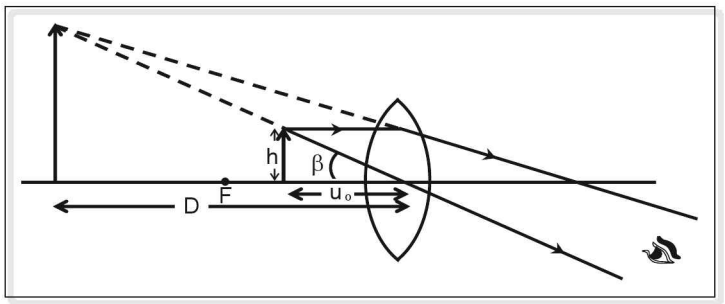


Figure 4.81

Now, the angle subtended by the object on the lens and on the eye is 'β' and the angular magnification now is given by $m = \frac{\beta}{\alpha_0}$. But $\beta = \frac{h}{u_0}$ and $\alpha_0 = \frac{h}{D}$.

$$\therefore m = \frac{h/u_0}{h/D} = \frac{D}{u_0}$$

The lens formula is $\frac{1}{f} = \frac{1}{v} - \frac{1}{u}$ where f, u and v are focal length of the lens, object distance and image distance respectively. Using the sign convention we get here, v = –D, u = –u.

$$\text{So, } \frac{1}{f} = \frac{1}{-D} - \frac{1}{-u_0}$$

$$\Rightarrow \frac{1}{u_0} = \frac{1}{f} + \frac{1}{D} \text{ or } \frac{D}{u_0} = \frac{D}{f} + \frac{D}{D} \Rightarrow \frac{D}{u_0} = 1 + \frac{D}{f}$$

As $\frac{D}{u_0} = m$, the magnification $m = 1 + \frac{D}{f}$.

Uses

1. Used as reading glass.
2. To observe weaving patterns and to read vernier scale.

Compound microscope

A simple microscope can produce images having magnification up to a certain limit. It is not possible for a simple microscope to produce images of greater magnification. When images of greater magnification are required, we use another convex lens in combination with the one used in a simple microscope. The combination of these two lenses which is used to obtain images of greater magnification is called 'compound microscope'.

Construction and working

The two lenses that are used in a compound microscope are each fixed to the outer ends of two hollow tubes that can slide one over the other such that their principal axes coincide. The lens that faces the object is called an 'objective' and is of smaller focal length when compared to the focal length of the other lens which faces the eye and is called the 'eye piece'.

A screw is provided so that on rotating, the distance between the objective and the eye piece can be adjusted.

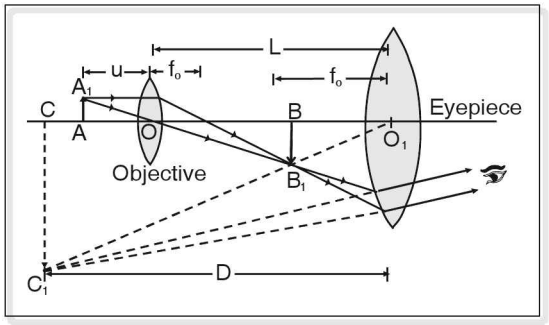


Figure 4.82

An object AA_1 is placed beyond the principal focus of the objective and its image BB_1 is formed between the objective and the eye-piece. The first image formed (BB_1) is inverted, real and magnified with respect to the object. This first image acts as an object to the second convex lens i.e., eye piece. The distance of the first image 'BB_1' from the eye piece is less than the focal length of the eye piece. So it is further magnified and a third and final image 'CC_1' is formed beyond 'BB_1'. The final image is inverted with respect to the object but virtual with respect to the first image. It is magnified when compared to the first image.

Magnification

The magnification of the compound microscope is the ratio of the size of the final image to the size of the object.

$$m = \frac{CC_1}{AA_1} = \frac{CC_1}{BB_1} \times \frac{BB_1}{AA_1} = m_e \times m_o$$

Where, m_o = magnification produced by the objective.
m_e = magnification produced by the eye piece.

When the final image is formed at infinity

$$\therefore m = m_o \times m_e = L/f_o \times D/f_e$$

Where, L = length of the tube = distance between the lenses
D = least distance of distinct vision
f_o = focal length of the objective
f_e = focal length of the eye piece

Telescopes

When an object is very far from the eye, the angle it subtends at the eye is very small and so it appears to be small. To make the object appear nearer to the eye, the angle it subtends at the eye should be increased. To make the distant objects appear nearer, the optical instrument used is called a telescope.

There are two types of telescopes, viz., the astronomical telescopes and the terrestrial telescopes. The astronomical telescopes are used to observe the celestial bodies and the terrestrial telescopes are used to observe distant objects on the earth's surface.

Astronomical telescope

An astronomical telescope has the same construction as that of a compound microscope, but the focal length of the objective is larger when compared to the focal length of the eye piece.

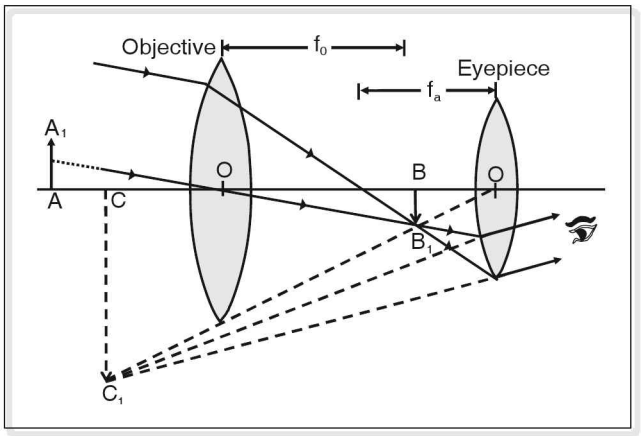


Figure 4.83

Working

A parallel beam of light rays from a distant object (AA_1) almost parallel to the principal axis of the objective are incident on the objective and an image (BB_1) of the object is formed between the objective and the eye piece. This first image which is diminished, inverted and real acts as an object for the eye piece. The distance between the first image and the eye-piece is less than the focal length of the eye piece. So, a magnified and a virtual image (CC_1) is formed on the same side of the eye piece to which the first image lies. The final image is erect with respect to the first image but is inverted with respect to the object. The angle subtended by the final image at the eye piece and so at the eye is larger than that of the angle subtended by the object at the eye. So the final image is magnified and appears to be having larger size than that of the object.

If the first image formed by the objective (i.e., BB_1) is formed at the principal focus of the eye piece, the final image is formed at infinity. This situation is called 'normal adjustment'. In the normal adjustment position, the length of the telescope (L) (which is the distance between objective and eye piece) is equal to the sum of the focal lengths of the objective and the eye piece ($L = f_o + f_e$). In the normal adjustment position, the magnification obtained by the telescope is given by $m = \frac{f_o}{f_e}$

Terrestrial telescope

In an astronomical telescope, the final image achieved is inverted with respect to the object. Due to this reason, it cannot be used to view distant objects on the earth's surface. Thus the astronomical telescope is modified by adding one more lens between the objective and the eye piece. The purpose of this lens is to make the image erect with respect to the object and so it is called an image erecting lens (EL) or simply erecting lens.

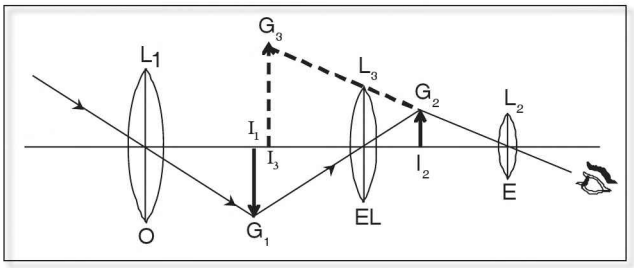


Figure 4.84

The first image (I_1G_1) formed by the objective is real, inverted and diminished and is formed at the centre of curvature of the erecting lens. This image (I_1G_1) acts as an object for the erecting lens and a second image (I_2G_2) is formed at the centre of curvature of the erecting lens between the erecting lens and the eye piece. The size of the second image (I_2G_2) is equal to that of I_1G_1. It is erect with respect to the original object and inverted with respect to I_1G_1. I_2G_2 acts as an object for the eye piece. The distance between I_2G_2 and eye piece is less than the focal length of the eye piece and so a third and final image I_3G_3 is formed. This image is erect with respect to I_2G_2. So the final image is erect with respect to the object.

test your concepts

Very short answer type questions

1. Define power of a lens.
2. _____ mirror can form virtual and diminished image.
3. What is luminous flux?
4. According to Huygen's theory, the colours of light are due to difference in the _____.
5. What is magnification?
6. What is spontaneous emission?
7. What is the condition required for constructive interference to take place?
8. Snell's law is not applicable when the angle of incidence is equal to _____
9. Which spherical mirror is used as rear view mirror?
10. Which type of lens is used to correct hypermetropia?
11. What is the principle of superposition?
12. When an ink spot which is at the bottom of a glass ($\mu = \frac{3}{2}$) slab of thickness 9 cm is viewed from top, the image of the ink spot appears to be raised by _____ cm.
13. What is refraction?

14. What is presbyopia?

15. Why do different electromagnetic radiation have different properties?

16. In a human eye colour is perceived by _____.

17. When is the angle of incidence greater than the angle of refraction?

18. An optical fiber used for telecommunication works on the principle of _____

19. What is meant by refrangibility?

20. Astigmatism can be verified by a suitable _____ lens.

21. Define refractive index.

22. The S.I. unit of luminous intensity is _____.

23. Define candela.

24. How does lateral displacement vary with wavelength?

25. For an object placed at a certain distance in front of a convex mirror the magnification is 0.8. If the focal length of the convex mirror is 40 cm, then the distance of the object from the mirror is _____ cm.

26. What is coherent scattering?

27. The angle of incidence of a ray of light passing through the centre of curvature of a concave mirror is _____.

28. What are the factors on which angle of deviation depends?

29. Both particle and wave nature are attributed to light according to _____ theory and _____ theory respectively.

30. What are composite colours?

Short answer type questions

31. Draw a ray diagram to obtain the image position when the object is beyond 2F of a convex lens.

32. Derive the expression for the magnification of an astronomical telescope in normal adjustment.

33. What is line emission spectrum?

34. State the condition for total internal reflection.

35. Explain briefly about a terrestrial telescope.

36. What is photoelectric effect?

37. What is astigmatism? Explain how it can be corrected.

38. What are total reflecting prisms?

39. What are the uses of lasers?

40. What are the postulates of Newton's corpuscular theory?

41. What are infrared rays? Name the source of IR rays and write their uses.
42. What is a refracting periscope?
43. Explain the construction of wavefronts using Huygen's principle.
44. What is Raman effect?
45. What are optical fibres?

Essay type questions

46. Distinguish between ordinary light and a laser.
47. Explain the construction and working of a compound microscope with a ray diagram. Derive the expression for magnification when image is formed at infinity.
48. State the new cartesian sign convention used in reflection by spherical mirrors
49. Explain the formation of mirages.
50. Derive lens formula.

CONCEPT APPLICATION



Concept Application Level—1

Direction for questions 1 to 7: State whether the following statements are true or false.

1. Interference takes place whenever two or more wave motions combine in space.
2. In the minimum deviation position of a prism, the refracted ray is always parallel to its base.
3. A person suffering from myopia cannot see distant objects clearly.
4. The focal length of a given lens depends on the surrounding medium.
5. Prismatic periscopes use the principle of total internal reflection.
6. A dentist uses a convex mirror to view the inner parts of a patient's mouth.
7. The solar spectrum in general is an absorption spectrum.

Direction for questions 8 to 14: Fill in the blanks.

8. A yellow coloured dress would appear to be _____ when viewed through a green filter.
9. Due to _____ of laser, large energy can be concentrated into an extremely small bandwidth.
10. The combination of a convex lens of power + 2 D and a concave lens of power –2 D, acts as a _____
11. The minimum length of a plane mirror required to view the full image of a person 6 feet tall is _____.

12. A bird is flying 12 m above the surface of a pond. The apparent position of the bird as viewed by a fish 1m below the surface is _____. $\left({}_{a}\mu_{w} = \frac{4}{3} \right)$

13. An object is placed on the principal axis of a concave mirror at a distance of 60 cm. If the focal length of the concave mirror is 40 cm then the magnification obtained is equal to _____

14. The power of a biconcave lens having focal length 25 cm is _____.

Direction for question 15: Match the entries in column A with appropriate ones from column B.

15.

A.	Plane mirror	( )	a.	Primary colours
B.	Convex lens	( )	b.	4π steradian
C.	Spectroscope	( )	c.	Incoherent scattering
D.	Atmospheric scattering of light	( )	d.	Magnification always one
E.	RB RBG	( )	e.	Terrestrial telescope
F.	Opt Optic fibre	( )	f.	Burglar alarm
G.	Sol Solid angle in a sphere	( )	g.	Virtual and magnified image
H.	Photo electric effect	( )	h.	Blue colour of effect the sky
I.	Raman effect	( )	i.	Observing and studying spectrum
J.	Erecting lens	( )	j.	Total internal reflection

Direction for questions 16 to 30: For each of the questions, four choices have been provided. Select the correct alternative.

16. A mirage takes place due to _____.
(1) reflection (2) refraction
(3) total internal reflection (4) Both (2) and (3)

17. Raman effect is an example of _____.
(1) coherent scattering of light. (2) incoherent scattering of light
(3) coherent interference of light (4) incoherent interference of light

18. The optical phenomenon which Newton's theory of light failed to explain is _____.
(1) interference (2) polarization
(3) diffraction (4) All the above three

19. An observer moves towards a stationary plane mirror at a speed of 4 m s^{-1}. The speed of image with respect to mirror is _____.
(1) 4 m s^{-1} towards the mirror (2) 8 m s^{-1} towards the mirror
(3) 4 m s^{-1} away from the mirror (4) 8 m s^{-1} away from the mirror

20. The process in which the electrons in the excited atoms are released on their own from their higher energy state to the ground state is called _____.
(1) forced emission (2) population inversion
(3) spontaneous emission (4) None of these

21. The refractive index of diamond is 2.4. If the velocity of light in air or vacuum is 3×10^8 m s^{-1}, then the velocity of light in diamond is _____ $\times 10^8$ m s^{-1}.
(1) 1.25 (2) 12.5 (3) 0.8 (4) 125

22. A convex mirror is used
(1) by a dentist.
(2) for shaving.
(3) as a rear view mirror in vehicles.
(4) as a light reflector for obtaining a parallel beam of light.

23. When an equilateral glass prism is in minimum deviation position,
(A) the refracted ray is parallel to the base of the prism.
(B) the angle of incidence is equal to the angle of emergence.
(1) Only A is true (2) Only B is true
(3) Both A and B are true (4) Both A and B are false

24. The colour of a green leaf, when white light passing through a yellow filter is incident on it, is _____.
(1) black (2) yellow (3) red (4) green

25. Spectrometer is used to measure _____.
(1) angle of prism (2) refractive index of material of a prism
(3) the dispersion (4) All the above

26. The expression for the magnification of a spherical mirror in terms of focal length (f) and the distance of the object from mirror (u) is

(1) $\frac{-f}{u-f}$ (2) $\frac{f}{u+f}$ (3) $\frac{-f}{u+f}$ (4) $\frac{f}{u-f}$

27. According to Newton, different colours of light are due to the difference in _____ of the corpuscles.
(1) mass (2) nature (3) shape (4) size

28. If the speed of light in medium – 1 and medium – 2 are 2.5×10^8 m s^{-1} and 2×10^8 m s^{-1} respectively then the refractive index of medium – 1 with respect to medium – 2 is _____.

(1) $\frac{3}{2.5}$ (2) $\frac{2}{2.5}$ (3) $\frac{2.5}{3}$ (4) $\frac{2.5}{2}$

29. A coin is placed at the bottom of a swimming pool. To obtain a vertical shift of 20 cm of the image of the coin, the required height of the water column is _____ ($\mu_{water} = \frac{4}{3}$).
(1) 20 cm (2) 2 m (3) 80 m (4) 80 cm

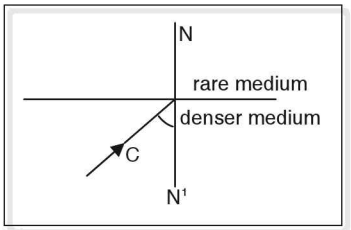
30. The apparent vertical shift of the image of a coin placed at the bottom of a water tank having constant depth of water is proportional to _____.
(1) μ (2) $\frac{1}{\mu}$ (3) $\mu - 1$ (4) $\mu + 1$

Concept Application Level—2

31. The critical angle of the material of a prism ABC is 40°. For a ray of light incident on AB, to emerge from the face AC, what should be the maximum value of angle A?

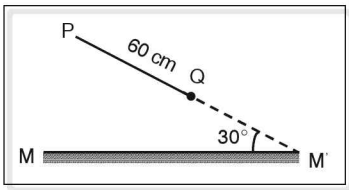
32. Two lenses, one convex and another concave, of focal lengths 10 cm and 20 cm respectively are taken together and used as a single lens. If a point object is held at a distance of 20 cm from the compound lens, find the position of its image.

33. A white light source is placed in a denser medium. For a ray of white light, shown in the diagram below, the angle of incidence equals the critical angle for green colour light. Which colours pass into rarer medium? Which colours are reflected back? Explain.

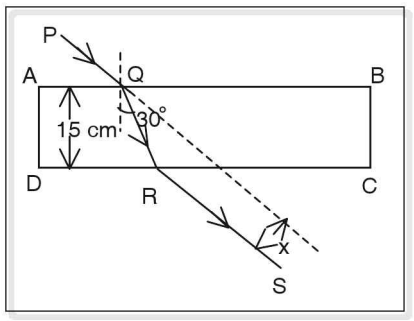


34. Two thin biconvex lenses L_1 and L_2 having focal lengths 10 cm and 15 cm respectively are placed coaxially. What distance (d) should be maintained between them so that a parallel beam of light parallel to their axis, incident on the lens L_1 emerges from L_2 undeviated? If the distance between them is increased by 'd' at what distance from L_1, on the axis, does the emergent beam converge?

35. An ant moves with a uniform speed along a straight line path from P to Q, 60 cm apart, in front of a mirror and takes two minutes to cover this distance. The line segment PQ is inclined to the mirror MM' at an angle of 30°. What is the rate at which the distance between the ant and its image changes?



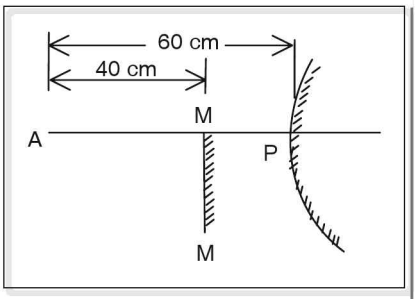
36. ABCD represents a glass slab and PQRS indicates the path of a light ray passing through the slab. If the refractive index of the glass is $\sqrt{3}$, find 'x'.



37. When an object can be seen with one eye, why is it that we have a pair of eyes?

38. A lens of focal length 20 cm produces a real image of a stationary object having magnification 2. Find the distance by which the lens should be moved so that it produces a virtual image of the same magnification.

39. An object A is placed on the principal axis of a convex mirror at a distance of 60 cm in front of it. A plane mirror is inserted between the object and the convex mirror at a distance of 40 cm from the object with the reflecting surface of the plane mirror facing the object. If the images formed by the two mirrors coincide, find the focal length of the convex mirror.



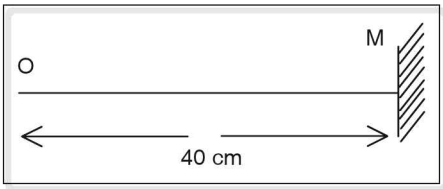
40. If an object is placed at a distance of 20 cm from the pole of a concave mirror, the magnification of its real image is 3. If the object is moved away from the mirror by 10 cm, find the decrease in its image magnification.

41. Why are red coloured lights are used in traffic signals to stop the vehicles?

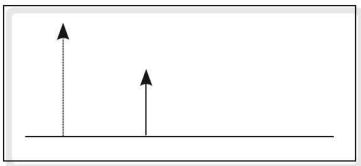
42. A convex lens forms a real image on a screen when an object is placed at 10 cm from the screen. On shifting the lens by 2 cm, a real image is again formed on the screen. Find the focal length of the lens.

43. A bird flying above a pond starts moving vertically down towards the water below at a speed of v m s^{-1}. Find its apparent velocity as viewed by a fish located at a depth of d m below the surface of water.

44. An object (O) and a plane mirror (M) are placed in the positions as shown in the figure. If the object and the mirror move simultaneously towards right in a straight path at uniform speeds of 2 cm s^{-1} and 5 cm s^{-1} respectively, find the shift in the image of the object at the end of 10 seconds.



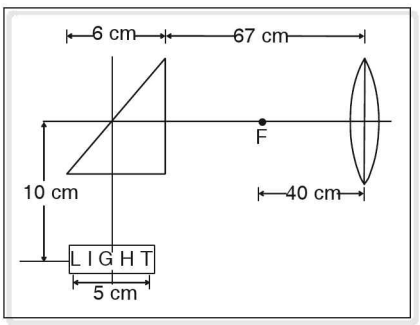
45. An object and its image are as shown in the diagram below.



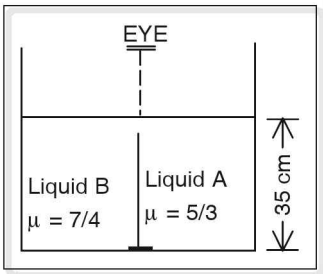
If the object image distance is 4 cm and the magnification is 3, find the ratio u: v : f. What is the type of the lens used? Draw the ray diagram to show the position of the lens and the principal focii.

Concept Application Level—3

46. An isosceles right angled prism and a convex lens (focal length = 40 cm) are placed as shown in the adjoining figure; and a 5 cm wide strip with the letters 'LIGHT' written on it is placed as indicated. What is the nature and the position of the image formed after refraction through the prism and the lens. Given the critical angle of the glass used for the prism and the lens is 42°.



47. A circle is painted at the bottom of a trough exactly in line with the interface of two liquids as shown in the figure. The trough is divided into two parts by a very thin glass sheet. Viewed along the interface from above, the circular ring would appear to be split into two halves.



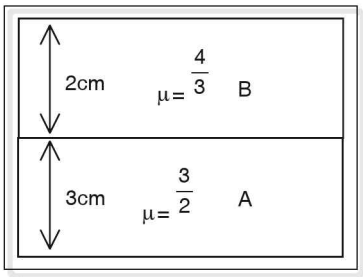
(i) Give reasons for the apparent split in the circular ring.

(ii) Find the difference in the levels of the two images.

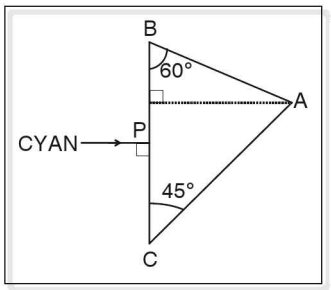
(iii) How can the two parts of the ring be made to appear at the same level?

48. A point object is placed at a distance of 10 cm from a convex lens of focal length 5 cm. A convex mirror is placed at 4 cm from the lens on the side of the lens opposite to the object such that the center of curvature of the convex mirror coincides with the image formed by the convex lens. Show the position of the new image with the help of a ray diagram. Find the focal length of the convex mirror.

49. The figure shows an arrangement of two transparent slabs A and B with their respective thickness and refractive indices as shown. Find the vertical shift of the image of a point object placed at the bottom of A due to both the slabs. Also find the effective refractive index of the compound slab.



50. The refractive index of the glass prism shown in the figure is 2 for green light. A ray of cyan colour light is incident at P on the face BC of the prism. From which side would the light emerge? Given the critical angle for blue colour is greater than 15°.



Very short answer type questions

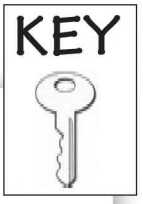
1. Reciprocal of focal length
2. Convex
3. Amount of radiant visible energy emitted per seconds
4. wavelength
5. Ratio of image distance to object distance
6. The returning of excited electrons back to their normal states.
7. Phase difference is 0 or $2n\pi$
8. 0°
9. Convex
10. Convex lens (converging)
11. Overlapping of two waves when they meet in same part of medium
12. 3
13. Bending of light when passed from rarer to denser medium
14. Hypermetropia due to ageing.
15. Difference in λ and ν
16. cones
17. When light enters from rarer to denser medium.
18. total internal reflection
19. Ability of light to get retracted when passed from one medium to another.
20. cylindrical
21. Ratio of sin i to sin r
22. candela
23. One lumen of luminous flux emitted for one second
24. Inversely
25. 10
26. Rayleigh scattering
27. 0°
28. Angle of prism, angle of incidence, μ, colour of light
29. quantum, electromagnetic
30. Colours obtained by mixing primary colours

Short answer type questions

31. Refraction though lens
32. $m = \frac{f_o}{f_e}$
33. Consists bright lines against a dark back ground.
34. (i) Definition of critical angle.
    (ii) $i > C$
35. View distant object on the earth's surface.
    Image erecting lens
36. Photo electrons
37. Defect in cornea, no proper focussing, cylindrical lens
38. Right angled isosceles prism
39. Surgery, holography, industry, communications, isotopes, etc.
40. Particle nature (light corpuscles)
    Reflection, Refraction different sizes of the corpusescles.
41. Invisible radiation, more wavelength, produces heat, sun. Missile system infrared photography etc.
42. Rotation through 180°
43. Wave nature
    - Longitudinal mechanical waves
    - Primary source, secondary source
    - Wave fronts (spherical, cylindrical)
    - Reflection, refraction, dispersion
44. Scattering of light by liquids
    Incoherent scattering.
45. (i) Thin glass fibre
    (ii) Works on principle of total internal reflection.

Essay type questions

46. Laser exhibits more directionality, coherence, intensity, monochromacity etc, than ordinary light.
47. Objective, eye piece
    $m = m_o \times me = L/fo \times D/fe$
    Ray diagram, find the values of L, fo, fe and obtain magnification (m).
48. (i) How to obtain ray diagram?
    (ii) How are f, u and v related?
49. Total internal reflection.
50. Ray diagram, similar triangles, ratio of proportional sides.



Concept Application Level—1

True or false

1. False
2. False
3. True
4. True
5. True
6. False
7. True

Fill in the blanks

8. black
9. high monochromatic nature.
10. glass slab
11. 3 feet
12. 15 m
13. 2
14. –4D

Match the following

15. A : d
    B : g
    C : i
    D : h
    E : a
    F : j
    G : b
    H : f
    I : c
    J : e

Multiple Choices

16. Choice (4)
17. Choice (2)
18. Choice (4)
19. Choice (1)
20. Choice (3)
21. Choice (1)
22. Choice (3)
23. Choice (3)
24. Choice (4)
25. Choice (4)
26. Choice (4)
27. Choice (4)
28. Choice (2)
29. Choice (4)
30. Choice (2)

Concept Application Level—2, 3

Key points

31. (i) Consider two cases where angle of incidence is 0° and 90°.

    Then the angle of refraction (r_1) at first face will be 0° and C respectively, where C is the critical angle.

    Find the corresponding angles of refraction (r_2) at other face of the prism.

    Find the angle of the prism when angle of refraction r_2 is less than critical angle and above critical angle.

    $A = r_1 + r_2$.

    Get the condition, when $\angle A < C$ and $\angle A > 2C$.

    Then find what happens when $\angle A$ is in between C and 2C.

    (ii) 80°

32. (i) Find the focal length of the compound lens.

(ii) Take the focal lengths of the convex and the concave lenses as positive and negative respectively.

(iii) Find whether the compound lens is a converging one or a diverging one.

(iv) Use the lens formula to find the position of the image.

(v) infinity

33. (i) When white light passes through a prism, it undergoes dispersion.

Here violet, indigo and blue colors are more deviated than green, yellow, orange and red colours ($\mu_{violet} > \mu_{red}$).

When white light ray passes from denser medium to rarer medium, and the angle of incidence is equal to the critical angle for green colour light, how does green light emerge out of the interface separating the two media?

From $\mu = \frac{1}{\sin C}$, find which colours will be refracted and which colours will be reflected into the same medium.

(ii) Red, orange and yellow pass into rarer medium; blue, indigo and violet reflect back into denser medium.

34. (i) Is the emergent beam a parallel beam parallel to the axis?

(ii) What should be the position of the object to get the image at infinity?

(iii) Does the first image formed serve as an object to the lens L_2?

(iv) 25 cm, 74 cm

35. (i) Find the value of uniform speed of the ant.

Shift the mirror horizontally and coincide it with the point Q.

Apply the trigonometric formula.

Find the distance of the ant from the mirror.

The object distance is equal to the image distance.

Then find the distance between the ant and its image, when ant is at P and Q.

Find the difference between the distance of the ant and the image, when present at P and Q.

Divide the obtained distance by the time taken.

Then, from this find the rate at which the distance between the ant and its image changes.

(ii) 0.5 cm s^{-1}

36. (i) Find angle of incidence i from $\mu = \frac{\text{Sin i}}{\text{Sin r}}$.

(ii) Find the angle of deviation, d = i – r.

(iii) Find the lateral shift using the formula for it.

(iv) $5\sqrt{3}$ or 8.66 cm

37. With one eye closed and looking with one eye, try to estimate the distance of an object. What difficulties do you have?

38. (i) Find the nature of the lens with the help of given information.

(ii) Use lens formula.

(iii) Find the object distance in the two cases and then find the distance moved by the lens.

(iv) 20 cm

39. (i) Find the object-distance with respect to the convex mirror.

Is the object-distance and the image-distance equal in the case of a plane mirror?

Find the distance between the plane mirror and the convex mirror.

Find the distance between the convex mirror and the image formed in the plane mirror.

Determine the image-distance for the convex mirror.

Find the focal length of the convex mirror by using the mirror formula $\left(\frac{1}{f}=\frac{1}{v}+\frac{1}{v}\right)$.

(ii) 30 cm

40. (i) Using mirror formula, express magnification 'm' in terms of focal length 'f' and object distance 'u'.

(ii) Use the expression to find the magnification in both the cases.

(iii) Find the new object distance and then find the magnification.

(iv) Difference in magnification gives the answer.

(v) 2

41. What is Reyleigh's scattering principle?

In all the seven colours, which colour deviates the least?

Which colour has more wavelength?

From Rayleigh's scattering formula,

$I_{max} = \frac{1}{\lambda^4}$, find whether red colour scatters more or less?

Can a least scattering colour be used as danger sign in traffic signals?

42. (i) Case 1:

Take the object distance from convex lens as 'x' cm.

Then, the image distance will be (10 – x) cm.

Substitute the values in lens formula $\frac{1}{f}=\frac{1}{v}-\frac{1}{u}$ and find the value of focal length (f) of the lens in terms of 'x'. ---- (1)

Case 2:

When the lens is shifted by 2 cm, take the object distance from convex lens as (x+2) cm.

Find the image distance from the lens. (10 – x – 2)

Substitute the values in lens formula $\frac{1}{f}=\frac{1}{v}-\frac{1}{u}$ to find the focal length (f) of the lens in terms of 'x' ---- (2). (use sign convention)

Is the focal length of a lens constant for a given medium?

Compare (1) and (2), and find x. Then, find the value of 'f'.

(ii) 2.4 cm

43. (i) Which is the observer – bird or fish?

Which is the object?

What is the speed of the bird?

What is the distance travelled by the bird in one second?

If the real distance of the bird is RD_1 (x + d) and after one second the real distance of the bird becomes RD_2 from the free surface of the water, where RD_2 = (x – v + d), then find the relation between RD_2 and RD_1.

What is the relation between real depth, apparent depth and refractive index?

What is the apparent depth of the bird as viewed by the fish?

Apparent velocity is equal to the rate of change in apparent depth.

Then find the value of apparent velocity of the bird as viewed by the fish.

(ii) μv

44. (i) Represent the positions of the object, mirror and the image along X- axis.

(ii) Difference in the final and initial positions of the image.

(iii) 80 cm

45. (i) From the given problem, the distance between the object and image is 4 cm.

The distance between object and image = x = 4 cm.

Find the image distance, v.

v = u + x. ---- (1)

What is the formula for magnifraction?

$\left(m = \frac{v}{u}\right)$ ---- (2)

Find value of u, v from (1) and (2)

Find the focal length (f) of the lens by using lens formula,

$\frac{1}{f} = \frac{1}{v} - \frac{1}{u}$.

To find the type of lens used, draw a straight line passing through the two arrow tips and draw another straight horizontal line from tip of arrow of the smaller line.

Draw the refracted ray and produce backward.

In the ray diagram, locate the lens, object and image

(ii) u : v : f = 2 : 6 : 3

46. Find the critical angle of the glass used for prism.

Find the angle of incidence of light on prism.

Will the letter LIGHT undergo total internal refection in an isosceles right-angled prism at hypotenuse side of the triangle?

Are the emergent light rays from the prism parallel?

Find whether the object distance can be taken as infinity.

Find the focal length of the convex lens from given information.

Find the image distance for convex lens.

Find the nature of the image formed due to convex lens.

47. (i) $\mu = \frac{\text{Real depth}}{\text{apparent depth}}$

(ii) 15 cm, 1 cm

15 : 14

48. (i) What is the value of 'u' and 'f' for convex lens?

(Follow sign convention.)

What is the distance between convex lens and convex mirror?

Find the value of the image distance, 'v' for convex lens by using the lens formula, $\frac{1}{f} = \frac{1}{v} - \frac{1}{u}$.

Here, v = 4 + 2f (where f = focal length of convex mirror).

Find the value of 'f'.

(ii) 3 cm

49. (i) $\Delta h = \Delta h_1 + \Delta h_2$.

(ii) $\Delta h = h\left(1 - \frac{1}{\mu}\right)$.

(iii) 1.5 cm

50. Find the refractive index of the glass prism for green light from the given information.

What colours are present in cyan?

Find the critical angle for green colour light, using the formula $\mu = \frac{1}{\sin C}$.

Bending power of blue colour light is more than that for green colour. Then will the critical angle for blue be less or more?

At point P, find the angles of incidence and refraction.

Find the angle of incidence of the light rays on the face AC with the help of a ray diagram.

Find whether it undergoes total internal reflection.

Then, will the light ray incident on the face AB?

Find the angle of incidence on the face BC.

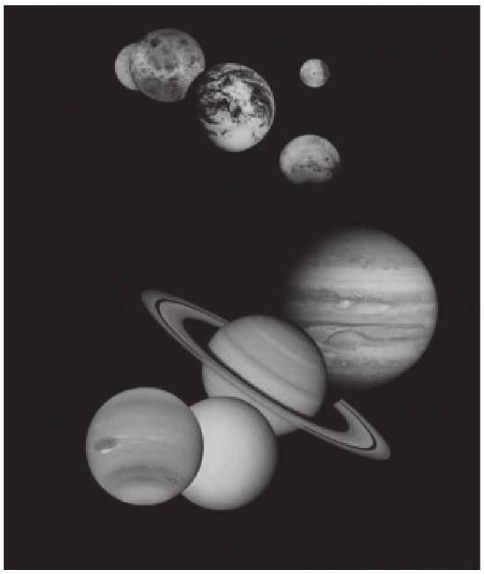
Is this angle equal to the critical angle for green colour?

Then, find whether green colour grazes out from the face AB.

From geometry, show that blue colour emerges from the face AC.

5

Hydrostatics



INTRODUCTION

We observe that a balloon filled with hydrogen gas goes up but a balloon filled with air falls down. A small iron nail sinks in water, but huge ships made of iron float. A stone weighs less in water than in air. To seek answers to these questions we need to understand how liquids and gases at rest react to external forces. This involves the study of a branch of physics called hydrostatics. In previous classes you have studied many laws, terms, etc. associated with hydrostatics.

Thrust and Pressure

Thrust

Force acting normally on a body is called thrust.

For example, the weight exerted by a book placed on table is normal to its surface. Hence it is called thrust. Since thrust is a force, its units are same as those of force.

Units of thrust:

S.I. – newton (N)

C.G.S. – dyne

Gravitational units:

S.I. kg_f or kg_{wt}

C.G.S. g_f or g_{wt}

Thrust is a vector quantity.

The dimensional formula of thrust is $[M^1L^1T^{-2}]$.

Pressure

Consider a force of 20 N acting normally on a surface whose area is 2 m^2. Thrust acting on a unit area is $= \frac{20\ N}{2\ m^2} = 10\ N\ m^{-2}$

We say that pressure of 10 N m^{-2} acts on the surface.
Thus, pressure is thrust acting on a unit area.

$\therefore$ Pressure $= \frac{\text{Thurst}}{\text{Area}}$

i.e., $P = \frac{F}{A}$

Units of pressure:

S.I. unit: N m^{-2}

1 N m^{-2} is called pascal (Pa) in honour of the French Scientist Blaise Pascal. Pressure is said to be 1 pascal if a force of 1 N acts normally on a surface of area 1 m^2.

C.G.S. unit of pressure is dyne cm^{-2}:

Gravitational unit is $kg_f\ m^{-2}$ or $kg_{wt}\ m^{-2}$

1 kg_f = 9.8 N

$\therefore$ 1 $kg_f m^{-2}$ = 9.8 N m^{-2} = 9.8 Pa

$$1\ Pa = 1\frac{N}{m^2} = \frac{1\times10^5 dyne}{(10^2)^2 cm^2}$$

$$= \frac{1\times10^5 dyne}{10^4 cm^2}$$

= 10 dyne cm^{-2}

Pressure is a scalar quantity.

The dimensional formula of pressure is $[ML^{-1}T^{-2}]$.

Pressure in fluids

Any substance which can flow and has no definite shape is called fluid. Thus liquids and gases together are referred to as fluids.

Solids exert pressure on a surface on which they rest due to their weight. Liquids and gases also have weights. Hence they also exert pressure. However, fluids exert pressure in all directions.

A fluid enclosed in a vessel exerts pressure at all points and in all directions. Fluid molecules are in constant random motion, due to which they collide with the walls of the container. In doing so, their momentum changes in a very short interval of time. This acts as thrust on the walls due to the fluid molecules.

The thrust on a unit area at a point gives rise to fluid pressure at that point. Since molecules move randomly, on an average equal number of molecules move in each direction. Thus fluid pressure at any point inside it is same in all directions. The pressure exerted by a fluid on the walls of the container is called lateral pressure.

Mathematical expression for pressure in fluids

Consider a vessel containing a liquid of density 'd' and a coin of area of cross section 'a' placed at the bottom. A liquid column of height 'h' and area of cross section 'a' forms imaginary cylinder above the coin. This column exerts its weight on the coin giving rise to liquid pressure.

Volume of imaginary liquid column = area of cross section × height

$V = ah$

Mass of liquid column = volume × density

$m = Vd$

Substituting for V,

$m = ahd$

Weight of the liquid column = mass × acceleration due to gravity

$\therefore mg = ahdg$

$\therefore$ thrust exerted by the liquid column = $ahdg$

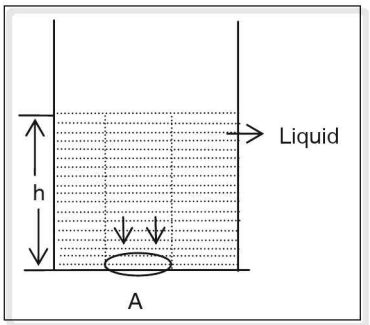


Figure 5.1

$$\text{Pressure due to liquid column (P)} = \frac{\text{Thurst}}{\text{Area}}$$

$$P = \frac{ahdg}{a}$$

$$P = hdg$$

Thus, pressure at a point inside a liquid is proportional to the depth of the point below the free surface of the liquid and its density.

☛ *Example*

Calculate the pressure at the bottom of a tank due to a water column of height 200 m. (Take g = 10 m s^{-2})

Solution

Given

h = 200 m

d = 10^3 kg m^{-3} (density of water)

g = 10 m s^{-2}

P = hdg

P = $200 \times 10^3 \times 10 = 2 \times 10^6$ N m^{-2}

= 2×10^6 Pa

☛ *Example*

A liquid column of vertical height 1.2 m exerts a pressure of 19992 Pa at the bottom. Find the density of liquid (Take g = 9.8 m s^{-2}).

Solution

$P = hdg$

$\therefore\ d = \frac{p}{hg}$

$= \frac{19992}{1.2 \times 9.8} = 1700 \text{ kg m}^{-3}$

Factors on which the pressure at a point in a fluid depends

The pressure at a point inside a liquid is directly proportional to the following three factors

1. The depth of the point below the free surface or the height of the fluid column above it.
2. The density of fluid.
3. The acceleration due to gravity.

The pressure at a point in fluid does not depend on

1. The shape and size of the vessel in which the fluid is contained.
2. The area of the surface on which it acts.

Let P_o be the atmospheric pressure acting on the liquid surface. The total pressure acting at a point inside the liquid is

P = atmospheric pressure + pressure due to the liquid column
$P = P_o + hdg$

☛ *Example*

A rectangular plate 10 cm × 5 cm is placed horizontally below the water surface in a trough at a depth of 2 m. If the atmospheric pressure is 1.013×10^5 N m^{-2}, calculate the total thrust on the plate.

Solution

Pressure at a point 2 m below the water surface = atmospheric pressure + pressure due to the water column of height 2 m.

Pressure due to water = hdg
$= 2 \text{ m} \times 10^3 \text{ kg m}^{-3} \times 10 \text{ m s}^{-2}$
$= 0.2 \times 10^5 \text{ N m}^{-2}$

Total pressure = $(1.013 \times 10^5 + 0.2 \times 10^5)$ N m^{-2}
$= 1.213 \times 10^5$ N m^{-2}

Area of the plate = 10 cm × 5 cm = 50 cm^2 = 50×10^{-4} m^2

Total thrust acting on the plate = total pressure × area = $1.213 \times 10^5 \times 50 \times 10^{-4} = 606.5$ N.

Laws of Liquid Pressure

Consider a piston fixed to an air tight cylinder which contains a spring (from its base to piston). When the external pressure on the piston increases, it is pushed into the cylinder. The spring fixed between the piston and the base of the cylinder helps the piston recover its original position as soon as the external pressure is removed. Insert this cylinder in a container containing a liquid. As depth increases, the piston is pushed more into the cylinder.

1. The pressure exerted by the liquid increases as we go deeper in the liquid.
2. Select a point at a particular depth in the liquid. Place the cylinder at this depth and rotate it by 360°. We find that the depth of the piston inside the cylinder remains the same. This shows that the pressure at a point in a liquid is the same in all directions.
3. Move the cylinder along the same horizontal level at a particular depth. We find that the depth of the piston inside the cylinder remains same. This shows that the pressure is the same at all points in a horizontal plane in a stationary liquid.
4. If the cylinder is immersed to the same depth in another liquid with higher density, the piston gets compressed more.

This shows that the pressure at a point in a liquid increases with increase in the density of the liquid.

Take an apparatus consisting of a number of tubes of different shapes and sizes connected together at the bottom as shown in figure (5.2).

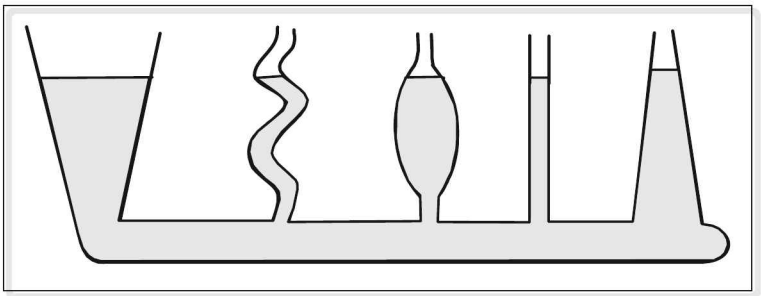


Figure 5.2

Pour a liquid through any one of the tubes. We find that the liquid level in all the tubes is same. This shows that a liquid seeks it own level.

Following is one of the consequences of liquid pressure.

1. The wall of a dam is made thicker at the bottom. The pressure exerted by the water increases with depth. At greater depth, higher pressure is exerted by the water on the wall of the dam. Therefore the thickness of wall is more at the bottom to prevent it from collapsing. (Figure 5.3).

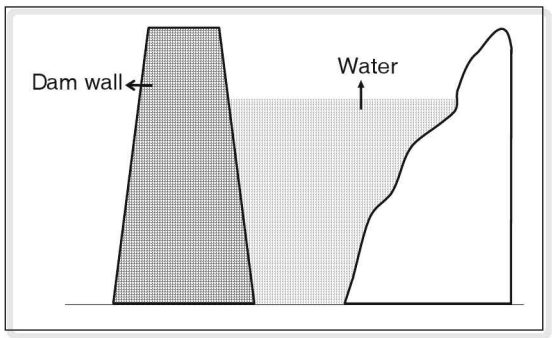


Figure 5.3

☛ *Example*

A rectangular container of dimensions 50 cm × 30 cm × 100 cm is filled completely with water. Determine

1. the thrust at the bottom of the vessel,

2. the pressure at a depth of 10 cm from the free surface and
3. the force experienced by a metal sheet held vertically at a depth of 10 cm.

Solution

Volume of the container $= l \times b \times h$

$$= \frac{50}{100}\text{m} \times \frac{30}{100}\text{m} \times \frac{100}{100}\text{m}$$

$$= 0.15 \text{ m}^3$$

Thrust at the bottom of the container = weight of water in the container

$$= mg = \text{volume} \times \text{density} \times g$$

$$= 0.15 \text{ m}^3 \times 10^3 \text{ kg m}^3 \times 10 \text{ m s}^{-2}$$

$$= 0.15 \times 10^4 \text{ N}$$

2. Pressure at a point inside a liquid is given by,

$$P = hdg$$

$$= \frac{10}{100} \times 1000 \times 10 = 1000 \text{ N m}^{-2}$$

3. The net force acting on the metal sheet will be zero because the lateral thrust exerted by the water on each face of the sheet is equal and opposite.

Transmission of pressure in liquids—Pascal's law

When pressure is applied at a point on a confined fluid, it is transmitted to other parts of fluid. For example, if you press a balloon filled with air or water at a point, it bulges out at some other point. This shows that a liquid can transmit pressure. Pascal summarised this in a statement called Pascal's law.

Pascal's Law

Pressure applied anywhere in a confined fluid is transmitted equally and undiminished in all directions throughout the fluid.

Consider a vessel as shown in the figure (5.4).

It is filled with water and has four identical water tight pistons P, Q, R and S. Let a force of 1 N be applied to piston P which moves inwards. The remaining three pistons are found to move outwards. To prevent them from moving out, 1 N force has to be applied to each one of them. This clearly shows that pressure applied at one point has been transmitted equally and undiminished in all directions.

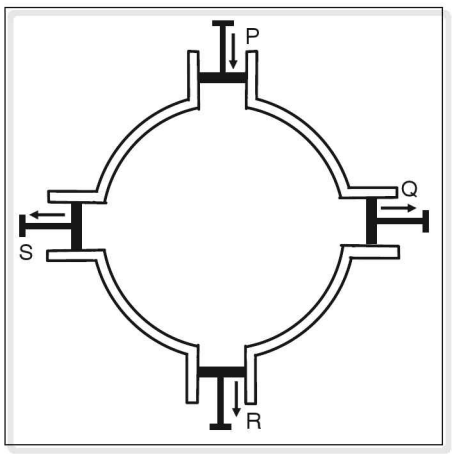


Figure 5.4

Applications of pascal's law

The working of hydraulic machines such as hydraulic press, hydraulic jack and hydraulic brakes is based on Pascal's law.

Principle of hydraulic machine

Consider two cylindrical vessels P and Q connected by a horizontal tube R. The vessels are filled with water or any other liquid and fixed with air tight pistons M and N respectively. Let A_1 and A_2 be the area of cross section of the two pistons respectively such that $A_2 > A_1$. A weight is kept on piston M which exerts a force F_1 on it. This force is acting on area A_1 hence the pressure P_1 due to it is

$$P_1 = \frac{F_1}{A_1} \times A_2 \quad \text{--- (1)}$$

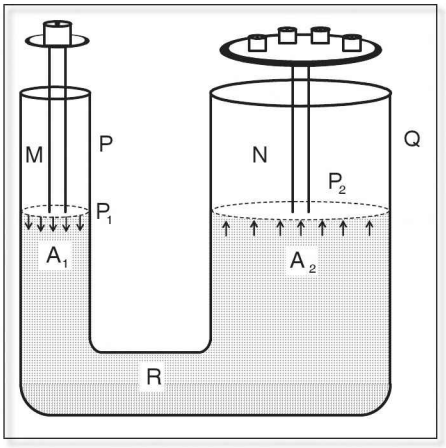


Figure 5.5

According to Pascal's law, this pressure is transmitted equally and undiminished throughout the liquid. Hence the pressure acting on the piston N in upward direction is P_1 and the piston moves up. To prevent it from moving up a force F_2 should be applied to it in downward direction by putting certain weight on it.

F_2 = pressure on N × area of cross section

$$F_2 = P_1 \times A_2$$

$$= \frac{F_1}{A_1} \times A_2 \quad \text{(from equation (1))}$$

$$\frac{F_1}{A_1} \times A_2$$

Since $A_2 > A_1$

$\therefore F_2 > F_1$.

Thus a small thrust applied to the smaller piston can be used to exert a larger thrust on the bigger piston.

Examples of hydraulic machines

1. Bramah press (hydraulic press)

Its working is based on Pascal's law.

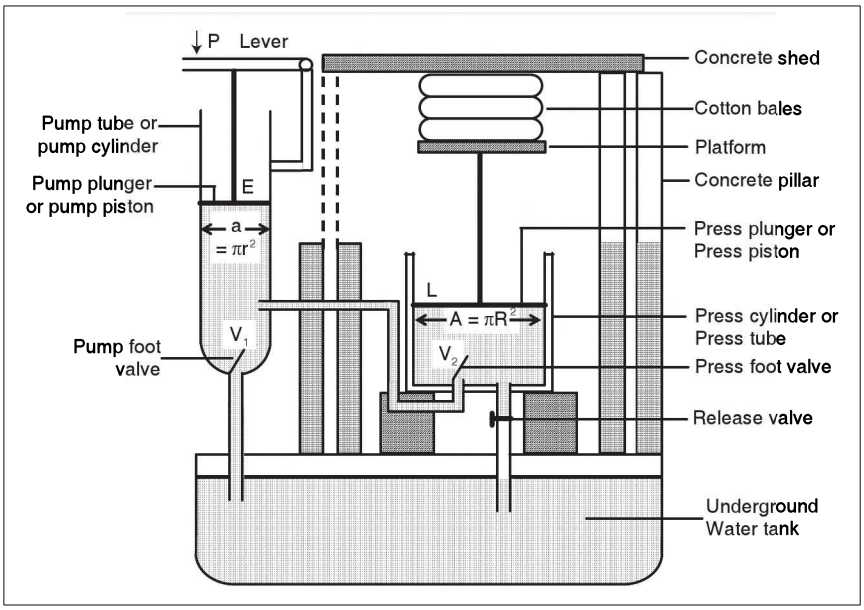


Figure 5.6 Hydraulic press

Uses of bramah press

It is used in the following applications

1. For pressing books, cotton bales, etc.
2. For squeezing the oil out of oil seeds
3. For punching holes in metals
4. For extracting juice of sugar cane, sugar beet, etc.

☛ *Example*

The ratio of area of cross section of the two pistons in a hydraulic machine is 1 : 15. Find the force that should be applied on the narrow piston to overcome a force of 120 N on the bigger piston.

Solution

$A_1 : A_2 = 1 : 15$

$F_2 = 120$ N

$F_1 = ?$

$$\frac{F_1}{A_1} = \frac{F_2}{A_2} \quad \text{---------} \quad \text{(Pascal's law)}$$

$$F_1 = F_2 \times \frac{A_1}{A_2} = \frac{120}{15}$$

$F_1 = 8$ N

2. Hydraulic brakes

To stop a heavy vehicle moving with high velocity, a large force has to be applied. These vehicles have hydraulic brakes which are hydraulic machines working on Pascal's law.

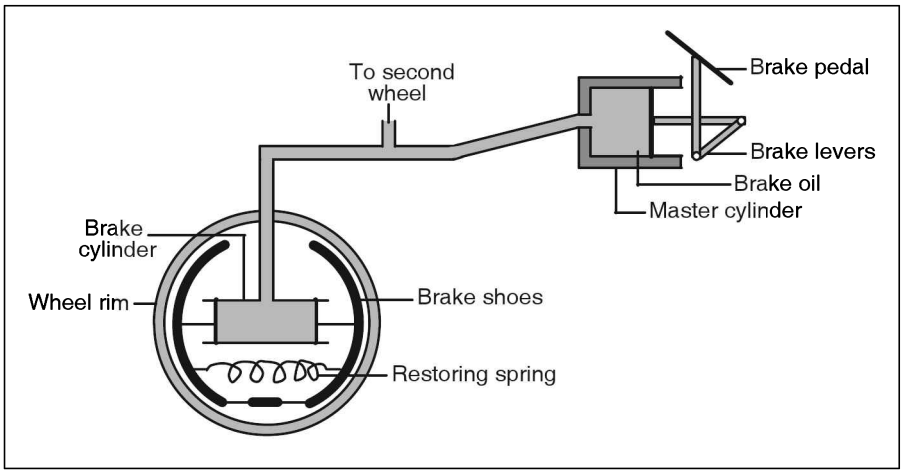


Figure 5.7

Atmospheric Pressure

Our earth is surrounded by air to a height of about 300 km. The envelope of air around the earth is called atmosphere. Atmosphere exerts thrust on the earth due to its weight. This thrust acting on a unit area of earth is called atmospheric pressure. Considering the extent of the atmosphere, the pressure exerted by it is huge. Its average magnitude is approximately 10×10^3 kg_f m^{-2} or 10^5 Pa in SI system or 10^3 g_f cm^{-2} in CGS system.

That atmosphere exerts its pressure on the earth can be demonstrated by the following experiment. Take a tin can and half fill it with water. Heat the can till water begins to boil. The steam formed drives away the air in the can. Now close the can with air tight stopper. Stop heating it and pour some cold water over the can. As the steam inside it condenses to water, the pressure inside the can decreases. As large atmospheric pressure is acting on it from outside, we find that the can gets crushed when cold water is poured over it.

Some consequences of atmospheric pressure

1. We are able to consume soft drinks and juices using a straw due to the atmospheric pressure.
2. Hand pump or lift pump can draw under ground water making use of the atmospheric pressure.

Measurement of atmospheric pressure

A device used to measure atmospheric pressure is called a barometer. Three types of barometer are used commonly.

They are:

1. Simple barometer or Torricellian barometer,
2. Fortin's barometer and
3. Aneroid barometer.

Simple barometer

It was first constructed by Italian scientist Torricelli. The atmospheric pressure is measured in terms of height of the mercury column taken in a glass tube. Pressure due to any liquid column is given by
$P = hdg$

At a given place and for a given liquid, acceleration due to gravity, 'g' and the density of the liquid, 'd' are constant. Therefore pressure can be expressed in terms of the height of a liquid column.

A simple barometer consists of a glass tube of uniform narrow bore of length 1m filled completely with dry and pure mercury. This tube is inverted into a trough containing pure mercury.

Depending upon the atmospheric pressure acting on the mercury in the through, some mercury from the tube flows into the trough. The empty space created at the top of the tube is called torricellian vacuum. At the sea level it is found that the length of the mercury column in the tube is approximately 76 cm.

The pressure exerted by 76 cm of mercury is taken as
1 atmospheric pressure.
1 atm = 76 cm of Hg = 0.76 m of Hg
$= 0.76 \text{ m} \times 13.6 \times 10^3 \text{ kg m}^{-3} \times 9.8 \text{ m s}^{-2}$
(density of mercury $= 13.6 \times 10^3 \text{ kg m}^{-3}$)

$= 1.01292 \times 10^5$ Pa

1 atm = 1.013×10^5 Pa

Other units which are used to measure the atmospheric pressure are

1 torr = 1mm of Hg

1 bar = 10^5 Pa

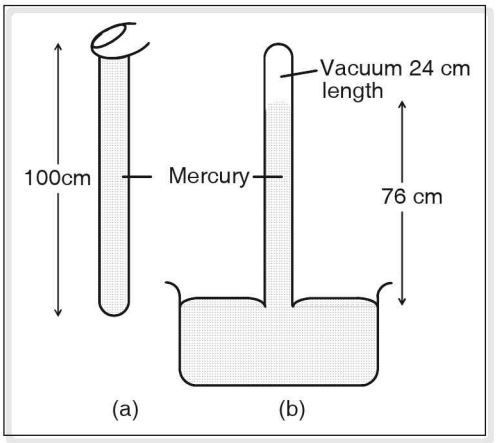


Figure 5.8 Simple barometer

The vertical height of the barometric liquid

1. is independent of the size and shape of the tube and
2. remains unaffected even if the tube is tilted.

The factors which affect the barometric height are

1. the presence of moisture in mercury
2. the presence of impurities in mercury
3. the altitude of a place above the sea level
4. temperature
5 humidity

Mercury is the most preferred barometric liquid as its density is very high. Hence the height of mercury column even at the sea level is only 76 cm. If water is used as a barometric liquid, the height of water column would be approximately 10.34 m. If the pressure due to water and mercury columns is $h_w d_w g$ and $h_{Hg} d_{Hg} g$ respectively, then $h_w d_w g = h_{Hg}\, d_{Hg}\, g$ (since both water and mercury columns are supported by the same atmospheric pressure).

$$\frac{h_w}{h_{Hg}} = \frac{d_{Hg}}{d_w}$$

$$\therefore\ h_w = \frac{13.6 \times 10^3 \times 0.76\text{m}}{1000}$$

$10.336 \cong 10.34$ m

Defects of simple barometer

1. The free surface of mercury in the trough does not always coincide with the zero mark of the scale attached to the barometric tube. Mercury from the tube flows into the trough and vice versa depending on the changes in the atmospheric pressure and temperature. Hence the tube has to be raised or lowered whenever the reading of the atmospheric pressure is to be taken.
2. The barometer is exposed to the surroundings. The mercury can get contaminated, which affects the barometric readings.
3. The mercury can spill over when the barometer is moved from one place to another.
4. It is quite bulky.
5. It is difficult to keep the glass tube vertical.

In order to overcome these defects, Fortin's barometer is used.

Fortin's barometer

It consists of a brass enclosure which contains the glass test tube inverted into a glass vessel having a leather cup. Both are filled with dry and pure mercury. At the bottom of brass enclosure, there is a screw. By adjusting this screw the leather cup can be raised or lowered and the mercury level in the glass vessel is made to be in level with zero mark. An ivory pointer is fitted in the brass tube such that when its tip just touches the mercury level in the vessel, the mercury level would correspond to the zero mark. A vernier scale is also provided at the top of the tube for accurate readings.

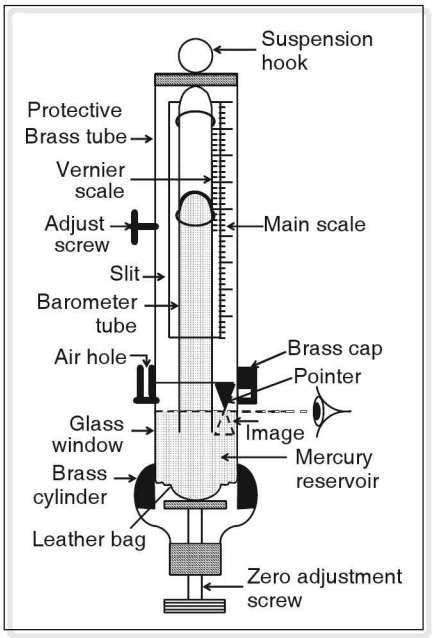


Figure 5.9 Fortin's barometer

Aneroid barometer

It contains no liquid. It is light and portable and can be placed in any orientation.

It consists of an evacuated box with a diaphragm at the top. At the centre of the diaphragm toothed rod is attached. A circular toothed wheel fits into the teeth of the rod so that as the atmospheric pressure changes, the diaphragm may get pressed down or raised which lowers or raises the rod. The toothed wheel in turn rotates, and the needle pointer pivoted to it deflects on a circular scale calibrated to give atmospheric pressure.

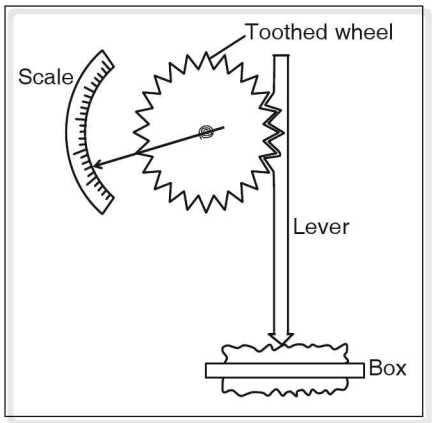


Figure 5.10 Aneroid barometer

Manometer

It is a device used to measure the pressure of gas in a container. It consists of a U -shaped glass tube of uniform cross section filled with a liquid. The ends of the U-tube are open. When exposed to the atmosphere, the liquid level in both the arms is same. One arm of the tube is connected to the gas container. If the pressure exerted by the gas is less than the atmospheric pressure, the level of the liquid in that arm goes up as shown in figure 5.11(b).

If P_o = atmospheric pressure

P = pressure exerted by the gas

$P + hdg = P_o$

where h = difference in liquid levels in the two arms, and

d= density of the liquid

$\therefore P = P_o - hdg$

If $P > P_o$, then $P = P_o + hdg$ (figure 11 c).

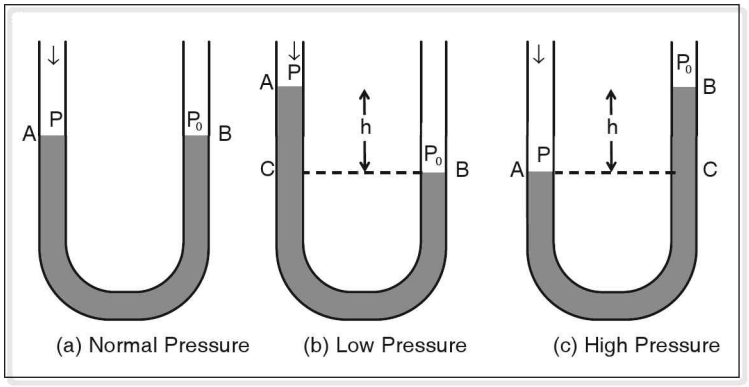


Figure 5.11 Manometer

Variation of atmospheric pressure with altitude

The atmospheric pressure decreases with increase in the altitude from the sea level due to two factors.

1. As the altitude increases, the height of the air column exerting the atmospheric pressure decreases
2. The density of air at higher altitude is less than the density of air at the surface of the earth.

Taking the average density of air as 1.29 kg m^{-3}, we can prove that near the surface of the earth 105 m of air column on an average exerts the same pressure as 1 cm of mercury.

The variations in the atmospheric pressure with altitude is shown in the following graph.

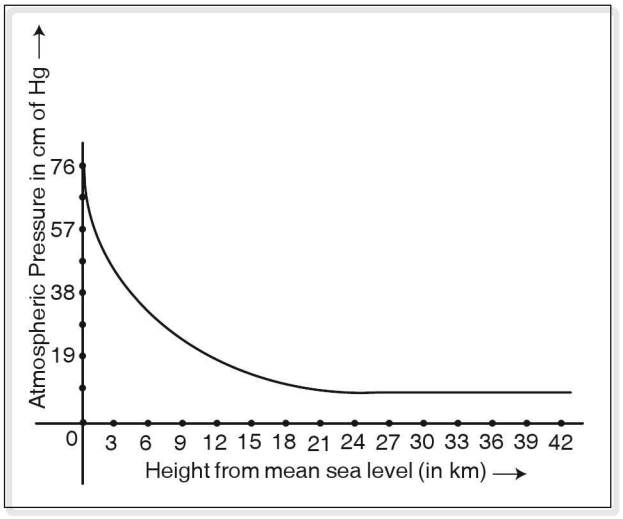


Figure 5.12 Variation of atmospheric pressure with altitude

Altimeter

A barometer measuring atmospheric pressure can also be used to determine the altitude. It is called altimeter. An aneroid barometer whose scale is calibrated in terms of altitude above sea level is commonly used as altimeter.

Upthrust

When a solid is immersed completely in a liquid, it experiences liquid pressure from all directions. The thrust due to pressure on its sides is called lateral thrust, on the top it is called downward thrust and at the bottom upward thrust.

Figure(5.13) shows pressure exerted by a liquid on a body of area of cross section 'A' and height 'h'. The top and base of the body are at the depth h_1 and h_2 respectively such that $h_2 - h_1 = h$

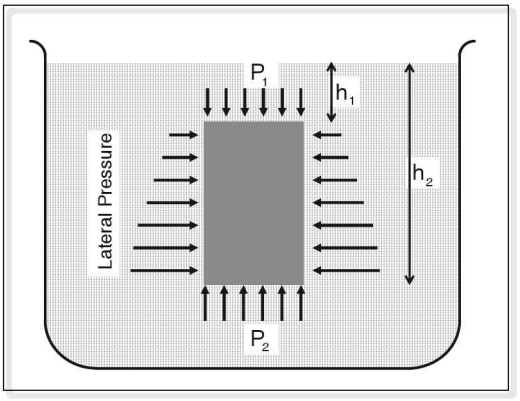


Figure 5.13 Upthrust

From the figure,

$P_1 = h_1 d_\ell g$

$P_2 = h_2 d_\ell g$

where d_ℓ = density of the liquid

$P_2 > P_1$ as $h_2 > h_1$

Downward thrust acting on the top surface = $AP_1 = Ah_1 d_\ell g$

Upward thrust acting on the bottom surface = $AP_2 = Ah_2 d_l g$

$\therefore$ Net thrust (F) acting on the body is in upward direction, and is given by

$F = Ah_2d_lg - Ah_1d_lg$

i.e., $F = A \times (h_2 - h_1)\, d_lg$

$= Ahd_lg \quad (\because h_2 - h_1 = h)$

$= Vd_lg \quad (\because Ah =$ volume (V) of the body inside the liquid)

volume of the body inside the liquid is also equal to the volume of the liquid (V_l) displaced by it

$\therefore F = V_ld_lg$

$\therefore$ This net force acting on the body due to the fluid in upward direction is called upthrust or buoyant force.

$F = m_lg$ ($\because V_ld_l =$ mass of the liquid displaced by the immersed part of the body)

$F = W_l$ ($\because m_lg =$ weight of the liquid displaced)

Thus, the upthrust acting on a body immersed partially or completely in a liquid is equal to the weight of the liquid displaced by it.

Factors affecting the upthrust

The factors on which the upthrust acting on a body, either partially or completely immersed in a liquid depends are

1. volume of the liquid displaced which is equal to the volume of the immersed part of the solid, and
2. density of the liquid in which the solid is immersed.

☛ *Example*

An iron rod of volume 100 cm^3 and density 7.8×10^3 kg m^{-3} is completely immersed in water. Calculate the upthrust acting on it. (Take g = 10 m s^{-2}, density of water = 10^3 kg m^{-3})

Solution

Upthrust (F) $= V_ld_lg$

V_l = volume of water displaced by iron rod

Hence V_l = volume of the rod

$\therefore V_l = 100\ cm^3 = 10^{-4}\ m^3$

$\therefore F = 10^{-4}\ m^3 \times 1000\ kg\ m^{-3} \times 10\ m\ s^{-2} = 1$ N.

Archimedes' principle

When a body is immersed completely or partially in a liquid, it is acted upon by two forces.

1. weight of the body, w acting vertically down and
2. upthrust due to liquid acting vertically up.

As a result of these opposing forces, the body appears to weigh less in the liquid than in air.

The apparent loss in weight of the body when immersed either partially or completely in a liquid is equal to the upthrust acting on it. The more the upthrust, the more is the apparent loss in weight of the body.

This fact was discovered by the famous Greek scientist Archimedes and is known as the Archimedes' principle which is stated below.

"When a body is immersed completely or partially in a liquid, it experiences an apparent loss in weight, which is equal to the weight of the liquid displaced by it".

Archimedes' principle is applicable to bodies immersed in liquids as well as in gases. However, the upthrust due to the atmosphere is negligible since the density of air is far less than that of a solid, and for all practical purposes we may consider the weight of the body in air as its true weight. Thus the difference in the weights of a body in air and in liquid gives the magnitude of the upthrust acting on it, i.e., the weight of the liquid displaced.

☛ *Example*

A body weighs 250 g in air and 235 g when completely immersed in water. Calculate

(a) the apparent loss in weight of the body and

(b) buoyant force acting on it.

Solution

Let W_1 = weight of the body in air = 250 g_f

W_2 = weight of the body in water = 235 g_f

Apparent loss in weight = $W_1 - W_2$

= 250 g_f – 235 g_f = 15 g_f

The apparent loss in weight of a body is also equal to the upthrust acting on it.

$\therefore$ upthrust = 15 g_f

☛ *Example*

A cube of a metal whose density is 8.2 g cm^{-3} and side 7 cm is tied to a thread and completely immersed in a liquid of density 1.2 g cm^{-3}. Calculate the tension in the string.

Solution

Tension in the string = net downward force acting on the cube.

Volume of the cube = 7 cm × 7 cm × 7 cm = 343 cm^3

Mass of the cube = Volume × density

= 343 cm^3 × 8.2 g cm^{-3}

= 2812.6 g

Weight of the cube = 2812.6 g_f

This weight acts in the downward direction

Upthrust = Volume of cube × density of liquid × acceleration due to gravity

$= 343 \text{ cm}^3 \times 1.2 \text{ g cm}^{-3} \times \text{g cm s}^{-2}$

$= 411.6 \text{ g}_f$

$\therefore$ tension in the string $= 2812.6 - 411.6$

$= 2401 \text{ g}_f$

Relative density

Density of a substance is defined as the ratio of its mass and volume.

$$d = \frac{m}{v} \text{ kg m}^{-3} \text{ or g cm}^{-3}$$

Relative density (R.D.) of a substance is the ratio of the density of the substance to the density of water at 4°C.

R.D. is also called specific gravity.

$$\text{R.D.} = \frac{\text{Density of the substance } (d_s)}{\text{density of water } (d_w) \text{ at } 4°C}$$

$$= \frac{\text{mass of unit volume of the substance}}{\text{mass of unit volume of water at } 4°C}$$

$$= \frac{\text{mass of the substance } (m_s)}{\text{mass of equal volume of water at } 4°C \ (m_w)}$$

$$\text{R·D.} = \frac{\text{weight of the substance } (W_s)}{\text{weight of equal volume of water at } 4° \text{ C } (W_w)}$$

Let W_1 = weight of the substance in air

W_2 = weight of the substance when it is completely immersed in water

$W_1 - W_2$ = apparent loss in weight of the substance in water.

= Weight of the water displaced

Here the volume of the water displaced is equal to the volume of substance

$$\text{Thus, relative density} = \frac{\text{Weight of the body in air}}{\text{Apparent loss of weight in water}}$$

$$\therefore \text{ R.D. of substance} = \frac{W_1}{W_1 - W_2}$$

Relative density being a ratio of same physical quantities, has no unit.

☛ *Example*

The R.D. of iron is 7·8. Calculate its density in S.I. system of units.

Solution

$$\text{R.D.} = \frac{\text{density of iron}}{\text{density of water}}$$

$\therefore$ density of iron = R.D. $\times$ density of water

$= 7.8 \times 1000 \text{ kg m}^{-3}$

$= 7.8 \times 10^3 \text{ kg m}^{-3}$

☛ *Example*

A solid weighs 60 g in air and 52 g when completely immersed in water. Calculate the volume and relative density of the solid (Take density of water = 1 g cm^{-3}).

Solution

Loss in weight of solid = 60 g_f – 52 g_f = 8 g_f

$\therefore$ Weight of the water displaced = 8 g_f

i.e., mass of water displaced = 8 g

$$\text{Volume of the water displaced} = \frac{8 \text{ g}}{1 \text{ g cm}^{-3}} = 8 \text{ cm}^3$$

Volume of the solid is equal to the volume of water displaced by it.

$\therefore$ volume of the solid = 8 cm^3

$$\text{R.D. of solid} = \frac{\text{weight of solid in air}}{\text{loss of weight in water}} = \frac{60 \text{ g}_f}{8 \text{ g}_f}$$

R.D. = 7·5

Relative density by balancing columns

The pressure at a point inside a liquid is given by

$P = h\, d\, g$

Let us consider columns of two liquids whose densities are d_1 and d_2. Consider two points in them at the depth of h_1 and h_2 respectively such that pressure due to the liquid at these points is same i.e., $P_1 = P_2$.

$\therefore h_1 d_1 g = h_2 d_2 g$

$$\frac{d_2}{d_1} = \frac{h_1}{h_2}$$

$d_2 = \frac{h_1}{h_2} \times d_1$. If the first liquid in water then, let

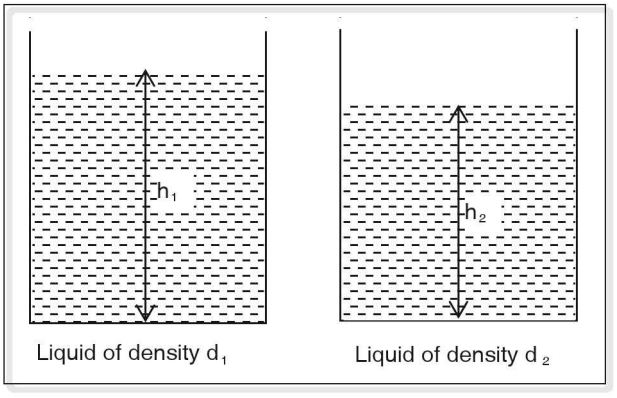


Figure 5.14

d_1 = density of water = 1000 kg m^{-3}

$\therefore$ density of other liquid is

$$d_2 = \left(\frac{h_1}{h_2} \times 1000\right) \text{ kg m}^{-3} \text{ in S.I. system}$$

or

$$d_2 = \frac{h_1}{h_2} \times 1 \text{ g cm}^{-3} \text{ in CGS system}$$

Thus, $\frac{d_2}{d_1} = \frac{h_1}{h_2}$

If d_1 is the density of water at 4°C,

$$\text{R.D. of liquid} = \frac{h_1}{h_2}$$

This leads us to the principle of balancing columns which states that when two liquid columns of different densities and heights balance each other, the pressure exerted by them is equal.

The Hare's apparatus which is used to find relative density of liquid works on the principle of balancing liquids.

Hare's apparatus

It consists of an inverted, U-shaped glass tube of uniform section. The top of glass tube has an outlet which is connected to a rubber tube provided with pinch cock. The uniform glass tube is fixed to a vertical stand with the help of clamps such that it can be easily raised or lowered. The two ends of the U-tube are immersed in two beakers kept on a horizontal stand. One of the beakers contains water while the other contains liquid whose specific gravity is to be determined. A vertical scale is fixed between the two arms of the U-tube so that the liquid levels in the beaker and the arms of the U-tube can be measured. The pinch cock is removed and air is sucked out from the rubber tube. The liquids rise in the arms of the U-tube. The pinch cock is replaced again.

The levels a_2 and a_1 of water in the beaker and in the tube respectively are noted. Difference in these two levels $(a_1 - a_2)$ gives the height of the water column in the tube. Let it be h_1. Similarly, $h_2 = b_1 - b_2$ gives the height of the liquid column in the second tube.

Applying the principle of balancing columns

$$\text{Relative density of the liquid} = \frac{\text{height of the water column}}{\text{height of the liquid column}}$$

$$\text{R.D.} = \frac{h_1}{h_2}$$

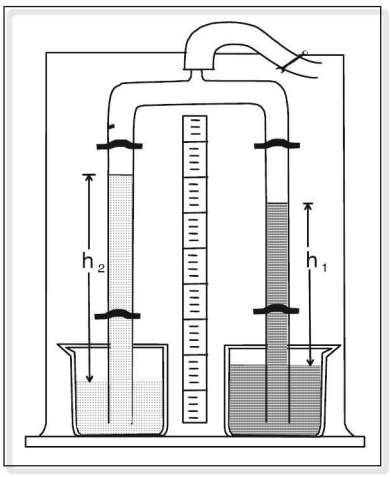


Figure 5.15 Hare's apparatus

☛ *Example*

Level of water in the tube and beaker in the Hare's apparatus are 40·4 cm and 5·3 cm and corresponding levels of liquid are 35 cm and 5 cm. Find the specific gravity of the liquid.

Solution

a_1 = level of water in the tube = 40.4 cm
a_2 = level of water in the beaker = 5.3 cm
$h_1 = a_1 - a_2 = 40.4 - 5.3 = 35.1$ cm
b_1 = level of liquid in the tube = 35 cm
b_2 = level of liquid in the beaker = 5 cm
$h_2 = b_1 - b_2 = 35 - 5 = 30$ cm

$$\text{R.D.} = \frac{h_1}{h_2} = \frac{35.1}{30} = 1.17$$

☛ *Example*

Using Hare's apparatus, R.D. of a liquid was found to be 1.25. If the height of the liquid in one of the tube was 32 cm, find the height of the water in the another tube.

Solution

R.D = 1.25
$h_2 = 32$
$h_1 = ?$

$$\text{R.D} = \frac{h_1}{h_2}$$

$$1.25 = \frac{h_1}{32}$$

$$h_1 = 32 \times 1.25 = 40 \text{ cm}$$

☛ *Example*

In a Hare's apparatus, the levels of the water in the tube and the beaker are 38.3 cm and 4.7 cm respectively. The specific gravity of liquid is 1.05 and its level in the tube is 35.5 cm. Find its level in the beaker.

Solution

a_1 = 38.3 cm
a_2 = 4.7 cm
$h_1 = a_1 - a_2 = 38.3 - 4.7 = 33.6$ cm
b_1 = 35.5 cm
b_2 = ?

$$R.D = \frac{h_1}{h_2} = \frac{a_1 - a_2}{b_1 - b_2}$$

$$1.05 = \frac{33.6}{35.5 - x}$$

$$(35.5 - x) = \frac{33.6}{1.05} = 32$$

$\therefore x = 35.5 - 32 = 3.5$ cm

Floatation

Consider a body immersed in a liquid.
Two forces act on it simultaneously. They are

1. The weight (W) of the body acting vertically down at its centre of gravity (C.G)
2. The upthrust or buoyant force (F_B) equal to the weight of the liquid displaced acting vertically upwards.

The point where the total upthrust due to the liquid displaced by the immersed part of the body acts is called the centre of buoyancy (B).

The weight of the body acts vertically down whereas the buoyant force acts vertically up. Depending upon their magnitudes, following situations arise.

1. $W > F_B$

   If the weight of the body is greater than the weight of the liquid displaced, the net force acting on the body is in the downward direction. Hence the body sinks

   $W = V\,d\,g$

   where V = volume of the immersed part of the body in the liquid

   d = density of the body.

   $F_B = Vd_l\,g$

   V = volume of the liquid displaced by the immersed part of the body

   d_l = density of the liquid

   $W > F_B$

   $Vdg > Vd_lg$

   $\therefore d > d_l$

   Thus a body sinks in a liquid if its density is greater than that of the liquid.

2. $W = F_B$

   If the weight of the body immersed in a liquid is equal to the weight of the liquid displaced by it, then the net force acting on the body is zero. Hence the body remains in static equilibrium wherever it is placed inside the liquid.

Thus the condition for a body to be in static equilibrium when completely immersed in a liquid is

$W = F_B$

$Vdg = Vd_1 g$

$d = d_1$

3. $W < F_B$

If the weight of the body immersed in a liquid is less than the weight of the liquid displaced by it, the net force acting on the body is in the upward direction. Thus the body moves up in the liquid till some part of it emerges out of the liquid surface. As the volume of the body inside the liquid now decreases the upthrust also decreases, until the upthrust acting on the body is equal to the weight of the body. In this condition the body is in equilibrium and it floats on the liquid surface with certain fraction of it below the surface. Such a body is called a floating body.

Therefore, the condition for floating is

$W < F_B$

$Vdg < Vd_1 g$

$d < d_1$

Thus a body floats in a liquid if its density is less than that of the liquid.

From the above discussion it is clear that only those bodies whose weight is less than the weight of the liquid displaced when completely immersed in the liquid, rise upto the liquid surface. However, when such bodies float in a liquid, their weight acting vertically down is equal to the buoyant force acting on them vertically up.

This forms the law of floatation, which can be stated as below.

"For a solid floating in a liquid, its weight acting vertically down at its centre of gravity is equal to the weight of the liquid displaced by the immersed part of the solid acting vertically up at its centre of buoyancy".

In the floating condition, the apparent weight and the apparent density of the solid are zero and the body is said to be weightless.

Laws of floatation

1. The weight of the floating body is equal to the weight of the liquid displaced. (Net force acting on the body is zero)
2. The centre of gravity of the floating body and the centre of buoyancy lie in the same vertical line. (Net turning effect is zero)

Mathematical expression for the law of flotation

Consider a wooden cylinder floating in a liquid of density d_1. Consider

V = total volume of the wooden cylinder

h = total height of the cylinder

V_1 = volume of the part of the cylinder inside the liquid

h_1 = height of the cylinder inside the liquid

a = area of cross section of the cylinder

d = density of the liquid

d_l = density of the cylinder.

Mass of cylinder $m = V \times d$

$= a\ h\ d$

Weight of cylinder $W = mg = ahdg$

Volume of liquid displaced $V_1 = ah_1$

Mass of liquid displaced $= V_1 \times d_l$

weight of liquid displaced $= ah_1\ d_l g$

By the law of flotation,

weight of the cylinder = weight of the liquid displaced by the cylinder

$ah\ dg = ah_1\ d_l g$

$hd = h_1\ d_l$

$$\frac{d}{d_\ell} = \frac{h_1}{h}$$

If the liquid used is water, the relative density $= \dfrac{h_1}{h} = \dfrac{\text{length of the cylinder in the water}}{\text{total length of the cylinder}}$

Also,

$Vdg = V_1\ d_l\ g$

$$\therefore \frac{d}{d_\ell} = \frac{V_1}{V}$$

for water $d = \dfrac{V_1}{V} = \dfrac{\text{volume of the cylinder in the water}}{\text{total volume of the cylinder}}$

Stability of floating bodies

Consider a stationary ship floating in water. Since it is floating it satisfies the following two conditions.

1. The weight of the ship is equal to the weight of the water displaced by the immersed part of the ship.
2. The centre of gravity of the ship and the centre of buoyancy (B), which is the C.G. of the displaced water, lie on the same vertical line.

This vertical line which connects the C.G. of ship and centre of buoyancy (B) is called the centre line. (Figure 5.16).

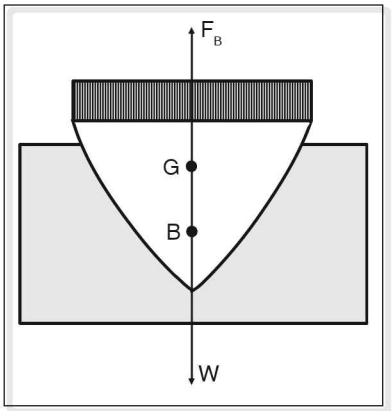


Figure 5.16

If the ship is caught in turbulent waters or storm, it may tilt. Now the second condition mentioned above will not be satisfied as the position of 'B' changes to B_1. In fact a vertical line from B_1 intersects the centre line at a point. This point is called the meta centre (M).

The weight of the ship and the upthrust acting on it through the new centre of buoyancy are in opposite directions. Their lines of action being parallel, they set up a couple. This couple can be either clockwise or anti-clockwise depending on the position of M with respect to the C.G (G). This determines the state of equilibrium of the ship.

1. **Stable equilibrium:** If 'G' lies below M, i.e., if the ship is tilted in the anti-clockwise direction, the couple set up is in the clockwise direction which restores the ship to its original position. (Figure 5.17).

   Stable equilibrium is possible if the base of the ship is made heavy, which lowers its C.G.

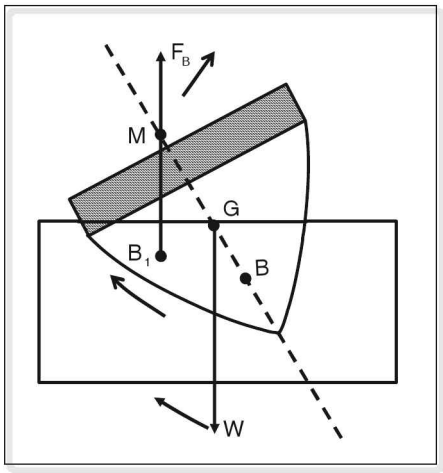


Figure 5.17

2. **Unstable equilibrium:** If 'G' lies above M, and if the ship is tilted in the anti-clockwise direction the couple set up is also in anti clockwise direction which increases the tilt of ship and eventually sinks it. This condition is said to be unstable equilibrium. (Figure 5.18).

   This condition of 'G' lying above 'M' is possible if the ship is heavy at the top. To avoid this the bottom of the ship is usually filled with sand bags and weights called ballast. There is a line marked on the lateral side of the ship called 'water line' and every ship has to float up to this line for safe sailing and stability.

3. **Neutral equilibrium:** If 'G' and 'M' coincide as shown in the (figure 5.19) even when the ship is tilted anti-clockwise, it is not acted upon by any couple. In this condition the ship is said to be in neutral equilibrium as it neither tilts further nor does it come back to its original position. (Figure 5.19).

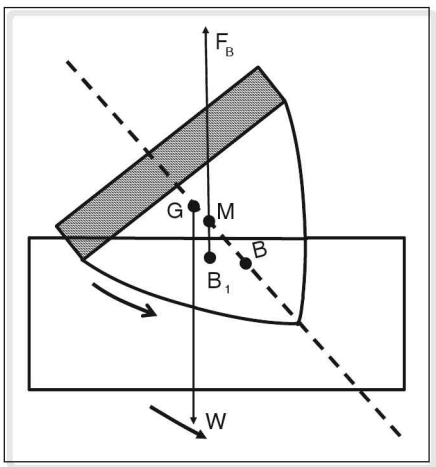


Figure 5.18

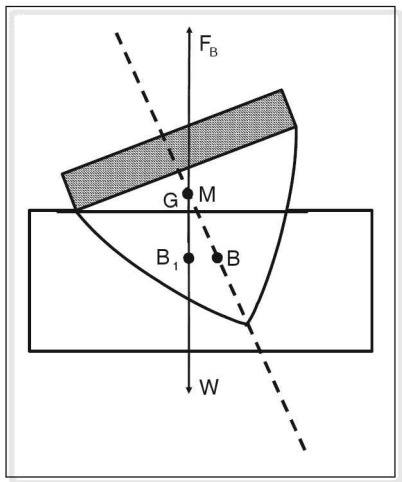


Figure 5.19

☛ *Example*

The densities of water and ice are 1 g cm^{-3} and 0·917 g cm^{-3}. Find the fraction of an iceberg floating above the water surface and the fraction below it.

Solution

Let V = the total volume of the iceberg.

V_1 = volume of the iceberg above the water surface

$\therefore V - V_1$ = volume of the iceberg below the water surface.

By the law of floatation, weight of the iceberg = weight of the water displaced by the immersed part of the iceberg.

$V \times 0.917 \times g = (V - V_1) \times 1 \times g$

$V_1 = (1 - 0.917)V$

$$\frac{V_1}{V} = \frac{0.083}{1}$$

$$= \frac{1}{12} \text{ (approx)}$$

Thus $\frac{1}{12}$th part of the total volume of the iceberg floats above the water surface.

$\therefore$ $\frac{11}{12}$th part of iceberg is submerged in water.

☛ *Example*

A block of wood floats on water with $\frac{2}{5}$th of its volume above the surface. If it is made to float in brine solution of R.D. 1.20, what fraction of the wood is below the surface of the brine solution?

Solution

Let V = Volume of the block

d = density of the block

1 g cm^{-3} = density of the water

volume of the block immersed = $V - \frac{2}{5}V$

$$= \frac{3}{5}V$$

By the law of floatation

$$Vdg = \frac{3}{5}V \times 1 \times g$$

$$\therefore d = \frac{3}{5} = 0.6 \text{ g } cm^{-3}$$

Density of the brine solution = 1.20 g cm^{-3}

$\therefore V \times 0.6 \text{ g cm}^{-3} = V_1 \times 1.20 \text{ g cm}^{-3}$

$$V_1 = \frac{0.6}{1.20} V = 0.5 V$$

Thus 50% of total volume of the wooden block is submerged in the brine solution.

Plimsoll lines

Ships are used to transport passengers and cargoes. As load in ship is increased, it sinks more in water. Ship may sink completely, if loaded beyond a permissible limit. But this limit on load cannot be calculated as ship travels through waters of different densities. Consider a heavily loaded ship travelling through cold waters of North Atlantic sea. Due to the high density of the sea water there, the upthrust acting on the ship is high and the ship stays afloat. When it enters warm tropical fresh water, it will sink more due to the low density of the sea water there. Hence it is better to know how far a ship would sink in particular waters than to measure its load. To specify upto which level a ship may be allowed to sink or to determine the maximum load for a ship, Samuel Plimsoll in 1876 suggested that ships should be marked with load lines which would suggest how much a ship when loaded would sink in waters of a particular region. These lines are called Plimsoll lines. The line for tropical fresh water which has lowest density is at the top and the line for North Atlantic seas with water having maximum density is at the bottom A ship is loaded till it sinks to the appropriate line. This ensures that irrespective of its destination, it would float at the safe level.

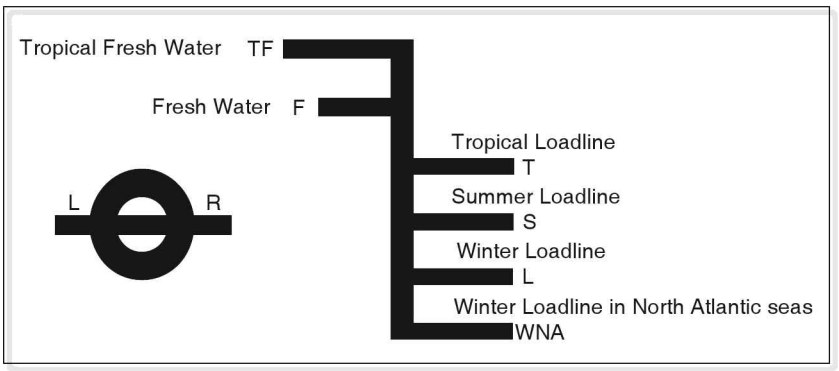


Figure 5.20 Plimsoll lines

Hydrometer

It is a device used to find relative density of a liquid. It works on the law of floatation. When a body floats in a liquid, its weight is equal to that of the liquid displaced.

There are two types of hydrometers. They are

1. variable immersion hydrometer, and
2. constant immersion hydrometer.

In a variable immersion hydrometer, the weight of the hydrometer is kept constant so that its depth of immersion varies depending on the density of liquid in which it floats. In a constant immersion hydrometer, the depth of immersion is kept constant in different liquids by changing its weight suitably.

The common hydrometer

It is a variable immersion hydrometer. It consists of a uniform narrow glass stem which terminates into a broader glass bulb also called floatation or buoyancy bulb. A gravity or stability bulb is attached to the glass bulb at the bottom. Gravity bulb is filled with mercury or lead shots so as to make the bottom heavier and give stability to the hydrometer so that it floats upright in a given liquid. The glass stem is graduated suitably to measure the density of the liquid.

The common hydrometer is of two types. Hydrometers that are used

1. to measure liquids having density less than that of water and
2. to measure liquids having density greater than that of water.

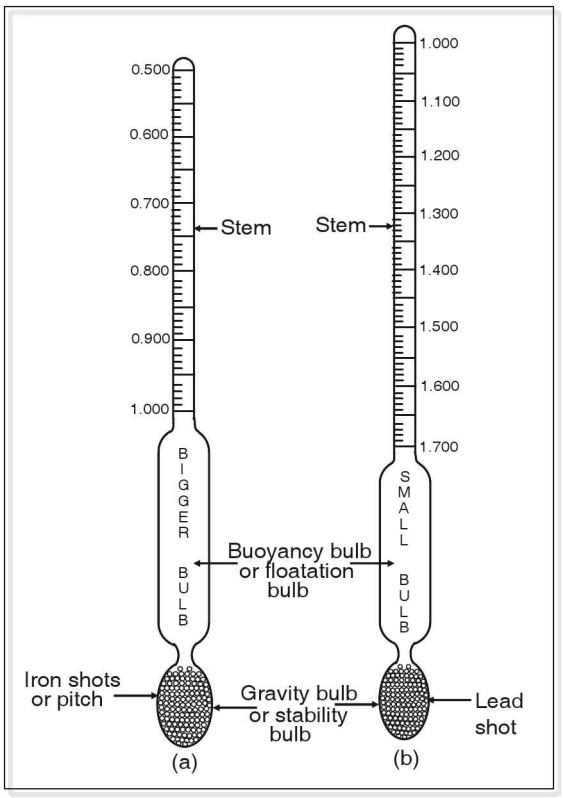


Figure 5.21 Hydrometer (a) Hydrometer for lighter liquids, (b) Hydrometer for heavier liquids

Uses of common hydrometers

Hydrometers are commonly used to measure densities of various liquids in pathological labs, research centres, factories, breweries, etc.

Some hydrometers are used to measure density of only one liquid. The lactometer, for example, is used to measure density of milk. If the density shown by lactometer is different from that of pure milk, the milk is said to be adulterated with water.

In a lactometer (figure 5.22), floatation and gravity bulbs are fused together and only one bulb is present. The top reading gives the density of water, whereas the bottom reading shows the density of pure milk (1.045 g cm^{-3}).

Acid battery hydrometer is used to measure the density of acids used in batteries.

Alcohol meter is used to measure relative density of alcohol. Hydrometers are also used to find the strength of brine solution and variations in density of waters of different seas and oceans.

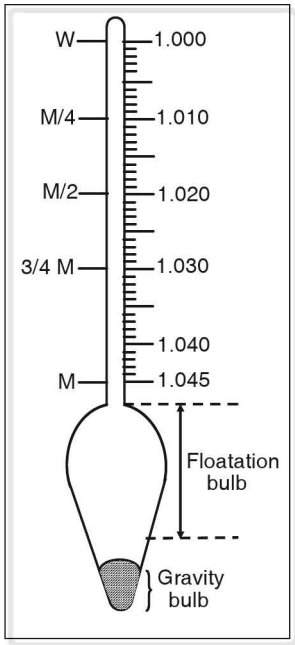


Figure 5.22 Lactometer

Test tube float

It consists of a flat bottomed uniform test tube. Its bottom is made heavy by adding mercury, lead shots or sand to it. This gives it stability. The test tube

is graduated in centimeters to measure the depth of immersion in liquids. Test tube float is used to find specific gravity of liquids and works on the law of floatation.

1. Test tube float as **variable** immersion hydrometer.

   In two similar tall jars pour water and a liquid whose R.D is to be determined to the same level. Tie a long twine near the mouth of the test tube float and lower it in water. Avoid touching the tube to the sides of beaker. Take care that air bubbles do not stick to the sides of the float. Note down the height h_1 of the tube inside the water.

   Remove the test tube out of water and wipe it dry with a clean cloth. Holding the twine immerse it in the liquid till it floats freely. Do not add or remove lead shots from the test tube.
   Note the height h_2 of tube inside the liquid.

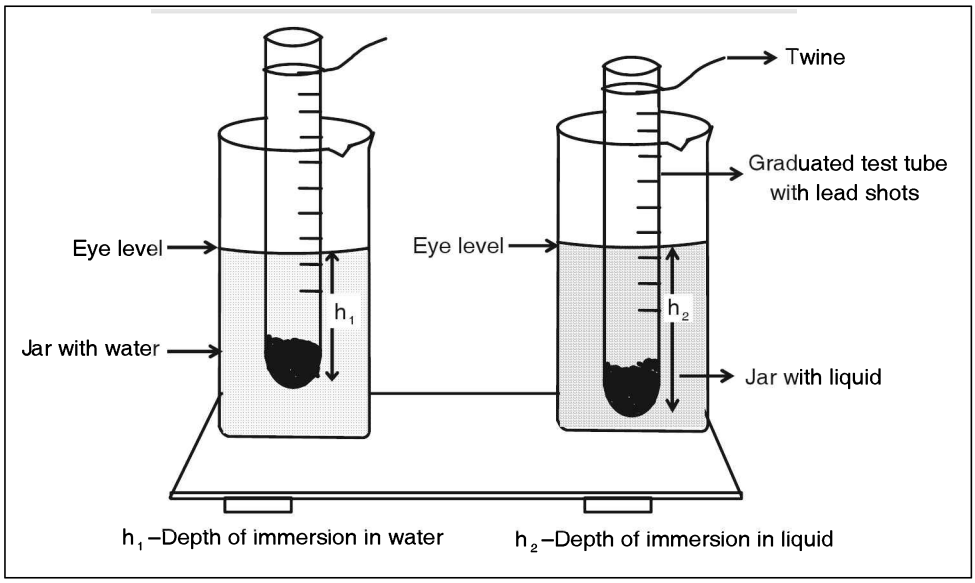


Figure 5.23 Test tube float as a variable immersion hydrometer

Since the weight of the float is same in both the cases, the weight of water displaced and weight of liquid displaced are equal according to the law of floatation.

Let weight of water displaced = ah_1d_1g

weight of liquid displaced = ah_2d_2g

Where, a = area of cross section of the test tube, d_1 and d_2 = density of water and liquid respectively.

According to the law of floatation

$ah_1d_1g = ah_2d_2g$

$$\therefore \frac{d_2}{d_1} = \frac{h_1}{h_2}$$

$$\text{R.D.} = \frac{h_1}{h_2}$$

$$\text{R.D. of liquid} = \frac{\text{depth of immersion of the float in water}}{\text{depth of immersion of the float in liquid}}$$

☛ *Example*

The depth of immersion of the test tube float in water is 16.8 cm and in a liquid it is 15.8 cm. Calculate the specific gravity of the liquid.

Solution

Depth of immersion in water, $h_1 = 16.8$ cm

Depth of immersion in the liquid, $h_2 = 15.8$ cm

The relative density of the liquid $= \frac{h_1}{h_2} = \frac{16 \cdot 8}{15 \cdot 8} = 1 \cdot 063$

☛ *Example*

Relative density of a liquid is 0.78 and depth of immersion of the float in it is 17.5 cm. Find the depth of immersion when the tube floats in water.

Solution

R.D. of the liquid = 0.78

Depth of immersion in liquid, $h_2 = 17.5$ cm

Depth of immersion in water, $h_1 = ?$

$$\text{R.D.} = \frac{h_1}{h_2}$$

$$0.78 = \frac{h_1}{17.5}$$

$$h_1 = 0.78 \times 17.5$$

$$h_1 = 13.65 \text{ cm.}$$

2. Test tube float as constant immersion hydrometer

The test tube float is immersed in water and its depth of immersion h is noted. The float is removed from water, wiped dry and suspended from the hook of a spring balance and its weight W_1 is noted. The float is then floated in a liquid whose R.D. is to be determined. By removing or adding lead shots, it is made to float to the same depth 'h' as in water. Its weight W_2 is measured after removing it from the liquid and wiping it dry.

The volume of water displaced and the volume of liquid displaced by the float is same since its depth of immersion in both is same. According to the law of floatation, the weight of a body floating in a liquid is equal to the weight of the liquid displaced by it.

$\therefore$ Weight of water displaced = W_1

Weight of liquid displaced = W_2

The volume of water and liquid displaced are equal.

$$\text{R.D. of the liquid} = \frac{\text{Weight of the liquid displaced}}{\text{weight of equal volume of water}}$$

$$= \frac{W_2}{W_1}$$

$$\therefore \text{ specific gravity of the liquid} = \frac{\text{weight of the float in liquid}}{\text{weight of the float in water}}$$

☛ *Example*

The weight of a test tube float in water is 43 g_{wt} and in a liquid is 44 g_{wt}. If the depth of immersion of the float in both is same, find the specific gravity of the liquid.

Solution

W_1 = 43 g_{wt} = weight of the float for immersion in water

W_2 = 44 g_{wt} = weight of the float for immersion to same level in liquid

$$\text{R.D. of liquid} = \frac{W_2}{W_1} = \frac{44}{43} = 1 \cdot 023$$

☛ *Example*

The R.D. of a liquid is 1.027 and a float of weight 38 g_{wt} is allowed to float upto a certain depth in it. What additional weight should be added to the float if it is to float to the same depth in water as in liquid?

Solution

Weight of the float W_2 = 38 g_{wt}

Relative density of the liquid = 1.027

Weight of the float when it floats to the same level in water = W_1

$$\text{R.D} = \frac{W_2}{W_1}$$

$$1.027 = \frac{38}{W_1}$$

$$W_1 = \frac{38}{1.027} = 37 \ g_{wt}$$

Since $W_1 < W_2$, additional weight need not be added to the float, but 1 g_{wt} of mass has to be removed from float.

test your concepts

Very short answer type questions

1. What is torricellian vacuum?
2. Hydraulic press works on the principle of _____.
3. What is a test tube float? What are the different types of test tube floats?
4. Apparatus used to determine the relative density of liquids is _____.
5. Define hydrostatics.
6. Define and write units of thrust and pressure.
7. Define fluids.
8. What do you understand by barometric liquid and barometric height?
9. Plimsoll lines are marked on ships to protect the ships from sinking into water due to _____.
10. An object weighs W_1 in air and W_2 when half of its volume is immersed in water. The relative density of the object is _____.
11. Write the expression for pressure exerted by a liquid. What do the terms in the expression stand for?
12. Pressure of the gas enclosed inside a cylinder can be measured using _____.
13. State Archimedes' principle.
14. A balloon of volume 'v' and density 'd' is moving in upward direction in air. If the density of air is 'D', then the effective upward acceleration of the body is
15. State Pascal's law.
16. Density of air at higher altitudes is _____ than that at the surface of earth.
17. Define centre of buoyancy.
18. Define metacentre.
19. Give the average magnitude of the atmospheric pressure in SI and CGS units.
20. Define stable, unstable and neutral equilibrium.
21. What is a hydrometer? Mention different types of hydrometers,
22. A cylindrical container is filled with water upto the brim. If the pressure exerted by the water at the bottom of the container is 100 Pa the height of the container is _____ cm (take $g = 10\ m\ s^{-2}$).
23. What is a barometer? Mention different types of barometers.
24. Why does Fortin's barometer give more accurate readings?
25. The upthrust acting on a body when immersed in two different liquids are 10 N and 15 N respectively. The ratio of densities of liquids is _____.
26. Define manometer.
27. Define altimeter.
28. Define upthrust.
29. Define relative density or specific gravity.

30. The ratio of the area of cross-sections of the smaller and bigger pistons of a hydraulic press is 1 : 4. When a 100 N force is applied on the smaller piston the number of weights each of weight 1 kg_{wt} that can be placed on the bigger piston is _____ ($g = 10\ m\ s^{-2}$).

Short answer type questions

31. Why are the walls of a dam broader at the base?

32. Obtain an expression for the upthrust experienced by a body when immersed in a liquid and state the factors affecting it.

Factors affecting the upthrust.

(1) volume of the immersed part of the body

(2) density of the liquid

33. Mention a few defects of a simple barometer.

34. Obtain the expression for pressure exerted by a liquid

35. Given some uses of Bramah press.

36. State the laws of floatation.

37. Why is mercury preferred as a barometric liquid?

38. What is lateral pressure? Why is the net lateral pressure acting on a body immersed in a liquid zero?

39. Mention some consequences of atmospheric pressure.

40. What are the advantages of an aneroid barometer over simple barometer?

41. State the forces acting on a body when it is immersed partially or completely in a liquid.

42. Enlist the factors on which the pressure at a point in fluids (liquids or gases) depends. Also, state the factors which do not affect it.

43. State the factors which affect the vertical height of a barometric liquid.

44. Write a note on variation in atmospheric pressure with altitude.

45. State laws of liquid pressure.

Essay type questions

46. Explain how Hare's apparatus is used to find R.D of a liquid.

47. Explain how a test tube float can be used as constant and variable immersion hydrometer.

48. What is relative density? Show that $R.D = \frac{W_1}{W_1 - W_2}$, where W_1 is the weight of the body in air and W_2 is its weight in water.

49. Write a note on manometer.

50. If w = weight of a body

F_B = buoyant force

Explain the three conditions which arise when the body is immersed in a liquid.

CONCEPT APPLICATION

Concept Application Level—1

Direction for questions 1 to 7: State whether the following statements are true or false.

1. A floating body is said to be in stable equilibrium if its metacentre coincides with the centre of gravity.
2. Hydraulic press works on the principle of Bramah's press.
3. The principle of working of hydrometer is based on 'law of floatation'.
4. The magnitude of pressure measured in SI system is greater than that measured in CGS system.
5. The upward thrust exerted on the body by the liquid acts vertically upwards through the centre of gravity of the displaced liquid.
6. The vertical height of the mercury column in a barometer remains unaffected even if the tube is tilted.
7. Hare's apparatus works on Archimedes principle.

Direction for questions 8 to 14: Fill in the blanks.

8. For greater mechanical advantage the diameter of the press piston of a hydraulic lift must be ____ the diameter of the pump piston.
9. A solid when immersed in water displaces $\frac{1}{2}$ litre of water. The same solid floats with $\frac{3}{4}$th of its volume below the surface of a given oil and displaces 800 g of oil. Then the relative density of body and oil is _____ and _____ respectively.
10. When a truck is loaded, heavy objects should be placed at _____.
11. The readings on a variable immersion hydrometer are marked in increasing order from _____ to _____ on its stem.
12. Hand pump lifts the underground water making use of _____.
13. The fraction of volume of a floating body inside water, gives _____ of the body.
14. The densities of two liquids are in the ratio of 1 : 3 and their volumes are in the ratio of 2 : 1. If the liquids are filled inside two identical containers, the ratio of pressure exerted by them at the bottom of the containers is _____.

Direction for question 15: Match the entries in column A with appropriate ones from column B.

15.

Column A		Column B	
A. Pascal's Law	( )	a.	altitude
B. Archimedes' principle	( )	b.	used to measure the hollowness of a body
C. Stability of a floating body	( )	c.	hydraulic brake
D. Test tube float	( )	d.	pressure of a gas inside a container

(Continued on following page)

E.	Liquids and gases	( )	e.	height of centre of gravity
F.	Barometer	( )	f.	less than or equal to the density of the liquid
G.	Manometer	( )	g.	variable immersion hydrometer
H.	Density of a floating body	( )	h.	atmospheric pressure
I.	Altimeter	( )	i.	fluids

Direction for questions 16 to 30: For each of the questions, four choices have been provided. Select the correct alter native.

16. If a block of wood is floating in a river water, then the apparent weight of the floating block is
(1) equal than weight of the displaced water
(2) zero
(3) greater than the weight of the displaced water
(4) equal to the actual weight of the block

17. When the left arm of a mercury manometer is connected to a cylinder filled with a gas the level of the mercury in the right arm rises by 2 cm. If the pressure of the gas in the container is 110160 Pa, the atmospheric pressure is _____ cm of Hg. (take g = 10 m s^{-2})
(1) 79 (2) 85 (3) 77 (4) 81

18. A test tube of mass 'm' sinks up to mark 'X' in water and the same test tube sinks below the mark 'X' in a given liquid. Then the density of water is
(1) less than the density of the given liquid.
(2) more than the density of the given liquid.
(3) equal to the density of the given liquid.
(4) product of depths of immersion in water and the given liquid.

19. When a lactometer is used in a given sample of milk 'A' it floats in the milk such that half of the length of its stem is immersed in the milk. When it is used in another sample of milk 'B', the length of the stem immersed in the milk is one fourth of the total length. Then which of the following statement(s) is(are) true?
(A) Water content in A is more than that in B.
(B) Density of the sample A is less than the density of the sample B.
(1) A is true, but B is false (2) A and B are false
(3) A and B are true (4) A is false but B is true

20. Find the odd choice among the given ones.
The weight of the floating body is equal to the
(1) weight of the displaced liquid.
(2) buoyant force acting on the floating body.
(3) the weight of the liquid that occupies the portion of the volume of the body that is inside the liquid.
(4) the weight of the liquid whose volume is equal to the volume of the part of the body that is above the liquid surface.

21. The pressure at a point inside a liquid is directly proportional to _____.
(1) the depth of the point below the free surface (2) the density of the fluid
(3) acceleration due to gravity (4) All the above

22. The mercury level in a simple barometer falls gradually, at constant temperature indicates that
(1) weather gradually changes to summer. (2) weather gradually changes to rainy.
(3) weather changes to an anti–cyclone one. (4) weather gradually changes to winter.

23. A plastic ball moves from the bottom of a container filled with water to the top. Then
(1) its velocity increases (2) its acceleration remains same
(3) its velocity remains same (4) both (1) and (2)

24. A body is floating in a given liquid with half of its volume above the surface of the given liquid. If the body is slightly pushed into the liquid, then the body,
(1) sinks in the liquid and settles at the bottom.
(2) come back to its position after some time.
(3) starts oscillating in vertical direction.
(4) it jumps into the air and falls on the liquid.

25. A block of wood floats separately in river water and then in ocean water. Then the weight of the displaced river water is
(1) equal to the weight of the displaced ocean water.
(2) less than the weight of the displaced ocean water.
(3) more than the weight of the displaced ocean water.
(4) not dependent on the density of the wooden block.

26. In pressure cooker, the food is cooked faster, because the boiling point of water in the cooker ___.
(1) is above 100°C
(2) is below 100°C
(3) is equal to 100°C
(4) depends upon the amount of the volume present inside the cooker.

27. Choose the wrong statement among the following.
(1) Weight of a body measured by using a spring balance in air is its apparent weight.
(2) Apparent weight of a floating body varies with the density of the liquid.
(3) Volume of a floating body is equal to the volume of the liquid displaced by it
(4) None of the above

28. A bucket containing liquid is suspended from a spring balance. When an iron sphere is suspended from a thread and immersed in the liquid contained in the bucket without touching the walls of bucket, then the amount of increase of weight of the bucket is equal to
(1) the actual weight of the iron sphere.
(2) the weight of the displaced liquid.
(3) the apparent weight of the iron sphere in the liquid.
(4) None of the above

29. A barometer which measures atmospheric pressure more accurately is _____.
(1) Fortin's barometer (2) mercury barometer
(3) anaeroid barometer (4) Torricelli barometer

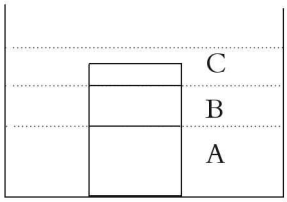
30. If the centre of gravity of the body lies above the meta centre, then the body is said be in_____.

(1) stable equilibrium (2) unstable equilibrium
(3) neutral equilibrium (4) Both (2) and (3)

Concept Application Level—2

31. While using Hare's apparatus, ratio of specific gravities of two liquids A and B was found to be 1 : 4 respectively. The differences in the level of liquids A and B in the tube were found to be 16 cm. The difference in the level of liquids A and B in the beaker was found to be 5 cm. Find the levels of liquids A and B in the tube, given the level of one of the liquids in the beaker is 3.0 cm.

32. Three identical cubes of side 2 cm and density 4 g cm^{-3} are placed one over the other in a container filled with three immiscible liquids A, B and C as shown in the figure. If the densities of A, B and C are 3 g cm^{-3}, 2 g cm^{-3} and 1 g cm^{-3} respectively, determine the net force exerted by the cubes at the bottom of the container.

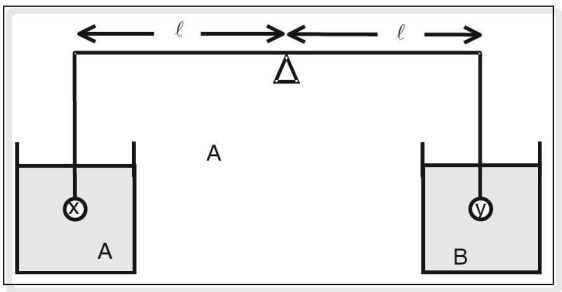


33. A sphere with a cavity in it is made of a material of mass 200 g and density 10 g cm^{-3}. When it is immersed in water its mass is found to be 170 g. Determine the volume of the cavity present inside.

34. A wooden cube of volume 1000 cm^3 floats in water with $\frac{2}{5}$ th of its volume above the surface of water level. If the given cube is made to float on a liquid of relative density 0.8 then find the amount of extra force required to completely submerge it the given liquid.

35. A sphere of volume 'V' is immersed in two immiscible liquids (mercury and water) taken in a vessel. If half of the volume of the sphere is in mercury and the other half of its volume is in the water, find the relative density of the sphere. (The densities of mercury and water are 13.6 g cm^{-3} and 1 g cm^{-3} respectively.)

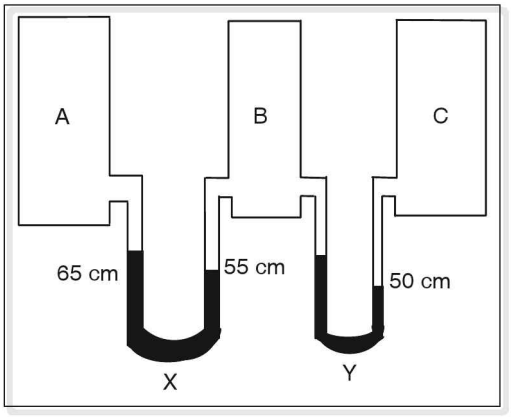
36. Two spheres X and Y of masses 1 kg and 1.1 kg are taken. They are made up of same material and are balanced in two liquids A and B as shown in the figure below. If the ratio of density of this material and liquid 'A' is 9. Given that specific gravity of liquid 'A' is 2.0. Find the value of density of liquid 'B'.



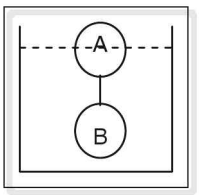
37. The length of a rounded bottomed test tube is 20 cm. The test tube is made to float, in turn, in two different liquids whose densities are 2 g cm^{-3} and 7 g cm^{-3} respectively. The corresponding depths of immersion of the test tube in the above mentioned liquids are 15 cm and 5 cm respectively. Find relative density of the liquid in which the test tube immerses by 10 cm.

38. A balloon of volume 1000 m^3 is filled with hot air of density 0.4 kg m^{-3}. The mass of the balloon is 120 kg and a weight 'w' is attached to it. The balloon is coming down to the ground. Find for what value of the weight 'w' do the balloon come down with zero acceleration (neglect the air resistance, take the density of cold air as 1.3 kg m^{-3} and g as 10 m s^{-2}).

39. Three cylinders A, B and C are filled with gases at different pressures. Two mercury manometers X and Y are connected to the three cylinders as shown in the figure. When the right arm of the manometer X is disconnected from the cylinder 'B', the level of the mercury in its left arm rises by 23 cm. If the initial level of the mercury in the limbs of the manometers is 60 cm, determine the pressure exerted by the gases in the cylinders A, B and C (atmospheric pressure = 76 cm of Hg).



40. A common hydrometer has a long uniform stem. When floating in pure water, 4.5 cm of stem lies below the water surface. The same hydrometer immerses up to 1.5 cm in a liquid of specific gravity 2.0. Find the specific gravity of the liquid in which hydrometer immerses up to 0.5 cm.



41. A container is filled with a liquid of density 4 g cm^{-3}. Two objects A and B of equal volume are held in equilibrium in the liquid as shown in the figure. The density of the objects A and B are 0.5 g cm^{-3} and 6 g cm^{-3} respectively. What fraction of the total volume of the object A is immersed in the liquid?

42. The pressure exerted by a force acting perpendicular on a horizontal surface is 200 Pa. Calculate the pressure exerted when the surface is rotated such that it makes an angle of 60° with its original position without altering the line of action of the force about the horizontal axis.

43. In a Hare's apparatus, one limb is immersed in a solution of specific gravity 0.5 and other in water. When the air is sucked out, the levels of water in the beaker and the tube are 10.0 cm, and 21.5 cm respectively. Find the level of solution in the tube, if the level in the beaker is 14.7 cm.

44. An ice cube has a stone of 500 g placed on its top, is floating in water with its lateral sides placed vertically. It displaces 5 kg of water. Suddenly, the stone slips into water. Because of this ice cube rises by $\frac{1}{10}$ th of its length above the water level. What is the density of the ice cube?

45. The length of a test tube is 20 cm. The test tube is made to immerse in two different liquids whose densities are 2 g cm^{-3} and 6 g cm^{-3} respectively. The depths of immersion of the test tube in the above mentioned liquids are 15 cm and 5 cm respectively. Find relative density of the liquid in which the test tube immerses by 10 cm.

Concept Application Level—3

46. In a faulty barometer, some inert gas is entrapped above the mercury level in 1 m long glass tube. The height of mercury column is x cm and the length of the space above the mercury is 10 cm. When the tube is pulled out of the trough by 10 cm, the space above mercury level is doubled. Now, 5 cm of the tube lies below the mercury level in the trough. Find the correct atmospheric pressure.

47. One end of a rod of mass 10 kg and length 20 m is tied at the bottom of a water tank by means of a rope with 15 m of its length below the surface of water. Determine the tension in the rope. The specific gravity of the rod is 0.5.

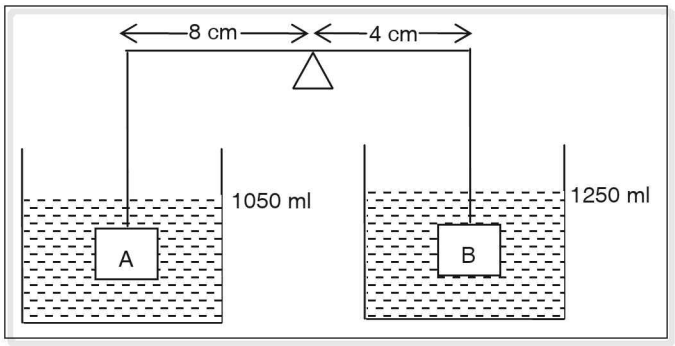
48. A student used a constant immersion hydrometer to determine the concentration of hydrochloric acid. A test tube of weight 50 g_f with 10 lead shots each of weight 5 g floats in water to a mark x on it. To keep the test tube float at the same mark in the HCl solution, he added 2 lead shots into it. Determine the volume percentage of the solution.

(Given density of pure water is 1g cm^{-3} and density of HCl is 1.18 g cm^{-3}

Volume percentage = $\frac{\text{volume of solute}}{\text{volume of solution}} \times 100$.

49. A cube made up of a metal weighs 308 g in air and 260 g in water but a cavity is present inside it. Find the volume of the cavity if the relative density of the metal is 11.

50. A uniform rod of weight 1 kg is held in equilibrium position as shown in the figure given below. If the initial volumes of the water in the containers are equal and the density of objects A and B are 3.5 g cm^{-3} and 4 g cm^{-3} respectively, determine the weight of objects A and B.



key points for selected questions

Very short answer type questions

1. The empty space created at the top of the tube is called torricellian vaccum.

2. Pascal's law

3. Test tube float is a flat bottomed uniform test tube used to find specific gravity of liquids and works on the law of floatation.
Test tube float are variable immersion hydrometer and constant immersion hydrometer.

4. Hare's apparatus and hydrometer.

5. Hydrostatics : Study of the fluids at rest.

6. Thrust: The force acting normally on a body is thrust.
Units: S.I. → N
C.G.S. → dyne
Pressure : Pressure is thrust acting on a unit area.
Units: S.I. → N m^{-2}
C.G.S. → dyne cm^{-2}

7. Fluids: Any substance which can flow and has no definite shape is called fluid.

8. Liquid used in barometer.
The height of the liquid column in the glass tube of a barometer.

9. Change in densities of sea water.

10. $\frac{W_1}{2(W_1 - W_2)}$

11. p = hdg
h = height of the liquid column,
d = density of the liquid and
g = acceleration due to gravity

12. Gas manometer

13. (i) When a body is immersed completely or partially in a liquid, it experiences an apparent loss in weight, which is equal to the weight of the liquid displaced by it.

(ii) When two liquid columns of different densities and heights balance each other, the pressure by them is equal.

14. $g\left(\frac{D-d}{D}\right)$

15. Pascal's law: Pressure applied anywhere in a confined fluid is transmitted equally and undiminished in all directions throughout the fluid.

16. Less than

17. The centre of gravity of the displaced liquid.

18. The intersection point formed by the centre line and the vertical line from new centre of Buoyance is called metacentre.

19. S.I. : 1 atm = 1.013×10^5 pa

C.G.S. 1 atm = 1.013×10^6 dyne cm^2

20. A body is in stable equilibrium, when its position changes on the application of external force, but once the external force is removed, it regains its original position.

A body is in unstable equilibrium, when its position changes on the application of an external force such that it does not regain its original position, on removing the force.

A body is in neutral equilibrium, if its position does not change by the application of an external force.

21. Device used to measure relative density of a liquid.

(i) Variable immersion Hydrometer

(ii) Constant immersion Hydrometer

22. 1

23. A device used to measure atmospheric pressure simple; fortins and aneroid barometer.

24. Vernier scale is provided.

25. 2 : 3

26. Device used to measure the pressure of gas in a container.

27. An aneroid barometer whose scale is calibrated in terms of altitude above sea level is called altimeter.

28. The net force acting on a body due to the fluid in upward direction.

29. Ratio of density of a substance to the density of water at 4°C.

30. 40

Short answer type questions

31. (i) Lateral pressure acts on the wall

(ii) Pressure increases with depth

(iii) At the base, very high lateral pressure acts

32. (i) Definition of upthrust

(ii) Solid body of area of cross section 'A' and height 'h' immersed completely in a liquid of density d_l

(iii) Expression for pressure due to liquid at the top of body and expression for pressure at the bottom

(iv) Net pressure – upward

(v) Net upward thrust

(vi) Upthrust = $v_l d_l g$

33. (i) Bulky

(ii) Mercury level in the trough not stable

(iii) Open and exposed to the surroundings

(iv) Risk of contamination of mercury

(v) No support to glass tube

(vi) Risk of spillage of mercury

34. (i) liquid of density d in a container

(ii) coin placed at its bottom

(iii) imaginary cylindrical liquid column of height 'h' above the coin

(iv) weight of the liquid column acting on the coin
(v) thrust due to liquid column
(vi) pressure due to liquid column

35. (i) Hydraulic press
(ii) Hydraulic brakes

36. (i) The weight in air of the floating body is equal to the weight of the liquid displaced.
(ii) the centre of gravity of the floating body and the centre of buoyance lie in the same vertical line.

37. (i) p = hdg
(ii) As 'd' increases, for same 'p', h decreases
(iii) Density of mercury is highest in liquids
(iv) Height of mercury column is 76 cm even at sea level
(v) Length of the tube needed is also less

38. (i) Definition
(ii) Same lateral pressure at given depth on all sides of body
(iii) Direction of resultant lateral pressure

39. (i) To draw underground water by hand pump or lift pump.
(ii) Helps to consume soft drinks and juices using a straw.

40. (i) Light and portable
(ii) Contains no liquid
(iii) Can be used as altimeter
(iv) Can be kept in any orientation

41. (i) weight
(ii) upthrust

42. (i) p = hdg
(ii) Pressure at a point in fluids depends on vertical height of the fluid column (h), density of the fluid(d) and acceleration due to gravity (g).
(iii) Pressure does not depends on the size, shape, structure of the container in which fluid is stored.

43. (i) Moisture in mercury
(ii) Impurities
(iii) Altitude of the place
(iv) Temperature
(v) Humidity

44. (i) Increase in altitude–decrease in height of air column
(ii) Increase in altitude–density of air decreases
(iii) For 105 m change in altitude, atmospheric pressure changes by 1 cm of Hg.
(iv) Graph showing the variation

45. (i) Pressure is same at all points in a horizontal plane in a stationary liquid. Pressure at a point in a liquid
(ii) is the same in all directions.
(iii) depends on the density of the liquid.

Essay type questions

46. (i) Working principle: principle of balancing columns
(ii) Construction
(iii) R.D. = $\frac{h_1}{h_2}$

47. (i) Explain what is test tube float
(ii) Principle
(iii) Mention its use
(iv) Explain the difference between constant immersion and variable immersion

48. (i) Definition
(ii) Archimedess' principle

49. (i) Construction
(ii) Principle
(iii) Applications and uses

50. (i) $W = vdg$: $F_B = vd_lg$
(ii) $W > F_B$ net force is in downward direction
(iii) $W > F_B$: $d > d_l$
(iv) $W = F_B$ body is in equilibrium $d = d_l$
(v) $W < F_B$: net force acts in upward direction
$W > F_B$: $d > d_l$

KEY

Concept Application Level—1

True or false

1. False
2. False
3. True
4. True
5. True
6. True
7. False

Fill in the blanks

8. greater than
9. 1.6 and 2.15
10. bottom
11. top and bottom
12. atmospheric pressure
13. relative density
14. 2 : 3

Match the following

15. A : c
    B : b
    C : e
    D : g
    E : i
    F : h
    G : d
    H : f
    I : a

Multiple choice questions

16. Choice (2)
17. Choice (3)
18. Choice (1)
19. Choice (3)
20. Choice (4)
21. Choice (4)
22. Choice (2)
23. Choice (4)
24. Choice (2)
25. Choice (1)
26. Choice (1)
27. Choice (3)
28. Choice (2)
29. Choice (1)
30. Choice (2)

Concept Application Level—2,3

Key points

31. (i) $\frac{h_A}{h_B}=\frac{d_B}{d_A}$

    (ii) h_A = 31 cm

    h_B = 15 cm

32. (i) Determined the upthrust acting on each part.

    Net force = Total weight – total upthrust.

    (ii) 0.4704 N

33. (i) Weight of the body in air – weight of the body in water = apparent loss of weight = upthrust= $V_{liquid}\ d_{displace} \times d_{liquid} \times g$

    $d = \frac{1}{2}$

    (ii) 10 cm^3

34. (i) Find the volume of a wooden cube (V_c) from given data.

    Find the volume of the water displaced from the given information.

    Here, weight of the cube is equal to the weight of the water displaced.

    Find the density of the cube (d_c) from law of floatation.

    That is, $V_c . d_c . g = \frac{1}{2} V_c . d_w . g$ ----- (1)

    Then, find the weight of the cube = $V_c . d_c . g$

When the same cube is made to float in the liquid of density 0.8 g cm^{-3}, find the volume of the displaced liquid (V_c) from the law of floatation.

$V_c . d_c . g = v_L . d_L . g$ ----- (2)

Find the value of "V_L" from (2)

Take the extra force required to completely submerge it in the given liquid as (F).

Then, weight of the cube + F = (100 cm^3) (0.8g cm^{-3})g ----- (3)

Find the value of 'F' from the 3rd equation.

(ii) 2 N

35. (i) By law of floatation, the weight of the floating body is equal to the weight of the liquid displaced, If d is the relative density of the sphere.

$$Vdg = \frac{v}{2} \times 13.6 \times g + \frac{v}{2} \times 1 \times g$$

(ii) 7.3 g cm^{-3}

36. (i) Find the net force acting on the spheres x and y, which is equal to [weight(w) – upthrust (u)]

By principle of moments

$(W_x - U_x)l = (W_y - U_y)l$ ---- (1)

substitute the below data in equation (1).

$W_x = V_x d_x g; W_y = V_y d_y g;$

$U_x = V_A d_A g = VU_x d_A g;$

$U_y = V_B d_B g - V_y d_B g$

Take $d_x = d_y = d;$

Given $\frac{d}{d_A} = 9;$

$d_A = 2$ g $cm^{-3}; \Rightarrow d = 18$ g cm^{-3}

Take, $V_x = \frac{m_x}{d};$

$V_y = \frac{m_y}{d}$. Substitute the value of W_x, V_x, W_y and U_y in ----- (1)

Then find the value of d_B.

(ii) 3.45 g cm^{-3}

37. (i) Let V be the volume of the hemispherical part of the test tube, x be the height of this part and the length of the test tube submerged in a liquid of density d then

[V + A (l – x)] = weight of the liquid displaced

= weight of the test tube

= constant.

(ii) 3.1

38. (i) Find the weight of hot air present in the balloon.

= (Volume of balloon) × (density of hot air) × g. ----- (1)

Find the upthrust acting on the balloon due to cold air.

= (volume of balloon) (density of cold air) × g ----- (2)

Now, find the weight of the balloon and also the weight attached to it from the given data.

Find the net forces acting on the balloon.

By the law of floatation i.e., Wt. of balloon + W + Weight of hot air = upthrust ----- (3)

Then, the sum of all forces acting on the balloon is zero because the balloon comes down with zero acceleration.

Substitute the appropriate value in (3) and get the value of 'W'.

(ii) 780 kg_{wt} or 7800 N.

39. (i) P_A = atmospheric pressure – Δl

$P_B = P_A + \Delta l_1$

$P_C = P_B + \Delta l_2$

Where Δl, l_1, Δl_2 are the change in levels of the mercury columns

(ii) 20 cm of Hg, 30 cm of Hg, 50 cm of Hg

40. (i) Let the volume of the floatation bulb $= V\ cm^3$,
and the area of cross–section of the stem $= a\ cm^2$.
By the law of floatation the weight of the floating body

= Weight of the displaced liquid

= Upthrust acting on common hydrometer.

Here, take the density of the liquid as 'd_L' where Hydrometer immerses only upto 0.5 cm.

Then, W = (Volume of the displaced liquid) (density of the liquid) (acceleration due to gravity)

$W = (V + a.l)d_L g$.

Obtain the expression as,

$W = [V + a(4.5)]1.g = [V + a\ (1.5)].2.g = [V + a(0.5)]d_L.g$. Determine the value of 'V' in terms of a.

Obtain the value of d_L by solving the expression.

(ii) 3.

41. (i) When the object floats freely

$V_{in} \times d_l\ g = V_A \times d_A\ g$

Where V_{in}, V_A are the volume immersed and total volume of the object A respectively d_l, d_A are the density of the liquid and the object A respectively.

The extra force acting on the object A = weight of the object B – upthrust acting on it

(ii) 5/8

42. (i) Pressure $= \dfrac{\text{Thrust}}{\text{Area}}$

Thrust is the force acting perpendicular to the surface

(ii) 100 Pa

43. (i) Find the densities of water and the given liquid.
Find the levels of water in the beaker and the tube from the given data.
Find the height of the water column in the tube.

By using the formula, $h_A d_A g = h_B d_B g$ ---- (1)

Find the height of the liquid column i.e., h_2.

If the level of liquid in the beaker is 14.7 cm, then find the level in the tube.

(ii) 37.7 cm.

44. (i) By the law of floatation,

[mass of cube (ice) + mass of stone]g

= [mass of displaced water]g ---- (1)

Then, find the mass of the ice cube from (1).

When a stone weighing 500g placed on its top, slips into water, the ice cube rises by $\dfrac{1}{10}$ th of its length.

Because of this, the mass of the water displaced changes by 500 g.

The change in the volume of the ice cube is equal to $\left(a\ \dfrac{1}{10}\right)$

Then, this value is equal to the change in the volume of the is equal water displaced.

Then $\left(a\ \dfrac{1}{10}\ al\right) = 500\ cm^{-3}$ ----- (2)

Then find the volume of the cube (V_i) from (2)

Here, mass of the cube = 4500 g.

This means, $V_i . d_i = 4500$ ----- (3)

Substitute 'V_i' from (2) in (3).

Find the value of d_i.

(ii) 0.9 g cm^{-3}

45. (i) Take the area of cross–section of the test tube as 'a'.

Then find the volumes of the two displaced liquids by the immersion of the test tube

Use the law of floatation.

(i.e.,) weight of the floating body (W) = Weight of the displaced liquid.

Here, $W = V_T.d_T.g$ ----- (1)

$\Rightarrow V_T.\, d_T.g = V_L.d_L.g$ ----- (2)

Find the density of the unknown liquid using equation (2).

(ii) 3.

46. (i) Part I:

Find the length of the Hg column and entrapped gas in a one metre tube from the given data.

Find the volume of entrapped gas (V_1).

Then atmospheric pressure, (P_a) = Pressure due to entrapped air (P_1) + Pressure due to 'x'cm of Hg column ----- (1)

Part II:

When a test tube is pulled out by 10 cm from the trough, find the length of the air column.

Find the volume of the entrapped gas (V_2).

Now, P_a = Pressure due to 'x' cm of Hg column + pressure due to entrapped air (P_2) ----- (2)

From Boyle's law.

$P_1V_1 = P_2V_2$ ----- (3)

Substitute the values of P_1 and P_2 that are obtained in (1) and (2) into (3).

Solve the equation thus obtained to get the value of 'P_a'.

(ii) 75 cm of Hg.

47. (i) T + mg = (upthrust)

Upthrust = (mass/unit length) × length immersed $\times \left(\frac{\text{density of liquid}}{\text{density of the body}}\right)$

(ii) 49 N

48. (i) Weight of the test tube + Weight of lead shots = Weight of liquid displaced

(ii) 55.56%

49. (i) Find the weight of the cube in air (W_1) and in water (W_2).

Find the apparent loss in the weight of the cube (i.e.,.,) ($W_1 - W_2$) = ----- (1)

But ($W_1 - W_2$) = upthrust acting on the cube.

$\Rightarrow W_1 - W_2 = V.\, d_{water}.g$ ----- (2)

From (2), find the volume of the cube(V_1)

Find the relative density(d) of the metal from the given data.

Find the volume of the metal (V_2) from using the formula.

$$V_2 = \frac{W_1}{d}$$

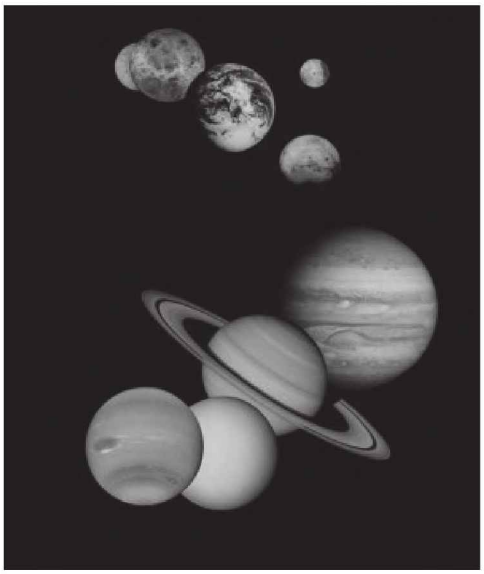
Now, the volume of the cavity $= V_1 - V_2$.

(ii) 20 cm^3.

50. 175 g_{wt}, 1000 g_{wt}

6

Wave Motion and Sound



INTRODUCTION

Sound and vibrations

Sound is a form of energy causing the sensation of hearing through the vibrating ear drum. A loud thunder or high intensity sound produced by a low-flying jet plane results in the rattling of window panes and doors. We know that work is done at the expense of energy and the rattling of doors and windows is caused by the sound, thus indicating that sound is a form of energy.

The source of sound energy is the mechanical energy possessed by vibrating bodies. In other words, sound is produced by vibrating bodies. This can be demonstrated by a simple experiment.

Vibrating bodies produce sound

A tuning fork is clamped to a wooden base and the prongs are gently struck with a leather or rubber hammer. As the sound is produced from the tuning fork a pitch ball suspended from a string is brought into contact with the tuning fork. It is found that as soon the pitch ball comes in contact with the tuning fork, it instantaneously moves away. On the other hand the pith ball does not jump away when it comes in contact with a tuning fork that does not produce sound. The jumping away of the pith ball is indicative of the vibrations of the tuning fork and this experiment illustrates that a tuning fork produces sound only when it vibrates.

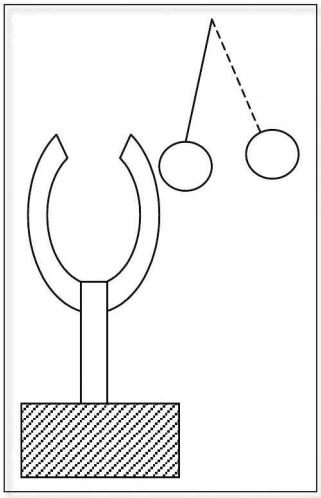


Figure 6.1

Do all the vibrating bodies produce audible sound. It has been found that bodies vibrating with a frequency in the range 20 Hz to 20 kHz only can

cause a sensation of hearing in humans, and this range is called audible range. Animals, birds and insects in general are sensitive to sound with a frequency outside this range also. If the frequency of the vibrations is less than 20 Hz it is called infrasonic sound and the sound produced by bodies vibrating with a frequency greater than 20 kHz is called ultrasonic.

Propagation of sound

A little earlier we had mentioned that sound is a sensation produced in the brain through the vibrations of the ear drum—a thin membrance in the ear. For the vibrations of the tuning fork to be transmitted to the ear the intervening material medium plays an important role. In other words, material medium is required for the propagation of sound energy. This is demonstrated in the bell-jar experiment. The sound produced by the electric bell located inside the bell-jar can be heard from the outside only when there is air inside the jar. It is observed that the sound produced by the electric bell is not heard when the jar is completely evacuated. No wonder we can only see the sun but not hear the roaring solar flames!

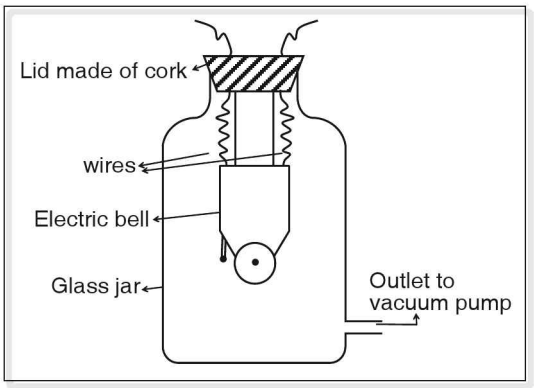


Figure 6.2 Bell-jar experiment

The important properties of a material medium that help in the proportion of sound energy are the elasticity and inertia of the medium.

Wave Motion

Consider the row of bicycles parked in a cycle-stand. If the bicycle at one end is pushed, progressively all the bicycles would fall and within a few seconds the bicycle at the other end also would fall. Without physically going to it, the last bicycle is successfully pushed down. This is the principle of wave motion, in which the vibrations of the particles of the intervening material medium are utilized to transmit the energy without the actual translatory motion of the source causing the disturbance, just as the intermediate bicycles help in pushing the last bicycle in the row.

A wave is a disturbance created at a place in a medium and travels in all possible directions in the medium. When a wave propagates from one place to another, the particles of the medium do not undergo any translatory motion but have only vibratory or oscillatory motion.

Types of waves

Waves may be classified in three different ways. The first classification is based on the requirement of a medium for the wave propagation (mechanical and electromagnetic waves); the second classification is based on the direction of the vibration of the particles of the medium with respect to the direction of the wave propagation (transverse and longitudinal waves); and the third classification is based on the limitation of the propagation of the waves (progressive waves and stationary waves).

Mechanical and electromagnetic waves

Mechanical waves are also called elastic waves and they require a material (an elastic) medium for their propagation.

☛ *Examples* Sound waves, water waves, etc.

Electromagnetic waves are the waves which do not require any material medium for their propagation. They can propagate even through vacuum.

☛ *Examples* Light waves, X-rays, radio waves, etc.

Transverse and longitudinal waves

If the direction of the vibration of the particles of the medium is perpendicular to the direction of the propagation of the wave, the wave is called a transverse wave.

☛ *Example*

Ripples on water surface, a wave pattern set in a stretched string.

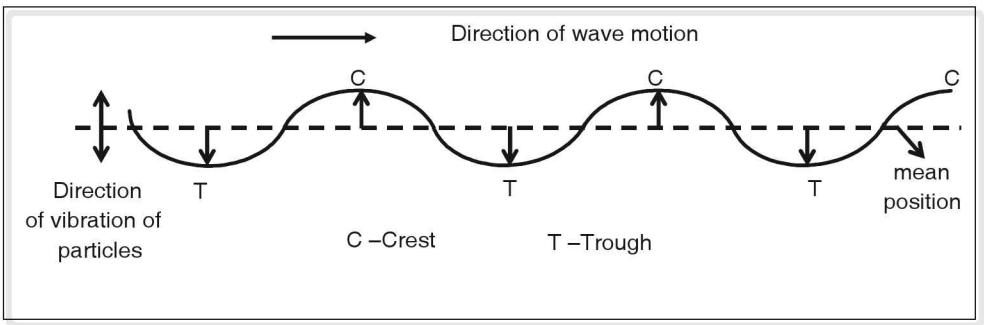


Figure 6.3 A transverse wave

Given below are the definitions of standard terms describing the properties of a wave.

(a) Crest: The position of the maximum upward displacement of a vibrating particle from the mean position in a transverse wave is called a **crest**.

(b) Trough: The position of the maximum downward displacement of a vibrating particle from the mean position in a transverse wave is called a **trough**.

(c) Amplitude: The maximum displacement of a vibrating particle from the mean position (either upward or downward) in the wave is called **amplitude** and is measured in metres or centimetres.

(d) Phase: The state of a vibrating particle in the wave with regard to its displacement from the mean position and the direction of vibration is called **phase** and is measured in radians. If the phase of any two particles in a wave is equal, i.e., they have a phase difference of 2π radians, they are said to be **in phase**; else they are said to be **out of phase**.

(e) Wavelength: The distance between any two successive particles of the wave which are in phase is known as **wavelength (λ)** and is measured in metres or centi-metres.

(f) Time period: The time taken by a vibrating particle to complete one full vibration or the time taken to produce one complete wave pulse is known as **time period (T)** and is measured in seconds.

(i) Frequency: The number of vibrations made by a vibrating particle in the wave in one second or the number of waves produced per second is known as **frequency** and is measured in hertz (Hz).

If the direction of the vibration of particles of the medium is parallel to the direction of the wave propagation, the wave is called a longitudinal wave.

☛ *Example*

Sound waves in air or in a gas; waves set in a long spring.

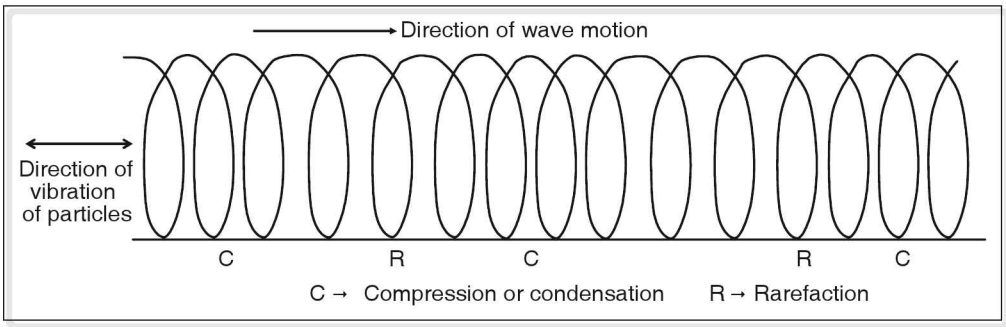


Figure 6.4 A longitudinal wave set in a spring

When a long spring is placed horizontally on a table and one end is vibrated as shown in figure (6.4), there are certain regions formed in the spring where the rings of the spring are much closer to each other than they would be when they are at rest position. Such regions are called **compressions** or **condensations**. Similarly, there are certain regions formed in the spring where the distance between the rings is more than the distance when they are at rest position. These regions are known as **rarefactions**.

Progressive and stationary waves

When a wave starts from a point in a medium and propagates in all the possible directions and never returns, the wave is called a **progressive wave**. When a progressive wave strikes a hard surface and is reflected, the incident wave and the reflected wave superpose forming a standing wave, also called a stationary wave. In such a wave, the amplitude of each particle varies with the position along the direction of the propagation of the incident wave.

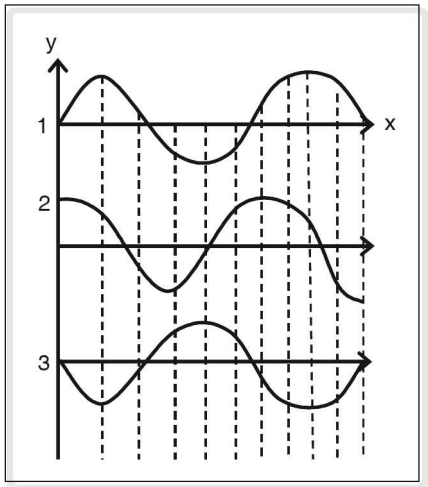


Figure 6.5

Phase difference between waves (1) and (3) is π, i.e., if the shape of the wave in (1) is described by $y = \sin x$, the shape of the wave in (3) is given by $y = \sin(x + \pi) = -\sin x$.

∴ When wave in (1) reaches a maximum, wave in (3) reaches a minimum and vice-versa.

Phase difference between (1) and (2) is $\frac{\pi}{2}$, i.e., if the wave in (1) is given by

y = sinx, the wave in (2) is given by $y = \sin\left(x + \frac{\pi}{2}\right) = \cos x$.

∴ If a wave on reflection undergoes a phase change π, the shape of the wave before and after reflection at the wall would appear as below.

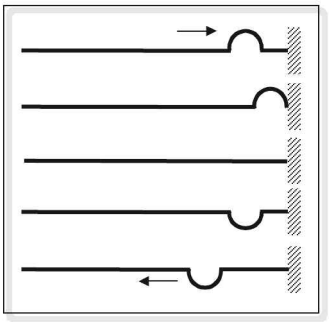


Figure 6.6

Stationary wave formation in a stretched string

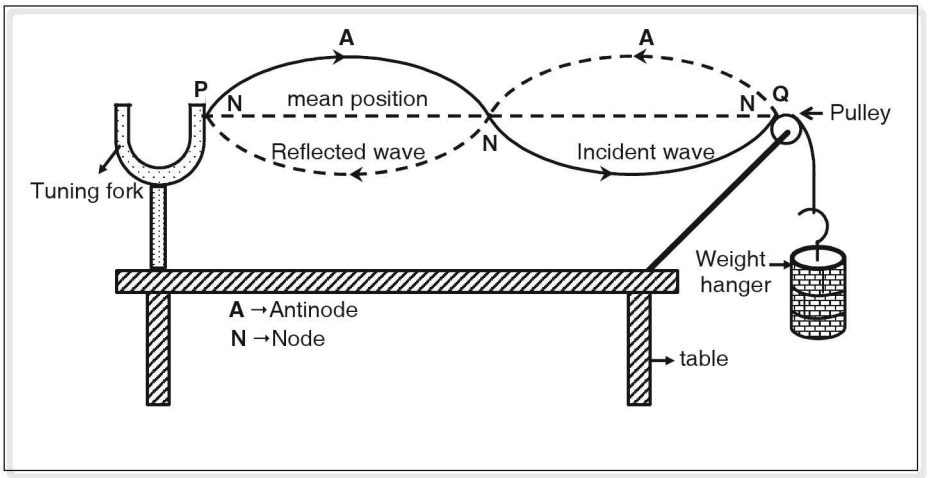


Figure 6.7 Formation of stationary waves in a string

Fix a tuning fork of some frequency and a smooth frictionless pulley at the ends of a table as shown in figure (6.7). One end of a string is tied to one of the prongs of a tuning fork and the other end passed over the pulley. To this free end of the string, a weight is attached. The weight attached provides the tension in the string. When the tuning fork is excited with a rubber hammer, the vibrating tuning fork sends a wave in the string which propagates from 'P' to 'Q'. At 'Q' the wave is reflected. The incident wave and the reflected wave superpose to form a stationary wave as shown in the figure. The positions where the amplitude of the vibrating particles is zero or minimum are shown as 'N' in the figure and are called '**nodes**'. Similarly the positions where the amplitude of the vibrating particles is maximum are shown as 'A' in the figure and are called '**antinodes**'. When the wave is reflected at 'Q', a phase change of π radians takes place. The string between the points 'P' and 'Q' appears as loops due to persistence of vision.

Differences between progressive and stationary waves

These waves are produced at a point and progress in all the possible directions till they die out due to damping.	These waves are limited between two positions and appear to be standing.
These waves cause crests and troughs or compressions and rarefactions.	These waves cause nodes and antinodes.
The amplitude of all the particles in the wave is equal.	The amplitude of different particles in the wave is different. It is maximum at antinodes and minimum or zero at nodes.
There is a continuous phase difference between particles in the wave.	All the particles between two successive nodes in the wave are in phase.
The distance between any two adjacent crests (or compressions) or two adjacent troughs (or rarefactions) is one wavelength (λ)	The distance between any two successive nodes or two successive antinodes is one half of the wavelength ($\lambda/2$).
The energy associated with these waves is kinetic in nature and is transmitted from one place to another.	The energy associated with these waves is potential in nature and it is stored in the medium between two fixed points.
In a progressive wave, there is no instant when all the particles cross the mean position together.	In a standing wave, all the particles cross their mean positions together.
In a traveling wave, there is no particle which is always at rest.	In a standing wave, the particles at nodes are always at rest.

Sound waves

Sound waves are longitudinal waves due to the compressions and rarefactions produced in the material medium surrounding the source.

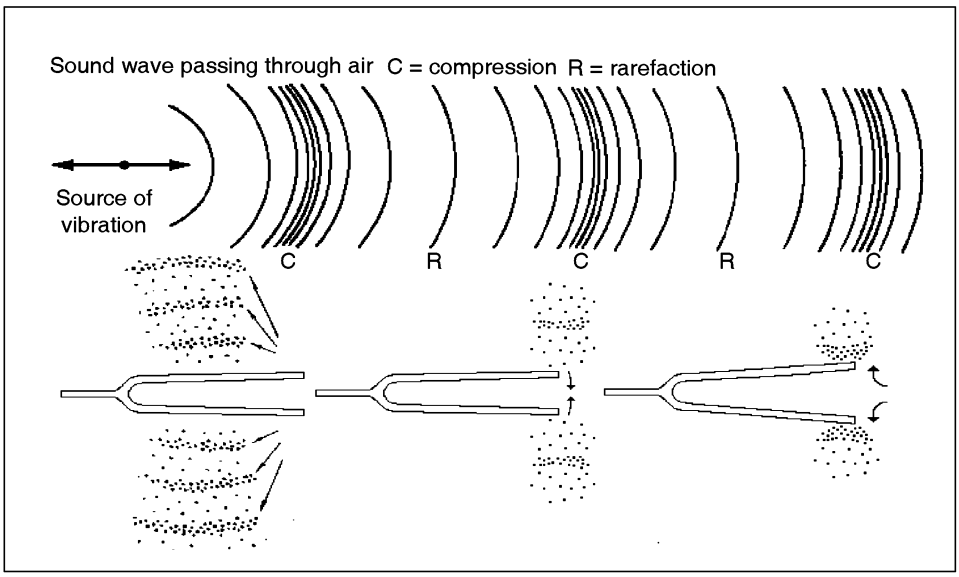


Figure 6.8

Sound waves can propagate in solids, liquids as well as gases. Put your ear to one end of a long pipe, say 100 m, and ask your friend to gently tap the other end. Almost immediately you would be able to hear the sound at your end.

Propagation of sound through liquids

Place a trough of water on a hollow wooden box (sound box) open at one end. Fix the stem of a tuning fork to a metallic disc. The prongs of the tuning fork are set to vibrate and the disc is gently made to come in contact with the water surface as shown in the figure. It is found that immediately the sound of the tuning fork with greater intensity is heard from the sound box.

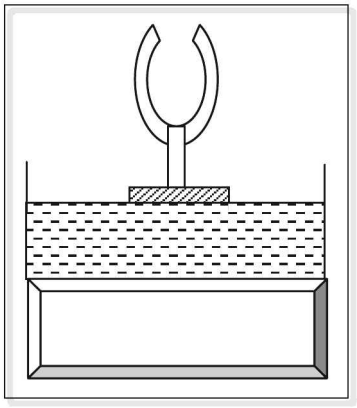


Figure 6.9

Simple Harmonic Motion (SHM)

The transmission of sound is by longitudinal wave propagation caused by the vibration of the particles of the medium resulting in compressions and rarefactions. These vibrations are periodic and oscillatory in nature, i.e., the motion of each particle repeats itself along the same path at regular intervals of time (periodic motion) and at the same time the motion is to and fro along the path (oscillatory). The oscillations of the bob of a simple pendulum, the vibrations of a body freely suspended from a spring, to and fro motion of the air molecules during the propagation of sound are all examples of harmonic motion.

A harmonic motion is categorized as a simple harmonic motion if it satisfies the following conditions.

1. The acceleration of the body is always directed towards its mean position. In other words a restoring force acts on the body directed always towards the mean position. In the oscillations of a simple pendulum tangential component of gravitational force is always directed towards the mean position.
2. The magnitude of acceleration (a) is directly proportional to the displacement (x) of the body from its mean position. Since the acceleration at any given position is directed towards the mean position and the displacement is considered away from the mean position, we have

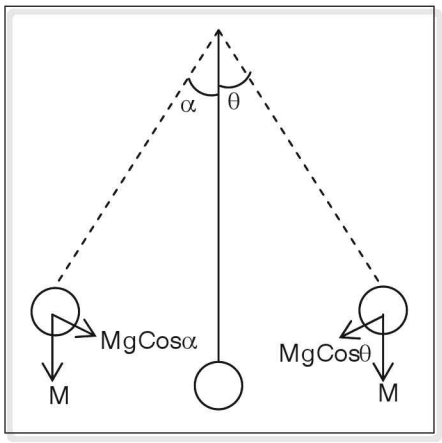


Figure 6.10

$$\bar{a} \propto -\bar{x}$$

A simple harmonic motion is particularly characterized by (1) constant time-period (T), which also implies a constant frequency ν (Greek letter 'nu') (2) a constant amplitude and (3) a constant total mechanical energy, i.e., the sum total of kinetic energy and potential energy is same at every point on its path. The bob of a simple pendulum has kinetic energy at its mean position which is conserved as potential energy at its extremes.

Graphical representation of SHM

When we plot the displacement time graph of a body suspended freely from a spring executing free oscillations, we get a curve as shown in the figure.

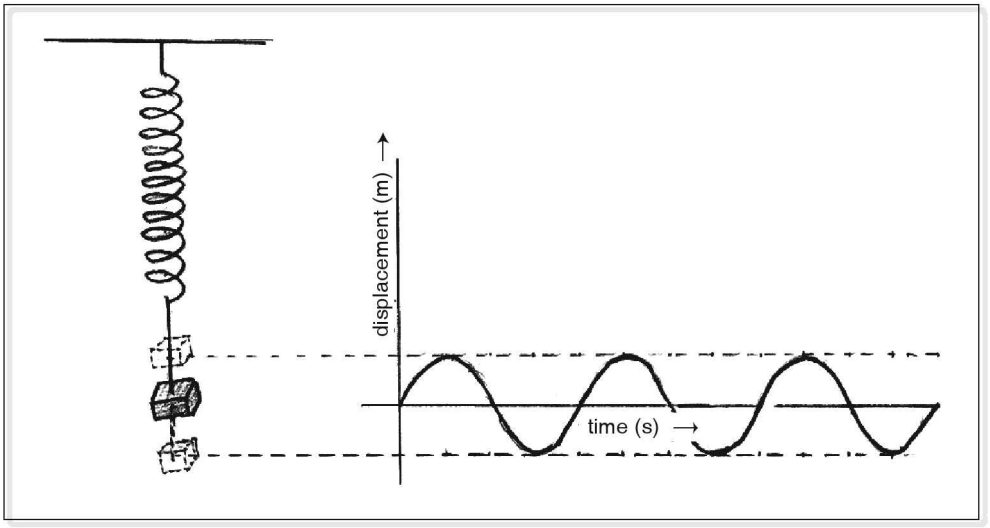


Figure 6.11

We find that the curve is in the form of a wave, and displacement–time graph of the particles in a transverse wave (water surface) and the pressure–time graph of a longitudinal wave are also similar to this.

(a) Displacement time graph for a longitudinal wave propagation.

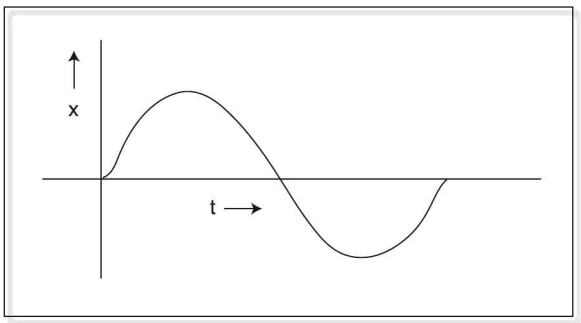


Figure 6.12

(b) Pressure time graph for sound wave.

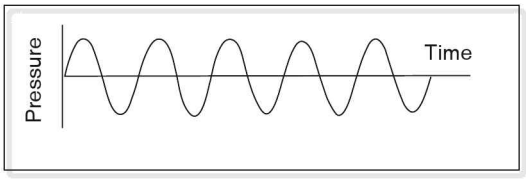


Figure 6.13

The crest in a transverse wave corresponds to a compression in a longitudinal wave, and the trough corresponds to the rarefaction.

If T is the time-period of oscillations and λ is the wavelength of the wave propagating then the velocity (v) of the wave propagation is given by

$$v = \frac{\lambda}{T}$$

Since $\nu = \frac{1}{T}$, $v = \nu\lambda$

Velocity of sound

The velocity of waves in a medium was given by Sir Isaac Newton as $V = \sqrt{\frac{E}{d}}$ where, 'V' is the velocity of the waves, 'E' and 'd' are the elasticity and the density of the medium respectively. The elasticity factor changes with the state of matter, i.e., solids, liquids or gases and we get the expression for the velocity of sound in the corresponding medium accordingly.

(a) Velocity of sound in solids

The elasticity factor in solids is given by Young's modulus (Y) and so the expression for the velocity of sound in solids is given as $V = \sqrt{\frac{Y}{d}}$, where 'Y' and 'd' are Young's modulus of elasticity and density of the given solid medium respectively.

(b) Velocity of sound in liquids

The elasticity factor in liquids is taken as its bulk modulus of elasticity (K) and so the expression for the velocity of sound in liquids is given as $V = \sqrt{\frac{K}{d}}$, where 'K' and 'd' are the bulk modulus of elasticity and the density of the liquid respectively.

(c) Velocity of sound in gases

The elasticity factor in a gas is its pressure. Thus the Newton's formula for the velocity of sound in a gas changes to $V = \sqrt{\frac{P}{d}}$ where 'P' and 'd' are pressure and density of the given gas. On theoretical calculation, at normal temperature, velocity of sound in air (a mixture of gases) using Newton's formula was obtained as 280 m s^{-1} whereas experimental value was found to be nearly 330 m s^{-1}. Owing to the large difference between theoretical and experimental values of velocity of sound in air, Newton's assumptions in arriving at his formula for velocity of sound in a given gas were reviewed and a correction was made by Laplace.

Thereby the formula for the velocity of sound in a gas was given as $V = \sqrt{\frac{\gamma P}{d}}$, where '$\gamma$' is a constant for a given gas and is defined as the ratio of the specific heat capacity of the given gas at constant pressure (C_p) to its specific heat capacity at constant volume (C_v).

Factors affecting the velocity of sound in a gas

1. Absolute temperature of gas

The velocity of sound in a gas is directly proportional to the square root of its absolute temperature. Mathematically, $V \propto \sqrt{T}$ where 'V' is the velocity of sound in a gas and 'T' is its absolute temperature.

If 'V_1' and 'V_2' are the velocities of sound in a gas at absolute temperatures 'T_1' and 'T_2' respectively, then,

$$\frac{V_1}{V_2} = \sqrt{\frac{T_1}{T_2}}$$

2. Density of a gas

The velocity of sound in a gas is inversely proportional to the square root of its density. Mathematically, $V \propto \frac{1}{\sqrt{d}}$ where 'V' is the velocity of sound in a gas and 'd' is its density. If 'V_1' and 'V_2' are the velocities of sound in two gases having densities 'd_1' and 'd_2' respectively, other parameters of the gases being constant, then $\frac{V_1}{V_2} = \sqrt{\frac{d_2}{d_1}}$

3. Molecular weight of a gas

The velocity of sound in a gas is inversely proportional to the square root of its molecular weight. Mathematically, $V \propto \frac{1}{\sqrt{M}}$, where 'V' and 'M' are the velocity of sound in a gas and the molecular weight of the given gas respectively. If 'V_1' and 'V_2' are the velocities of sound in two gases whose molecular weights are 'M_1' and 'M_2' respectively, the other parameters of gases being constant, then $\frac{V_1}{V_2} = \sqrt{\frac{M_2}{M_1}}$.

Factors that do not affect velocity of sound in a gas

1. Pressure of a gas

When the pressure of a gas is changed at a constant temperature, its density also changes accordingly such that the ratio 'P/d' is always a constant. For a given gas, the value of 'γ' is a constant. So the variation in the pressure of a gas does not affect the velocity of sound in it.

2. Amplitude

Amplitude is the maximum displacement of a vibrating particle from its mean position. The amplitude of a sound wave in a medium does not affect its velocity.

3. Frequency

Frequency is the number of waves produced per second. The velocity of sound in a gas or any medium does not depend on its frequency.

4. Wavelength

Wavelength is the distance between two successive particles in a wave which are in phase. The velocity of a wave in a medium does not depend on the wavelength of the wave.

The velocity of sound in air

Air is a mixture of gases. So the factors that affect the velocity of sound in gases also affect the velocity of sound in air. If V_0 and V_t are the velocities of sound in air at 0°C and t°C respectively, then they are related as $V_t = V_0\left(1+\frac{t}{546}\right)$.

Another factor (to be considered) which affects the velocity of sound in air is humidity. Humidity is the amount of water vapour present in air. The presence of humidity decreases the density of air. So higher the humidity, less is the density of air and more is the velocity of sound in it. Thus, the velocity of sound in air increases as the humidity in air increases.

Relation between velocity, wavelength and frequency of sound

In one time period (T), a sound wave advances in a medium by one wavelength (λ). So the velocity of sound in the given medium is given by $V = \frac{\lambda}{T}$.

But as $\frac{1}{T} = n$, the frequency of sound, we have, $V = n\lambda$

Reflection of sound

Echo

Like the light waves, sound waves also undergo reflection on striking a hard surface. The reflected sound is known as echo.

When a person makes a sound in front of a big wall, he hears two sounds, one produced by him and the other is the reflected sound. The person can hear the two sounds distinctly if they reach his ear with a time gap of at least 0.1 s. This is known as **persistence of hearing**. Consider a person standing at 'P' at a distance 'd' from a wall (reflector of sound) as shown in figure (6.14).

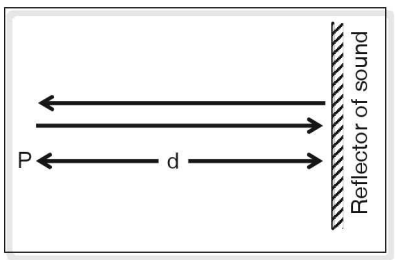


Figure 6.14 Echo

The person produces a sound and hears an echo after 't' seconds. Then the velocity of sound in air is given by the expression, $V = \frac{2d}{t}$. If the velocity of sound is known, the distance between the person and

the reflector is given by $d = \frac{Vt}{2}$. Since the persistence of hearing (the minimum time gap between two sounds reaching the ear) is 0.1 second, the minimum distance required to hear an echo $= \frac{V(0.1)}{2} = \frac{V}{20}$ m.

Applications of reflection of sound

1. Echo can be used to find out the velocity of sound in air.
2. The reflection of sound can be used in instruments like megaphone and in hearing aids.
3. The reflection of sound is used in sound boards and whispering galleries.
4. The reflection of sound is an important phenomenon used in SONAR, which is a special technique to find the depth of ocean beds.
5. Bats produce ultrasonics and can even sense their reflections from objects. With the help of this phenomenon, they can move easily in darkness.

Determination of velocity of sound in air by echo method

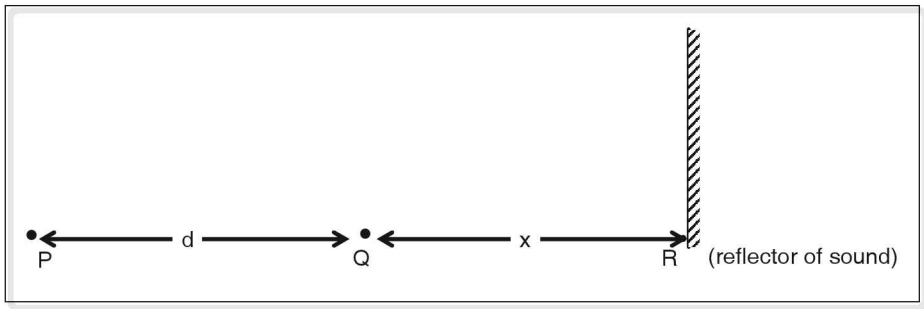


Figure 6.15 Determining the velocity of sound in air by echo method

Consider a person standing at position 'P' in the front of a wall (reflector of sound) (R). He produces a sound and hears the echo after 't_1' seconds. He then moves to a new position 'Q' towards the reflector by a distance 'd' in a direction normal to it. Now, he produces the sound again and hears the echo after 't_2' seconds.

If the velocity of sound in air is 'V', then $V = \frac{2(d+x)}{t_1}$ and also $V = \frac{2(x)}{t_2}$

$\therefore Vt_1 = 2d + 2x$ and $Vt_2 = 2x$

$\therefore Vt_1 = 2d + Vt_2$

$\therefore V(t_1 - t_2) = 2d$

or $V = \frac{2d}{t_1 - t_2}$

Reverberation

Consider a source and an observer of sound placed in a closed room as shown in figure (6.16). The observer hears the sound that travels not only directly from the source, but also indirectly from a series

of reflections taking place at the walls of the room. So the observer can hear the sound even after the source has stopped producing it due to multiple reflections taking place at the walls of the enclosure. This phenomenon is known as '**reverberation**'. Reverberations are often heard in empty halls, theatres and auditoriums.

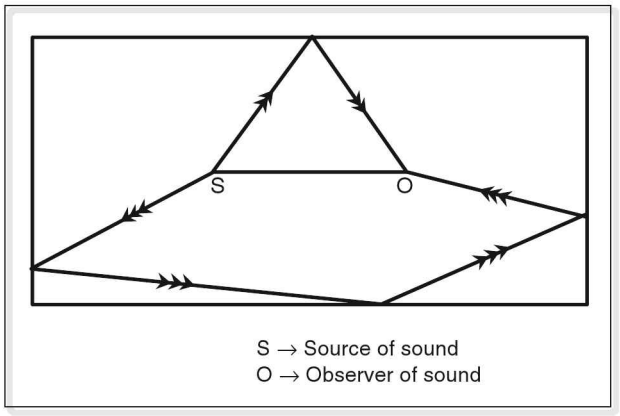


Figure 6.16 Reverberation

Acoustics

Taking the effects of echo and reverberation into consideration in designing an auditorium and the interior of buildings is called acoustics. The time required to reduce the intensity of sound to one millionth of its initial value is called reverberation time. This time plays an important role in designing an auditorium. Generally a reverberation time of 0.5 is considered as optimum value for speeches whereas for good acoustic effect in a music hall a reverberation time of 2 to 2.5 is taken into consideration. Reflecting surfaces like walls and ceiling are covered with good sound absorbing materials and the floor is carpeted. Curtains and sound absorbing hangings are also used for better acoustic effect.

Natural vibrations

If a body is set into vibrations and allowed to vibrate on its own, the vibrations of the body are known as '**natural**' or '**free vibrations**'. The frequency with which the body vibrates is known as natural frequency and the corresponding time period is known as natural time period.

Examples of natural vibrations

1. The strings of musical instruments that vibrate freely when plucked.
2. Oscillating pendulum that is suspended freely.
3. A blade clamped at one end and plucked at the other end.
4. A tuning fork when hit by a rubber hammer.

Damped vibrations

When a pebble is thrown on the surface of still water, ripples are formed and these ripples travel in all possible directions. As the ripples move away from the source of disturbance, we observe that the amplitude of the waves decreases continuously and finally the wave vanishes. This 'damping' of the wave takes place because the medium itself absorbs the energy of the wave that is propagated in the medium.

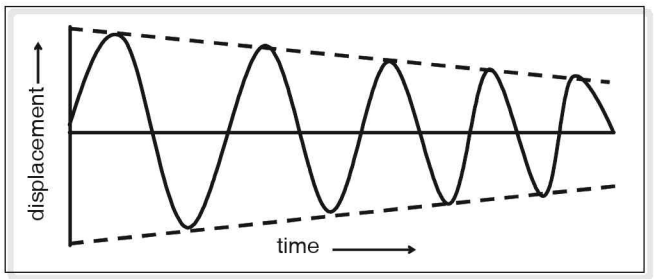


Figure 6.17 Displacement versus time graph for damped vibrations

The vibrations whose amplitude progressively decreases with time or with position or with both as the wave is propagated in a medium are called '**damped vibrations**'.

☛ *Examples of damped vibrations*

1. When a simple pendulum is disturbed from its mean position and left alone, the amplitude of the pendulum decreases with time and eventually stops. These oscillations are damped oscillations.
2. If a blade is clamped at one end and plucked at the other end, the amplitude of the vibrating blade decreases with time. These vibrations of the blade are damped vibrations.
3. When the string of a musical instrument like veena or sitar is plucked once, the amplitude of the vibrations in the string progressively decreases with time.

Forced vibrations

If a body is set into vibrations by the influence of another vibrating body or an external periodic force, the vibrations executed by the former are known as forced vibrations.

☛ *Examples for forced vibrations*

1. The vibrations executed by the pendulum of a clock are forced vibrations as they are caused under the influence of a main spring which is wound.
2. Musical instruments with strings contain hollow sound boxes. When a string of an instrument like veena is plucked, the sound produced by the vibrating string causes the air in the sound box to vibrate at the same frequency as the string and a loud sound is heard. The vibrations of air in the sound box are forced vibrations.

Resonance

Consider two simple pendula 'A' and 'B' of different lengths ($\ell_A > \ell_B$) suspended from an elastic, rigid support like a thin rod as shown in figure (18). Since the frequency of a simple pendulum is inversely proportional to the square root of its length, the frequency of 'A' (n_A) is less than that of 'B'(n_B). When only pendulum 'B' is made to oscillate, we observe that after some time, pendulum 'A' also begins to oscillate under the influence of the oscillations of 'B'. Then the vibrations of 'B' are free vibrations and that of 'A' are forced vibrations. The frequencies 'n_A' and 'n_B' mentioned earlier are the natural frequencies of 'A' and 'B' respectively. Now since 'A' is oscilliating under the influence of 'B', it will oscillate with the frequency of 'B' and not with its own natural frequency.

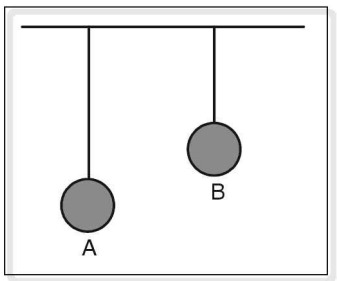


Figure 6.18

Similarly, if 'B' is at rest and 'A' is set into oscillations, after some time, we observe that 'B' also oscillates with a frequency equal to the natural frequency of 'A' and not with its own natural frequency. So, here the oscillations of 'A' are free oscillations and those of 'B' are forced oscillations.

The forced oscillations take place due to the transfer of energy from one pendulum to another. If 'A' has forced oscillations, it will derive energy from 'B'. Since the natural frequencies of the two pendula

are not equal, the transfer of energy from pendulum 'B' to pendulum 'A' is not maximum. Thus, the amplitude of 'A' is not maximum. This maximum transfer of energy from one pendulum to another takes place when their natural frequencies are equal, i.e., when the lengths of the two pendula are equal.

Consider two pendula 'C' and 'D' of equal length suspended from an elastic rigid support as shown in figure (19). As they have equal length, their natural frequencies are also equal. If one of them, say 'C' is set into oscillations, the other being at rest; the second pendulum 'D' too oscillates after sometime. The pendulum 'D' oscillates with the maximum possible amplitude because both C and D have the same natural frequency and so, energy transmitted from 'C' to 'D' is the maximum. This phenomenon is known as '**resonance**'.

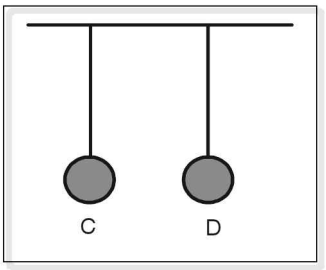


Figure 6.19

Thus, resonance can be defined as "the phenomenon in which when one of the two bodies having equal natural frequencies is set into vibration, then under its influence the other body also vibrates with increased amplitude".

Resonance in sound boxes

Consider two hollow boxes 'P' and 'Q' with one end open placed adjacent to each other, such that their open ends face each other. Two tuning forks 'A' and 'B' having equal frequencies are fixed on boxes 'P' and 'Q' respectively. A pith ball suspended from a hook is in contact with the fork 'B' as shown in figure (6.20).

When the fork 'A' is struck gently with a rubber hammer, it vibrates and after some time, we observe that the pith ball which is in contact with fork 'B' flies outward, indicating that fork 'B' too vibrates with maximum amplitude. At the same time, a loud booming sound is heard. This is due to the resonance between the two forks 'A' and 'B'.

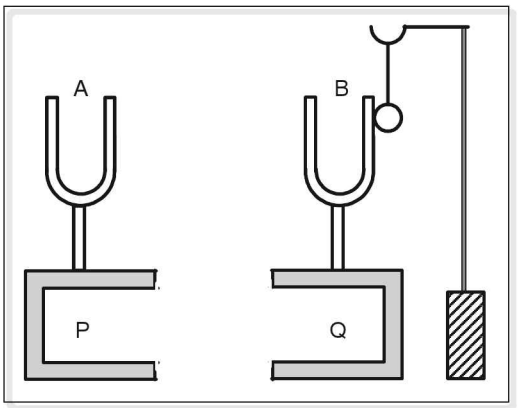


Figure 6.20 Resonance in Sound boxes

The vibrations of the fork 'A' produce forced vibrations in the box 'P', which produce forced vibrations in the air in the box 'Q'. These vibrations of air lead to vibrations in the box 'Q' and through it to the fork 'B'. Since the two forks, 'A' and 'B' are of equal frequency, maximum energy transfer takes place and thus 'B' is in resonance with 'A'. The vibrations of 'B' which are forced vibrations and in resonance with those of 'A', are also known as '**sympathetic vibrations**'.

Conditions for resonance

(a) The vibrations of a body causing resonance should be strong enough to set forced vibrations in the other body.

☛ *Example*

A single soldier marching on a rope bridge cannot cause sympathetic vibrations in the bridge, but a large group of soldiers marching to rhythmic steps on the bridge could cause the bridge to collapse.

(b) The natural frequency of the given body which is in resonance must be equal to either the natural frequency or an integral multiple of the natural frequency of the vibrating body that causes the resonance.

☛ *Example*

The resonance in air columns takes place only at certain lengths of the air columns that correspond to the resonating lengths.

Resonance—examples and applications

1. We observe a rattling sound in a bus when the natural frequency of the vibrations of the body of the bus equals the natural frequency of the vibrations of its engine. When the speed of the bus changes, the frequency of the vibrations of the engine changes. When this frequency of the vibrations of the engine equals the natural frequency of the vibrations of the body of the bus, both are in resonance and the body of the bus vibrates with larger amplitude producing a rattling sound.
2. When a group of soldiers march across a rope bridge, they are asked to march out of step. This is because, when they march in step across the bridge, the natural frequency of the rhythmic marching of soldiers may become equal to the natural frequency of the vibrations in the bridge. If this happens, resonance takes place between the marching soldiers and the bridge which then vibrates with maximum amplitude leading to its collapse.
3. A wine glass kept before a loud speaker vibrates to the musical notes emitted by the speaker. When the frequency of the musical note is equal to the frequency of the vibrating wine glass, resonance takes place and the glass vibrates with larger amplitude. This may topple the glass.
4. The basic principle involved in the transmission of radio and television programmes is resonance between the frequency of signals emitted from the station and the frequency of the radio or the television set. The resonance that takes place here is not of sound waves but of electromagnetic waves.

Resonance in air columns

Air column

A long tube about 40 cm in length and 6 cm in diameter, open at both the ends or closed at one end forms an 'air column'. If it is open at both the ends, it is called an 'open end tube'. If it is closed at one end, it is called a 'closed end tube'.

Resonating air column

When a tuning fork is excited with a rubber hammer and is held at the open end of an air column such that the vibrating prongs are perpendicular to the length of the air column, as shown in figure (6.21), the air in the tube vibrates (forced vibrations). If the natural frequency of the air column matches with that of the tuning fork, a loud sound is produced. This happens as the vibrating air column is in resonance with the given tuning fork. The air column in this condition is known as 'resonating air column'.

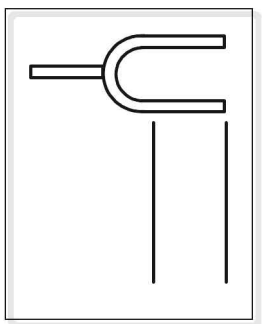


Figure 6.21

Closed end resonating air column

Consider a long cylindrical glass jar of length about 50 cm and 10 cm diameter, filled with water to more than $\frac{3}{4}$ th of its height as shown in the figure (6.22). Let a glass tube of smaller diameter (say 6 cm) and length approximately about 40 cm and open at both ends be inserted into the jar as shown in the figure (22). The length of the smaller tube above the water level in the jar can be adjusted, thus varying the length of the air column above the water surface in the smaller tube.

When a vibrating tuning fork is held horizontally above the open end of the smaller tube, it sets vibrations in the air in the tube, thus producing forced vibrations in the air column. The other end of the smaller tube being immersed in water jar, the water surface acts as a reflecting surface to the waves set in the air column. Thus, the smaller tube acts as a closed end tube.

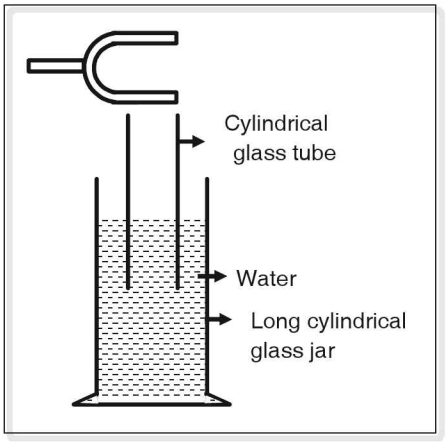


Figure 6.22

The compressions and rarefactions formed at the open end by the vibrating tuning fork travel towards the water surface, get reflected and again travel back towards the open end. So the incident and reflected waves superpose each other to form stationary longitudinal waves in the air column. At the open end of the air column, the particles are free to vibrate. So this end corresponds to a rarefaction. The closed end (water surface) is rigid and the air particles at this location are not free enough to vibrate longitudinally (along the length of the tube). So, at the closed end, a compression is formed.

The length of the air column can be adjusted and when it is equal to '$\frac{\lambda}{4}$' (where 'λ' is the wavelength of the sound wave) resonance occurs between the tuning fork and the air column. This can be confirmed by a loud (booming) sound heard. The length of the air column is the '**first resonating length**' of the air column. It is observed that resonance takes place whenever the length of the air column is an odd integral multiple of '$\frac{\lambda}{4}$'. The general expression for resonating length of an air column is given by $\ell_n = \frac{(2n-1)\lambda}{4}$ where n = 1, 2, 3, - - - . Thus, for a given tuning fork of certain frequency, the resonating lengths of air column are 1st resonating

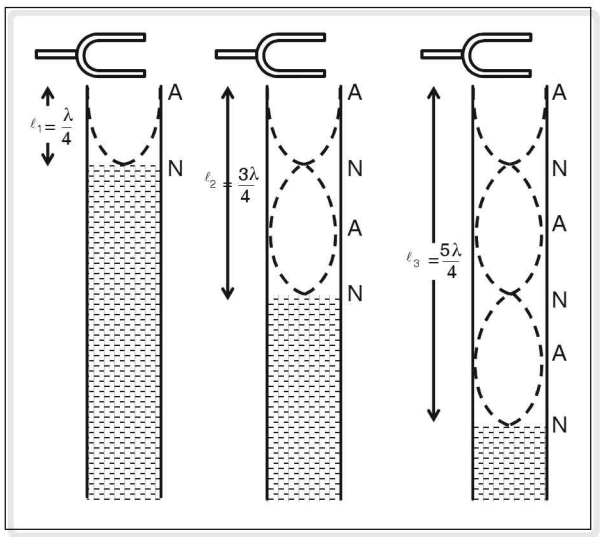


Figure 6.23 Resonating air columns

length, $\ell_1 = \frac{(2\times1-1)\lambda}{4} = \frac{\lambda}{4}$; 2nd resonating length, $\ell_2 = \frac{(2\times2-1)\lambda}{4} = \frac{3\lambda}{4}$; 3rd resonating length, $\ell_3 = \frac{(2\times3-1)\lambda}{4} = \frac{5\lambda}{4}$ and so on.

When the reflection of a sound wave takes place at the water surface in the closed tube, phase change does not take place at the instant of reflection. Thus, a compression is reflected as compression and a rarefaction is reflected as rarefaction. As these reflected compressions or rarefactions reach the open end of the tube, phase change occurs. The reflected compression becomes rarefaction and the reflected rarefaction becomes compression on reaching the open end. This occurs due to the volume changes of air taking place in the tube.

As a reflected rarefaction reaches the open end, due to low pressure, some of the air is pushed into the tube making a compression. Similarly, as a reflected compression reaches the open end, due to excess pressure, some of the air is pushed out of the tube creating a rarefaction. When such a rarefaction reaches the mouth of the tube, another rarefaction is produced by the tuning fork at the mouth and so these two rarefactions superpose giving the maximum amplitude and an antinode is formed at the mouth of the tube.

Similarly, when a compression reaches the mouth of the tube, another compression is produced by the tuning fork at the mouth of the tube and so these two compressions superpose giving the maximum amplitude and so an antinode is formed at the mouth of the tube.

Thus "whenever resonance occurs between the vibrating air column and the tuning fork, an antinode is formed at the mouth of the tube and a node is formed at the closed end of the tube".

Resonance—velocity of sound in air

Consider the glass jar and the glass tube shown in figure (6.22). Hold a vibrating tuning fork of frequency 'n' horizontally over the mouth of the tube, having an air column of less than 10 cm. Gradually, increase the length of the air column by raising the tube and at one particular length of the air column, say 'ℓ_1', we hear a booming sound indicating that the length 'ℓ_1' of the air column is in resonance with the given tuning fork.

This is the first resonating length and so $\ell_1 = \frac{\lambda}{4}$. On further increasing the length of the air column, holding the vibrating tuning fork at the mouth of the tube, we get another booming sound for the air column of length, say 'ℓ_2' such that this is the second resonating length of the air column. So $\ell_2 = \frac{3\lambda}{4}$.

Thus $\ell_2 - \ell_1 = \frac{3\lambda}{4} - \frac{\lambda}{4} = \frac{2\lambda}{4} = \frac{\lambda}{2}$ $\therefore \lambda = 2\,(\ell_2 - \ell_1)$. If 'n' is the frequency of the tuning fork, and 'v' is the velocity of sound in air, we have $v = n\lambda = n2(\ell_2 - \ell_1)$ $\therefore v = 2n(\ell_2 - \ell_1)$. From the above expression, we can find out the velocity of sound in air at room temperature.

Resonance in open organ pipe

Resonance of sound can take place in open organ pipes also. An open organ pipe is a tube with both ends open, and in this case both the ends correspond to antinodes. For fundamental frequency one node is formed at the centre of the pipe and the length of the tube (ℓ) corresponds to half the wavelength, being the distance between two consecutive antinodes.

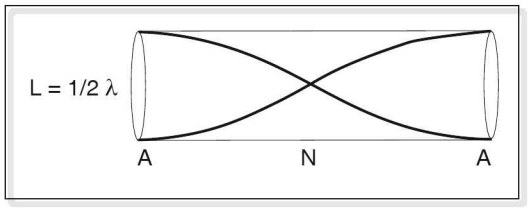


Figure 6.24

Thus we have,

$$\ell = \frac{\lambda}{2} \text{ or } \lambda = 2\ell$$

since $v = \upsilon\lambda$,

$$\upsilon = \frac{v}{\lambda} = \frac{v}{2\ell}$$

This frequency is called the fundamental frequency and is denoted as $\upsilon_o = \frac{v}{2\ell}$

For the first overtone or second harmonic, two nodes are formed inside the tube. In this case the length of the tube corresponds to λ, the wavelength, as shown in the figure.

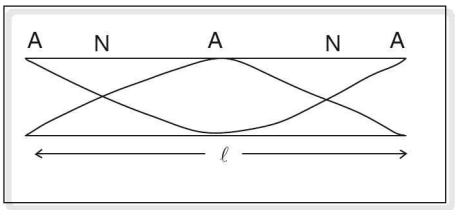


Figure 6.25

$\therefore \ell = \lambda$ and

$$\upsilon_1 = \frac{v}{\ell} = 2\upsilon_o$$

Hence, υ_1, the frequency of the first overtone is twice the fundamental frequency, υ_o.

The second overtone or the third harmonic is produced when three nodes are formed inside the tube as shown in the figure. A total of four antinodes are formed such that ℓ, the length of the tube corresponds to $\frac{3\lambda}{2}$. Thus we have

$$\ell = \frac{3\lambda}{2} \text{ or } \lambda = \frac{2\ell}{3} \text{ and}$$

$$\upsilon_2 = \frac{v}{\lambda} = \frac{3v}{2\ell} = 3\upsilon_o$$

Therefore frequency of the second overtone is thrice the fundamental frequency.

Thus the frequencies in an open organ pipe are in the ratio 1 : 2 : 3 indicating that all harmonics are present, whereas in a closed-end pipe the ratio of the frequencies is 1 : 3 : 5..... showing that only the odd harmonics are present.

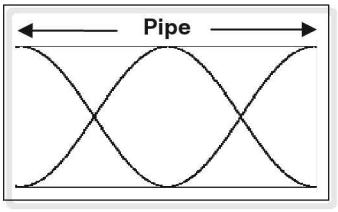


Figure 6.26

Vibrations in stretched strings

A string is a thin wire or a thread which is inextensible. When such a string is stretched by applying a force that acts along the length of the wire, the force applied is known as 'tensile force' or 'tension'. When a stretched wire is plucked, stationary waves perpendicular to its length are produced.

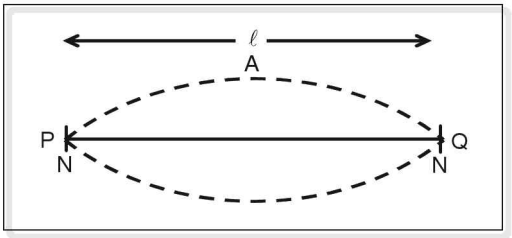


Figure 6.27 Stationary waves in strings; string vibrating in fundamental mode

Consider a string stretched between two points 'P' and 'Q'. When the string is plucked at the middle, a transverse stationary wave is set up in the string with nodes (N) at the fixed end points 'P' and 'Q' and an antinode (A) at the middle of the string, as shown in the figure (6.27).

The distance between two successive nodes or antinodes in a stationary wave is equal to half the wavelength $\left(\frac{\lambda}{2}\right)$. If '$\ell$' is the length of the vibrating string between the fixed ends 'P' and 'Q', then $\ell = \frac{\lambda}{2}$ or $\lambda = 2\ \ell$. If 'n' is the frequency of vibration and 'v' is the velocity of the wave set in the string then, $v = n\lambda \Rightarrow v = n\ (2\ell)$.

But the velocity of the wave set in the string is given by $v = \sqrt{\frac{T}{m}}$ where 'T' is the tension in the string and 'm' is mass per unit length (linear density) of the string.

$$\therefore \sqrt{\frac{T}{m}} = n(2\ell) \quad \text{or} \quad n = \frac{1}{2\ell}\sqrt{\frac{T}{m}} \rightarrow (1)$$

If the string described above is considered to be placed on a table and stretched between two fixed points 'P' and 'Q' and a vibrating tuning fork of frequency 'n' is kept in between the fixed points 'P' and 'Q' as shown in figure (6.28), resonance takes place in the stretched string between 'P' and 'Q', and the tuning fork. A stationary wave is set up in the string as shown in figure (6.27). Then the length 'l' of the string between the fixed points 'P' and 'Q' is called the resonating length.

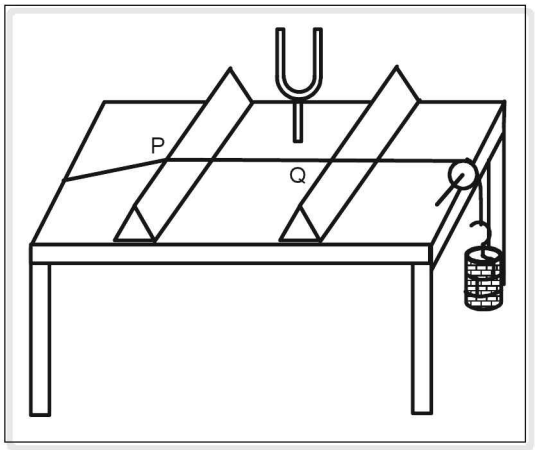


Figure 6.28

This resonating length of the string is said to be vibrating in fundamental mode and the corresponding frequency of the vibrating string, given by $n = \frac{1}{2\ell}\sqrt{\frac{T}{m}}$, is known as the fundamental frequency or first harmonic.

The length 'ℓ' of the stretched string is found to be in resonance with another higher frequency 'n_1' with the tension (T) and linear density (m) being constant. Then, the string is found to be vibrating in two segments, with two antinodes (A) and three nodes (N) as shown in figure (6.29). Then, the length of the vibrating stretched string is $\ell = \lambda$ and so frequency

$$n_1 = \frac{1}{\ell}\sqrt{\frac{T}{m}} = \frac{2}{2\ell}\sqrt{\frac{T}{m}} = 2\left(\frac{1}{2\ell}\sqrt{\frac{T}{m}}\right) = 2n \rightarrow (2).$$

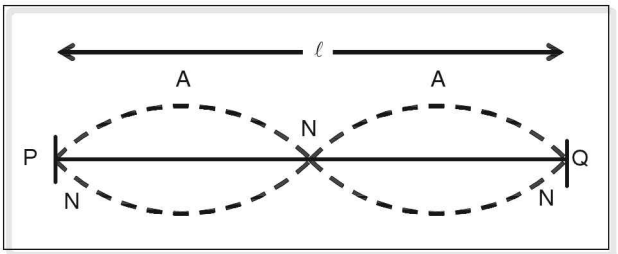


Figure 6.29 String vibrating in second harmonic

So, this frequency 'n_1' is double the fundamental frequency and is called 'second harmonic' or 'first overtone'. We get similar resonating lengths whenever the length of string $\ell = \frac{P\lambda}{2}$ where P = 1, 2, 3 - - - -.

So, the corresponding frequencies obtained are tabulated as follows.

1.	$\frac{\lambda}{2}$	$\frac{2\ell}{1}$	$n = \frac{1}{2\ell}\sqrt{\frac{T}{m}}$	One	Fundamental frequency or 1st harmonic
2.	$\frac{2\lambda}{2}$	$\frac{2\ell}{1}$	$n_1 = \frac{1}{\ell}\sqrt{\frac{T}{m}} = \frac{2}{2\ell}\sqrt{\frac{T}{m}} = 2n$	Two	2nd harmonic or 1st overtone
3.	$\frac{3\lambda}{2}$	$\frac{2\ell}{3}$	$n_2 = \frac{1}{\left(\frac{2\ell}{3}\right)}\sqrt{\frac{T}{m}} = \frac{3}{2\ell}\sqrt{\frac{T}{m}} = 3n$	Three	3rd harmonic or 2nd overtone
K	$\frac{K\lambda}{2}$	$\frac{2\ell}{K}$	$n_{k-1} = \frac{K}{2\ell}\sqrt{\frac{T}{m}} = Kn$	K	Kth harmonic or (K – 1)th overtone

Laws of Vibrating Stretched Strings

The expression for the fundamental frequency of a stretched vibrating string is given by

$n = \frac{1}{2\ell}\sqrt{\frac{T}{m}}$. From this, three laws are derived for the frequency of a vibrating stretched string and are as follows.

1. **Law of length:** The frequency of a vibrating stretched string, vibrating in a fundamental mode, is inversely proportional to its length, when its tension and linear density are constant. Mathematically, $n \propto \frac{1}{\ell}$ when 'T' and 'm' are constant. $\Rightarrow n\ell$ = constant.

   If 'ℓ_1' and 'ℓ_2' are the lengths of two strings vibrating in the fundamental frequency having their linear density and tension constant and 'n_1' and 'n_2' are their fundamental frequencies respectively, then $n_1 \ell_1 = n_2 \ell_2$

2. **Law of tension:** The frequency of a vibrating stretched string, vibrating in a fundamental mode, is directly proportional to the square root of its tension, when length and linear density remain constant. Mathematically, $n \propto \sqrt{T}$, when 'ℓ' and 'm' are constant.

   $\Rightarrow \frac{n}{\sqrt{T}}$ = constant.

   If 'T_1' and 'T_2' are the tensions of two strings whose length and linear density are constant and vibrating in the fundamental mode and 'n_1' and 'n_2' are their fundamental frequencies respectively, then $\frac{n_1}{\sqrt{T_1}} = \frac{n_2}{\sqrt{T_2}}$

3. **Law of linear density:** The frequency of a vibrating stretched string, vibrating in a fundamental mode, is inversely proportional to the square root of its linear density, when its length and tension are constant. Mathematically, $n \propto \frac{1}{\sqrt{m}}$ when 'ℓ' and 'T' are constant.

   $\Rightarrow n\sqrt{m}$ = constant.

   If 'm_1' and 'm_2' are the linear densities of two strings vibrating in fundamental mode having equal length and tension and 'n_1' and 'n_2' are their fundamental frequencies respectively, then $n_1\sqrt{m_1} = n_2\sqrt{m_2}$.

Sonometer

It is a hollow wooden box about one metre long on which a peg is fixed at one end and a smooth frictionless pulley is fixed at the other end on the top surface which is also called as a sonometer table. Wedge-shaped wooden pieces known as bridges are placed on the table and they are movable. At one end of the Sonometer table, along its length, a metre scale is fixed. A string is tied at one end to the peg and other end is passed over the two bridges and the pulley, and attached to a weighted hanger. The weights attached to the hanger provide tension in the string. On the lateral walls of the sonometer box, holes are provided.

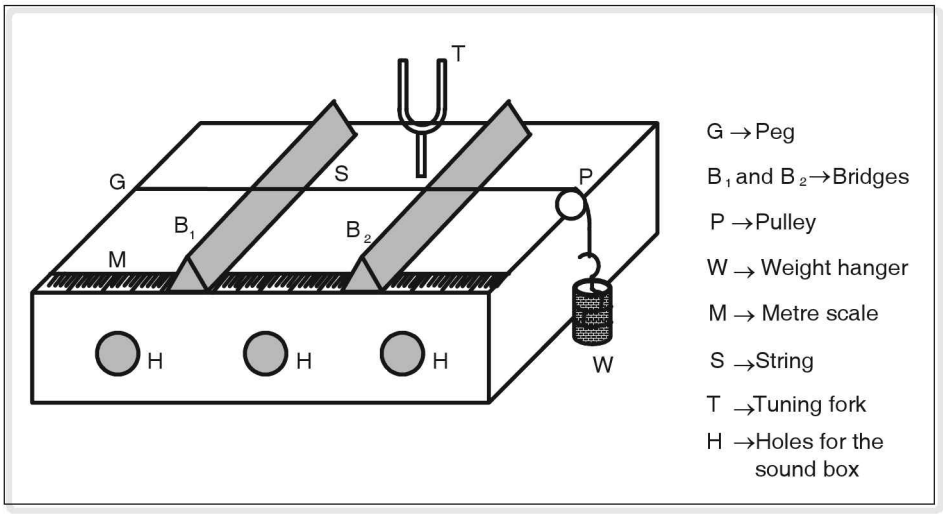


Figure 6.30 Sonometer

Sonometer is used to verify the laws of transverse vibrations in stretched string. Excite a tuning fork of frequency 'n_1' and keep it on the sonometer table between the bridges as shown in figure (6.30). Adjust the distance between the bridges to change the length of the vibrating string between the bridges (from minimum to maximum). At a particular length, say 'ℓ_1', the wire between the bridges vibrates with maximum amplitude and this can be confirmed by the flying away of a small paper rider of light weight and inverted V–shape (Λ) placed on the string between the bridges. Then, the length 'ℓ_1' of the wire is in resonance with the tuning fork. The procedure is repeated by taking another tuning fork of frequency 'n_2' and the resonating length 'ℓ_2' is obtained. The first law of vibrating stretched strings is verified by showing $n_1\ell_1 = n_2\ell_2$.

Music and noise

Sound that causes a pleasing sensation to ears is called music. Sound that causes an unpleasant or irritating sensation to ears is called noise. A sound cannot be strictly classified into either music or noise, as a sound that causes a pleasing sensation to one person can cause an unpleasant sensation to another person. But based on certain characteristics, we can broadly classify any sound note into either music or noise.

Characteristics of musical sound

1. Pitch

It is a special characteristic of sound based on which we can differentiate between two sounds having equal loudness. The pitch of a given sound can be measured by its frequency.

☛ *Example*

Female voices in general are high pitched compared to male voices though both the female voice and the male voice may have equal loudness.

2. Intensity

Intensity is a measure of loudness of sound. It is defined as the energy that passes through unit area per unit time.

The following are the factors on which the intensity of sound depends.

(a) Amplitude: The intensity of a sound wave varies directly as the square of its amplitude. Mathematically, intensity (I) $\propto$ (amplitude)^2.

(b) Distance from source: The intensity of a sound wave varies inversely as the square of the distance of the point of observation from the source.

Mathematically, intensity (I) $\propto \frac{1}{(\text{distance})^2}$

(c) Density: The intensity of a sound wave varies directly as the density of the medium through which it is propagated.

Mathematically, intensity (I) $\propto$ density (ρ).

Note: The intensity level of sound is measured in units of decibels (dB).

A normal human being can bear sound having an intensity level of maximum 80 dB. Sounds having intensity level less than 10 dB and more than 180 dB are inaudible.

3. Quality or timbre

The quality or the timbre of a musical sound is a special characteristic by means of which two sounds having equal pitch or loudness can be differentiated. Two musical notes having equal loudness and pitch, one originating from one instrument, say veena; and the other produced from another instrument, say violin; can be identified separately even though they are produced simultaneously. The wave form of the sound note produced signifies its quality.

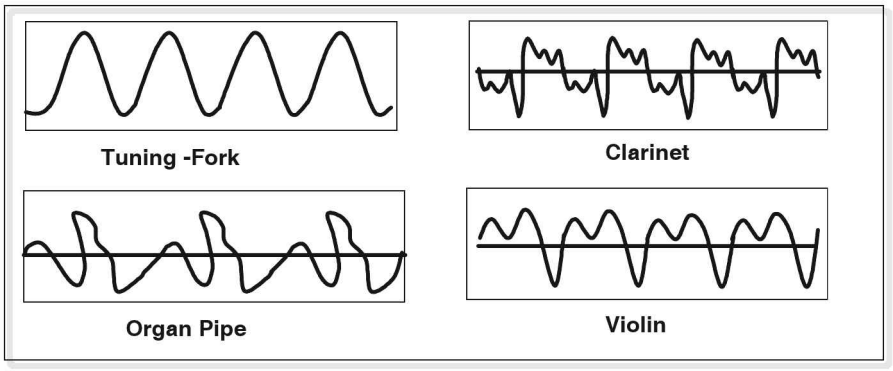


Figure 6.31

Differences between music and noise

1. Music produces a pleasing effect whereas noise produces an irritating effect in our ears.
2. Music has variations in amplitude or frequency at regular intervals of time whereas in noise, the variations in amplitude or frequency are abrupt.

3. Musical sound has a definite wave form which repeats itself in regular intervals of time; whereas noise does not have any regular or definite wave form.

Drum

It is an instrument that produces sound by vibrating membrane. It is a hollow cylindrical box to which the processed skin of animals (leather) is attached at sides like a membrane.The membranes are stretched and then excited at any point on the membrane to produce sound.

Bell

A cone shaped metallic object, which produces sound when excited is known as a bell. It has a complex mode of vibration.When a bell is struck at its edge, it is divided into an even number of vibrating sectors and it produces transverse stationary waves.

Pollution due to noise

Noise in general disturbs the regular work, it causes strain, irritation. Thus noise is undesirable and uncomfortable and can be considered as pollution.

The undesirable unaudible, harsh and irritating sound produced in the surroundings is called noise pollution.

The noise which can be called pollution depends on its (1) loudness (2) duration.

Effects of noise pollution

1. It disturbs the normal work, conversations.
2. It produces irritation, strain, headache etc., thus everybody can loose their concentration.
3. Very loud noise may cause temporary or permanent deafness.
4. Noise may also create night blindness and colour blindness.

Sources of noise pollution

Grinders, mixers, vacuum cleaners using at home, loud speakers, exploding crackers in various functions, hawkers in streets, various machines used in factories create a large amount of noise in our surroundings. Along with these various transportation means such as trains, buses, air craft, ships etc., also create noise pollution.

Controlling of noise pollution

Nowadays, in modern society it is not easy to eliminate the noise pollution but by taking some measures and precautions it can be decreased.Various means that can be followed are:

1. The volume levels of various audio, video systems should be controlled.
2. The factories and various industries should be constructed very far from residential areas.
3. It is better to ban the various musical entertainments in public places.
4. The vehicles that produces loud sound should not be allowed in residential areas.
5. All machines, automobiles, generators, motors should be designed such that no noise can be produced.

Recording and reproduction of sound

Sound can be recorded and can be reproduced, when it is desired. Microphone is an instrument which can convert sound vibrations into electrical signals. These electrical signals can be converted back to sound by using loud speakers. The recording of sound can be done either on magnetic tapes or photosensitive films.

Recording and reproduction of sound by using magnetic tapes

The tape is very thin, flexible and usually made of polyvinyl chloride. One side of it is coated with ferric oxide (Fe_2O_3) paste. Before recording any sound the molecules of magnetic paste align such that the net magnetic moment is zero.

Sound recording

The major part of the sound recording equipment is the recording head. Recording head is a soft iron ring having a narrow gap. It is wound by an insulated conducting wire. The two ends of which are connected to an amplifier which in turn is connected to a microphone. The unrecorded magnetic tape is made to run in front of the narrow gap of recording. Microphone converts the sound vibrations created in front of it into electrical signals which are amplified by the amplifier.

When these signals are made to pass through the wire wound around the recording head, it acts as an electromagnet. Because of the varying electrical signals the magnetic field created in the narrow gap also varies and accordingly the molecules of ferric oxide realign. The amount of realignment depends on the strength of magnetic field produced.

Sound reproduction

In sound reproduction process equipment, the major part is playback head which is similar to the recording head used in the recording process. The ends of insulated conducting wire wound around the play back head is connected to loud speaker through an amplifier. During the reproduction the recorded tape is made to run in front of the narrow gap of the playback head. The varying magnetic impressions recorded on the tape create a varying magnetic field in the gap of the head. This produces electrical signals in the wire wound around the head. These signals are amplified by amplifier and sent to a loud speaker which converts electric signals into sound vibrations.

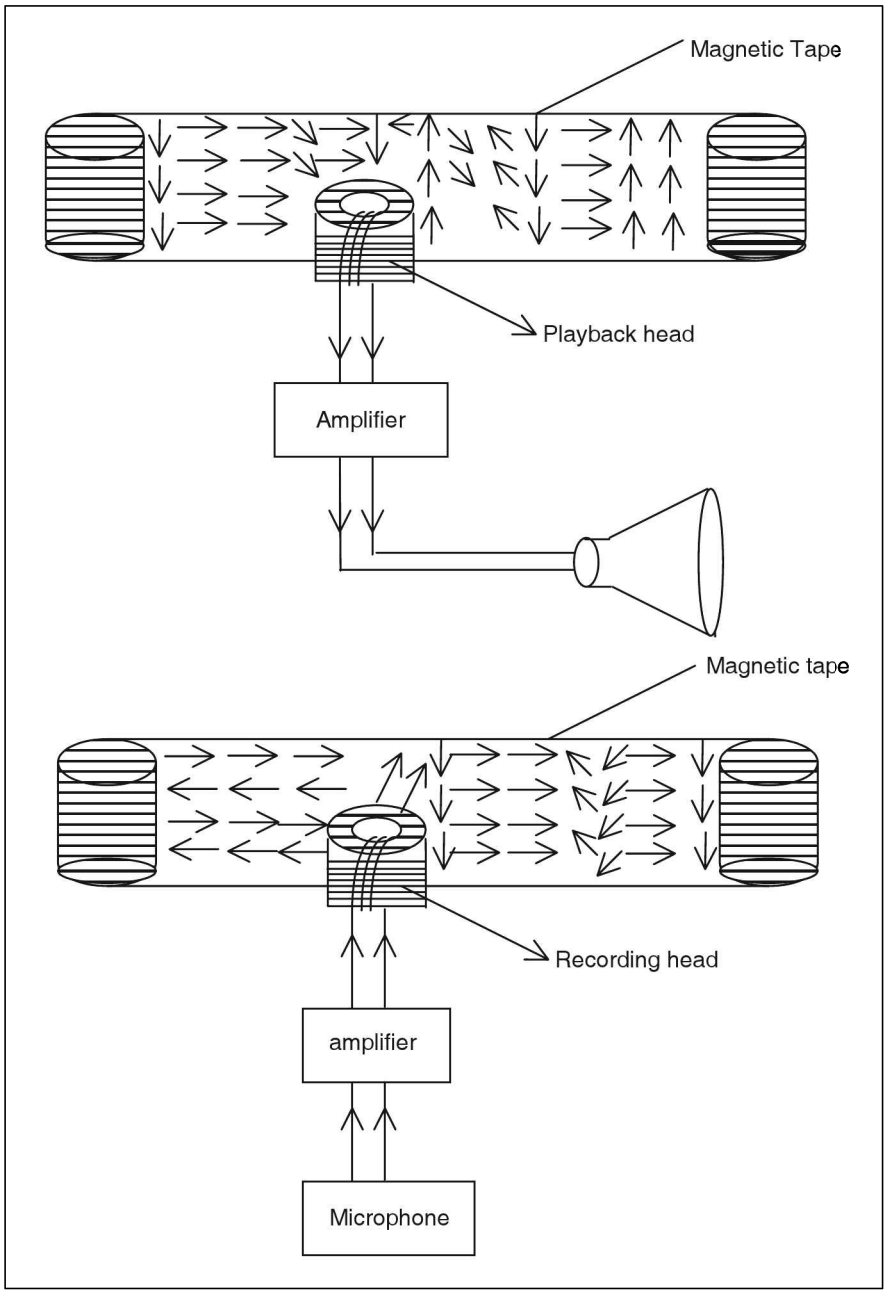


Figure 6.32

Recording and reproduction of sound by using films

The sound is recorded on the photosensitive film in the form of zig-zag patches called sound track. In this method a phosphor-bronze loop is suspended between the concave poles a magnet. The loop is attached by a small mirror. The mirror reflects the light coming from an intense source towards a narrow slit behind which the photosensitive film is made to run at constant speed.

The sound to be recorded is converted into electrical signals by a microphone and is passed through a phosphor-bronze loop. Due to these signals the loop and the mirror deflect in different orientations. Depending on the amount of deflection of the mirror a zig-zag patch of different widths are formed on the film.

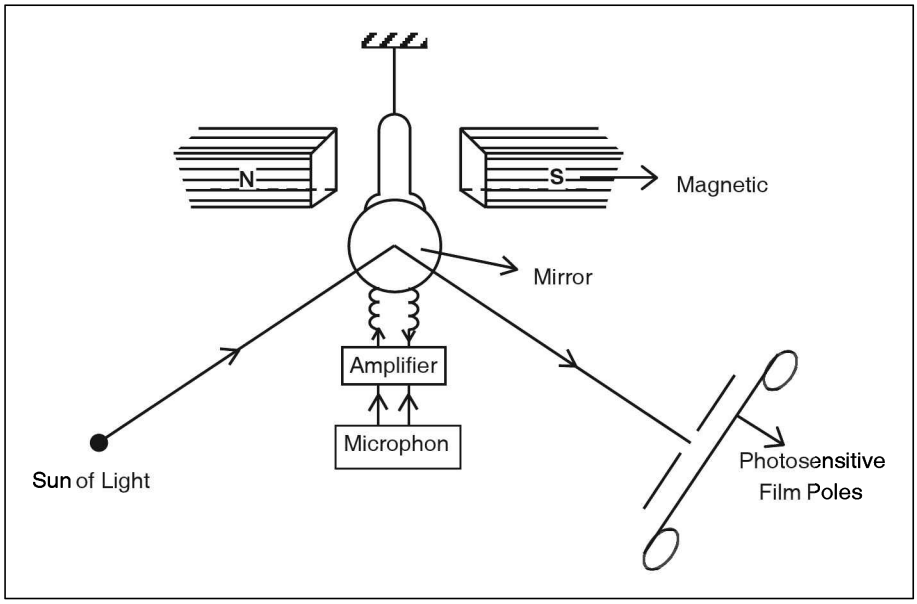


Figure 6.33

Sound reproduction

The sound track is exposed to the light coming from an intense source after passing through a converging lens system. The emergent light is allowed to fall on a photo cell, which works on the principle of photo electric effect. Because of variations in the intensity of light emerging from film, photo cell produces varying electric signals. These are converted into sound by using a loud speaker.

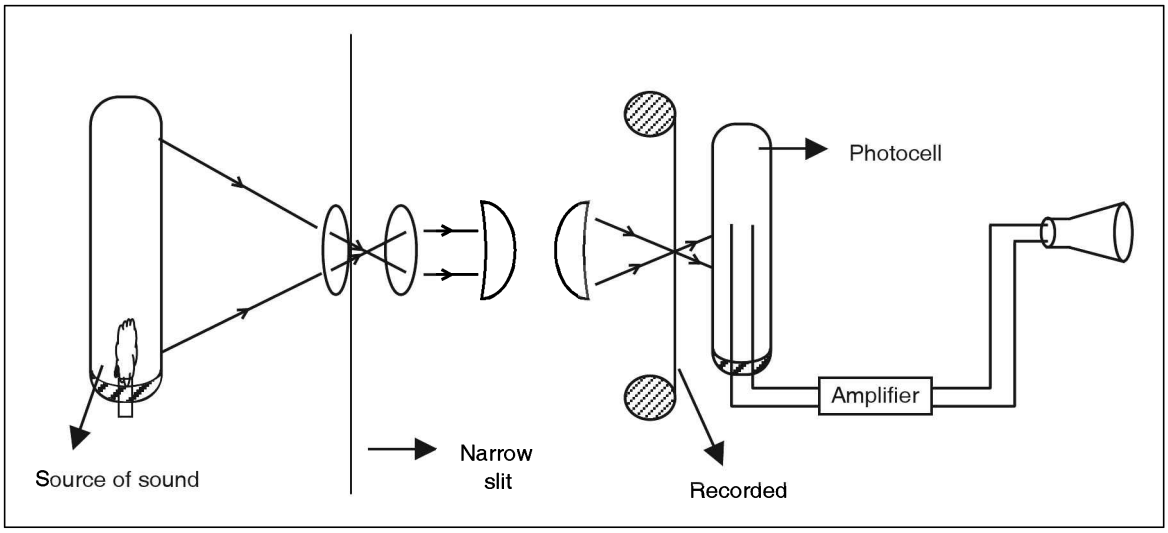


Figure 6.34

☛ *Solved Examples*

1. The ratio of specific heat at constant pressure to that constant volume (γ) of a gas is 1.5. The density of gas is given as 2×10^{-3} g cm^{-3}. Given that the pressure is equivalent to 80 cm of Hg. Find out the velocity of sound in the gas (taking g = 1000 cm s^{-2}).

Solution

$\gamma = 1.5$

P = 80 cm of Hg

$= 80 \times 13.6 \times 1060$

$= 1088 \times 10^3$ dynes cm^{-2}

$d = 2 \times 10^{-3}$ g cm^{-3}

Velocity of sound, $V = \sqrt{\frac{\gamma P}{d}}$

$$= \sqrt{\frac{1.5 \times 1088 \times 10^3}{2 \times 10^{-3}}}$$

$= 28.5 \times 10^3$ cm s^{-1}

$= 285$ m s^{-1}

2. The Young's modulus of a solid is given by 6.25×10^{11} dynes cm^{-2}. If the velocity of sound in the solid is 2500 m s^{-1}, find its density.

Solution

Velocity of sound in a solid

$$V = \sqrt{\frac{Y}{d}}$$

$$\Rightarrow \text{density, } d = \frac{Y}{V^2} = \frac{6.25 \times 10^{11}}{(250000)^2}$$

$$= \frac{625 \times 10^9}{625 \times 10^8} = 10 \text{ g cm}^{-3}$$

3. The velocity of sound in oxygen is 100 m s^{-1}. Find the velocity of sound in hydrogen.

Solution

Velocity of sound in a gas,

$$V \propto \frac{1}{\sqrt{\text{Molecular weight M}}}$$

$$\frac{V_1}{V_2} = \sqrt{\frac{M_2}{M_1}}$$

$$\frac{V}{V_{O_2}} = \sqrt{\frac{M_{O_2}}{M_{H_2}}} = \sqrt{\frac{32}{2}} = \sqrt{16} = 4.$$

$V_{H_2} = 4 \times 4V_{O_2}$

$= 4 \times 100 = 400 \text{ m s}^{-1}$

4. A person standing at a distance 200 m from a hill blows a horn. After what time can he hear the echo? (Given that the velocity of sound in air is 300 ms^{-1}).

Solution

Let 'd' be the distance of reflection from source then $V = \frac{2d}{t}$

't' is time taken to receive the echo.

$$t = \frac{2d}{v} = \frac{2 \times 200}{300} = \frac{400}{300} = \frac{4}{3} \text{s}.$$

5. A person trying to determine the velocity of sound in air by the echo method produces a sound at a certain time and hears the echo after 10 seconds. If the reflector is at a distance of 1700 m. Find the velocity of sound in air.

Solution

Velocity of sound, $V = \frac{2d}{t}$

$$= \frac{3400}{10} = 340 \text{ m s}^{-1}$$

6. In a resonating air column method the first and second resonating lengths are recorded as 25 cm and 75 cm respectively. If the velocity of sound is 300 m s^{-1}, find the frequency of tuning fork used.

Solution

$V = 300 \text{ m s}^{-1}$

$\ell_1 = 25 \text{ cm} = 0.25 \text{ m}$

$\ell_2 = 75 \text{ cm} = 0.75 \text{ m}$

$V = 2n(\ell_2 - \ell_1)$

$$n = \frac{v}{2(\ell_2 - \ell_1)} = \frac{300}{2(0.75 - 0.25)} = \frac{300}{2 \times 0.5}$$

$$= \frac{300}{\ell} = 300 \text{ Hz}.$$

7. In a sonometer experiment the mass suspended from the string is 1 kg. The resonating length for a given tuning fork is noted as 20 cm. If the mass suspended is increased to 4 kg, find the resonating length of the string for the same frequency.

Solution

In a stretched string

$$n = \frac{1}{2\ell}\sqrt{\frac{T}{m}}$$

If n is a constant and it is the linear mass density in kg m^{-1}.

$$\frac{\sqrt{T}}{\ell} = \text{constant}$$

$$\therefore \frac{\sqrt{T_1}}{\ell_1} = \frac{\sqrt{T_2}}{\ell_2}$$

$T_1 = 1 \text{ kg} \times 10 = 10 \text{ N}$

$T_2 = 4 \text{ kg} \times 10 = 40 \text{ N}$

$$\frac{\ell_2}{\ell_1} = \sqrt{\frac{T_2}{T_1}} = \sqrt{\frac{40}{10}} = \sqrt{4} = 2$$

$\ell_2 = 2\,\ell_1 = 2 \times 20 = 40 \text{ cm}$

test your concepts

Very short answer type questions

1. Define echo.
2. What are forced vibrations? Give an example.
3. In resonance, the body undergoing forced vibrations vibrates with _____ amplitude.
4. What is meant by the fundamental frequency of a vibrating stretched string?
5. How does humidity in air affect the velocity of sound in it?
6. What is the expression for the velocity of sound in air using resonating air column method?
7. How does the intensity of sound vary with its amplitude and density of a medium?
8. The energy carried by a progressive wave is the sum of _____ and _____ of the vibrating particles of the medium.
9. Define audible range.
10. What are damped vibrations? Give an example.
11. Define transverse and longitudinal waves.
12. When a tuning fork of 512 Hz is used, the first and the second resonating lengths of closed air column are obtained as 16 cm and 48 cm respectively. What is the velocity of sound in air in metre per second?

13. What is the characteristic of sound that distinguishes between two sounds of equal amplitude and frequency?

14. What is the expression for the velocity of a stationary wave set in a stretched string?

15. Define progressive and stationary waves?

16. What are infrasonics and ultrasonics?

17. The time taken to reduce the intensity of sound to_____ of the initial value is called reverberation time.

18. What is resonating air column?

19. Mention Newton's formula for the velocity of sound in a medium, and Laplace correction to Newton's formula.

20. What is the pitch of a sound?

21. How does the intensity of sound vary with the distance of the point of observation from the source of sound?

22. What is the principle used in a mega-phone?

23. How does the velocity of sound in a gas vary with absolute temperature?

24. Define reverberation.

25. Hertz is the unit of _____.

26. How many antinodes are obtained when a closed end resonating air column vibrates in the fundamental mode?

27. How does the velocity of sound in a gas vary with its molecular weight, and pressure?

28. If the fundamental frequency of a vibrating stretched string is 200 Hz, what are its 1st and 2nd overtones?

29. What are natural vibrations? Give an example.

30. By changing either the wavelength or the frequency of a sound wave, is it possible to change the velocity of sound in a medium?

Short answer type questions

31. Explain resonance with an example?

32. A string is stretched by a tension of 200 N. If its linear density is 2×10^{-2} kg m^{-1}, what is the velocity of waves set in the string?

33. Explain the formation of stationary waves in a stretched string.

34. In a resonating air column method, the first and the second resonating air columns are 16 cm and 48 cm respectively. If the frequency of a tuning fork used is 500 Hz, find the velocity of sound in air.

35. The first resonating length of a closed end air column is 30 cm. At approximately, what length of the air column do we obtain the second resonance?

36. Differentiate between progressive and stationary waves.

37. The fundamental frequency of an air column in a closed end tube is 30 Hz. Find the frequency of the third harmonic.

38. If the velocity of sound in air is 300 m s^{-1}, express the audible range in terms of wavelength.

39. Explain the factors on which the velocity of sound in a given gas depends.

40. The velocity of the sound in a gas is 300 m s^{-1}. The density of gas and the ratio of specific heats of gas are given by 0.002 g cm^{-2} and 2 respectively. Find the pressure of the gas.

41. If the velocity of sound in air is 330 m s^{-1}, what is the minimum distance between the source or the observer of sound and the reflector to hear an echo?

42. What are the factors which do not affect the velocity of sound in a gas?

43. The molecular weights of two gases A and B are given as 36 units and 16 units respectively. If the velocity of sound in gas A is 300 m s^{-1}, find the velocity of sound in B.

44. The velocity of sound in a gas at 0°C is x m s^{-1}. At what temperature of the gas, does the velocity of sound become 2x m s^{-1}?

45. How do you determine the velocity of sound in air using an echo?

Essay type questions

46. Explain the formation of transverse stationary waves in stretched strings and their different modes of vibration.

47. Explain resonance in sound boxes.

48. Explain resonance in air columns.

49. Explain the laws of transverse vibrations in stretched strings.

50. Describe an experiment to find out the velocity of sound in air using resonating air columns.

CONCEPT APPLICATION



Concept Application Level—1

Direction for questions 1 to 7: State whether the following statements are true or false.

1. The frequency of a note produced by a stretched string is directly proportional to the square root of linear mass density of the string.

2. The phenomenon of resonance is used in receiving the radio and television programmes.

3. In order to transmit sound waves, the medium should possess elasticity and inertia.

4. In simple harmonic motion, the acceleration is inversely proportional to the displacement of the body from its mean position.
5. The frequency of damped vibrations remains same but the amplitude decreases gradually.
6. In a standing wave, the amplitude is maximum at nodes.
7. In an open pipe, the pressure variation at the ends of the pipe is maximum.

Direction for questions 8 to 14: Fill in the blanks.

8. At a given temperature, the velocity of sound in hydrogen is _____ than its velocity in oxygen.
9. γ is the ratio of molar specific heat capacity of the given gas at ____ to its specific heat capacity at ____ .
10. The pitch of a given sound can be measured by its _____.
11. The pitch of the sound produced by a stretched string gets higher as the tension_____.
12. In a standing wave, the distance between two successive antinodes is 15 cm. The distance covered by it in producing 4 waves is ____ m.
13. A rope is vibrating with a frequency 20 Hz. If the velocity of the wave in the rope is 20 m s^{-1}, then the distance between a crest and the successive trough is _____.
14. If d is the distance between the source of sound and reflector, t_1, t_2 are times at which echoes are heard, the expression for determination of velocity of sound in air by echo method is _____ .

Direction for question 15: Match the entries in column A with appropriate ones from column B.

15.

A.	Elastic waves	( )	a.	nodes and antinodes
B.	Infrasonic sound waves	( )	b.	reflection of sound
C.	Phase	( )	c.	laws of transverse vibrations in a stretched string
D.	Stationary waves	( )	d.	water waves
E.	SHM	( )	e.	Sound track
F.	Newton's formula	( )	f.	frequency less than 20 Hz
G.	Whispering galleries	( )	g.	measured in radians
H.	Damped vibrations	( )	h.	Isothermal change
I.	Sonometer	( )	i.	constant time period
J.	Photosensitive films	( )	j.	decrease in amplitude

Direction for questions 16 to 30: For each of the questions, four choices have been provided. Select the correct alternative.

16. A man moving towards a vertical cliff at a constant velocity of u m s^{-1}, fires a gun and hears the echo after t s. If he was at a distance of d m from the cliff when he fired the gun, the velocity of sound in air is _____ m s^{-1}.

(1) $\frac{2d}{t}$ (2) $\frac{2d}{t} + u$ (3) $\frac{2d}{t} - u$ (4) $\frac{d}{t}$

17. For a body executing simple harmonic motion,

(1) acceleration of the body is always directed towards the mean position.

(2) acceleration of the body directly proportional to the displacement from the mean position.

(3) the total mechanical energy is constant at any point on its path.

(4) All the above

18. In a resonating air column, the first booming sound is heard when the length of air column is 10 cm. The second booming sound will be heard when length is

(1) 20 cm (2) 30 cm (3) 40 cm (4) None of the above

19. The general expression for the resonating length of an air column in organ pipes with one end closed is given by _____.

(1) $\frac{(2n-1)\lambda}{2}$ (2) $\frac{n\lambda}{4}$ (3) $\frac{(2n-1)\lambda}{4}$ (4) $\frac{(n+1)\lambda}{2}$

20. If the amplitude of a sound is doubled, then its

(1) velocity is doubled (2) frequency is doubled

(3) wavelength is doubled (4) None of the above

21. In a resonating air column, the second resonating length is 60 cm. The first resonance occurs when the length of the air column is _____ cm.

(1) 30 (2) 10 (3) 20 (4) 60

22. In a closed end organ pipe, if ℓ_1 and ℓ_2 are two successive resonating lengths, then $2(l_2 - l_1)$ is ____.

(1) $\frac{\lambda}{2}$ (2) λ (3) 2λ (4) None of these

23. In a stationary wave,

(1) energy is carried away throughout the infinite medium

(2) energy is confined to a limited region of the medium

(3) different particles in a loop have different amplitudes

(4) Both (2) and (3)

24. The energy of the wave that passes through unit area per unit time is called as _____.

(1) pitch (2) quality (3) intensity (4) amplitude

25. If the velocity of sound in a gas is 200 m s^{-1} when the temperature is 127°C, then the velocity of sound in the same gas when temperature is increased by 900°F is _____ m s^{-1}.

(1) $200\sqrt{3}$ (2) 300 (3) $100\sqrt{5}$ (4) 450

26. A normal human being can hear sound having an intensity level of maximum _____.

(1) 50 dB (2) 80 dB (3) 100 dB (4) 150 dB

27. The ratio of densities of two media is 5 : 2. Then, the ratio of intensities of a given sound wave in the two media is

(1) 25 : 4 (2) 2 : 5 (3) 5 : 2 (4) 1 : 1

28. If the tension in a string is doubled and its linear density is halved, then the velocity of the wave in the string

(1) is quadrupled (2) is halved (3) remains constant (4) is doubled

29. When a compression propagates in a resonating tube, the phase change when it is reflected from the water surface is

(1) π radians (2) $\frac{\pi}{2}$ radians (3) $\frac{3\pi}{2}$ radians (4) zero radians

30. A man blows a whistle and hears an echo after 1.2 seconds. The distance between the man and the reflector of the sound is _____ m. (take velocity of sound in air is 330 m s^{-1}).

(1) 250 (2) 190 (3) 298 (4) 198

Concept Application Level—2

31. Two wires of same material and area of cross section each of length 30 cm and 40 cm are stretched between two ends with tensions 10 N and 20 N respectively. The difference between the fundamental frequencies of two wires is 4.0 Hz. Find the linear mass density of the wire.

32. A person speaks at one end of a metal pipe 200 m long and a person at other end hears two sounds at an interval of 0.5 seconds. If the velocity of sound in air is 300 m s^{-1}, answer the following questions.

(a) In which of the two media, air or pipe, does sound travel faster?

(b) Why are two sound notes heard at two different instants?

(c) Find the speed of sound in the metal pipe.

33. A block of mass 3 kg is suspended from the ceiling of a lift through a string having a linear density of 32.4×10^{-3} kg m^{-1}. Find the speed with which a wave can move on the string, if the lift accelerates upward at a rate of 2 ms^{-2}. Take g = 10 m s^{-2}

34. For a certain organ pipe, three successive resonant frequencies are observed at 225 Hz, 375 Hz, 525 Hz respectively. If the velocity of sound in air is 330 m s^{-1}, find whether the pipe is closed at one end or open at both the ends. Also find the length of the pipe.

35. A ball floating on water is pushed slightly into water and released. Explain in detail what type of motion it would execute.

36. A sonar fixed to a stationary submarine is used to detect a shoal of fish. It sends waves in east direction and the reflected waves from the shoal of fish are observed at the end of 8 seconds. After 20 minutes, again the waves are sent in east direction and the reflection is observed at the end of 2.0 seconds. Find the speed and direction of motion of the shoal of fish.

Take the velocity of sound in water as 1500 m s^{-1}.

37. A gas of gram molecular weight 48 g was found to transmit sound with a velocity of 290 m s^{-1} at a temperature of 27°C. If R = 8.3 J K^{-1} $mole^{-1}$, find the constant γ.

38. A person drops a stone in a well of depth 80 m and hears the sound of splash after 4.25 seconds. If the acceleration due to gravity is 10 m s^{-2}, find the velocity of sound in air.

39. The velocity of sound in a gas at 0°C is 335 m s^{-1}. If its γ value is 1.4, find the molecular weight of the gas. [Hint: for an ideal gas, $\frac{p}{\rho} = \frac{RT}{M}$ and R = 8.3 J K^{-1} $mole^{-1}$].

40. Find the temperature in °C at which sound travels in hydrogen with the same velocity as in oxygen at 1007°C.

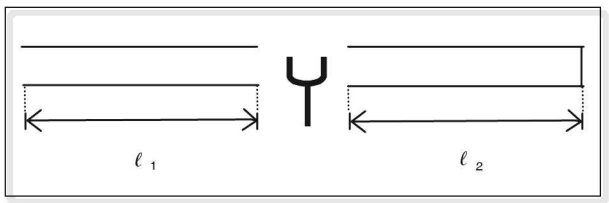
41. A man stands in front of a cliff and finds that when he claps at regular time interval of 1.2 seconds, he does not hear the echo distinctly. Then he moves closer to the cliff by 32 m and finds that he cannot hear the echo distinctly when he claps at an interval of 1 second. Find the velocity of sound waves.

42. A sonometer wire of length 'l' under tension 'T' is vibrating with its fundamental frequency. If its length is decreased by 45% and tension is increased by 21%, find the percentage change in its frequency.

43. Explain why the pitch of sound increases gradually when a vessel is filled by a water tap.

44. Empty vessels make more vibrations than the vessels filled with a liquid when they are disturbed. Explain.

45. A vibrating tuning fork is held between an open organ pipe and a closed organ pipe of lengths ℓ_1 and ℓ_2 respectively as shown in the figure.



If the first overtones are produced in each organ pipe, find the ratio of the lengths of the two organ pipes.

Concept Application Level—3

46. A straight wire is made up of two copper wires welded in the middle. This wire is used as a sonometer wire. Stationary waves are set up in the wire with tension T and welded joint as node. If one half of the wire vibrates with 3 loops and the other half with 6 loops, find the ratio of the velocities of the waves in one half to that in the other half of the wire. Is the thicknesses of each half of wire same? If no, find the ratio of the radii of each half of the wires.

47. Why is the sound heard when a piece of paper or cloth is torn? Why does the pitch of sound increase when the speed of tearing is increased?

48. Find the ratio of velocities of sound in hydrogen to oxygen at STP. The γ values of hydrogen and oxygen are same.

49. The vibration of a string of length 20 cm and mass 2 g stretched between two ends and those in a closed end tube of length 40 cm are compared. The difference in the frequencies of the tube in its fundamental mode and the string in its first overtone is 10 Hz. On decreasing the tension in the string, the difference in their frequencies decreases, then find the tension in the string. (Velocity of sound in air is 320 m s^{-1})

50. A source of sound at rest produces sound waves, which move with velocity v. Wind blows with velocity 0.2v towards an observer standing at a distance from the source. Find the percentage change in the wavelength and frequency of the sound as heard by the observer.

Very short answer type questions

1. The reflected sound is known as echo.

2. If a body is set into vibrations by the influence of another vibrating body or an external periodic force, the vibrations executed by the former are known forced vibrations.

☛ *Example:* The vibrations executed by the pendulum of a clock are forced vibrations as they are caused under the influence of a main spring which is wound.

3. greater

4. $n = \frac{1}{2\ell}\sqrt{\frac{T}{m}}$

5. If humidity in air increases then the velocity of sound also increases.

6. $V = n\lambda$

$\therefore V = 2n(l_2 - l_1)$

7. Intensity (I) $\propto$ (amplitude)^2
Intensity (I) $\propto$ density (ρ)

8. Potential energy, kinetic energy

9. 20 Hz to 20,000 Hz.

10. The vibrations whose amplitude progressively decreases with time or with position or with both as the wave is propagated in a medium are called 'damaged vibrations'.

Example: The oscillations of a disturbed simple pendulum.

11. If the direction of the violation of the particles of the medium is perpendicular to the direction of the propagation of the wave, the wave is called a transverse wave. If the direction of the vibration of the particles of the medium is parallel to the direction of the wave propagation, the wave is called 'a longitudinal wave'.

12. Frequency of the tuning fork, n = 512 Hz

First resonating length, l_1 = 16 cm

Second resonating length, l_2 = 48 cm

$(l_2 - l_1)$ = 32 cm

$V = 2n(l_2 - l_1) = \frac{2 \times 512 \times 32}{100} = 327.68 \text{ m s}^{-1}$

13. Timbre or quality.

14. $V = \sqrt{\frac{T}{m}}$

15. When a wave starts from point in a medium and propagates in all the possible directions and never returns, the wave is called a progressive wave. When a progressive we strikes a hard surface and is reflected, the incident wave and the reflected wave superpose forming a standing wave, also called a stationary wave.

16. If the frequency of the vibrations is less than 20 Hz, it is called infrasonics and if greater than 20,000 Hz it is called ultrasonic.

17. One millionth

18. If the natural frequency of the air column matches with that of the tuning fork, a loud sound is produced due to resonance. The air column in this condition is known as 'resonating air column'.

19. Velocity of sound in air is given by:-

$V = \sqrt{\frac{p}{d}}$ (By Newton)

$V = \sqrt{\frac{\gamma p}{d}}$ (By laplace)

20. Pitch is a special characteristic of sound based on which we can differentiate between two sounds having equal loudness. It is measured by sound frequency.

21. Intensity (I) $\propto \frac{1}{(\text{density})^2}$

22. Multiple reflections of sound.

23. $V \propto \sqrt{T}$

$\frac{v_1}{v_2} = \sqrt{\frac{T_1}{T_2}}$

24. The observer can hear the sound even after the source has stopped producing it due to multiple reflections taking place at the walls of the enclosure. This phenomenon is known as 'reverberation'.

25. frequency

26. One

27. $V \propto \frac{1}{\sqrt{M}}$

$\frac{V_1}{V_2} = \sqrt{\frac{M_2}{M_1}}$

The variation of the pressure of ages does not affect the velocity of the sound in it.

$V = \sqrt{\frac{\gamma p}{d}}$;

$\frac{p}{d}$ is always constant.

28. $n_0 = 200$ Hz
First overtone, $n_1 = 3n_0 = 2 \times 200 = 400$ Hz
Second overtone, $n_2 = 3n_0 = 3 \times 200 = 600$ Hz

29. The frequency with which the body vibrates is known as natural frequency.
Example: A tuning fork when hit by a rubber hammer.

30. Velocity of sound depends only on the nature of the medium.

Short answer type questions

31. The phenomenon in which when one of the two bodies having equal natural frequencies is set into vibration, then under its influence the other body also vibrates with increased amplitude.

Example: Resonating air columns.

32. $V = \sqrt{\frac{T}{m}} = \sqrt{\frac{200N}{2 \times 10^{-2}\ kg\ m^{-1}}} = \sqrt{10^4}$

$= 100\ m\ s^{-1}$

33. (i) Tuning fork of same frequency, string, frictionless pulley, rubber hammer, weights.

(ii) When tuning fork is excited with another hammer, sends a wave in the string.

(iii) At pulley the wave is reflected.

(iv) The incident wave and reflected wave superpose to form stationary wave.

34. $V = 2n(l_2 - l_1)$

Ans: 320 m s^{-1}

35. $\frac{\lambda}{4} = 30$ cm then $\frac{3\lambda}{4} = 90$ cm

36. (i) A progressive wave starts from a point in a medium and propagates in all the possible directions and never returns.

(ii) the incident wave and reflected wave suppose forming a stationary wave.

37. $n_p = (2p - 1)n_0$

Ans: 150 Hz

38. $V = n\lambda$

If n = 20 Hz,

$\lambda = \frac{300\ m\ s^{-1}}{20\ Hz} = 15$ m

If n = 20,000 Hz,

$\lambda = \frac{300\ m\ s^{-1}}{20,000\ Hz} = 15 \times 10^{-3}$ m.

Audible Range in terms of wave length is 15×10^{-3} m to 15 m

39. (i) Absolute temperature of gas

(ii) Density of a gas

(iii) Molecular weight of a gas

40. $V = \sqrt{\frac{\gamma P}{\rho}}$

Ans: 9×10^5 Pa

41. $d = \frac{v}{20} = \frac{330}{20} = 16.5$ m

42. (i) Pressure of a gas
(ii) Amplitude
(iii) Frequency
(iv) Wave length

43. $\frac{V_1}{V_2} = \sqrt{\frac{M_2}{M_1}}$

Ans: 450 m s^{-1}

44. $\frac{X}{2X} = \sqrt{\frac{273}{T_2}}$

$T_2 = 273 \times 4 = 1092K = 819°C$

45. (i) A person makes a sound in front of a big wall at a distance 'd'.
(ii) hears two sounds distinctly with time interval t seconds.
(iii) velocity of sound, $V = \frac{2d}{t}$

Essay type questions

46. (i) Stretched string with both ends tied to fixed ends.
(ii) Formation of standing waves when exited.
(iii) Fundamental frequency and overtones.

47. (i) Take two hollow boxes 'p' and 'q'.
(ii) open ends face each other.
(iii) Two tuning forks 'A' and 'B' of equal frequencies are fixed on boxes 'p' and 'q' respectively.
(iv) If 'A' is disturbed to vibrate produced forced vibration in 'p' and 'q' boxes.
(v) B also vibrates, because of the same frequency maximum energy transfer takes place.

48. (i) Resonating air column—Definition
(ii) Closed end resonating air column
(iii) Formation of stationary longitudinal waves in the air column.
(iv) Expression for resonating length of the air column
(v) Phase change at the open end of the tube
(vi) Resonance in open organ pipes
(vii) Formation of nodes and antinodes
(viii) Expression for fundamental frequency
(ix) Harmonics overtones

49. (i) Law of length:-

$n \propto \frac{1}{\ell} \Rightarrow n_1\ell_1 = n_2\ell_2$

(ii) Law of tension:-

$n \propto \sqrt{T} \Rightarrow \frac{n_1}{\sqrt{T_1}} = \frac{n_2}{\sqrt{T_2}}$

(iii) Law of linear density:-

$n \propto \frac{1}{\sqrt{m}} \Rightarrow n_1\sqrt{m_1} = n_2\sqrt{m_2}$

$n = \frac{p}{2\ell}\sqrt{\frac{T}{m}}$

50. (i) Two boys standing on two distant hills at night.
(ii) Sound and flash being produced when a boy shoots a gun.
(iii) Determination of time taken by sound to travel from the source (to the other boy.)
(iv) Calculation of the velocity of sound.
(v) Repetition of the same procedure after the boys interchange their places.
(vi) Calculating velocity of sound in opposite direction.
(vii) Taking the average of two values.
(viii) Producing a stationary wave by using a tuning fork.
(ix) Determination of the first resonating length, and the second resonating length of the air columns for the given frequency.
(x) Expression for velocity of sound in terms of resonating air column.

KEY

Concept Application Level—1

True or false

1. False
2. True
3. True
4. False
5. True
6. False
7. False

Fill in the blanks

8. more (or greater)
9. constant pressure, constant volume
10. frequency
11. increases
12. 1.2
13. 0.5 m
14. $\frac{2d}{t_1 - t_2}$

Match the following

15. A : d
    B : f
    C : g
    D : a
    E : i
    F : h
    G : b
    H : j
    I : c
    J : e

Multiple choice questions

16. Choice (3)
17. Choice (4)
18. Choice (2)
19. Choice (3)
20. Choice (4)
21. Choice (1)
22. Choice (2)
23. Choice (4)
24. Choice (3)
25. Choice (2)
26. Choice (2)
27. Choice (3)
28. Choice (4)
29. Choice (4)
30. Choice (4)

Concept Application Level—2,3

Key points

31. (i) For a stretched string. $n = \frac{1}{2\ell}\sqrt{\frac{T}{m}}$

    $n_2 - n_1 = 4$ Hz

    (ii) 6.4×10^{-3} kg m^{-1}

32. (i) Find the distance the sound travels to reach the observer from source of sound.

    Find the interval of time heard between the two sounds that passes through the pipe and air x.

    Find the velocity of sound in air from the given information.

    Find time taken by sound to reach from source to observer through air, by using formula time, $t_1 = \frac{\text{distance}}{\text{velocity}}$

    since u the velocity of sound in solids more than that of the air then, the time taken by sound to travel in solids is less than that in air.

    Time taken by sound to travel in metallic pipe, $t_2 = \left(t_1 - \frac{1}{2}\right)$ s.

    Find the time taken by sound to reach the observer through the metallic pipe.

Find the velocity of sound in metallic pipe by dividing distance by 't_2'.

(ii) Sound travels faster in pipe.

(iii) Speed of sound is different in the two media.

(iv) 1200 m s^{-1}

33. (i) $T = M(a + g)$

$$v = \sqrt{\frac{T}{m}}$$

(ii) 33.3 m s^{-1}

34. (i) Fundamental frequency for a closed pipe

$$n = \frac{v}{4\ell}$$

(ii) closed pipe, 110 cm

35. What is SHM?

When a wave passes through a medium, will the particles in the medium execute SHM?

When ball is pushed slightly into water, will it produce waves on the surface of the water?

Do the particles on the surface of water execute SHM?

Will the ball also execute the motion similar to the particles present on the water surface.

36. (i) Find the time taken by the reflected waves from the shoal of fish to reach the stationary submarine two times in an interval of 20 minutes form the given data.

Take them as t_1 and t_2.

Find the velocity of sound in water in given information.

Here, $v = \frac{2d}{t}$ ---- (1)

By using the formula, (1) for echo, find the distance of the shoal of fish from the submarine. (Take it as d_1 and d_2).

Now, the speed of the shoal of fish, v is equal to the distance moved by the shoal of fish in one second

Consider, $v = \frac{d_1 \sim d_2}{t}$

What is the direction in which sonar sends waves?

Find the direction of the motion of the fish

(ii) 3.75 m s^{-1} (west)

37. (i) $v = \sqrt{\frac{\gamma R T}{M}}$

(ii) 1.62

38. (i) What is the depth of the well?

What is the initial velocity of the stone?

What is the time (t_1) taken by the stone to reach the bottom of the well?

Find the value of "t_1" by using the formula, $s = ut + \frac{1}{2}gt^2$.

Take the time taken by the person to hear sound of splash as 't_2'.

Find the time interval between t_2 and t_1 by using the formula $\Delta t = (t_2 - t_1)s$.

Then, find the velocity of sound in air by dividing the depth of the well by Δt.

(ii) 320 m s^{-1}

39. (i) Find $\frac{P}{\rho}$ from $V = \sqrt{\frac{\gamma p}{\rho}}$

(ii) Substitute it in $\frac{P}{\rho} = \frac{RT}{M}$

(iii) 28 g

40. (i) $v \alpha \sqrt{T}$, $v \alpha \frac{1}{\sqrt{M}}$

(ii) –193 °C

41. (i) Case 1:

When a man claps at regular time interval of 1.2 seconds, he does not hear the echo distinctly.

Now, find the time taken to produce echo (t_1).

Consider the distance between man and cliff as 'd_1'.

Case 2:

The man moves closer to the cliff by 32 m and claps at regular interval of time of 1 second. He doesn't hear the echo distinctly. Find the time taken for producing an echo (t_2).

Now, consider the distance between man and cliff as "d_2".

Find the relation between d_1 and d_2. ----- (1)

$d_2 = d_1 - 32$

The formula for velocity of sound using echo method is:

$$V = \frac{2d_1}{t_1} = \frac{2d_2}{t_2} \text{ -------- (2)}$$

Substitute the value of 'd_2' interms of 'd_1'.

Calculate the value of 'v' from 2nd equation.

(ii) 320 m s^{-1}

42. (i) $n \alpha \frac{\sqrt{T}}{\ell}$

(ii) Percentage change in frequency =

$$\frac{n_2 - n_1}{n_1} \times 100$$

(iii) 100%

43. Which physical quantity is a measure of pitch? How does it vary in the case of vibrating air columns?

When a vessel is filled by a water tap, will the length of the vibrating air column decreases?

Will this decrease in the length affect the frequency of the vibrations?

Find the relation between the frequency and the length of the resonating column.

Now, find the relation between the pitch and frequency.

44. (i) How the frequency depends on length of the air column?

(ii) How the presence of water influence the frequency of vibrations.

45. (i) Take the velocity of sound in air as 'v'. Take the frequency of the tuning fork as "n".

When the first overtone (or) second harmonic is considered, then, $\ell_1 = \lambda$ and

$\ell_2 = 3\frac{\lambda}{4}$ ----- (1)

Obtain the value of 'λ' in terms of ℓ_1 and ℓ_2 and substitute it in the formula $v = n\lambda$. Find the ratio of 'ℓ_1' to 'ℓ_2'.

(ii) 4 : 3

46. (i) Take the length of one half of the wire as 'ℓ'.

Take tuning fork of frequency 'n', to excite the sonometer wire.

How are the velocities (v), frequency (n) and wavelengths (λ) in each half related?

Find the values of λ in terms of 'ℓ' for both the sides (halves).

Find the ratio of velocities of both the halves, using the formula,

$v = n\lambda$ ---- (1)

Does the tension (T) in two halves of the wire remain same?

Find the formula that shows the relation between v, T and linear density (m).

Is it, $v = \sqrt{\frac{T}{M}}$ ----- (2)

Now, Take linear densities of first and second half M_1 and M_2.

Linear density, M =

$$\frac{\text{mass of wire (m)}}{\text{length of the wire } (\ell)}$$

Then $\frac{v_1}{v_2} = \sqrt{\frac{M_2}{M_1}} = \sqrt{\frac{m_2}{m_1}}$.

Is the metal of the two halves of the wire same?

Will the density (d) of the two halves be equal?

Then, find the value of 'm' in terms of volume and density.

Here, volume (V) of the wire = (Area of cross – section) (length of the wire)

Then, $V = \pi r^2 \ell$.

Take the radii of two halves of the sonometer wire as 'r_1' and 'r_2'

Now, find the ratio of v_1 and v_2 interms of r_1 and r_2.

(ii) 1 : 2 or 2 : 1 (ratio of velocity)

(iii) 2 : 1 or 1 : 2 (ratio of radii)

47. Pitch of the sound depend on frequency.

48. (i) Find the values of temperature, pressure and volume at S.T.P.

Get the formula for the velocity (v) of the sound in the gas.

Are the values of 'λ' for Hydrogen and oxygen, equal?

Then, is $v \propto \frac{1}{\sqrt{d}}$ ---- (1)

Is density (d) of the gas directly proportional to the molecular weight of the gas at S.T.P?

Then, $\frac{v_1}{v_2} = \sqrt{\frac{M_2}{M_1}} = \sqrt{\frac{m_2}{m_1}}$ ----- (2)

Find the values of M_1 and M_2 (i.e.,) molecular weights of hydrogen and oxygen and substitute in (2).

Find the ratio of v_1 and v_2 from (2).

(ii) 4 : 1

49. (i) Frequency of first overtone of a stretched string = 2(fundamental frequency of the closed tube)

2. $n_2 > n_1$

(ii) 17.64 N

50. (i) When the source of sound and observer are at rest is there any apparent change in the frequency of the sound?

Will it result in Doppler effect?

Find the relation between the velocity of the sound (v_1) and its wavelength (λ_1), when frequency of the sound remains constant?

Find λ_1 in terms of v_1 and f.

Is the direction of the wind and sound same?

Does the velocity of the sound increase? (Take it as v_2)

Find the value of 'v_2' interms of 'v_1'

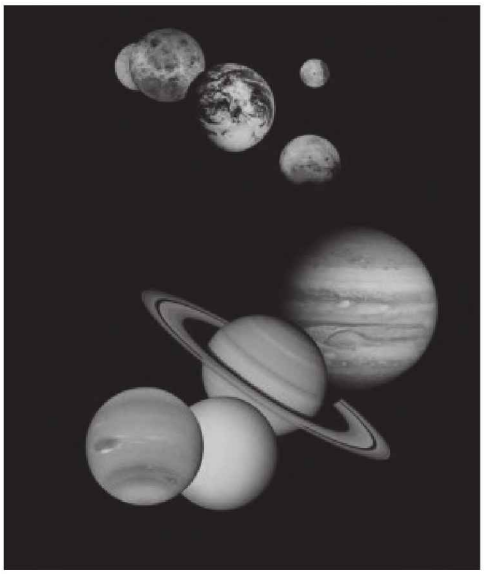
$V_2 = v + 0.2v$, Find λ_2.

Ratio of frequency is the ratio of velocity of sound with respect to the observer to that with respect to the source.

(ii) 20% change in wavelength, zero per cent change in frequency.

7

Electricity



INTRODUCTION

In the chapter heat, you have studied that heat is a form of energy. The other forms of energy are light, mechanical, chemical energy etc. Among all forms of energy, electric energy is the most useful and convenient for mankind. It can be converted to almost all forms of energy easily and can be stored and transmitted over long distances with very little wastage. Electric bulbs, motors, all phones, air conditioners, heaters, computers and heavy machinery all use different effects of electric energy. In fact it is difficult to imagine our lives without electric energy. The origin of electric energy lies in electric charges. The branch of physics which deals with the study of electric charges and their effects is called electricity.

In order to understand the topic of electricity better, we shall first recall some of the properties of electric charges which you have studied in detail in previous classes.

Electric charges—properties

1. Electric charge is of two types viz., positive and negative charge.
   Proton is said to be charged positively and electron is said to be charged negatively.
2. Like charges repel and unlike charges attract. Thus a proton repels a proton and attracts an electron.
3. The force of attraction or repulsion between two charges is given by Coulomb's law.

$$F = \frac{Kq_1q_2}{r^2}$$

where q_1 and q_2 are charges.

K = constant of proportionality.

r = distance between the charges.

4. Charge can neither be created nor destroyed. The charge from one body can be transferred to another body but the total charge of a system remains constant. This is called the law of conservation of charge.
5. Protons and electrons are elementary charged particles. Though the charge on them is opposite in nature, the magnitude of charge possessed by them is same i.e., 1.6×10^{-19} C. Charge on a body is always an integral multiple of this value. This is called quantization of charge.
6. When a body gains electrons, it becomes negatively charged. When it loses electrons it becomes positively charged. The positive charge being bound firmly in the nucleus does not participate in charging.

The study of electricity is classified into two parts.

1. Static electricity: It deals with electric charges at rest and their effects.
2. Current electricity: It deals with charges in motion and their effects.

This chapter deals with charges in motion i.e., current electricity.

Conductors and insulators

Current electricity is the study of charges in motion. Hence it is of our interest to know the substances which allow charges to flow through them easily and which resist it. Accordingly substances are classified as conductors and insulators.

1. Conductors

The substances which conduct electricity easily are called electric conductors.

Since positive charges cannot flow, the conduction of electricity is primarily due to flow of electrons. Hence conductors have large number of free electrons.

☛ *Example* All metals are generally good conductors of electricity.

2. Insulators

The substances which do not conduct electricity are called electric insulators.

☛ *Example* wood, rubber, mica insulators do not have free electrons.

Electric Current

The rate of flow of charges is known as electric current. It is denoted by I. If "q" is the net charge passing through any cross section of a conductor in a time "t", then

$$\text{Current (I)} = \frac{\text{Charge (q)}}{\text{Time (t)}}$$

Current is a scalar quantity. S.I. unit of current is ampere (A). Multiple units of current are kiloampere (1 kA = 10^3 A), Megaampere (1 MA = 10^6 A).

Submultiple units are 1mA (1 milliampere = 10^{-3} A) and microampere (1 μA = 10^{-6} A).

One ampere

The current passing through a conductor is said to be one ampere when the net flow of charge per second through its cross section is one coulomb.

The charge transferred is always an integral multiple of the fundamental charge (charge of an electron). This property is referred to as **quantization of charge**.
i.e., Q = ne

Since current (I) = $\frac{Q}{t}$ or $I = \frac{ne}{t}$

where, e = magnitude of charge of an electron
n = number of electrons

Electric potential

It is our common observation that water flows from a place at a higher level to a place at a lower potential or level irrespective of the amount of water at the two levels. Heat flows from a body at a higher temperature to a body at a lower temperature irrespective of the amount of heat contained by the two bodies. Similarly electric charges move from one point to another if there exists a difference in the electric potential irrespective of the concentration of charges at the two points. Thus electric potential determines the direction of flow of electric current. Charge flows from a body at higher potential to a body at lower potential. By convention a positively charged body is always at a higher potential as compared to a negatively charged body. If positive charges like protons or positive ions are free to move they would move from higher potential point to lower potential point. Usually the movement of charges implies movement of electrons. It is clear that electrons will move from lower potential to higher potential. This constitutes electronic current. Since the convention is that flow is always from higher to lower level, the conventional current is taken opposite to the electronic current. It should be remembered that conventional current implies the movement of positive charges whereas electronic current is the movement of electrons.

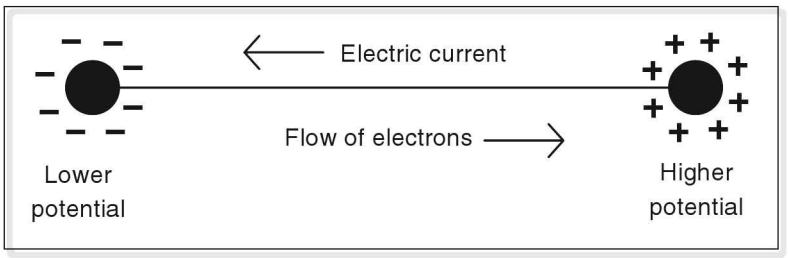


Figure 7.1

As discussed earlier, concentration of charges determines the electric potential at a point. This can be increased by bringing charges at that point. In doing so some work has to be done on the charge. This work is stored as electric potential at that point. The electric potential can be measured in terms of work done and is defined as the work done in bringing unit positive charge from infinity to a point in an electric field is called electric potential (V).

The electric potential is measured in volts (V). The electric potential at a point is said to be 1 volt if 1 joule of work is done in bringing unit positive charge from infinity to that point.

Potential difference

To bring a unit positive charge from infinity to a point in an electric field, some work has to done which is called electric potential. Within an electric field if a unit positive charge has to be moved from one point to another against the direction of the field, some work has to be done which is called as potential difference between the two points.

Like electric potential, the potential difference is also measured in volts (V). Let Q columbs charge be moved from point A to the point B in an electric field when work done is W. Let V be the potential difference between two points. Therefore the work done is moving unit positive charge

$= \frac{W}{Q}$.

By definition, this work done is also the potential difference.

$\therefore V = V_1 - V_2 = \frac{W}{Q}$

Where V_1 = Potential at point A.

V_2 = Potential at point B.

When 1 joule of work is done in moving unit positive charge from one point to another in an electric field, the potential difference between the two points is 1 volt.

$$V = \frac{1 \text{ joule}}{1 \text{ coulomb}}$$

$$= 1 \text{ volt} \left(\because \frac{\text{joule}}{\text{coulomb}} = \text{volt} \right)$$

Example

5 joule of work is done in moving 12.5×10^{18} electrons from one end to other end of a conductor. What is the potential difference between the two ends of conductor?

Solution

The charge on 6.25×10^{18} electrons is 1 C.

$\therefore$ When 12.5×10^{18} electrons move, the net charge transferred is 2 C.

$\therefore$ Q = 2 C.

Work done (W) in moving 2 C charge = 5 J. ---------- (given)

The potential difference (V) between the ends of conductor is

$$V = \frac{\text{Work done}}{\text{charge}}$$

$$= \frac{5 \text{ J}}{2 \text{ C}} = 2.5 \text{ J C}^{-1}$$

V = 2.5 volts.

☛ *Example*

Electric potential at a point in an electric field is 0.5 V when charge of 3 C was brought from infinity to that point. Calculate the work done.

Solution

Electric potential (V) = 0.5 volts
Charge (Q) = 3 C
Work done (W) = ?

$$V = \frac{W}{Q}$$

$$\therefore W = VQ$$

$$= 0.5\,V \times 3\,C = 1.5\,V\,C$$

$$= 1.5\,J.$$

☛ *Example*

The electric potential at two ends of a conductor is 10 V and 5 V respectively. 5 J of work needs to be done to move certain amount of charge from end at higher potential to end at lower potential. Find the amount of charge flowing between the two ends.

Solution

The potential difference between the two ends of conductor (V) = 10 V – 5 V.

V = 5 V

Work done (W) = 5 J

Charge (Q) = ?

$$\text{Potential difference} = \frac{\text{work}}{\text{charge}}$$

$$V = \frac{W}{Q}$$

$$= \frac{5\ J}{5\ V} = 1 \text{ coulomb.}$$

Current in conductors and electrolytes

Rate of flow of charge is called current. We have already seen that charge moves from one point to another if there exists a potential difference between them. Thus whether the current is flowing through a conductor or electrolyte, a potential difference is necessary for it. You have also studied in previous classes that cell or battery is used to provide potential difference.

Current in a conductor

A potential difference can be produced between a conductor in the form of metallic wire if it is connected between the two terminals of a cell. The end of the wire connected to the positive terminal is at a higher

potential while the end connected to the negative terminal is at a lower potential. A conventional current is set up between these two ends.

A conductor consists of a large number of free electrons which are in random motion just like molecules in air. When it is connected across a cell or battery, these free electrons begin to drift from lower potential to higher potential. This constitutes the electronic current. Thus the free electrons are responsible for the flow of electric current in conductors. However the direction of conventional current is taken opposite to that of electronic current i.e., from higher potential to lower potential.

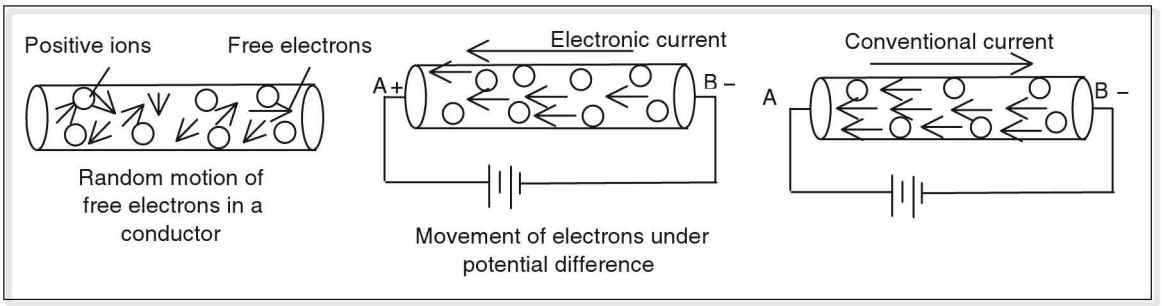


Figure 7.2

When the metallic wire is not connected to a battery, although the electrons are moving, the net flow in any particular direction is zero. Hence no current flows through it in the absence of potential difference.

Current in electrolyte

A liquid or a solution which conducts electricity is called an electrolyte.

For example, copper sulphate ($CuSO_4$), sodium chloride (NaCl) or common salt solution.

When a potential difference is created across an electrolyte, it undergoes chemical change and decomposes into its constituents. This process is called electrolysis.

Take some water in a beaker and add common salt (NaCl) to it. Insert two metallic rods A and B in the solution. Connect A to positive and B to negative terminal of a cell. A is at higher or positive potential and is called anode while B which is at lower or negative potential is called cathode.

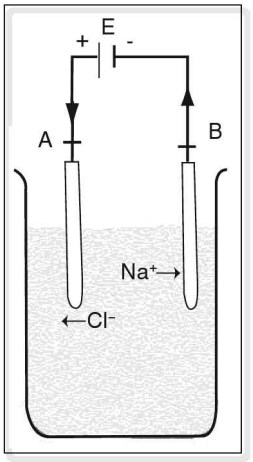


Figure 7.3

The solution dissociates into positive sodium ions (Na^+) and negative ions (Cl^-). The positive Na^+ ions move towards cathode B and gain electrons to form Na atoms. The negative Cl^- ions migrate towards anode A through electrolyte and form chlorine atoms by giving away electrons. The bubbles of chlorine can be seen coming from A.

Thus in an electrolyte both positive and negative ions move through it towards opposite electrodes. The rate of flow of charge is electric current. However, the conventional current is taken as movement

of positive ions through electrolyte towards cathode B. Through the external wire connecting the cell to the electrodes, electrons enter from cathode and move towards anode A. If a bulb is connected, it begins to glow confirming the flow of current. The conventional current is considered to flow from anode A to cathode B through external wire.

Thus in an electrolyte, the movement of ions is responsible for electric current.

Electrical resistance

Conductors conduct electricity owing to the presence of a large number of free electrons. When a potential difference is applied across a conductor these electrons begin to drift from the end at low potential to the end at high potential. While drifting they collide with other electrons as well as fixed positive ions. The fixed ions are formed from the atoms which give free electrons. Thus even as conductor conducts electricity, at the same time it offers some obstructions to the flow of charges.

The obstruction offered by a conductor to the flow of electric current is called its resistance.

As the collisions suffered by drifting electrons increase, the resistance offered by the conductor also increases. Thus the resistance depends basically on the arrangement of atoms and the configuration of electrons around the atoms of the conductor.

Units of resistance

The SI unit of resistance is ohm. It is denoted by Greek symbol Ω (Omega).
Higher units of resistance

1. kilo-ohm(kΩ) = $10^3\,\Omega$
2. mega-ohm(MΩ) = $10^6\,\Omega$
3. giga-ohm(GΩ) = $10^9\,\Omega$

Electrical conductance

The reciprocal of resistance is known as electrical conductance. It is denoted by G. It is a measure of ease with which a charge can flow in a conductor.

$G = \frac{1}{R}$. Unit of conductance is mho or ohm^{-1} or siemen.

Electrical conductivity

The reciprocal of resistivity is called conductivity.

It is denoted by σ. $\sigma = \frac{1}{\rho}$; The unit of conductivity is mho/metre or siemen/metre.

Super conductors

Conductors exhibit a phenomenon known as super conductivity at a very low temperature known as critical temperature. At critical temperature, a metal offers zero resistance to the passage of electricity. These type of conductors are called superconductors.

The temperature below which a conductor offers no resistance to electric current is known as its **critical temperature**.

Factors affecting resistance

1. Resistance of a conductor depends upon its nature.
2. Law of length: Resistance of a conductor is directly proportional to its length.

   Let ℓ and R be the length and resistance of conductor, then consider two conductors of same material such that ℓ_1 and ℓ_2 are their lengths and, R_1 and R_2 are their resistances respectively. Then

   $$R_1 \propto \ell_1$$

   $$R_2 \propto \ell_2$$

   $$\therefore \frac{R_1}{R_2} = \frac{\ell_1}{\ell_2}$$

3. Law of area of cross section

   Resistance of a conductor is inversely proportional to its area of cross section.

   Let a be the area of cross section of a conductor whose resistance is R.

   $$R \propto \frac{1}{a}$$

   Considering two conductors of area of cross sections a_1 and a_2 and resistances R_1 and R_2.

   Then,

$$R_1 \propto \frac{1}{a_1}$$

$$R_2 \propto \frac{1}{a_2}$$

$$\therefore \frac{R_1}{R_2} = \frac{a_2}{a_1}$$

$$= \frac{\pi r^2_2}{\pi r^2_1} \quad (\because a = \pi r^2)$$

$$\frac{R_1}{R_2} = \frac{a_2}{a_1} = \left(\frac{r_2}{r_1}\right)^2$$

Where r_1 and r_2 are the radii of the two conductors.

Combining the 2nd and 3rd laws, we get

$$R \propto \frac{\ell}{a}$$

$$R = \rho \frac{\ell}{a}$$

Where ρ is the constant of proportionality.

Specific resistance or resistivity

We have seen that

$$R = \rho \frac{\ell}{a}$$

The constant of proportionality ρ is called the specific resistance or resistivity of a given conducting material like copper, aluminum etc. It is a characteristic property of conducting material and independent of the size and shape of the conductor.

$$\rho = R \frac{a}{\ell}$$

Consider a conductor having unit area of cross section and unit length.

$$\rho = R \times \frac{1}{1}$$

$$\therefore \rho = R.$$

Thus specific resistance or resistivity can be defined as the resistance of a conducting material having unit area of cross section and unit length.

Unit of resistivity

$$\text{Unit of } \rho = \frac{\text{unit of R} \times \text{unit of a}}{\text{unit of } \ell}$$

$$= \frac{\text{ohm} \times (\text{metre})^2}{\text{metre}}$$

unit of ρ = ohm-metre = Ω m

☛ *Numericals*

1. The length of copper wire is 100 m and its radius is 1 mm. Calculate its resistance if resistivity of copper is 1.72×10^{-8} Ω m.

Solution

Length of copper wire (ℓ) = 100 m.

Area of cross section (a) = πr^2

Resistivity (ρ) of copper = 1.72×10^{-8} Ω m

Resistance offered by a conductor is given by

$$R = \rho \frac{\ell}{A}$$

$$= \frac{1.72 \times 10^{-8} \times 100}{3.14 \times 10^{-6}} = 0.55\ \Omega$$

In this numerical, length of copper wire is 100 m. If the length is 1000 m i.e., 1 km, the resistance offered by it would be 5.5 Ω which is very less. Thus copper is a good conductor of electricity.

2. The resistance of 1 m of nichrome wire is 6 Ω. Calculate its resistance if its length is 70 cm.

Solution

Given

1st case

Length of nichrome wire (ℓ_1) = 1 m = 100 cm

Resistance of nichrome wire (R_1) = 6 Ω

2nd case

ℓ_2 = 70 cm

R_2 = ?

By 1st law of resistance

$$\frac{R_1}{R_2} = \frac{\ell_1}{\ell_2}$$

$$\frac{6\ \Omega}{R_2} = \frac{100\ \text{cm}}{70\ \text{cm}}$$

$$\therefore R_2 = 4.2\ \Omega$$

3. Two wires made of German-silver are taken such that the length and area of cross-section of the second wire are twice and thrice respectively those of the first wire. If the resistance of the second wire is 12 r, find the resistance of the first wire.

Solution

1st case

ℓ_1 = length of German silver wire

a_1 = area of cross section

R_1 = resistance

$$R_1 = \rho \frac{\ell_1}{a_1} \quad (1)$$

2nd case

ℓ_2 = length of wire = $2\ell_1$

a_2 = area of cross section of wire = $3a_1$

R_2 = resistance of wire = 12 Ω

$$R_2 = \rho \frac{\ell_2}{a_2}$$

$$\therefore 12 = \rho \frac{2l_1}{3a_1} \qquad (2)$$

dividing (1) by (2)

$$\frac{R_1}{12} = \frac{\rho \ell}{a_1} \times \frac{3a_1}{\rho(2\ell)} = \frac{3}{2}$$

$$\therefore R_1 = 18\ \Omega$$

Effect of temperature on resistance

The resistance of conductors increases with an increase in temperature. Let the resistance of a conductor at 0°C be R_0.

Let the resistance of the conductor at t°C be R_t.

Then, $R_t = R_0(1 + \alpha t)$, where 'α' is known as the **temperature coefficient of resistance**.

Pure metals have positive temperature coefficient of resistance. The resistance of metals increases with an increase in temperature.

Alloys have a very less temperature coefficient of resistance. So the resistance of alloys like Manganin and Constantan vary very little with an increase in temperature. Because of this property they are used in making standard resistances.

Semiconductors like germanium, silicon and bad conductors like glass, pure water etc., have negative temperature coefficient. The resistance of these materials decreases with an increase in temperature.

Effect of impurities on resistance

(i) Resistance of conductors increases with addition of impurities.

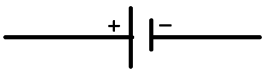
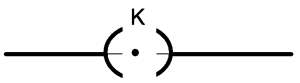
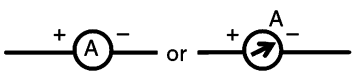
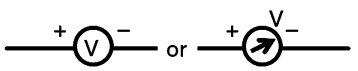
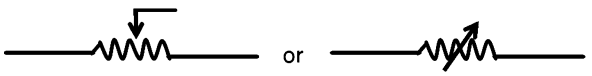
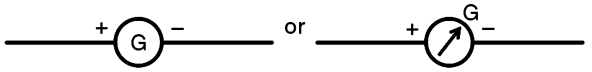
(ii) Resistance of semiconductors decreases with addition of selected impurities.

Electric circuits and circuit diagrams

In order to make use of electric energy, the current has to be passed through various devices and electric components. For this they need to be connected properly with respect to each other and electric source. This systematic arrangement of electric components and devices with the electric source is called electric circuit. It is also important that the circuit arrangement be represented schematically. The schematic representation of electric circuit is called circuit diagram. In order to draw a circuit diagram,

various electric components and devices are represented by suitable symbols.

The table below gives conventional symbols for some common components.

A cell	
A battery	
Alternating current source	
Plug key	
Switch	
Tapping Key	
Conducting Wire	
An ammeter	
A Voltmeter	
A resistance (fixed value)	
A variable resistance or rheostat	
Galvanometer	
Load	
Heater	
Bulb	

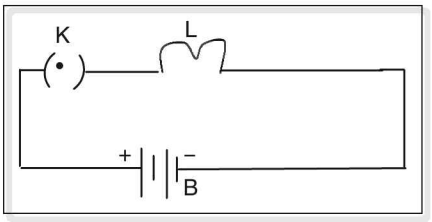


Figure 7.4

The above figure is a circuit diagram showing the connection of lamp (L) with a battery (B). Switch key (K) is used to connect or disconnect lamp with battery.

Ohm's Law

We have studied that when a potential difference is applied between the ends of a conductor, current begins to flow through it. George Simon Ohm, a German physicist performed a series of experiments to determine the exact relationship between the potential difference and electric current. On the basis of these experiments he formulated a law which is called a Ohm's law and can be stated as, the potential difference (V) applied between the ends of a conductor bears a constant ratio with the electric current (I) flowing through it, provided the physical state of the conductor remains the same. The factors which constitute physical state of conductor are its length, area of cross section, temperature and material.

According to Ohm's law

$$V \propto I$$

$$\frac{V}{I} = \text{Constant}$$

The constant in the above relation is called resistance (R) of the conductor.

$$\therefore \frac{V}{I} = R.$$

$$V = IR$$

If the potential difference is measured in volts and current in amperes, the resistance of conductor is measured in ohms.

$$\frac{\text{volt}}{\text{ampere}} = \text{ohm}.$$

The resistance of a conductor is said to be 1 ohm if the potential difference of 1 volt causes a current of 1 ampere to pass through the conductor.

$$\frac{1 \text{ volt}}{1 \text{ ampere}} = 1 \text{ ohm}.$$

Experimental verification of ohms' law

Voltmeter—ammeter method

Connect a battery (B), an ammeter (A), a resistance (R), a rheostat (Rh) and a voltmeter as shown in the figure.

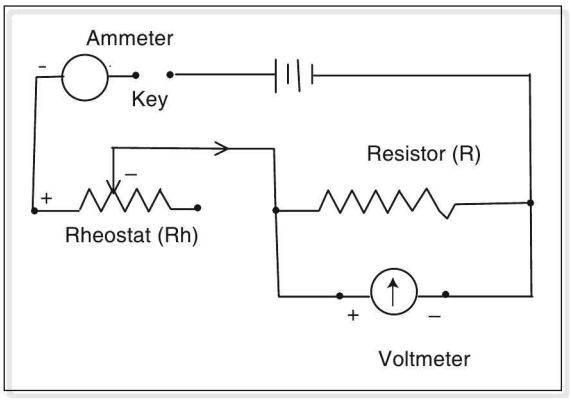


Figure 7.5

Move the slider position of the rheostat for different values of current in the circuit. The reading of current and voltage across the resistor for different positions of the slider switch are noted. In each case, the ratio of voltage to current is found.

It is observed that, the ratio V/I remains constant showing that the voltage is proportional to current ($V \propto I$).

Potentiometer method

Connect a potentiometer (a resistance wire of 1 m length) AB stretched between two brass terminals, as shown in the figure.

Close the key and record the voltage in the voltmeter connected parallel to A and jockey at different points.

Experimental set up of potentiometer method.

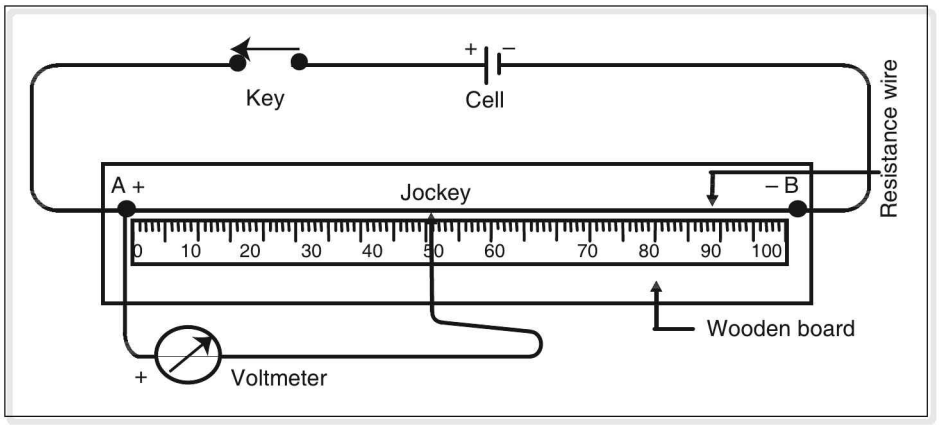


Figure 7.6

The ratio of the voltage to the corresponding length at each reading will be constant showing $\frac{V}{L}$ = constant, but $L \propto R$

$\Rightarrow \frac{V}{R}$ = constant

Graphical representation of Ohm's law

Plot voltage (V) on x-axis and current (I) on y-axis

Observation: A straight line passing through the origin is obtained (implies $V \propto I$).

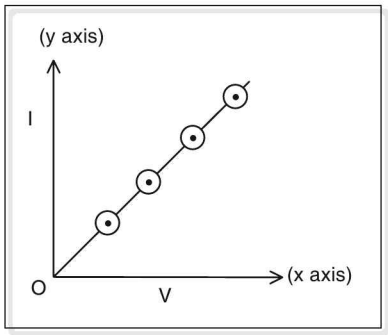


Figure 7.7

Ohmic and non-ohmic conductors

Ohmic conductors are the conductors which obey Ohm's law, hence they are referred to as linear conductors.

☛ *Example*

All metals.

V – I graph for ohmic conductors is a straight line passing through the origin
The conductors which do not obey Ohm's law are known as **non-ohmic conductors**.

☛ *Example*

Electrolytes, semiconductor devices like diodes, transistors etc.

V-I graph for non-ohmic conductors is not linear. (not a straight line).

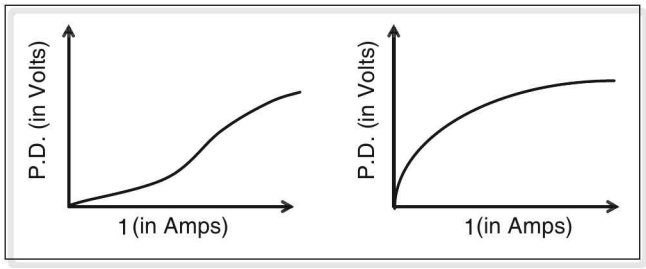


Figure 7.8

Resistances in series

Resistances are said to be connected in series, when there exists only one path for the flow of current and the current through each resistor is equal.

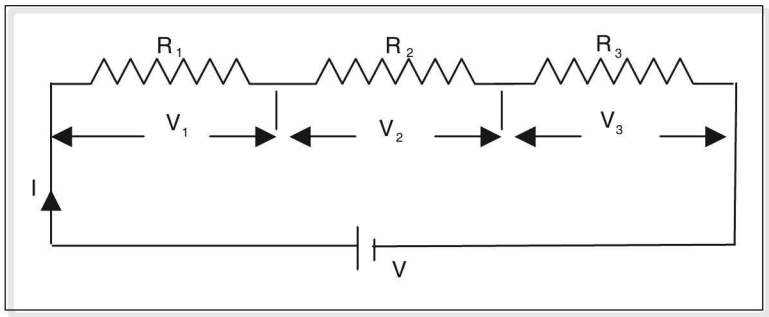


Figure 7.9

Resistances R_1, R_2 and R_3 are connected in series as shown above. Apply a voltage "V" across the series combination. Let V_1, V_2 and V_3 be the voltage drop across R_1, R_2 and R_3 respectively as shown in the figure. Let the current I flow through all the three resistances.

Therefore, the voltage V_1 across R_1 : $V_1 = IR_1$

the voltage V_2 across R_2 : $V_2 = IR_2$

the voltage V_3 across R_3 : $V_3 = IR_3$

The applied voltage $V = V_1 + V_2 + V_3 = IR_1 + IR_2 + IR_3$

If R_{eff} is the effective resistance, then $V = IR_{eff}$

$IR_{eff} = IR_1 + IR_2 + IR_3$

$IR_{eff} = I(R_1 + R_2 + R_3)$

$\therefore R_{eff} = R_1 + R_2 + R_3$.

☛ *Numericals*

1. An electric bulb whose resistance is 60 Ω is connected to a source of potential difference of 230 V. Find the current flowing through it.

Solution

Potential difference (V) applied to the bulb = 230 V
Resistance (R) of bulb = 60 Ω
Current (I) flowing through bulb = ?
According to Ohm's law
V = IR

$$\therefore I = \frac{V}{R}$$

$$= \frac{230\ V}{60\ \Omega}$$

$\therefore$ I = 3.83 A.

2. Study the following circuit diagram

Find the potential difference provided by the cell.

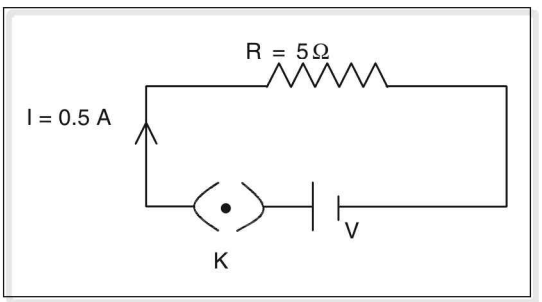


Figure 7.10

Solution

Current (I) flowing through the resistor = 0.5 A
Resistance (R) of the resistor = 5 Ω
Potential difference (V) = ?
According to Ohm's law
V = IR
= 0.5 A × 5 Ω
V = 2.5 V

Characteristics of series circuit

(i) Same current is flowing through all the resistances.

(ii) The effective resistance is the sum of the individual resistances. Effective resistance $R_{eff} = R_1 + R_2 + R_3$.

(iii) The applied voltage $V = V_1 + V_2 + V_3$

(iv) The maximum power is consumed by the resistor having the highest resistance, or the voltage drop is maximum across the highest resistance.

☛ *Example*

Find the effective resistance, when 1Ω, 10 Ω and 4 Ω resistances are connected in series.

Solution

When resistances are connected in series the effective resistance is given by

$R_{eff} = R_1 + R_2 + R_3$

$R_{eff} = 1 + 10 + 4 = 15$

Resistances in parallel

Resistances are said to be connected in parallel, when identical potential difference exists across all of them, i.e., the potential drop across each resistor is the same.

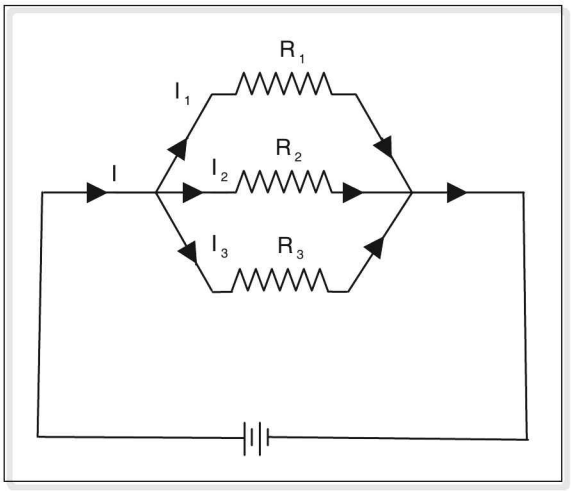


Figure 7.11

Let three resistors R_1, R_2 and R_3 be connected across the voltage source V in parallel. The current I drawn from the battery is divided into three parts I_1, I_2 and I_3 flowing through the resistors R_1, R_2 and R_3 respectively are as shown above.

The total current $I = I_1 + I_2 + I_3$

Let R_{eff} be the effective resistance.

Applying Ohm's law $I = \frac{V}{R_{eff}}$;

$I_1 = \frac{V}{R_1}$; and $I_2 = \frac{V}{R_2}$, $I_3 = \frac{V}{R_3}$

As, $I = I_1 + I_2 + I_3$; $\frac{V}{R_{eff}} = \frac{V}{R_1} + \frac{V}{R_2} + \frac{V}{R_3}$

$\therefore \frac{V}{R} = V(\frac{1}{R_1} + \frac{1}{R_2} + \frac{1}{R_3}); \Rightarrow \frac{1}{R} = \frac{1}{R_1} + \frac{1}{R_2} + \frac{1}{R_3}$

Characteristics of a parallel circuit

(i) The potential drop across each resistance is the same.

(ii) The reciprocal of the effective resistance is equal to the sum of the reciprocals of the individual resistances.

(iii) Current in the circuit $I = I_1 + I_2 + I_3 + \ldots\ldots$

(iv) The effective resistance is less than the least resistance in the circuit.

(v) Maximum power is consumed by the least resistance, or the maximum current flows through the least resistance.

☛ *Example*

Find the equivalent resistance when 3 ohm and 6 ohm resistances are connected in parallel.

Solution

When resistances are connected in parallel the equivalent resistance is given by

$$\frac{1}{R_{eff}} = \frac{1}{R_1} + \frac{1}{R_2}; \frac{1}{R_{eff}} = \frac{1}{3} + \frac{1}{6} = \frac{2+1}{6} = \frac{3}{6} = \frac{1}{2}$$

$$R_{eff} = 2\ \Omega$$

☛ *Example*

For the given circuit, find the effective resistance between

(i) A and B
(ii) B and C
(iii) A and C

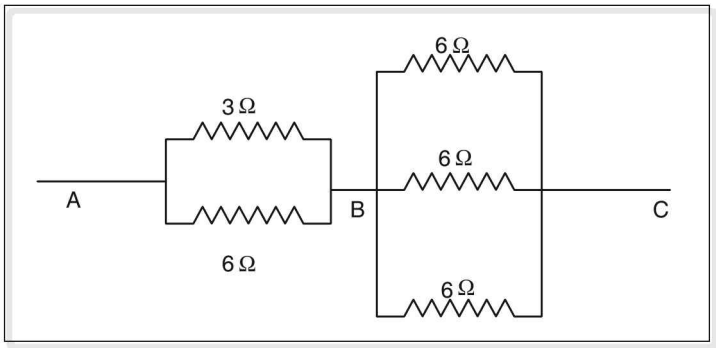


Figure 7.12

Solution

(i) Resistance between A and B is

$$\frac{1}{R_{AB}} = \frac{1}{R_1} + \frac{1}{R_2} \Rightarrow \frac{1}{R_{AB}} = \frac{1}{3} + \frac{1}{6} = \frac{2+1}{6} = \frac{1}{2}$$

$$R_{AB} = 2\Omega$$

(ii) Resistance between B and C is

$$\frac{1}{R_{BC}} = \frac{1}{R_1} + \frac{1}{R_2} + \frac{1}{R_3} = \frac{1}{6} + \frac{1}{6} + \frac{1}{6} = \frac{3}{6} = \frac{1}{2}$$

$$R_{BC} = 2\Omega$$

(iii) Resistance between A and C is obtained by the following simplified circuit. Now these two parallel branches are connected in series therefore the circuit simplifies as follows

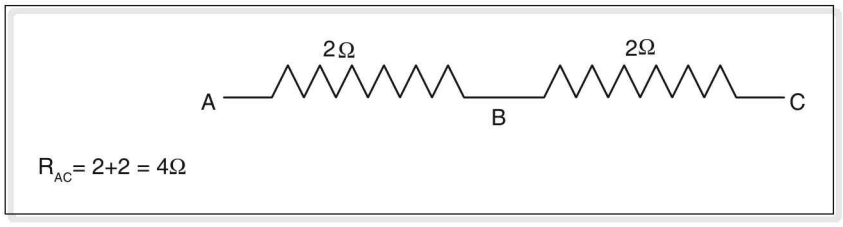


Figure 7.13

$$R_{AC} = 2 + 2 = 4\ \Omega$$

Colour coding of resistors

The value of resistance is usually coded by three bands of colour, followed by one band for the tolerance. The first two bands of colour indicate the first two digits. The third band gives the value of the multiplier, and the fourth band gives the value of tolerance.

The following chart gives the colour code of resistance

Black	0	0	10^0	-
Brown	1	1	10^1	-
Red	2	2	10^2	-
Orange	3	3	10^3	-
Yellow	4	4	10^4	-
Green	5	5	10^5	-
Blue	6	6	10^6	-
Violet	7	7	10^7	-
Grey	8	8	10^8	-
White	9	9	10^9	-
Gold	-	-	10^{-1}	5%
Silver	-	-	10^{-2}	10%
No colour	-	-	-	20%

(i) red, red, orange and no colour indicate the value of resistance, 22000 Ω ± 20% tolerance

(ii) Red, orange, orange and gold indicate the value of resistance 23000 Ω ± 5%

Bulbs in series

If the second terminal of a bulb is connected to the first terminal of the next bulb and so on, then the bulbs are said to be connected in series.

If a voltage (V) is applied across the bulbs in series, then,

(i) voltage applied is divided among the bulbs.

(ii) same current exists in all the bulbs.

(iii) the brightness of each bulb depends on the number of bulbs connected

(iv) failure of any bulb leads to a break in the circuit.

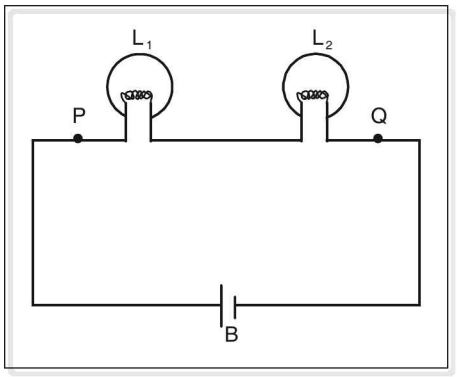


Figure 7.14

Bulbs in parallel

If the first terminal of all the bulbs are connected together and the second terminals of all the bulbs are connected together, then the bulbs are said to be connected in parallel.

If a voltage (V) is applied across the bulbs connected in parallel then,

(i) the voltage across each bulb remains the same.

(ii) the current is divided among the bulbs.

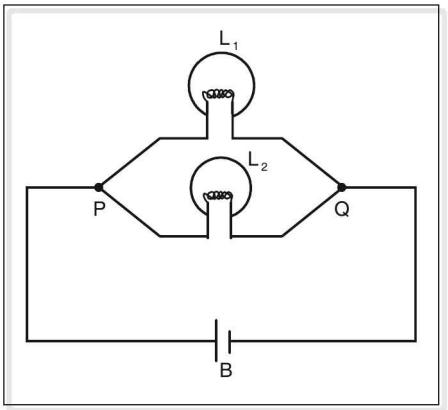


Figure 7.15

(iii) the power consumption of each bulb remains the same, even when the other bulb is removed.

(iv) the brightness of each bulb remains the same, even when the other bulbs is removed.

(v) the failure of a particular bulb has no effect on the working of the other bulbs in parallel.

EMF, terminal voltage and internal resistance of cell

An electric cell is a device which provides a constant potential difference between the ends of conductor connected across its two terminals. It converts chemical energy into electrical energy.

As a result of constant potential difference provided by a cell, a current flows through the circuit connected to it. A circuit is a path through which the current flows. If the path is broken due to a open switch or if there is an insulator within the path, current cannot flow and such a path is called open circuit. The closed continuous path made of conductor is called closed circuit.

Electromotive force (EMF) of a cell

When no current is drawn from a cell i.e., it is open circuit, the potential difference between the terminals of the cell is called electro-motive force (EMF).

The EMF is a characteristic of a cell and is different for different cells. For e.g., EMF of voltaic cell is 1.08 V and if leclanche cell is 1.5 V. It is denoted by symbol E. The EMF is equal to the work done in taking a unit positive charge through the external circuit connected to cell and electrolyte inside the cell.

Terminal voltage

The potential difference between the two terminals of the cell, when current is drawn from it, i.e., the cell is in closed circuit, is called its terminal voltage. It is denoted by 'V'.

Terminal voltage is equal to the work done in taking a unit positive charge around the external circuit connected across the terminals of the cell. The terminal voltage is always less than the EMF of cell.

The difference between E and V of a cell is called lost voltage. The voltage is lost because of the resistance offered by the cell itself to the flow of current.

Internal resistance of cell

When a current is drawn from a cell, in the external circuit it flows from anode to cathode and inside the cell through electrolyte, it flows from cathode to anode to maintain the continuity. The electrolyte offers certain resistance to the flow of current. This is called the internal resistance of cell. It is denoted by symbol r. If current I is drawn from the cell, the lost voltage or voltage drop across the internal resistance 'r' of the cell is given by

$v = Ir$

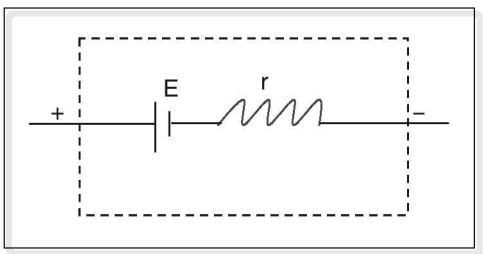


Figure 7.16

The internal resistance is considered to be in series with the cell.
EMF = terminal voltage + internal voltage drop.
$E = V + v$

The internal resistance of the cell depends upon the following factors:

(i) It is directly proportional to the distance between the two electrodes
(ii) It is inversely proportional to the area of cross section of the electrodes
(iii) It depends upon the nature, concentration and temperature of the electrolyte.

The terminal voltage depends on external resistance, internal resistance and emf of the cell, whereas emf is independent of external and internal resistance

Relation between emf, terminal voltage and internal resistance

When the cell is connected to an external resistance '**R**', a current '**I**' is drawn from the cell. The total voltage '**E**' of the cell is dropped across the internal resistance '**r**' and the external resistance '**R**'.

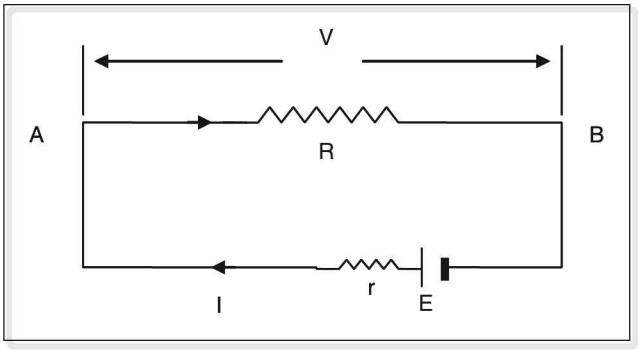


Figure 7.17

According to Ohm's law the voltage drop across the internal resistance 'r' is equal to Ir.
Similarly the voltage drop across the external resistance 'R' is equal to IR
Since both the resistances are connected in a series $E = Ir + IR$.

Let the voltage across the terminal AB be V. then,

$V = IR$ or $E = Ir + V$.

$$\Rightarrow E - V = Ir \text{ or } r = \frac{E - V}{I}$$

$$\text{Since } I = \frac{V}{R},\ r = \frac{E - V}{V/R} = \left(\frac{E - V}{V}\right) \times R = \frac{E - V}{I} \left(\because {}^{V}/_{R} = I\right).$$

The smaller the internal resistance the better is the cell. With the passage of time, the internal resistance of a dry cell increases.

Cells in series

When the negative terminal of a cell is connected to the positive terminal of the next cell and so on, the cells are said to be in series.

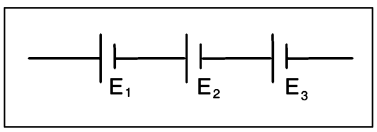


Figure 7.18

Cells connected in series are represented as

When several cells are connected in a series the total emf is the sum of the emf's of individual cells.

$$E = E_1 + E_2 + E_3$$

When the cells are connected in a series the total internal resistance is equal to the sum of the internal resistances of the individual cells.

$$r = r_1 + r_2 + r_3$$

Cells in parallel

When the negative terminals of all the cells are connected together and all the positive terminals are connected together then the cells are said to be in parallel.

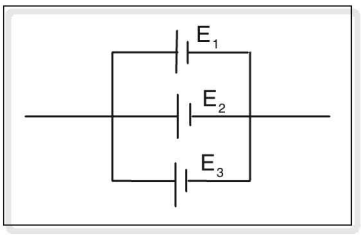


Figure 7.19

When several identical cells are connected in parallel the total emf is the same as the emf of individual cells.

When the cells are connected in parallel the equivalent internal resistance is given by "reciprocal of the total internal resistance is equal to the sum of the reciprocals of individual internal resistances".

$$\frac{1}{r_{eq}} = \frac{1}{r_1} + \frac{1}{r_2} + \frac{1}{r_3} \ldots\ldots$$

If cells are identical, then $r_{eq} = r/n$

Heating effect of electricity

When a potential difference is applied between the ends of a conductor, current begins to flow through it. The current flows as conductor contains large number of free electrons which begin to drift from the end at lower potential to the end at higher potential. In the process they constantly keep on colliding with atoms of conductor. The atoms gain energy and begin to vibrate more vigorously about their mean position. The average kinetic energy of atoms of conductor increases. This results in a rise in the temperature of the conductor and we say that the conductor has been heated. Thus flow of current has a heating effect on the conductor. The heat produced (Q) depends on the following factors.

1. Resistance of conductor: More the resistance of conductor more is the number of collisions between free electrons and atoms. Hence if a conductor offers more resistance to the flow of current, the heat produced in it will be more.
   $Q \propto R$
2. Current flowing through the conductor: Heat produced increases with the increase in the strength of the current flowing through conductor.
   $Q \propto I^2$
3. Time of flow of current: As the time of flow of current increases, the heat produced also increases.
   $Q \propto t$

Thus a conductor having larger resistance will generate more heat when a stronger current passes through it for a longer time.

If I is the current flowing a conductor of resistance R for time t, the heat produced (Q) is given by $Q \propto I^2Rt$

Electric energy or work

The energy spent in moving an electric charge through a conductor is called electric energy. This energy is provided by a cell which converts chemical energy into electric energy.

The work done in moving an electric charge through a conductor when some potential difference is applied at its ends is called electric work.

Let 'W' work is done in moving charge Q under potential difference V.

$$V = \frac{W}{Q}$$

or

$$W = VQ$$

But $Q = It$

$$\therefore W = VIt \quad (1)$$

According to Ohm's law

$V = IR$ Substitute. This in equation (1)

$\therefore W = (IR)It$

$W = I^2Rt$

Also $I = \frac{V}{R}$ Substitute. This in equation (1)

$$W = V\left(\frac{V}{R}\right)t$$

$$W = \frac{V^2}{R}t$$

$\therefore$ Electric work or energy can be expressed as

$$W = VIt = I^2Rt = \frac{V^2}{R}t = VQ$$

If the current is flowing through a heating element like heater, the electric energy is converted to heat energy.

Electric power

The rate at which electric work is done is called electric power.

or

The amount of electric work done is one second is called electric power.

Let a cell do work W in time t, the power delivered by it is

$$P = \frac{W}{t}$$

Units of electric power

If work done is measured in joule and time in second, power is measured in watt (W).

$$\text{Watt} = \frac{\text{Joule}}{\text{second}}$$

The power developed is 1 watt if 1 joule of work is done in 1 second.

$$1 \text{ watt} = \frac{1 \text{ joule}}{1 \text{ second}}$$

Higher units of power

1 kilowatt (kW) = 1000 W = 10^3 W

1 Megawatt(MW) = 10^6 W

1 Horsepower (hp) = 746 W

1 KW = 1.34 hp

$$P = \frac{W}{t} \quad \ldots . (1)$$

But W = VIt

$$\therefore P = \frac{VIt}{t}$$

P = VI

Power is also measured in volt-ampere

$W = I^2 R t$

Put in equation (1)

$$P = \frac{I^2Rt}{t}$$

$$\therefore P = I^2R$$

$$W = \frac{V^2}{R} t$$

Substitute in eqn. (1)

$$P = \frac{\frac{V^2}{R}t}{t}$$

$$P = \frac{V^2}{R}$$

$$\therefore P = \frac{W}{t} = VI = I^2R = \frac{V^2}{R}$$

Commercial units of electrical energy

$$\text{Electric power} = \frac{\text{electric work}}{\text{time}}$$

$$= \frac{\text{electric energy}}{\text{time}}$$

$\therefore$ electric energy = electric power × time

unit of electric energy = unit of electric power × unit of time = watt × second

The unit watt second gives the energy consumed by an electrical appliance, in t seconds whose power is P watt.

However for practical purposes, higher units of energy are used. The higher units of electrical energy are watt-hour (W h) or kilowatt-hour (kW h).

If an electric appliance of power P watt is used for 1 hour, the energy consumed by it is 1 watt hour.

1 watt hour = 1 watt × 1 hour

= 1W × (60 × 60) s = 3600 W s

1 Watt hour = 3600 J. (W s = J)

One kilowatt hour is the energy consumed in hour by an electric appliance whose power is 1 kilowatt.

1 kilowatt – hour = 1 kilowatt × 1 hour

= 1 kW × (60 × 60) s

= 1000 W × 3600 s

1 kilowatt hour = 3.6×10^6 J

1 kW h = 3.6 MJ

The electric meter installed at our homes measures the energy consumed in kilowatt hour.

☛ *Numericals*

1. Two resistors of 15 Ω and 4 Ω in series are connected to a cell of EMF 2V and internal resistance 1 Ω. Draw the circuit diagram and find the following:
   1. current in the circuit.
   2. voltage drop across each resistor.
   3. terminal voltage of the cell.
   4. lost voltage.

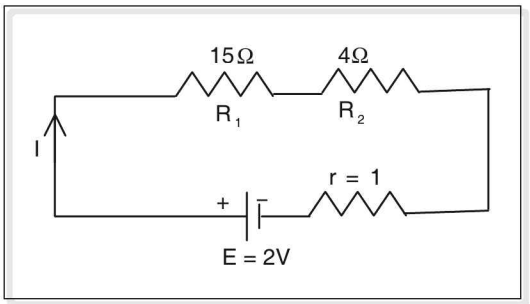


Figure 7.20

Solution

The circuit diagram based on the given information is as follows:

Current in the circuit

$$I = \frac{\text{EMF (E)}}{\text{Total circuit resistance (R)}}$$

The internal resistance of cell is in series with the external resistances.

$\therefore R = R_1 + R_2 + r = 15\ \Omega + 4\ \Omega + 1\ \Omega$
$R = 20\ \Omega$

$$I = \frac{2\ V}{20\ W}$$

$I = 0.1\ A$

2. Voltage drop or potential difference (V_1) across R_1

   $V = IR_1$ (Ohm's law)
   $V_1 = 0.1\ A \times 15\ \Omega$
   $V_1 = 1.5\,V$
   Similarly,
   $V_2 = IR_2 = 0.1\ A \times 4\ \Omega$
   $V_2 = 0.4\,V$

3. Terminal voltage

   It is equal to the total voltage drop in external circuit.
   $V = V_1 + V_2$
   $= 1.5\,V + 0.4\,V$
   $= 1.9\,V$

4. Lost voltage is the voltage drop across the internal resistance of cell.

   $v = Ir$
   $= 0.1 \times 1 = 0.1V$
   Also,
   $v = E - V$
   $= 2V - 1.9V = 0.1V$

2. A 20 V electric bulb is connected to a 20 V battery of negligible resistance. The resistance offered by the bulb is 5 Ω. Find:
   1. Current flowing through bulb.
   2. Power of bulb.
   3. Electric energy consumed in 3 hours.

Solution

Given,
Voltage of bulb (V) = 20 V
Resistance of bulb (R) = 5 Ω

Time (t) = 3 hours = $3 \times 60 \times 60 = 10800$ s

1. Current (I) through bulb = $\frac{V}{R}$

$$I = \frac{20\ V}{5\ \Omega} = 4\ A$$

2. Power (P) = $V \times I$
$= 20\,V \times 4\,A = 80\,W$

3. Electrical energy (W) = $V \times I \times t$
$= 80 \times 10800$
$= 864000\ J$

$$W = \frac{864000}{3.6 \times 10^{6}}\ kW\ h = 0.24\ kW\ h$$

3. The resistance of a electric heater is 200 Ω and it draws a current of 1A. Calculate its power and potential difference at its ends.

Solution

Given,
Resistance of heater (R) = 200 Ω
Current through heater (I) = 1 A
Power of heater (P) = ?
Potential difference (V) = ?
$P = I^2R$
$= (1)^2 \times 200$
$P = 200\,W$
$V = IR = 1A \times 200\ \Omega$
$V = 200\,V$

Watt hour meter

Watt hour is a unit in which the amount of electric energy consumed by a device is measured. A meter which measures the electric energy consumed by various electric appliances is called watt hour meter. It is commonly called a electric meter which is installed at every house by the state electricity board.

It is connected before the main switch so that the total current drawn by all the appliances in the house passes through it and it can give the reading of energy consumed. The meter readings at the beginning and end of month are noted. The difference in the readings gives the energy consumed in that month in watt hour. This reading when multiplied by



Figure 7.21 Kilowatt hour meter

rate the fixed by the state electricity board per unit (one kW h) gives the monthly bill which has to be paid by the owner of the house.

☛ *Example*

1. What should be the bill for the month of March for a heater of resistance 60.5 Ω connected to 220 V mains? The cost of energy is Rs 2 per kW h and the heater is used for 3 hours daily.

Solution

Given,

Resistance (R) of heater = 60.5 Ω

Potential difference (V) = 220 V.

Power (P) of heater is

$$P = \frac{V^2}{R}$$

$$P = \frac{220 \times 220}{60.5} = \frac{48400}{60.5}$$

$$= \frac{484000}{605}$$

P = 800 W.

Energy consumed by heater in a day = P × t

= 800 W x 3 h.

= 2400 W h.

There are 30 days in the month of March

∴ Energy consumed by heater in 30 days

= 2400 W h × 30 days

= 72000 W h = 72 kW h

Rs 2 is the charge for 1 kW h

For 72 kW h the charge will be

72 × 2 = Rs 144

☛ *Example*

Calculate the monthly bill if the following appliances are used as indicated below at the rate of Rs 5 per unit.

(i) A heater of 1000 watt for 2 hours daily.

(ii) Ten 60 watt bulbs for 5 hours daily.

Solution

Energy consumed by 1000 watt heater for 2 hours = 1000 × 2 = 2000 watt hour

Energy consumed by ten 60 watt lamp for 5 hours = 10 × 60 × 5 = 3000 watt hour

Total energy consumed per day = 2000 + 3000 = 5000 watt hour = 5 kW h

Total energy consumed for the month = 5 × 30 =150 kW h

Cost per kw h = Rs 5

Cost for 150 kw h = 150 × 5 = Rs 750

Wattage

The amount of electrical energy consumed by an electrical appliance per unit time is known as wattage.

Wattage (P) = I^2R or $\frac{V^2}{R}$ or Vi

Unit of wattage is watt (W).

Multiple units are kW (10^3 W) and MW (10^6 W).

Joule's Law

If W is the mechanical work done by a body (in joule) and Q is the quantity of heat produced by it (in calories) then is $W \propto Q$ or $W = JQ$ where J is the mechanical equivalent of heat or Joule's constant. J = 4.18 J cal^{-1}.

Joule's Law

We have seen that when electric current passes through a resistor, the electric energy spent or electric work done in moving the charges is converted into heat energy. In fact work done in any form i.e., mechanical, chemical etc., can be converted into heat energy. Joule conducted a series of experiments and found that the same work done in different forms produces the same quantity of heat in all cases. This is called Joule's law.

Joule showed that there is a definite relation between work done and heat produced by it.

Let W is the mechanical work done and Q is the heat produced in calories.

According to Joule's law

$W \propto Q$

$W = JQ$

where J is the constant of proportionality also called mechanical equivalent of heat

$$J = \frac{W}{Q}$$

The value of is 4.18 J cal^{-1} if work done and heat produced are measured in joules and calories respectively. If both are measured in joules, the value of J becomes unity.

Determination of the value of J using the Joule's calorimeter experiment

Take a calorimeter with a stirrer having a mass of w_1 gm.

Take some quantity of water in a calorimeter, let the mass of the calorimeter with stirrer and water be w_2 gm.

Place the calorimeter in a wooden box, insert a thermometer and note the initial reading of water let it be t_1°C.

Place a heater or coil (H) immersed in water, the other two ends of the heater are connected to a plug key, battery, an ammeter and a rheostat, a voltmeter is connected in shunt as shown.

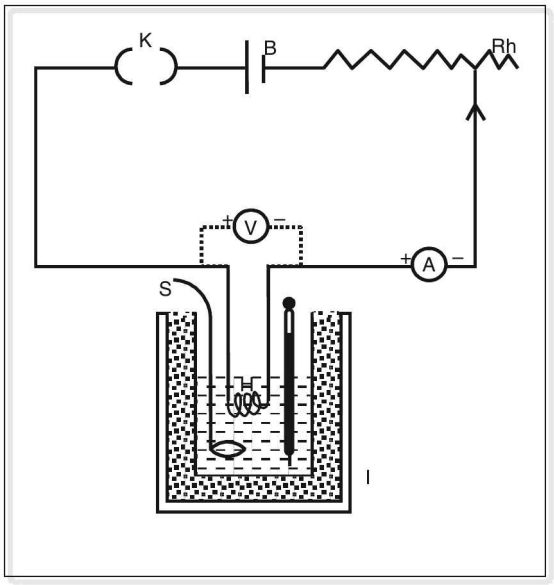


Figure 7.22 Joule's experiment

When the plug key (k) is closed, adjust the rheostat allowing a current of I ampere through the heater. Allow the current for a period of 30 min. Unplug the key and note the temperature. Let it be t_2°C.

The electrical work (w) done by heater = VIt

Heat developed by the heater (Q) is given by the sum of the heat gained by the calorimeter with stirrer (Q_1) and water (Q_2)

$\Rightarrow Q = Q_1 + Q_2$

$Q = w_1 s (t_2 - t_1) + ms_w (t_2 - t_1) = (t_2 - t_1) [w_1 s + m]$

where w_1 = mass of the calorimeter with stirrer

s = specific heat of copper

s_w = specific heat of water = 1 g cm^{-3}

mass of water, $m = (w_2 - w_1)$

According to Joule's law

$W = JQ$

$$J = \frac{W}{Q} = \frac{VIt}{(t_2 - t_1)[w_1 s + m]}$$

Experimentally it is found that J = 4.18 J cal^{-1}

Applications of heating effect of electric current

1. **Electric heater:** It uses nichrome wire as a heating coil. The spiral coil of nichrome wire is placed in a groove in a porcelain frame in a zig zag manner. When current is passed through it, the wire gets

red hot due to high resistance. The heat generated is used for cooking or heating purposes. Porcelain being an insulator of heat, does not become hot.

2. **Electric iron:** The coil of high resistance nichorme wire is enclosed in mica sheets which are placed in a heavy metal block. The coil is connected to power supply and transfers its heat to the metal block which is used to iron clothes. The heating coil is embedded between mica sheets as mica is a good conductor of heat but bad conductor of electricity.

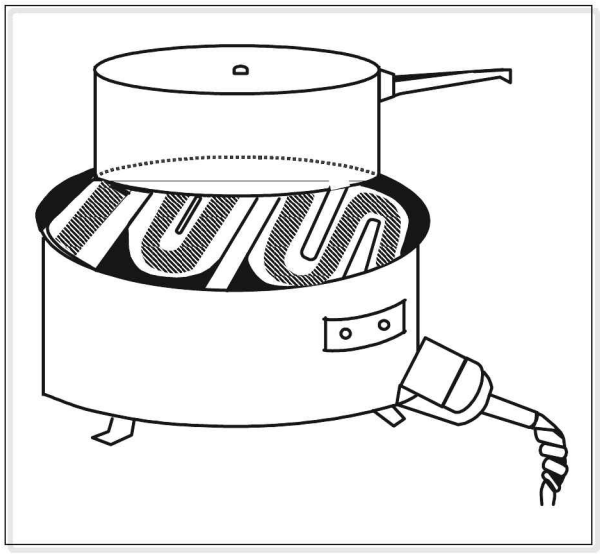


Figure 7.23 An electric stove

Lighting effect of electricity

When current passes through a conductor, it gets heated up. On being sufficiently heated, it may begin to emit light. This is called lighting effect of electricity.

For example electric bulbs and tubes.

Thus electric lamps work on lighting effect of electricity. Following types of electric lamps are commonly used.

1. Filament lamp.
2. Flourescent lamp.
3. Discharge lamp.
4. Electric arc lamp.

1. Incandescent or filament lamp

Incandescence is a phenomenon in which an object emits light when heated to a high temperature.

This phenomenon is used in electric lamps which are called incandescent lamps. The element which is heated is connected in lamps in the form of thin filament so as to offer high resistance to electric current. Hence it is also called filament lamp. Tungsten filament is most suitable for this purpose because of its following characteristics.

1. Its melting point is very high about 3665°C.
2. It offers high resistance. Hence it gets heated quickly.
3. On heating it becomes white hot and glows brightly for long time.

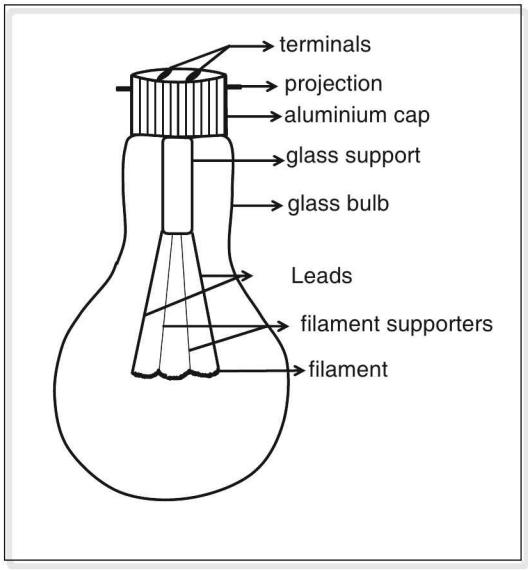


Figure 7.24 An electric bulb

The temperature of long thin tungsten filament can rise upto 3000°C and it is usually operated for a long period of time. Hence it evaporates and condenses on the inner side of the glass surface of the bulb forming a layer on it. This decreases the life and efficiency of lamp. To prevent this, the bulb is evacuated and filled with gas at low pressure. The rate of evaporation decreases and filament can reach a higher temperature.

In olden days carbon filament was used in bulbs. The melting point of carbon is 1600°C which is low considering the heat generated in the bulb. On evaporation, black layer of carbon gets deposited on the inner side of bulb. Life of carbon filament bulb is short.

Discharge lamps

They are so called because they use discharge tubes. Discharge tube is used to study conduction of electricity through gases. It consists of a glass tube having two electrodes fused to it and filled with one or more gases at very low pressure. A high potential is applied across the electrodes which ionises the gases. The ions get accelerated towards electrode with opposite polarity during which they collide with neutral gas atoms. If the pressure of gas is 1.65 mm of Hg, the tube begins to glow due to successive collisions. The nature of gas determines the color of light emitted.

1	Neon	Orange
2.	Nitrogen	Red
3.	Sodium	Yellow
4.	Mercury	White
5.	Chlorine	Green
6.	Hydrogen	Blue

Fluorescent lamps

Discharge lamp when coated with fluorescent material from inside is called fluorescent lamp.

A high voltage applied to coils heat them which results in emission of electrons. These electrons collide with the atoms of mercury vapours present in the tube. These collisions cause emission of invisible ultraviolet (uv) material which emits visible bright light.

Compact fluorescent lamps (CFL) use small fluorescent tubes. Electronic circuits control the intensity of current. Main advantage of CFL is that it consumes small amount of electrical energy.

The advantages of fluorescent lamps over incandescent lamps are listed below:

1. Fluorescent lamps have a life span which is almost 5 times the life span of incandescent lamps.

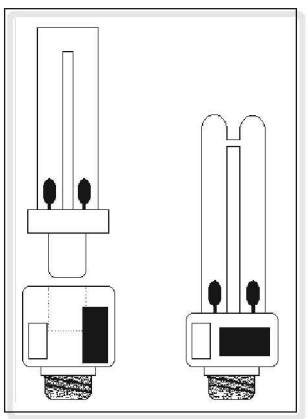


Figure 7.25

2. Consumes less electrical energy.
3. The shadow formed by fluorescent lamps is faint when compared to incandescent lamps.
4. The brightness of fluorescent lamp is more for same power consumed.

Electric arc lamps

They are used to produce high intensity light as in search lights or cinema projector.

It consists of two carbon rods having pointed tips at one end. The tips are separated by a very small gap. A potential of about 40 V to 60 V is applied across the carbon rods and a glow is produced between the pointed tips. The glow emits very intense light. Using concave mirrors, this light can be focused on a point or can be converted to a parallel beam.

Household distribution of current

Fuse

Fuse is a very important safety device in an electric circuit. It is a thin wire which melts and breaks the electric circuit due to (1) overloading (2) short circuit (3) fluctuation in power supply

Characteristics of fuse

1. It should have a low melting point. Generally a fuse wire is made from an alloy containing 50% lead and 50% tin. Its melting point around 200°C.
2. Since fuse should melt, before the excess of current reaches electric appliances and damages them, the temperature of fuse should rise quickly. This is possible if the resistance of fuse is very high.

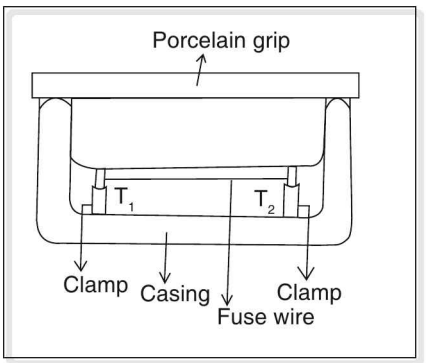


Figure 7.26

A fuse wire is stretched between two terminals T_1 and T_2 fixed to porcelain casing or holder which fits into porcelain grip or socket. The porcelain socket is connected to live wire. For any reason, if excessive current begins to flow the fuse wire gets heated up more quickly than conducting wires and melts, breaking the circuit. Thus the excessive current is prevented from flowing through electric appliances, which may damage them.

Current rating of fuse

Different electric circuits can withstand certain maximum amount of current without getting heated up and damaging themselves. For example, for a lighting circuit this limit is 5 A and for a heating circuit, it is 15 A. If current approaches this value, fuse wire will melt. Thus the current ratings of fuse should be less than the maximum current that can safely flow through the circuit.

Fuse should be connected in live wire of circuit. If it is connected to neutral wire, the fuse will melt when current exceeds its rating, but the appliance will still be connected to high potential of supply

through live wire. If a person touches the appliance, his body provides a path for current to flow to earth which is at zero potential. Thus person may get severe shock which may prove fatal. Hence fuse is connected to live wire so that if it melts, the appliance is no longer connected to the high potential of the supply and possibility of electric shock is eliminated.

Earthing

When the insulation of live wire in a circuit inside a electric appliance melts due to some reasons, the live wire may touch the metal casing of the appliance. Thus metal casing is connected to high potential. This metal casing of appliances such as mixer, fan, refrigerator, toaster etc., is continuously touched by us by our bare hands. Hence we may get a severe shock as current will flow through our body to earth which is considered as zero potential. To prevent this, metal casings of all appliances are connected to earth by a conducting wire. This is called earthing.

Thus earthing involves connecting metal body of electric appliance to thick copper wire which is connected to a copper plate buried deep in the earth. The copper plate is surrounded by a mixture of charcoal and common salt.

Earthing helps in two ways. When live wire comes in contact with metal casing, the current instead of flowing through the person touching it, flows through earth wire. Also when live wire touches wire i.e., during short circuit, the current from metal body flows through earth wire which offers very less resistance. Thus a large current is drawn which melts the fuse instantly. The appliance does not receive current any more thus preventing it from potential damage.

Household distribution of current

Electricity is supplied to houses through over head wires on poles or through underground cable. The electric cable consists of 3 wires viz live wire or phase wire, neutral wire and earth wire. The company which supplies electric power installs its fuse near the pole or before the kWh meter. If the house draws more than allotted current, this fuse which is called pole or company fuse melts and supply is cut off.

At the power station, neutral and earth wire are connected together so that both are at ground or zero potential. The live and electrical wire are connected to the input terminals of kWh meter and the earth wire to the body of meter. On the output side of kWh meter, fuse is connected in series with live wire. This fuse is called main fuse. After this, live and neutral wires are connected to the main switch.

The main switch is an iron clad double pole switch. It is operated by a single lever and can connect and disconnect both live and neutral wire simultaneously to the house. These two wires then carry power to whole house. The distribution of electric power in the house is done by two types of wiring system.

1. Tree type
2. Ring type

Tree system

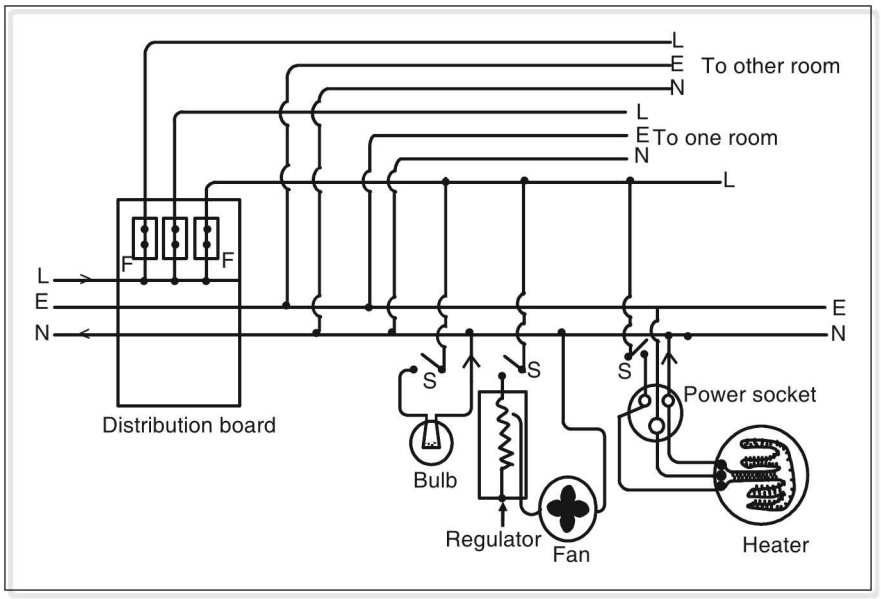


Figure 7.27

Phase wire to different rooms is taken from the main board, through a fuse on the main board. Neutral and earth wire is common to all the circuits.

The different circuits coming from the main board is connected in parallel.

If there is a short circuit in one particular circuit, then only the fuse of that circuit is broken and the power supply to other parts of the circuit will not be affected. This is the major advantage of parallel circuit.

Disadvantages of Tree system

A long length of wire is required as all the wires are to be drawn from the main board itself. It increases the initial installation cost. When the fuse of a particular line melts, it affects all the appliances connected to that circuit.

Ring system

It consists of a ring formed by all the three wires–phase, neutral and earth and runs through all the rooms and comes back to the main board. Each appliance is connected to this ring. Appliances which are likely to get short circuited are protected by a fuse near the appliance.

Advantage

As one ring feeds the entire house, it requires less length of wire. It reduces installation cost.
As there is an individual fuse for each appliance, only faulty appliance will not work and all the other appliances will not be affected. It is easier to install.

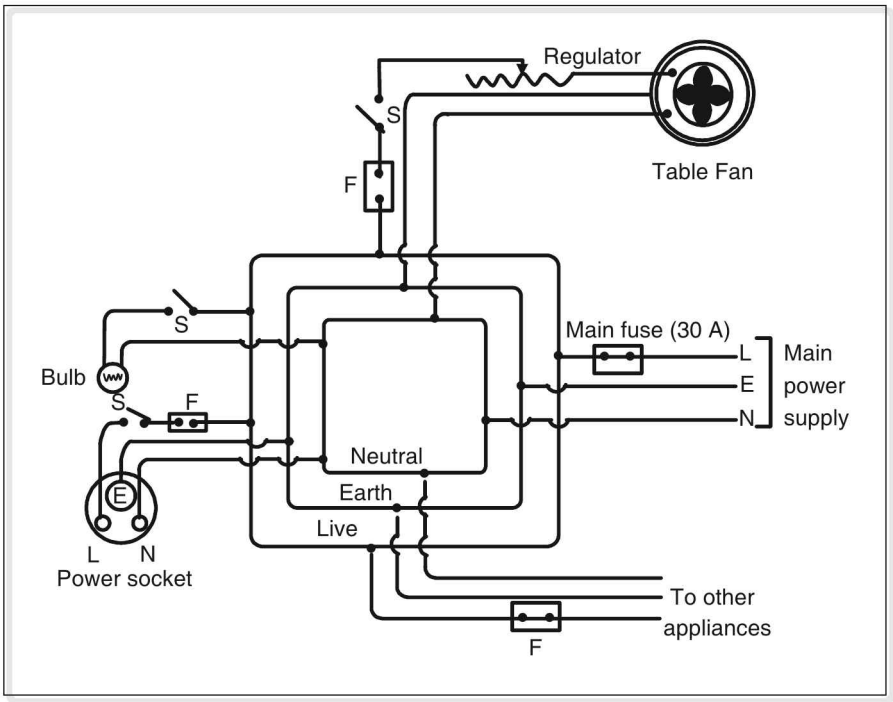


Figure 7.28

Switch

Switch is a device to connect or disconnect an electric appliance. Lever type switch is used for light, fan, heater etc.

Push switch is used in devices like calling bell.

Single pole double throw switch is used for staircase lights.

One staircase bulb can be switched on or off either from the first floor or from the ground floor.

The same switch is used near the room entrance and near the bed, so that we can switch on the light as we enter the room and switch it off when we go to bed.

The functioning of a staircase type switch is shown below.

We should always connect the switch in the live wire and not on the neutral wire.

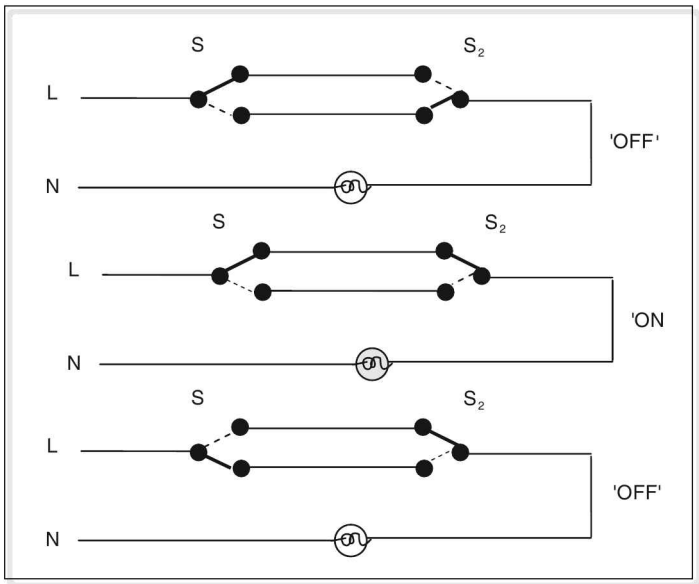


Figure 7.29

This is for safety reasons. Suppose the switch is connected to neutral, the live wire remains connected to the appliance even when the switch is in off position. This may prove fatal if proper earthing is not there. We should not touch the switch with a wet hand as it may act like a conducting path and one receives a fatal electric shock.

Socket

Sockets are elements of electric circuit from where we draw power to movable electric appliances like radio, electric iron, toaster etc. It has three terminals live, neutral, and earth. The bigger hole at the top is the earth terminal, Live terminal is on the right and neutral terminal is on the left.

Plugs

It is the counterpart of socket which is kept as the input terminal to electrical appliances. It has three pins. Live pin is on the right, neutral pin is on the left and the earth pin is thick and at the top.

Plugs and sockets are available for different applications like 5 A and 15 A. Proper sockets and plugs are to be used as otherwise it will overheat the socket/plug and burn.

Colour coding of wires

All electrical appliances are to be wired as per some standard. The old convention is Red wire for Live, Black wire for Neutral and Green wire for Earth.

According to the specifications of a recent international convention live wire is to be connected to brown, neutral wire to be connected to light blue and earth wire is connected to green or yellow wire.

Power transmission

Mainly we use alternating current as it is easy to step up and step down ac Voltage using a transformer. Now let us see how the power from the power generating station reaches us.

At the generating station electricity is produced at a potential of 11 kV

It is stepped up by a transformer to 220 kV and sent to cities located far away.

At the city outskirts it is stepped down to 11 kV.

At the street level it is further reduced to 440 V (three phase) (220 volt single phase) and supplied to the houses.

The schematic diagram of power distribution is shown below.

Let us now see why it is necessary to step up the voltage at the sending end and step down the voltage at the receiving end.
We know power = VI

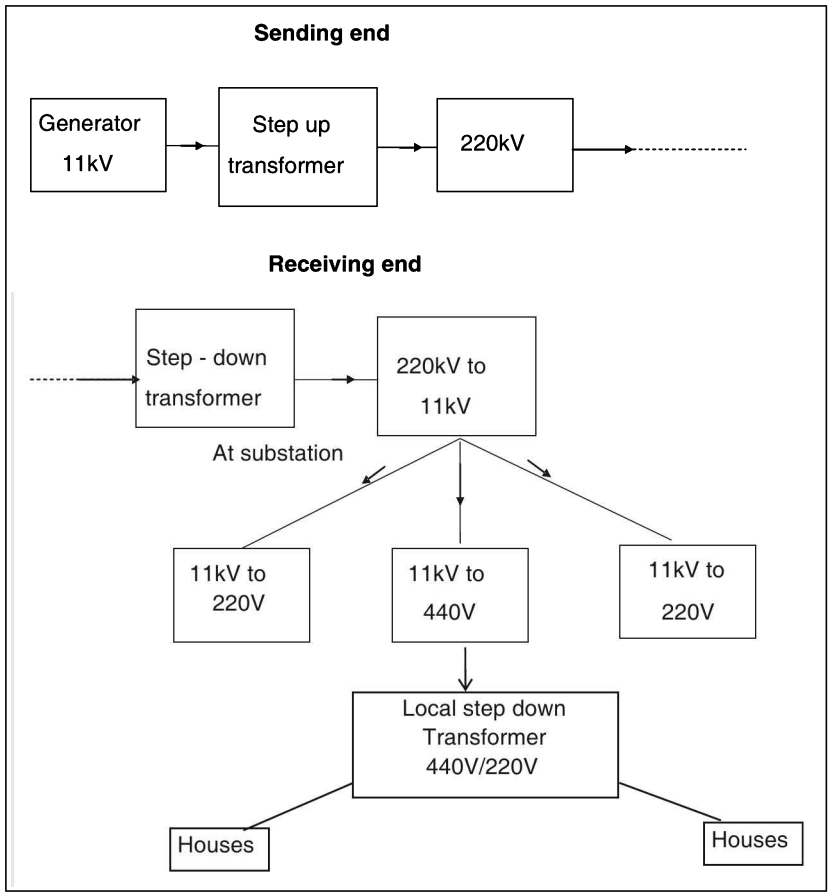


Figure 7.30

For example power required at Hyderabad can be sent from the power station at Nagarjunasagar in two ways. If we send at 220 volt a large current flows through the transmission wires and hence the heat loss due to transmission wire will be large.

This is because the power is I^2R. Since current is more for a given conductor the loss is more.

Further we require thick wires which cost more.

If we send at high voltage then less current would flow for a given power and hence the loss will be minimized.

Further more we can use thin wires as compared to earlier case.This reduces the cost of transmission wires.

Chemical effect of electricity

The most common chemical effect of current electricity is electrolysis. Just like solids, some liquids also conduct electricity. Such liquids are called electrolytes. When electric current passes through an electrolyte it decomposes into its constituents.

Electrolysis

The process in which an electrolyte undergoes decomposition into its constituents when an electric current is passed through it is called electrolysis.

The plates inside the electrolyte through which the electric current is passed from the external battery is called **electrode**.

The electrode which is connected to the positive terminal of the battery is called **anode**.

The terminal which is connected to the negative terminal of the battery is called **cathode**. The vessel in which the electrolyte is kept is called **voltameter**.

Electrolysis of water

Take some water in a glass vessel and add a few drops of sulphuric acid. Take two copper rods and connect the positive and the negative terminal of the battery to them.

When the current is passed through them the water molecules dissociate into hydrogen ions and oxygen ions.

Negatively charged oxygen ions are attracted by the positive electrode anode.

Positively charged hydrogen ions are attracted by the negative electrode cathode.

So Hydrogen and Oxygen ions are collected at the cathode and the anode respectively.

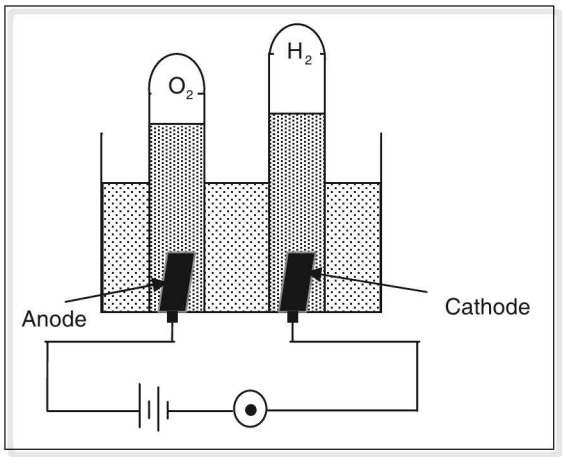


Figure 7.31

Electrolysis of copper sulphate

Here copper sulphate is the electrolyte and both the electrodes are copper rods.

When current is passed through the copper sulphate solution copper sulphate molecules split into copper ions Cu^{++} and sulphate ions SO_4^{--}

$$CuSO_4 \rightarrow Cu^{++} + SO_4^{--}$$

When current is passed, the positive copper ions are attracted by the cathode and a coating of copper is observed on the cathode. The negative sulphate ions are

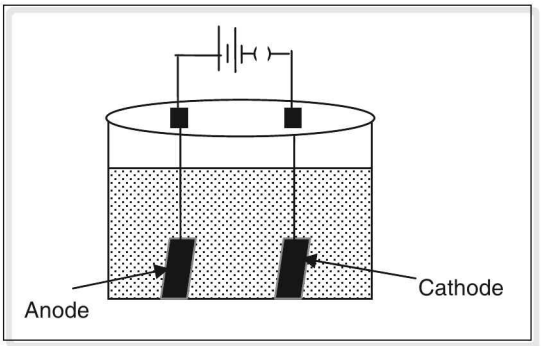


Figure 7.32

attracted by the anode and $CuSO_4$ molecules are formed, thereby maintaining the concentration of electrolyte.

Uses of electrolysis

Electroplating, extraction of metals from ore (i.e., Metallurgy), and purification of metals, electrotyping

Electroplating

The process of depositing a thin layer of desired metal over another metal by passing an electric current through some electrolyte is called electroplating.

The object to be electroplated is taken as cathode and the metal to be coated is kept as anode. Suitable electrolyte is used.

Some common electroplating applications

Copper plating

Cathode	:	Article to be plated
Anode	:	Copper rod
Electrolyte	:	Copper sulphate solution

Silver plating

Cathode	:	Article to be plated
Anode	:	A silver rod
Electrolyte	:	Silver nitrate solution $AgNO_3$

Nickel plating

Cathode	:	Article to be plated
Anode	:	Nickel rod
Electrolyte	:	Nickel ammonium sulphate solution $NiSO_4(NH_4)_2SO_4 6H_2SO_4$

For Chromium plating

Cathode	:	Article to be plated
Anode	:	Chromium rod
Electrolyte	:	Chromic acid solution H_2CrO_4

Laws of Electrolysis

Michael Faraday carried out a series of experiments and proposed two laws which are known as "Faraday's laws of electrolysis".

Following are the characteristics of electrolysis process.

1. The concentration of electrolyte remains unaffected during electrolysis.
2. The anode gains electrons as negative ions come to it and cathode loses electrons to positive ions which are attracted to it.
3. Number of electrons gained by anode is equal to number of electrons lost by cathode.
4. During electrolysis, mass of cathode increases while that of anode decreases by same amount.

Faraday's first law of electrolysis

The mass of the substance deposited at the cathode is directly proportional to the quantity of electricity. (i.e., charge) and the time for which the current is passed through the Electrolyte.

Larger current passing through a given electrolyte for a given time or the same current passing for a longer time will deposit more mass on the cathode.

$m \propto I$; or $m \propto t$; $m \propto It$
since $Q = It$, $m = ZQ$
The constant Z is known as **electrochemical equivalent** of the substance.

The electrochemical equivalent of a substance is defined as the mass of the substance which is produced during electrolysis by passing one coulomb of charge. The unit of Z is g C^{-1}.

e.c.e. (electro chemical equivalent)

The mass of ions liberated at the electrode when 1 coloumb of electricity is passed through the electrolyte is known as e.c.e. It is denoted by Z and $Z = m/q$.

Units of Z are g C^{-1} (or) g A^{-1} s^{-1}.

Faraday's second law of electrolysis

When the same quantity of electricity passes through different electrolytes, the masses of ions liberated at the respective electrodes are proportional to their chemical equivalents.

If E_1, E_2 and E_3, are the chemical equivalents and Z_1, Z_2 and Z_3, are the electro chemical equivalents of three substances, says Cu, Ag and Zn respectively and their masses liberated are m_1, m_2 and m_3, when the same quantity of electricity is passed through their respective electrolyte then according to Faraday's second law of electrolysis

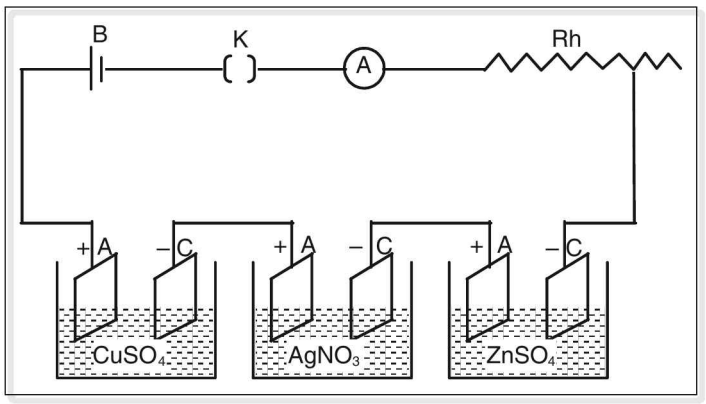


Figure 7.33

$$\frac{m_1}{E_1} = \frac{m_2}{E_2} = \frac{m_3}{E_3} \text{ or}$$

$$m_1 : m_2 : m_3 = E_1 : E_2 : E_3 \quad \text{---------- (1)}$$

This is one form of Faraday's second law. If we apply the first law then

$$m_1 = Z_1 q \qquad m_2 = Z_2 q \qquad m_3 = Z_3 q$$

$$m_1 : m_2 : m_3 = Z_1 : Z_2 : Z_3 \text{ ---------- (2)}$$

Combining equations (1) and (2) we get

$$Z_1 : Z_2 : Z_3 = E_1 : E_2 : E_3$$

Chemical equivalent (E)

The ratio of the atomic weight of an element to its valency is known as the chemical equivalent. It is denoted by E.

The amount of charge which on passing through an electrolyte liberates 1 gram equivalent of the substance at an electrode is called Faraday constant. The value of Faraday constant is 96500 C.

☛ *Numerical*

1. The e.c.e. of copper is 0.33×10^{-3} g C^{-1}. If 1A current is passed for 2 hr. through copper sulphate find the mass of copper deposited on cathode.

Solution

Given,
Current (I) = 1A
time (t) = 2 h = $2 \times 3600 = 7200$ seconds
e.c.e. (z) of copper = 0.33×10^{-3} g C^{-1}
mass (m) of copper deposited = ?
we have

$$Z = \frac{m}{It}$$

$\therefore m = Zit$
$= 0.33 \times 10^{-3} \times 1 \times 7.2 \times 10^{3} = 2.376$ g

2. During gold plating, 3 A current is passed for 1000 seconds and 2 g of gold is deposited. Find electro chemical equivalent of gold.

Solution

Given,
Mass (m) of gold deposited = 2 g

time(t) from which current is passed = 1000 s
e.c.e. of gold (Z) = ?

$$Z = \frac{m}{It} = \frac{2\ g}{3 \times 1000} = 0.000666\ g\ C^{-1}$$

Dry cell

A dry cell is a form of primary cell which is used widely in torches, transistor radios, electric toys, clocks, etc.

It consists of a zinc container at the centre of which a carbon rod is fixed. Carbon rod acts as anode while the zinc body acts as cathode. The dry paste of ammonium chloride (NH_4Cl) and zinc chloride ($ZnCl_2$) serves as an electrolyte. The electrolyte is mixed with powdered coke. The carbon rod is surrounded by graphite and manganese dioxide. The top of carbon rod is connected to a brass cap to improve electrical conductivity. The container is sealed with sealing wax to make it air and water tight.

Zinc atoms from the container lose two electrons when the cell is connected to a electric circuit

$$Zn \rightarrow Zn^{2+} + 2e^-$$

These electrons flow through external conducting wire and reach carbon electrode. This constitutes electric current. The conventional current is taken from carbon rod to zinc body.

The ammonium chloride in the electrolyte dissociates as

$$NH_4Cl \rightarrow NH^+_4 + Cl^-$$

The electrons lost by zinc combine with positive ammonium ions (NH^+_4)

$$2NH^+_4 + 2e^- \rightarrow 2NH_3 + H_2$$

The hydrogen gas produced is reduced by manganese dioxide.

$$H_2 + 2MnO_4 \rightarrow H_2O + MnO_3$$

The positive zinc ions combine with negative chlorine ions to form zinc chloride.

$$Zn^{2+} + 2Cl^- \rightarrow ZnCl_2$$

The powered carbon present in electrolyte facilities the movements of charges.

Dry cell gives an emf of 1.5 volt. Since it is primary cell, once the chemicals present in cell are used up, cell has to be replaced by a new cell.

There are many secondary cells or accumulators which are also in wide use nowadays in which chemical reactions can be reversed by passing current in reverse direction. For e.g., lead sulphuric acid battery is widely used in automobiles and industries. Cell phones are powered by battery which can be easily recharged.

test your concepts

Very short answer type questions

1. Define resistance and one ohm resistance.
2. How is equivalent resistance varies as resistances are connected in series and parallel?
3. What is a fuse and specify its uses?
4. What is Incandescence?
5. State laws of resistance and define resistivity.
6. What is electricity?
7. Find the total internal resistance of the cells when they are connected in
   (a) series and (b) parallel connections
8. The net charge flowing through a cross-section of conductor in unit time is called _____.
9. What is the use of kilowatt-hour meter?
10. Define static and current electricity.
11. What is the value of mechanical equivalent of heat (J).
12. The unit of resistance is _____.
13. Define conductance, conductivity and super conductors.
14. What are conductors and insulators? Give one example of each _____.
15. Define terminal voltage, internal resistance and lost voltage.
16. Unit of electric charge is_____.
17. What happens to the value of resistance when the length and the area of cross section are tripled?
18. What is the reason for conduction of electricity?
19. What is watt hour?
20. Electric potential is measured in _____.
21. On increasing the temperature what is the variation of the resistance of a metallic conductor and semi-conductor?
22. Define electric current.
23. Define electric energy or work.
24. Define electric potential, potential difference, and EMF
25. What is the relation between current (I), emf (E), terminal voltage (V) and internal resistance (r)?
26. Fuse should be connected to _____ wire of the circuit.
27. Define electric power and give its units in S.I.

28. What is meant by conventional and electronic current?

29. State Ohm's law and what are ohmic and non-ohmic conductors.

30. What is electro plating?

Short answer type questions

31. Give reasons why the terminal voltage is less than the emf of a cell.

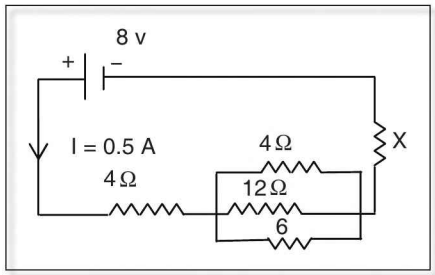
32. Why is it desirable to transmit high voltage and low current during the transmission of power?

33. How are the following systems connected?
(a) ring system
(b) tree system

34. Three resistors of 5 Ω, 3 Ω and 7 Ω are connected in series across a battery of potential difference 3 V. Find the current flowing through 3 Ω and 7 Ω resistor.

35. Explain the factors influencing the internal resistance of a cell.

36. Calculate the value of resistor X.

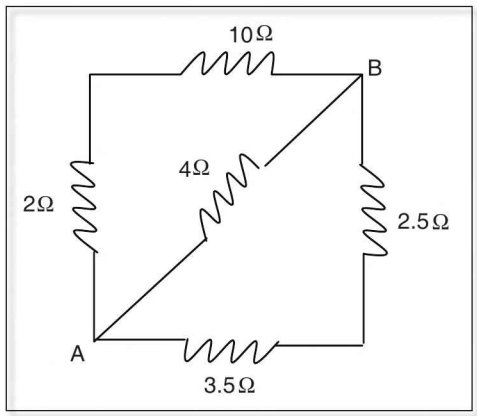


37. Explain how earthing is done.

38. Two heaters of 200 W each and two bulbs of 40 W each operate 4 hours daily on the supply of 200 V. If the cost of energy is 3 rupees per kW h. Find the bill for the month of April.

39. Give an expression for equivalent internal resistance when cells of different internal resistances are connected in series and parallel.

40. Find the equivalent resistance across A and B in the following circuit.



41. When does a fuse melt?

42. The length and area of cross section of a wire are 1m and 0.040 cm^2 respectively. If its resistance is 4 ohm, calculate its resistivity?

43. Derive $r = \frac{E - V}{I}$ where r is internal resistance, E – emf, V – terminal voltage, I – current.

44. What is the need for a step up and step down transformer at various stages of power transmission?

45. Derive 1 kW h = 36×10^5 J.

Essay type questions

46. State Joule's law. Explain an experimental method to determine the value of J.

47. State Faraday's first law and second law of electrolysis and verify it experimentally.

48. Derive an expression for equivalent resistance of resistors connected in series.

49. Explain the working of the following
    (1) Filament lamp
    (2) Flourescent lamp
    (3) Discharge lamp
    (4) Electric arc lamp
    (5) Drycells

50. Derive an equivalent emf when cells are connected in series and parallel.

CONCEPT APPLICATION



Concept Application Level—1

Direction for questions 1 to 7: State whether the following statements are true or false.

1. Conventional current always flows from a body having high charge density to body having low charge density.
2. At constant temperature, the resistance of a conductor changes according to the applied voltage.
3. The used up chemicals in a dry cell can be re obtained by charging it.
4. In an electrolyte, the movement of ions is responsible for electric current.
5. Electrotyping is an application of electrolysis.
6. Electric potential at a point in an electric field is defined as the work done in moving a unit positive charge from infinity to that point.
7. Earthing of electric appliances prevents electric shock.

Directions for questions 8 to 14: Fill in the blanks.

8. Power transmission is carried out at high _____ and low _____.
9. Switch is always connected to the _____ wire.
10. Rate at which electric work is done is called _____.
11. The process of depositing a thin layer of desired metal over another metal by passing an electric current through some electrolyte is called _____.
12. Conductance is the reciprocal of _____.
13. For a given cell, its terminal voltage depends on _____ and _____.
14. The amount of work done by the cell on a unit positive charge carrier to force it to go to the point of higher potential is called _____.

Direction for question 15: Match the entries in column A with appropriate ones from column B.

15.

A.	Electric potential	( )	a.	Transistor
B.	Commercial unit of electrical energy	( )	b.	ohm-meter
C.	Resistance	( )	c.	$E = E_1 + E_2$
D.	Superconductor	( )	d.	$Q = i^2Rt$
E.	1 ohm	( )	e.	Black
F.	Non-ohmic conductor	( )	f.	$v = W/Q$
G.	Resistivity	( )	g.	kW h
H.	Cells connected in series	( )	h.	$\frac{\rho L}{A}$
I.	Heat produced in a Conductor	( )	i.	Zero resistance
J.	Neutral wire	( )	j	1volt/1ampere

Direction for questions 16 to 30: For each of the questions, four choices have been provided. Select the correct alternative.

16. One kilowatt is equal to _____ horse power.
(1) 1.34 (2) 1.32 (3) 1.28 (4) 1.38

17. In a series circuit
(1) current flow is same in all the resistors of the circuit
(2) potential difference across each resistor in the circuit is same.
(3) Both (1) and (2)
(4) None of these

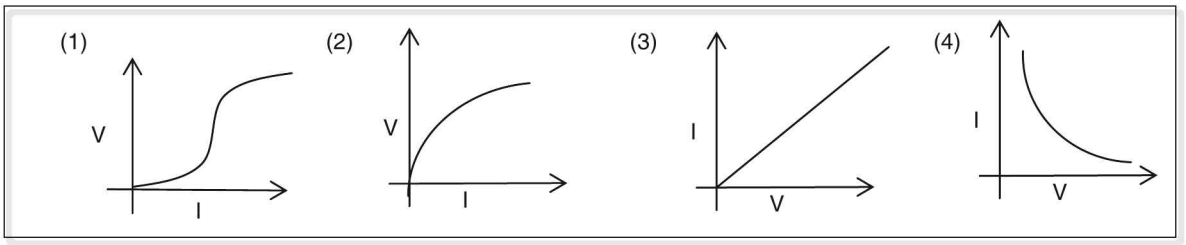
18. The colour code of a resistor is brown, black and brown. Then the value of resistance is _____.
(1) 10 Ω (2) 100 mΩ (3) 0.1 kΩ (4) 100 + 5%

19. The free electron density is more in _____
(1) conductors (2) insulators (3) semi conductors (4) electrolytes

20. Among the following statements,
(A) A discharge lamp uses a discharge tube which is filled with a gas at a very low pressure.
(B) Always white light is emitted independent of gas taken in the lamp.
(1) Only A is true (2) Only B is true
(3) Both A and B are true (4) Both A and B are false

21. The quantity of heat produced by a current in a conductor is directly proportional to _____ in the conductor.
(1) the time of current flow. (2) the square of magnitude of current.
(3) the resistance. (4) All the above

22. Which of the following graphs represents a ohmic conductor?



23. During the electrolysis process,
(1) mass of anode decreases (2) mass of cathode decreases
(3) amount of the electrolyte increases (4) Both (1) and (3)

24. Among identical spheres A and B having charges as –5 C and –16 C
(1) –5C is at higher potential. (2) –16C is at higher potential.
(3) both are at equal potential. (4) it cannot be said.

25. _____ is independent of size and shape of a conductor.
(1) Resistance (2) Conductance (3) Resistivity (4) all the above

26. The mass of the element deposited at the electrode when one coulomb of charge is passed through the electrolyte is called _____ of the element.
(1) equivalent weight (2) electrochemical equivalent
(3) mass equivalent (4) None of these

27. In a parallel circuit of bulbs,
(1) same current exists in all the bulbs
(2) voltage across each bulb remains the same
(3) failure of any bulb leads to a break in the circuit
(4) All the above

28. Resistance of a metallic conductor depends on _____.
(1) its length (2) its area of cross section
(3) temperature (4) All the above

29. Three resistors each of 10 Ω are connected in series to a battery of potential difference 150 v. The current flowing through it is _____ A.
(1) 45 (2) 5 (3) 15 (4) 20

30. The terminal voltage of a cell in open circuit condition is

(1) less than its emf. (2) more than its emf.

(3) equal to its emf. (4) depends on its internal resistance.

Concept Application Level—2

31. In a three pin plug, why is the earth pin longer and thicker than the other pins?

32. Given below are the characteristics of wires made of different materials.

wire A: low resistivity and high melting point.

wire B: low resistivity and low melting point.

wire C: high resistivity and low melting point.

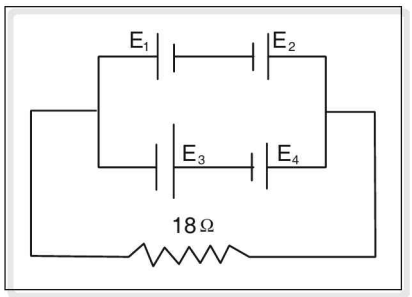
wire D: high resistivity and high melting point.

Which of the wires can be used as a (i) conducting wire (ii) fuse wire (iii) heating element?

33. How does the power-voltage rating of an electrical appliance decide the type of connecting wires to be used?

34. Two bulbs 60 W, 220 V and 100 W, 220 V are connected in series to a 440 V supply. Which of the two bulbs fuse? What happens if these bulbs are connected in parallel?

35. Four cells each of emf 1.9 V and internal resistance 1 Ω are connected to an external resistance of 18 Ω as shown in the figure. Find the terminal potential difference of this combination.



36. A technician has two resistance coils. By using them separately, in series or in parallel he is able to obtain resistances of 3 , 4 , 12 and 16 (not in the same order). What is the resistance of the two coils?

37. The heat generated in the coil of resistance 180 Ω which is connected to a silver voltmeter is 364.5 K J. If a steady current is passed through the circuit for 15 min, find the mass of Ag that gets deposited on the cathode. Relative atomic mass of silver is 108 g and Faradays constant is 96500 C.

38. A 200 W, 220V light bulb has a greater resistance than a 100W, 200V light bulb.

39. When a 3 Ω resistance coil is connected across the terminals of a battery, the current is 1.5 A. When a 5 Ω resistance is connected across the same battery, the current is 1 A. Find the internal resistance of the battery.

40. Even currents of very low magnitude can lead to death. Is heat produced by the electric current responsible for the death? Explain.

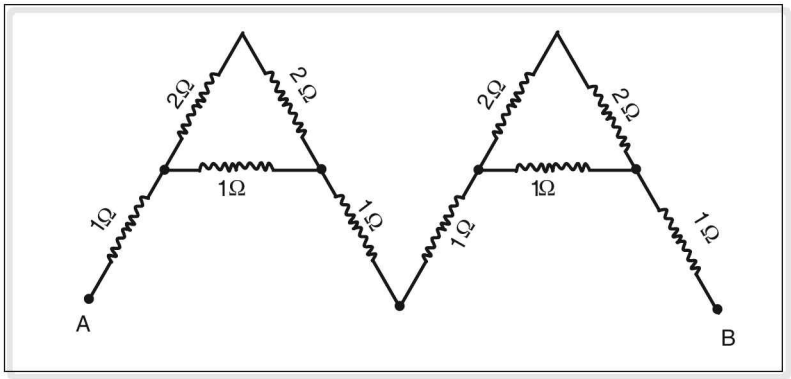
41. It is easier to start a car on a warm day then on a cold day.

42. The table given below given the current (in amperes) passing through conductors A and B for different values of potential difference (V). All values are measured at the same temperature.

V	I	V	I
4	2.8	4	9
5	3.5	5	11.25
6	4.1	6	13.5

Which among A and B is an ohmic conductor and also find the resistance of that conductor.

43. Calculate the effective resistance between A and B.

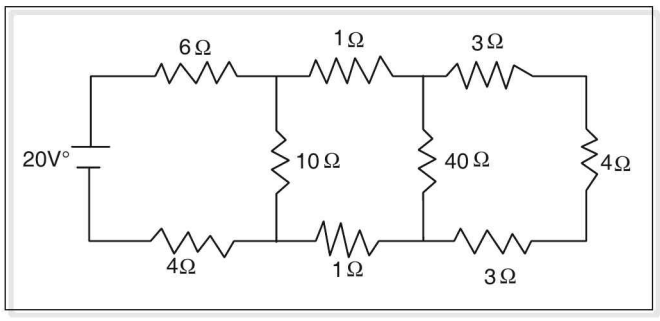


44. "A current carrying conductor is electrically charged." Explain with reason whether the statement is true or false.

45. A heater coil operates at 1200 W when connected to a 240 V supply. The resistivity and area of cross section of the coil are 10^{-5} Ω m and 0.4 mm^2 respectively. If the radius of each turn is 7 mm, find the number of turns present in the coil.

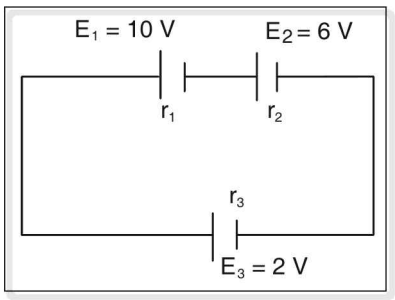
Concept Application Level—3

46. In the given circuit, find the power dissipated in the 6 Ω resistance.

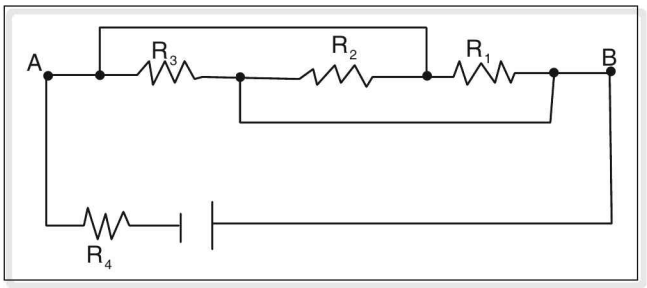


47. When electric heater is switched on it is noticed that the lamps in the room become dimmer for a short interval of time and then regain their original brightness. Explain

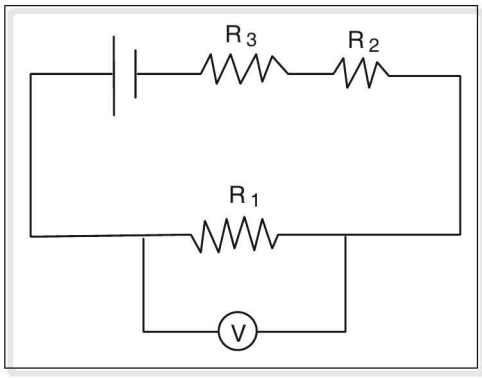
48. Three batteries with emf E_1, E_2, E_3 are connected as shown in the figure. If their terminal voltages are 4 V, 2 V, 6 V respectively, find the ratio of their internal resistances.



49. Find the current through the resistor R_4, if $R_1 = R_2 = R_3 = 20\ \Omega$, $R_4 = 40\ \Omega$. The emf of the battery is 4 V.



50. Three resistors R_1, R_2, R_3 of which $R_1 = 20\ \Omega$, $R_2 = 20\ \Omega$, $R_3 = 30\ \Omega$ connected to a battery of emf 24 V as shown in the figure. If the resistance of the voltmeter is 20 Ω, find the reading in the voltmeter.



Very short answer type questions

1. (i) The obstruction offered by a conductor to the flow of electric current is called its resistance.

(ii) The resistance of a conductor is said to be one ohm, if the P.D of one volt causes a current of 1A to pass through it.

2. $R = R_1 + R_2 + \ldots\ldots + R_3$ (series)

$\frac{1}{R} = \frac{1}{R_1} + \frac{1}{R} + \ldots\ldots + \frac{1}{R_n}$ (parallel)

3. Fuse is a thin wire which melts and breaks the electric circuit due to (i) overloading (ii) short circuit (iii) fluctuation in power supply.

4. Incandescence is a phenomenon in which an object emits light when heated to a high temperature.

5. $R \propto \ell \Rightarrow \frac{R_1}{R_2} = \frac{\ell_1}{\ell_2}$

$R \propto \frac{\ell}{a}$

$\Rightarrow \frac{R_1}{R_2} = \frac{a_2}{a_1}$

$R = \rho\frac{\ell}{a}$

Specific Resistance or Resistivity (ρ) can be defined as the resistance of a conducting material having unit area of cross section and unit length.

6. Study of electric charges and its effects.

7. (a) In series connection, total internal resistance (r) is equal to the sum of the internal resistances of the individual cells;

$r = r_1 + r_2 + r_3$

(b) In parallel connection, total internal resistance is equal to the sum of the reciprocals of individual internal resistances.

$\frac{1}{r_{eg}} = \frac{1}{r_1} + \frac{1}{r_2} + \frac{1}{r_3} + \ldots\ldots$

8. electric current.

9. Gives the reading of energy consumed.

10. (i) Charges at rest;

(ii) Charges in motion.

11. $J = \frac{W}{Q} = 4.18\ J\ cal^{-1}$

12. ohm.

13. (i) Reciprocal of resistance is known as electrical conductance.

(ii) Reciprocal of resistivity is called conductivity.

(iii) Super conductors are those which offers zero resistance to the passage of electricity at critical temperature.

14. (i) Those which conduct electricity easily. Example All metals.

(ii) Those which do not conduct electricity. Example wood

15. (i) The P.D. between two terminals of the cell, when current is drawn from it, i.e., the cell is in closed circuit, is called its terminal voltage (V).

(ii) The electrolyte offers certain resistance to the flow of current. This is called the internal resistance of cell (r).

(iii) The difference between EMF (E) of the cell and terminal voltage (V) of a cell is called lost voltage.

16. coulomb

17. $R_1 = \rho\frac{\ell_1}{a_1}$

$R_2 = \rho\frac{\ell_2}{a_2}$

$\ell_2 = 3\ell_1; \quad a_2 = 3a_1$

$\Rightarrow R_2 = \rho\ \frac{3\ell_1}{3a_1} = R_1$

18. It is primarily due to flow of electrons.

19. Watt hour is a unit in which the amount of electric energy consumed by a device is measured.

20. volt

21. (i) The resistance of conductors increases with an increase in temperature.
(ii) the resistance of semi conductors decreases with an increase in temperature.

22. The rate of flow of electric charges.

23. The energy spent or work done an moving on electric charge through conductor is called electric energy or electric work.

24. (i) Work done in bringing unit positive change from infinity to a point in an electric field is called potential (v).
(ii) The work done in moving a unit positive change from one point to another point against the direction of the field, is called as potential difference (P.D) between two points.
(iii) The P.D. between the terminals of the cell, when no current is drawn from the cell, i.e., it is open circuit is called EMF.

25. $E - V = Ir$

26. phase

27. The rate at which electric work is done is called electric power. In S.I., power = watt = $1\ J\ s^{-1}$.

28. (i) Positive charges flows from higher potential to lower potential.
(ii) If electrons moves from lower potential to higher potential, it constitutes electronic current.

29. (i) Ohm's law: The P.D. (V) applied between the ends of a conductor bears a constant ratio with the electric current (I) flowing through it, provided the physical state of the conductor remains the same.
(ii) Those which obeys Ohm's law are Ohmic conductor.
(iii) Those which do not obey Ohm's law are non Ohmic conductor.

30. The process of depositing a thin layer of desired metal over another metal by passing an electric current through some electrolyte is called electro plating.

Short answer type questions

31. EMF = terminal voltage + internal voltage drop.

32. To reduce the loss in power transmission.

33. Tree system
Phase wire to different rooms is taken from the main board, through a fuse on the main board. Neutral and earth wire is common to all the circuits.

34. $R = R_1 + R_2 + R_3, V = iR$
Ans: 5 A

35. (i) distance between the two electrodes.
(ii) area of cross section of the electrodes.
(iii) nature, concentration and temperature of the electrolyte.

36. $R = R_1 + R_2 + R_3, \frac{1}{R} = \frac{1}{R_1} + \frac{1}{R_2} + \frac{1}{R_3}$,
$V = iR$
Ans: $X = 10\ \Omega$

37. Earthing involves connecting metal body of electrical appliance to thick copper wire which is connected to a copper plate buried deep in the earth.

38. Energy = (power)(time)
Total cost = (Energy)(price)
Ans: Rs 172.8

39. (i) $r = r_1 + r_2 + r_3 + \ldots\ldots$ (series)

(ii) $\frac{1}{r} = \frac{1}{r_1} + \frac{1}{r_2} + \frac{1}{r_3} + \ldots\ldots$ (parallel)

40. $\frac{1}{R} = \frac{1}{R_1} + \frac{1}{R_2} + \frac{1}{R_3}$
Ans: $2\ \Omega$

41. If excessive current begins to flow the fuse wire gets heated up more quickly than the conducting wires and melts, breaking the circuit.

42. $R = \frac{\rho \ell}{a}$

Ans: 0.0016 Ω - cm.

43. Let "R" be external resistance. Then voltage drop across external resistance (R) is = IR
Voltage drop across the internal resistance (r) is = Ir.

44. (i) To increase and decrease the voltage.
(ii) By this loss in power transmission can be decreased.

45. $1\ kW\ h = 1 \times 10^3 \times \frac{15}{15} \times 3600\ s$
$= 36 \times 10^5 J$

Essay type questions

46. The same work done in different forms produces the same quantity of heat in all cases.

$W = JQ$

$J = \frac{W}{Q} = 4.18\ J\ cal^{-1}$

47. **1st Law** The mass of the substance deposited at the cathode is directly proportional to the quantity of electricity.

$m \propto I$

$m \propto It$

$m \propto st$

$m = Zit = ZQ$

$m_1:m_2:m_3 = z_1:z_2:z_3$

2nd Law

The same quantity of electricity passes through different electrolytes, the masses of ions liberated at the respective electrodes are proportional to the chemical equivalents.

$m_1:m_2:m_3 = E_1:E_2:E_3$

$E_1:E_2:E_3 = z_1:z_2:z_3$

48. $V = V_1 + V_2 + V_3; V = iR$

49. (i) Filament offers high resistance to electric current. Electric energy is converted into light and heat energy.

(ii) Discharge lamp is coated with fluorescent material from inside.
High voltages are applied to coils, heat them which results in emission of electrons. Collision of electrons with mercury vapours produces lights.

(iii) Glass tube having two electrodes fused to its. When high voltage is applied, gases inside the tube gets ionized.
Accelerated ions collide with neutral gas atoms that produce glow.

(iv) Produces high intensity light.
Carbon rods having pointed tips at one end and separated by small gap.
When 40 V to 60 V is applied across the carbon rods, glow is produced.

(v) Form of a primary cell.
Consists of zinc container (cathode) at the centre of which a carbon rod (anode) is fixed.
Dry paste of $NH_4Cl + ZnCl_2$ serves as an electrolyte.
Electrolyte is mixed with powdered coke.

$NH_4Cl \rightarrow NH_4^+ + Cl^-$

$2NH_4^+ + 2e^- \rightarrow 2NH_3 + H_2$

$H_2 + 2MnO_4 \rightarrow H_2O + MNO_3$

$Zn^{2+} + 2Cl^- \rightarrow ZnCl_2$

Here, chemical energy is converted in electrical energy.

50. (i) $E = E_1 + E_2 + E_3$ (series)

(ii) For identical cells that are connected in parallel, the total emf is same as the emf of individual cells.

KEY

Concept Application Level—1

True or false

1. False
2. False
3. False
4. True
5. True
6. True
7. True

Fill in the blanks

8. voltage, current
9. phase/live
10. electric power
11. electroplating
12. resistance
13. external resistance, internal resistance
14. emf

Match the following

15. A : f
    B : g
    C : h
    D : i
    E : j
    F : a
    G : b
    H : c
    I : d
    J : e

Multiple choice questions

16. Choice (1)
17. Choice (1)
18. Choice (3)
19. Choice (1)
20. Choice (1)
21. Choice (4)
22. Choice (3)
23. Choice (1)
24. Choice (1)
25. Choice (3)
26. Choice (2)
27. Choice (2)
28. Choice (4)
29. Choice (2)
30. Choice (3)

Concept Application Level—2,3

Key points

31. (i) The earth pin gets connected to the terminal first as compared to live or neutral pins.
    (ii) Thicker earth pin can never enter in live or neutral cylinder of the socket.

32. (i) A conducting wire, ideally, should not offer any resistance to the current. A conducting wire, sometimes carry a large current, which may melt it.
    A fuse wire should melt easily whenever excess current flows through it.
    Also, the heat energy required to melt the fuse wire will be produced when it offers high resistance to the current.
    A heating element is required to produce large amount of heat energy which is possible if it offers high resistance to the given current.
    Also, a heating element should not melt due to large amount of heat produced in it.
    (ii) wire A
    (iii) wire C
    (iv) wire D

33. The thickness of connecting wire carrying large current is more than the one carrying smaller current. An electrical appliance of high power voltage rating draws larger current.

34. (i) Use $R = \frac{V^2}{P}$ to find resistances R_1 and R_2 of both the bulbs.
    Determine the current flowing through bulbs by using $I = \frac{V}{R_1 + R_2}$.

Find the potential difference V_1 and V_2 across each bulb, by using $V_1 = IR_1$ and $V_2 = IR_2$. The bulb whose potential difference is greater than 220 V will fuse.
If the bulbs are connected in parallel across 440 V, the PD across them will be greater than the maximum value of 220 V.

(ii) 60 W bulb fuses when connected in series. Both bulbs fuse when connected in parallel.

35. (i) When cells are connected in series $E = E_1 + E_2$.

(ii) When cells are connected in parallel $E = \frac{E_1 r_2 + E_2 r_1}{r_1 + r_2}$

(iii) 10.8 V

36. (i) The maximum resistance of 16 Ω is obtained when the resistors are connected in series. The minimum value of 3 Ω is obtained when the resistors are connected in parallel.

(ii) 4 Ω and 12 Ω

37. (i) Heat generated $q = i^2Rt$

(ii) $m = Zit$

(iii) $m_1 : m_2 : m_3 = E_1 : E_2 : E_3$

(iv) 1.5 g

38. Determine the resistance of the two bulbs using the formula, $R = \frac{V^2}{P}$.
where R is the resistance of the filament, V is potential difference and P is the power.

39. (i) Terminal voltage $V = E - ir$

(ii) 1 Ω

40. Remember that nerves are the carriers of electric signals. What is their current carrying capacity?

41. A car starts due to current supplied by the storage battery. The ionic dissociation of the electrolyte present in the battery is more at higher temperatures than at lower temperatures.

42. (i) Definition of ohmic resistor.

(ii) conductor B, 0.5 Ω

43. (i) The equivalent resistance can be formed by successively simplifying the given circuit.
Find the equivalent resistance of the two 2 Ω resistors which are connected in series.
These together are connected in parallel to the 1 Ω resistor. This combination is in series with the other resistors.
Thus using the formulas for series and parallel connections, find the equivalent resistance in three steps.

(ii) 5.6 Ω

44. A body is charged when it loses or gains electrons. The number of electrons leaving from one end of the current carrying conductor is equal to the number of electrons entering through its other end.

45. (i) $R = \frac{V^2}{P}$

(ii) $R = \frac{\rho \ell}{A}$

(iii) Length of each turn = $2\pi r$

(iv) No. of turns = $\frac{L}{2\pi r}$

(v) 44

46. (i) Equivalent resistance $R = R_1 + R_2$, when connected in series.
$\frac{1}{R} = \frac{1}{R_1} + \frac{1}{R_2}$ when connected in parallel.

(ii) Power dissipated = V I

(iii) 9 W

47. Compare the resistances and current drawn by the heater and the bulb.

48. (i) $E_{net} = E_1 + E_2 - E_3$

(ii) $I = \frac{E_{net}}{r}$

(iii) $V_1 = E_1 - ir_1, V_2 = E_2 - ir_2, V_3 = E_3 + ir_3$

(iv) Find the ratio of r_1, r_2, r_3.

(v) 3 : 2 : 2

49. (i) R_1, R_2, R_3 in parallel are in series with R_4.

(ii) $I = \frac{V}{R_{eff}}$

(iii) 46.7 Ω, 0.08 A

50. (i) Resistor R_1 and resistance of voltmeter are in parallel.

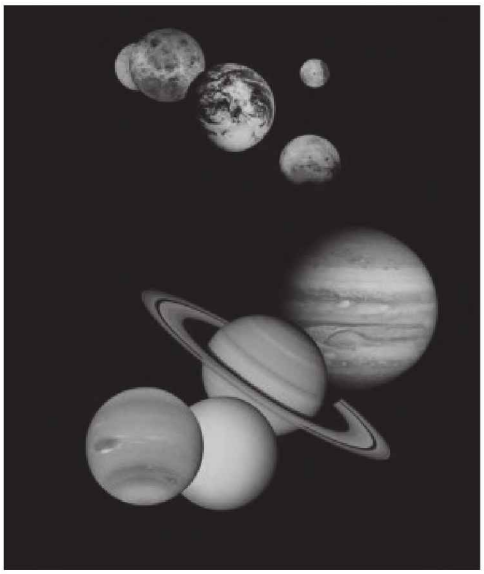
(ii) This effective resistance and R_2, R_3 are in series. Calculate the net effective resistance.

(iii) When resistors are in series $V_1 : V_2 : V_3 = R^1 : R_2 : R_3$

(iv) 4 V

8

Electromagnetism



INTRODUCTION

In 1820, Hans Christian Oersted discovered that current carrying conductor produces a magnetic field around it. Later Micheal Faraday showed that current can be induced in a conducting coil, when there is relative motion between the coil and a magnet. Thus electric current has magnetic effect and changing magnetic fields can induce electricity. Together, their study forms a very important branch of physics called electromagnetism. Large scale generation of electricity worldwide and most of the devices which work on it are based on the principles of electromagnetism.

In the previous chapter, you studied electricity. Before we begin the study of electromagnetism, let us understand some important aspects of magnetism.

Magnetism

Since ancient times it was known that ferrous oxide (Fe_3O_4) has the ability to attract substances like iron, nickel etc. Substances having this ability are called magnets. Fe_3O_4 is called natural magnet as it is naturally occurring while those substances in which this ability is artificially imparted are called artificial magnets.

The substances which are attracted towards a magnet and can become magnets are called magnetic substances. Example: iron, nickel, cobalt etc.

The substances which are not attracted towards a magnet and cannot be magnetized are called non-magnetic substances. Example: wood, paper, plastic etc.

Important properties of magnets

1. **Property of attraction:** Magnets attract small pieces of materials like iron, nickel, and cobalt. The property of a magnet to attract small pieces of iron seems to be concentrated in small regions at the ends of the magnet. These regions are called magnetic poles. The pole which points towards geographic north is called North pole of the magnet. The pole which points towards geographic south is called South pole of the magnet.
2. **Property of direction:** A freely suspended magnet always aligns itself in the North-South direction.
3. Like-poles of magnets repel each other and unlike-poles attract each other, just as like-charges repel and unlike-charges attract. Since a magnet can attract small pieces of iron and also the opposite pole of another magnet, property of attraction is not a sure test to find whether a given piece is a magnet or not. Repulsion is a sure test to confirm whether a given piece is a magnet or not.
4. **Property of induction:** A magnet can induce magnetism in substances like soft iron, cobalt, nickel etc.
5. Breaking a magnet successively into smaller pieces would still produce tiny magnets each with a north pole and south pole. The above phenomena is observed till we reach molecular stage.
6. Magnetic poles always exist in opposite pairs.

Magnetic field and field lines

The area surrounding magnet, in which its effect can be felt is called magnetic field.

Place a bar magnet on a cardboard. Randomly sprinkle iron filings on the cardboard and tap it lightly. The iron filling arrange themselves in a particular pattern.

Draw lines along these patterns. These lines form continuous, closed curves between the poles of the magnet. Place a magnetic needle near the north pole of bar magnet and mark the north pole end of needle. Now place the needle with its S-pole at this mark and mark the new position of its N-pole. It is found that the needle always traces a path formed by above mentioned curved lines. These lines are called magnetic lines of force. Thus a magnetic field can be represented by lines of force.

Magnetic lines of force are those lines in a magnetic field along which an imaginary free isolated north pole moves.

(i) Lines of force are continuous lines.

(ii) Outside a bar magnet the lines of force are directed away from North pole and towards the South pole.

(iii) Inside a bar magnet, the lines are directed from South pole and towards the North pole.

(iv) Lines of force do not intersect.

(v) The magnetic field is strong at places where the lines of force are crowded and weak where the lines of force are far apart.

(vi) The direction of magnetic field at a point is tangential to the line of force at that point.

Ewing's Molecular Theory of Magnetism

If we break a magnet into two parts each part becomes a magnet with a north pole at one end and a south pole at the other end. If we break it further, we observe the same behaviour. This will continue even if we break the magnet to its molecular level.

Each molecule of a magnetic substance is an independent magnet irrespective of whether the substance is magnetized or not, these tiny magnets are called molecular magnets.

In an unmagnetized state (figure 8.1(a), the molecular magnets are arranged in random direction. Therefore the net magnetic effect is zero.

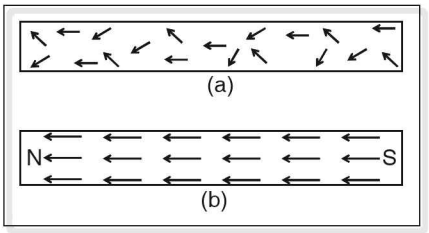


Figure 8.1

In the magnetized state (figure 8.1(b), these molecular magnets are arranged in an order. All the south poles are aligned in one direction and the north poles in the other direction. Therefore a strong magnetic field is created.

In general perfect alignment of all the magnets in the entire magnet is difficult to obtain. When this condition is obtained it is called the point of saturation and magnetizing the magnet further cannot increase the strength of the magnet. These postulates are known as Ewing's molecular theory named after the scientist who proposed them.

However, Ewing's molecular theory could not explain why the individual molecules of a magnetic substance like iron behave as tiny magnets. Also it failed to explain, why molecules of non-magnetic substances, like brass, do not behave like magnets. The theory could not explain the fact that substances like bismuth, copper are repelled by strong magnets.

Inverse Square Law of Magnetism

The ability of the pole of a magnet to attract or repel another magnetic pole is called its **pole strength**. It is denoted by 'm'. Its SI unit is ampere-metre.

Consider two hypothetical isolated magnetic poles of pole strengths m_1 and m_2 separated by a distance 'r'.

According to inverse square law of magnetism, given by Coulomb the force of attraction or repulsion between two magnetic poles is directly proportional to the product of their pole strengths and inversely proportional to the square of the distance between them and acts along the line joining the poles.

Mathematically,

$F \propto m_1 \times m_2$

and $F \propto \frac{1}{r^2}$,

or, $F = K \frac{m_1 m_2}{r^2}$

where K is the constant of proportionality. Its value depends on the magnetic properties of the medium surrounding the magnetic poles and the system of units used.

If we choose SI units, and if vacuum or air is the medium surrounding the poles then the value of K is found to be $\frac{\mu_0}{4\pi}$.

$$\therefore F = \frac{\mu_0}{4\pi} \frac{m_1 m_2}{r^2} \text{ ---------- (1)}$$

Where μ_0 is permeability of vaccum or air.

The value of μ_0 is $4\pi \times 10^{-7}$ henry metre^{-1}.

If any medium other than vacuum or air surround the poles then

$$F_1 = \frac{\mu}{4\pi} \times \frac{m_1 m_2}{r^2} \text{ ---------- (2)}$$

where μ is absolute magnetic permeability of the medium.

Some important terms

1. Magnetic permeability

Magnetic permeability (μ) of a medium is defined as its ability to allow the magnetic lines of force to pass through it or to allow itself to be influenced by magnetic field.

On dividing equation (2) by equation (1), we have

$$\frac{F_1}{F} = \frac{\mu}{\mu_o} = \mu_r \Rightarrow \mu = \mu_0 \mu_r$$

where μ_r is called **relative permeability** and it is defined as the ratio of magnetic force in a medium to the magnetic force in vacuum.

μ_r being a ratio has no units. For air or vacuum, $\mu_r = 1$.

2. Unit pole strength

The unit magnetic pole is defined as that pole which repels an identical pole at 1 metre distance with a force of 10^{-7} N.

By taking $\mu_0 = 4\pi \times 10^{-7}$ henry metre^{-1}, r = 1 m and $m_1 = m_2 = 1$(A–m), we have

$$F = \frac{\mu_0}{4\pi} \times \frac{m_1 m_2}{r^2}$$

$$= \frac{\mu_0}{4\pi} \times \frac{1 \times 1}{1^2} = \frac{4\pi \times 10^{-7}}{4\pi} = 10^{-7} \text{ N.}$$

3. Magnetic flux density "B"

The space around the magnet where the influence of the magnet can be felt is called **magnetic field**. Let us imagine this field as constituted by magnetic lines of force. The total number of lines passing through the given area is called **magnetic flux "φ"**. The unit of magnetic flux is weber (Wb). If a magnetic substance exits in this field then the lines of force pass through it. These lines of force are called magnetic induction lines as these lines induce magnetism in the magnetic substance.

Magnetic flux per unit area is called magnetic flux density or magnetic induction.

i.e., $B = \frac{\phi}{A}$

Magnetic induction can also be defined as the force experienced by a unit north pole kept in a magnetic field.

We know that the force experienced by two poles m_1 and m_2 is given by

$$F = \frac{\mu_o}{4\pi} \frac{m_1 m_2}{d^2}$$

Let m_1 be the pole strength of unit N pole.

$\therefore m_1 = 1\ A - m$

Let $m_2 = m$

$$\therefore F = \frac{\mu_0}{4\pi} \frac{m}{r^2}$$

By definition magnetic induction at a point in a magnetic field is the force experienced by a unit north pole placed at that point.

$$\therefore B = \frac{\mu_o}{4\pi} \frac{m}{r^2}$$

The SI unit of flux density is Wb m^{-2} or tesla (T). The CGS unit of flux density is gauss (G).

$1\ T = 10^4\ G$

Its other unit is $N\ A^{-1}–m^{-1}$

The force experienced by a north pole of pole strength 'm', kept in a magnetic field of induction 'B', is given by $F = mB$ or $B = \frac{F}{m}$

4. Intensity of magnetic field (H)

The intensity of magnetic field refers to the strength of a magnetic field at a point in C.G.S. units,

We know that

$$F = K\frac{m_1 m_2}{r^2}$$

where $K = \frac{1}{\mu}$ and μ is the permeability of free space which is equal to 1 in C.G.S system. In the above equation, if m_1 is unit north pole and $m_2 = m$.

$$F = \frac{m}{r^2}$$

This force is called intensity of magnetic field (H)

$$\therefore H = \frac{m}{r^2}$$

Thus intensity of magnetic field can be defined as the force acting on unit north pole independent of the medium.

In S.I. system,

$$H = \frac{m}{4\pi r^2}$$

Since $K = \frac{1}{4\pi}$ in S.I system, the S.I. units of H are ampere metre^{-1} (A m^{-1})

In C.G.S. system, its unit is oersted.

Both H and B represent magnetic field at a point, the difference being that H is independent of the medium.

H and B are related as follows:

$$B = \mu_o H$$

5. Magnetic moment "M"

The magnetic moment of a bar magnet is given by the product of its length and pole strength.

Let us take the length of the magnet as "2*l*" and magnetic pole strength of the magnet as "m". Then the magnetic moment of the magnet is given by $M = (2l) \times m$.

Unit of magnetic moment is ampere – metre^2(A m^2)

Magnetic field induction of a bar magnet on the axial line of the magnet

The line joining the north pole and south pole of the magnet is called the axial line.

The magnetic induction 'B' on the axial line is given by

$$B = \frac{\mu_o}{4\pi} \frac{2M}{d^3} \text{ N A}^{-1}\text{–m}^{-1}$$

Where μ_o is the permeability of free space, M is the magnetic moment of the magnet and 'd' is the distance of the point 'P' from the centre of the magnet. The direction of B is along the axis parallel to $\overrightarrow{SN}$ vector.

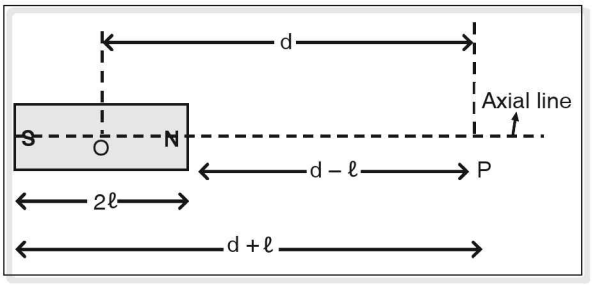


Figure 8.2

Magnetic field induction of a magnet on the equatorial line of the magnet

A line perpendicular to the axis and passing through the midpoint of the magnet is called the equatorial line.

The magnetic induction "B" on the equatorial line is given by $B = \frac{\mu_o}{4\pi} \frac{M}{d^3}$ N A^{-1}–m^{-1}, where μ_o is the permeability of free space, M is the magnetic moment of the given magnet and 'd' is the distance of the point 'P' on the equatorial line from the magnet.

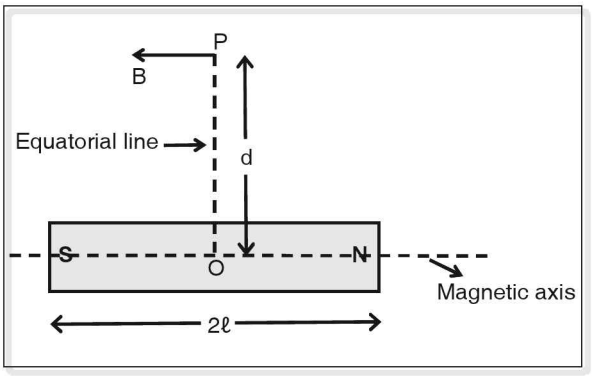


Figure 8.3

The direction of B is parallel to the axis of the magnet and is in the $\overrightarrow{NS}$ direction.

Calculation of magnetic moment M of a bar magnet

At the neutral points, the magnetic field intensity due to a bar magnet equals horizontal component of earth's magnetic field (B_H). Knowing B_H at a given place (which is 0.39×10^{-4} T in Andhra Pradesh) and distance 'd' of neutral point we can determine M.

Case I:

When neutral points lie on equatorial line,

$$B = \frac{\mu_o}{4\pi} \times \frac{M}{d^3}$$

Substituting $B_H = 0.39 \times 10^{-4}$ T, $\frac{\mu_o}{4\pi} = 10^{-7}$, and simplifying,

we get $M = 390 d^3$ A m^2.

Knowing the length of the magnet, we can find m.

The value of m can be calculated as

$$m = \frac{M}{2l} = \frac{390\ d^3}{2l} \text{ A m}$$

Case II:

When neutral points are on axial line,

$$B = B_H = \frac{\mu_0}{4\pi}\frac{2M}{d^3}$$

Substituting the values of B_H and $\mu_0/4\pi$, we get $\frac{2M}{d^3} = 390$, and we get $M = 195d^3$ A m^2

Knowing the value of d, the value of m can be calculated.

If we know the length of the bar magnet, we can find its pole strength m as

$$m = \frac{M}{2l} = \frac{195\,d^3}{2l} \text{ A m}$$

Origin of Magnetism–Modern Electron Theory

One of the major drawbacks of Ewing's theory is that it fails to explain why molecules of magnetic substances like iron behave like tiny magnets while those of non-magnetic substances like wood do not. The modern electron theory offers an explanation to this. A molecule consists of two or more atoms. Each atom consists of electrons which revolve around the nucleus in orbits and rotate or spin about an axis. Since electrons carry negative charge, the orbital and spin motion of electrons give rise to electric current which produces a magnetic field.

For substances like iron, the net magnetic moment due to a molecule is not zero hence they are magnetic in nature. The net magnetic moment due to a molecule is zero in wood hence it is non-magnetic.

Magnetic properties

(i) **Intensity of magnetisation "I":** It is the magnetic moment acquired per unit volume when a substance is placed in a magnetic field. Its unit is A m^{-1}.

(ii) **Magnetic susceptibility "χ":** It is the measure of capability of a medium to get magnetized. Susceptibility of a substance is defined as the ratio of intensity of magnetization "I" to the intensity of applied magnetic field "H". Soft iron has high susceptibility.

$$\chi = \frac{I}{H}$$

Magnetic susceptibility is a constant for a given substance and has no unit.

(iii) **Retentivity:** It is a property of magnetic material by virtue of which it retains the magnetic property even after removal of the applied magnetic field. Steel has high retentivity. Due to this reason steel is used for making permanent magnets.

Classification of magnetic materials

Based on the magnetic properties all materials can be classified as:

(i) Ferromagnetic substances

(ii) Paramagnetic substances
(iii) Diamagnetic substances

Diamagnetic substances

The resultant magnetic moment of an atom in a substance depends on the orientation of electron orbits and the axes about which electrons spin.

The substances in which the net magnetic moment of any atom in them is zero are called diamagnetic substances.

1. A diamagnetic substance is feebly repelled by a powerful magnet.
2. When suspended in a non uniform magnetic field it moves at right angles to the direction of the field from stronger region to weaker region of the field.
3. Its relative permeability (μ_r) is approximately equal to or less than 1 ($\mu_r \leq 1$).
4. Magnetic susceptibility is negative.
5. Example: air, water, copper, bismuth

Paramagnetic substances

Those substances in which the orientation of electron orbits and axes of their spins are such that the net magnetic moment of any atom is not zero, are called paramagnetic substances. Those atoms which have unpaired electrons acquire resultant magnetic moment.

1. Paramagnetic substances are feebly attracted by a strong magnet.
2. When placed in a non-uniform external magnetic field it aligns itself with the direction of the field and moves from weaker to stronger region of the field.
3. Its relative permeability is slightly greater than 1 ($\mu r > 1$).
4. Its magnetic susceptibility (χ) is small and positive.
5. Example: aluminium, platinum, chromium

Ferromagnetic substances

Like paramagnetic substances, the atoms of these elements have resultant magnetic moment. But unlike paramagnetic substances they achieve high degree of magnetic alignment due to a special effect called 'exchange coupling'.

The elements Fe, Co, Ni, Gd and Dy are examples of ferromagnetic substances.

Ferromagnetic substances are those in which the resultant magnetic moments of individual atoms align themselves parallel to each other due to a special effect which leads to spontaneous magnetisation.

(i) They are strongly attracted by magnets.
(ii) They align in the direction of the applied field.
(iii) Their relative permeability is very high ($\mu_r >> 1$).
(iv) Magnetic susceptibility (χ) is very high and positive. Therefore they can be easily magnetised and made into magnets.

Domain Theory

Domain theory explains the properties of ferromagnetic substances. According to domain theory, a large number of small local regions are formed in ferromagnetic substances, called **domains**. Within each domain, the magnetic dipoles of all the atoms align themselves parallel to each other giving rise to magnetism. The direction of magnetisation in each domain is different from that in the other domains as indicated by arrows in the figure shown below.

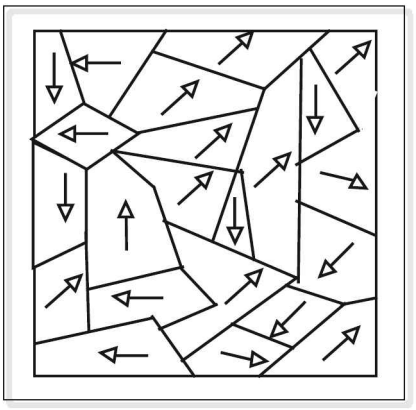


Figure 8.4

In the diagram, boundary lines indicate different domains and arrows indicate direction of magnetisation of respective domains. In the absence of applied magnetic field, the arrow of each domain point along different directions. As a result, the net magnetisation of the specimen as a whole is zero. This explains why an iron rod is not a magnet by itself.

When a ferromagnetic specimen is magnetised the domains expand and all arrows align parallel to each other in one direction. In an ideal magnet, there is a single domain with all the arrows pointing in one direction. This way the specimen becomes a magnet. Ferromagnetic substances exhibit retentivity because on demagnetising them, the original domains are not formed immediately, so that a certain amount of magnetization is still retained.

6. Terrestrial magnetism

Our earth behaves like a magnet. This is proved by the fact that a freely suspended bar magnet or a compass needle always come to rest in north–south direction. It is as if a huge bar magnet is buried at the centre of the earth with its south pole at the geographic north and vice versa. The axis of the earth's magnet makes an angle of about 17° with the geographic axis.

It is supposed that electric currents in the molten outer core rich in iron give rise to the earth's magnetism.

The study of the earth's magnetism and its various elements constitutes terrestrial magnetism.

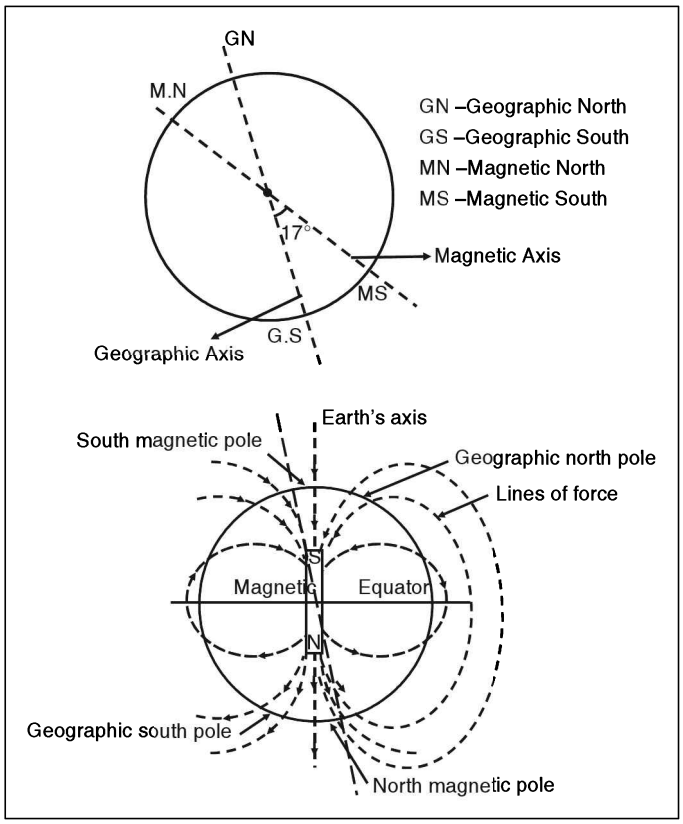


Figure 8.5

Elements of earth's magnetism

1. **Angle of dip or inclination**

   A magnetic needle suspended at its centre of gravity and free to rotate in vertical and horizontal planes comes to the rest in the direction of the earth's magnetic field. The angle made by the axis of magnetic needle with the horizontal is called angle of dip or inclination.

   The angle of dip is maximum at the magnetic poles and its value is 90°. At magnetic equator of the earth, its value is zero.

2. **Angle of declination**

   The vertical plane passing through a place and containing geographic north and south poles is called geographic meridian.

   The vertical plane passing through a place and containing the earth's magnetic north and south pole is called magnetic meridian.

   The angle between geographic and magnetic meridian at a place is called angle of declination.

Mapping–neutral points

The magnetic field due to a bar magnet consists of number of lines of force which emerge from its north pole and enter the south pole. At the same time the effect of the horizontal component (B_H) of the earth's magnetic field is also felt at a given place. The magnetic field induction (B) due to the bar magnet is large near it and decreases with the distance, whereas B_H remains constant in magnitude and direction at a given place. Thus within the magnetic field of a bar magnet, the lines of force represent the net effect of B and B_H. There are some points within the magnetic field where B and B_H are equal in magnitude and opposite in direction and cancel each other. These points where the resultant magnetic field is zero are called as neutral points. The location of neutral points also depend on the orientation of the bar magnet.

1. **North pole of the bar magnet facing the geographical north of the earth**

   At the geographical north pole lies the magnetic south pole of the earth and vice versa. The earth's magnetic lines of force are thus directed from geographical south to north. Hence the horizontal component (B_H) of the earth's magnetic field neutralizes the magnetic field induction (B) of the bar magnet at two diametrically opposite points lying on the equatorial line of the bar magnet.

   These points are the neutral points when N-pole of the bar magnet points towards geographic north. If a compass needle is placed at any of these points, it points in any random direction.

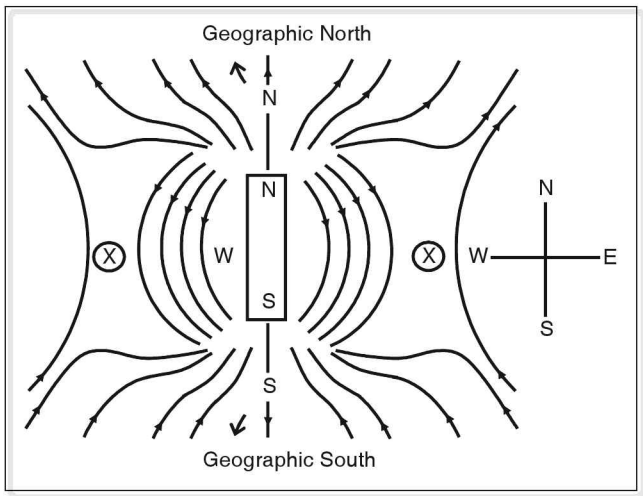


Figure 8.6

2. South pole of the bar magnet facing geographical north pole of the earth

In this orientation, the south pole of the bar magnet faces the magnetic south pole of the earth.

Along the axial line of the bar-magnet, the lines of force due to the magnetic fields of the earth and bar magnet are oppositely directed. Hence at two particular points along the axial line, one nearer to the south pole and the other closer to the north pole of the bar magnets, the B_H nullifies the B. These two points represent the neutral points.

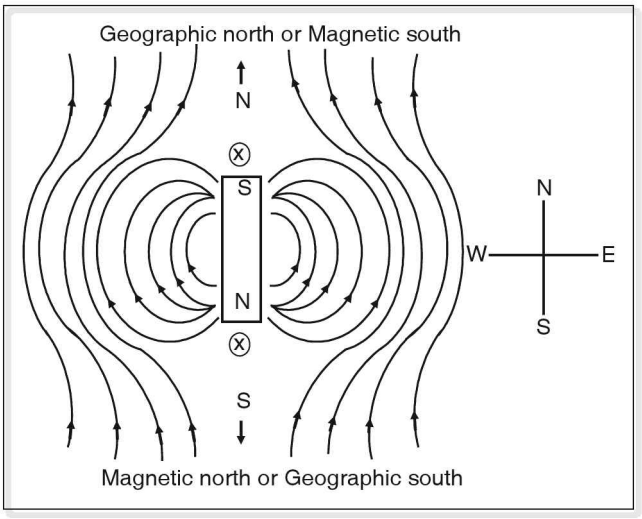


Figure 8.7

Oersted experiment

Oersted was the first to put forth the direct relation between electricity and magnetism. He conducted several experiments to determine the magnetic effect of a current carrying wire. The following describes the Oersted experiment conducted to establish that a current carrying wire acts as a magnet.

A long straight wire is connected to an external battery and an electric current is passed through it. When a magnetic needle is placed below the wire such that the wire is parallel to the axis of the magnetic needle and the current flows in the south to north direction, a deflection in the needle is observed. It is observed that the north pole of the needle is deflected westwards and as the magnitude of current is increased, the deflection increases till the north pole of the needle turns towards exact west. It is also observed that if instead of placing the magnetic needle below the wire, if it was placed above the wire, the north pole of the magnetic needle is deflected eastwards. From this experiment Oersted concluded the following facts.

1. Any current carrying wire produces a magnetic field around it, as it can deflect a magnetic needle placed near it.
2. The intensity of the magnetic field is proportional to the magnitude of the current passing through the wire.
3. The magnetic field setup acts at right angles to the direction of the flow of current.
4. The direction of the magnetic field depends upon the direction of the flow of current.

The direction of the magnetic field produced due to a current carrying wire may be determined using any one of the following rules.

Ampere's swimming rule

Imagine a man swimming along the conductor in the direction of current, facing a magnetic compass kept near and below the conductor. Then the north pole of the needle will be deflected towards his left hand.

Ampere's right hand thumb rule

Imagine you are holding the current carrying wire with your right hand, with the thumb pointing the direction of current. Then the direction of fingers encircling the wire show the direction of magnetic field.

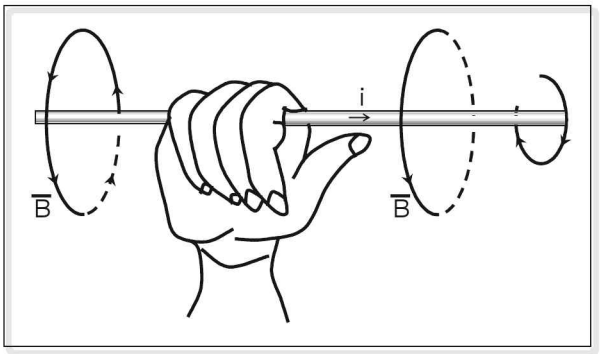


Figure 8.8

Maxwell's cork screw rule

Imagine a right handed cork screw held by your hand. Rotate the screw's head such that its tip advances in the direction of the current.Then the direction in which the head rotates gives the direction of the magnetic field.

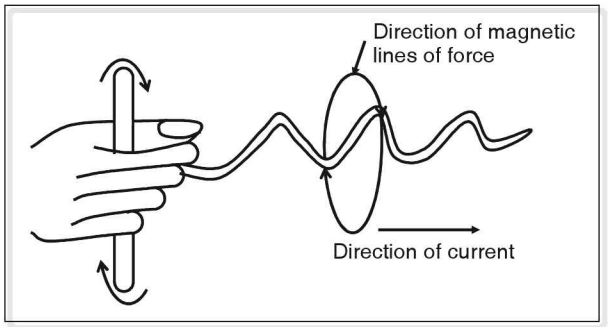


Figure 8.9

Magnetic field due to current

I. Magnetic field due to a long straight current carrying wire

The following experiment helps us study the magnetic field around a long straight current carrying wire.

Take a cardboard and fix a white paper over it. Make a small hole at its centre. Clamp this cardboard, to a stand in a horizontal position. Sprinkle some iron filings over it. Pass an insulated copper wire through the small hole at the centre of the cardboard and perpendicular to it.

A battery, a switch and a variable resistance are connected in series to the wire and when the circuit is closed, a current (i) flows through the wire. The iron filings on the cardboard form concentric circles around the wire showing that a current carrying conductor sets up a magnetic field around the conductor. The direction of the magnetic field at any point is in the plane of the paper and is always tangential to the circle passing through that point.

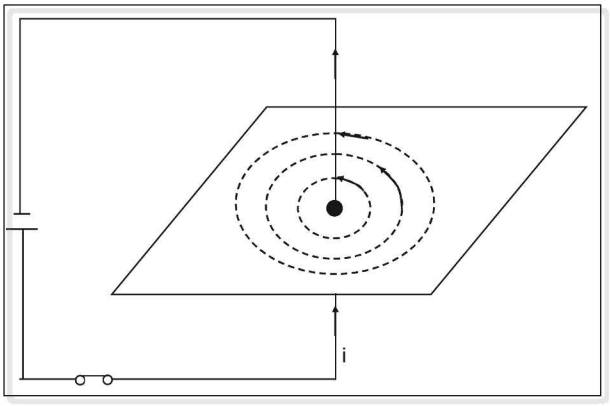


Figure 8.10

If the magnitude of the current in the wire is increased, then the number of concentric circles increases (i.e., the number of magnetic lines of forces increases).	The magnetic field induction $(\vec{B})$ is directly proportional to the current flowing through the wire. $\Rightarrow B \propto i$

(Continued on following page)

If a magnetic needle is brought near the conductor, the deflection in the needle is more and when it is moved away, the deflection decreases.	As the distance from the wire increases, magnetic field $(\vec{B})$ decreases. The magnetic induction (B) at a point is inversely proportional to the distance of the point from the wire. Hence, $B \propto \frac{1}{r}$

From these conclusions, we have,

$B \propto i$ and $B \propto \frac{1}{r}$

$\therefore B \propto \frac{i}{r} \Rightarrow B = \frac{Ki}{r}$

K is the constant of proportionality and may experimentally be verified to be equal to $\frac{\mu_0}{2\pi}$

(in S.I system), where μ_0 is called the permeability constant of free space or vacuum, whose value is given as

$\mu_0 = 4\pi \times 10^{-7}$ T m A $\because 1.26 \times 10^{-6}$ T m A^{-1}

Hence, $B = \frac{\mu_0 i}{2\pi r}$

☛ *Example*

Numerical

The figure shows two long parallel wires carrying currents i_1 and i_2 in opposite directions. What is the magnitude and direction of the resultant magnetic field at the point P?

The magnetic field B_1 at P due to i_1 is given as,

$$B_1 = \frac{\mu_0 i_1}{2\pi r_1} = \frac{\mu_0}{2\pi} \times \frac{15 \text{ A}}{0.05 \text{ m}} = \frac{150\mu_0}{\pi} \text{ T}$$

The magnetic field B_2 at P due to i_2 is given as

$$B_2 = \frac{\mu_0 i_2}{2\pi r_2} = \frac{\mu_0}{2\pi} \times \frac{30 \text{ A}}{0.1 \text{ m}} = \frac{150\mu_0}{\pi} \text{ T}$$

By the Right-hand-thumb rule, we find that both B_1 and B_2 are in the same direction at P. Thus, the resultant magnetic field at P is

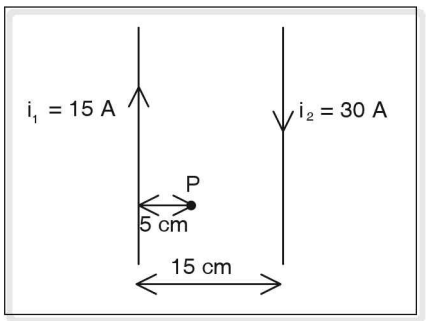


Figure 8.11

$$B = B_1 + B_2 = \frac{150\mu_0}{\pi} + \frac{150\mu_0}{\pi}$$

$$= \frac{300 \times 4\pi \times 10^{-7}}{\pi} = 1200 \times 10^{-7} \text{T}$$

$$= 1.2 \times 10^{-4} \text{T}$$

The direction of B is perpendicular to the plane of the conductors.

II. Magnetic field due to a circular current carrying wire

Let us now study the magnetic field due to a circular current carrying conductor. Take a copper wire, bend it in the form of a circular coil. Connect this coil to a battery with a rheostat and a switch in series. Pass this copper wire over two holes of a cardboard covered with a white paper. Sprinkle iron filings on the cardboard.

It is observed that iron filings arrange themselves in concentric circles as shown in the figure below.

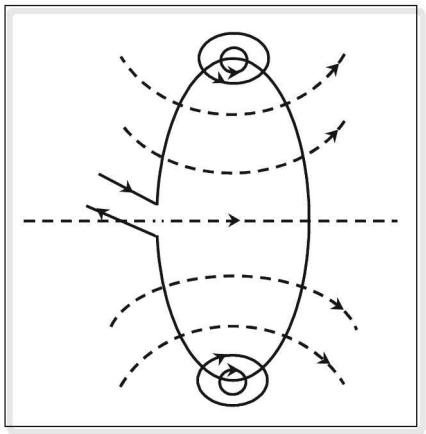


Figure 8.12

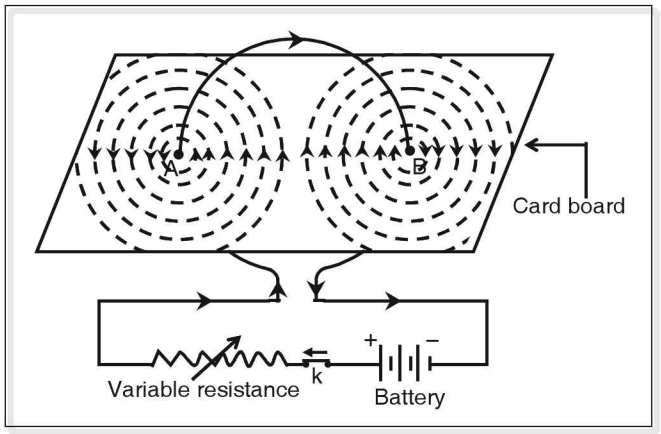


Figure 8.13

An increase in the magnitude of the current in the coil increases the number of magnetic lines of force.	Magnetic field induction is directly proportional to the current flowing in the wire.
It is observed that a magnetic compass shows more deflection if the circular coil is bent into a circle of less radius compared to a coil bent into a circle of larger radius.	As the radius of circular coil increases the magnetic induction decreases.

Near the conductor, the field is circular and the direction of the field is given by right hand thumb rule. As we move away from the wire towards the centre of the coil, the radius curvature of magnetic lines increases, i.e., they become straighter and almost parallel to each other. At the centre of the coil, the magnetic line of force is perpendicular to the plane of the coil.

The magnitude of the magnetic field at the centre of the circular current carrying wire is found to be:

1. directly proportional to the strength of the current passing through the wire and
2. inversely proportional to the radius of the circular coil.

∴ If B is the magnetic field at the centre, then

$B \propto i$, and

$$B \propto \frac{1}{r}$$

$\therefore B = K\frac{i}{r}$, where K is the proportionality constant and is equal to $\frac{\mu_0}{2}$ (in S.I. system)

$$\therefore B = \frac{\mu_0}{2}$$

μ_0 as defined earlier is the permeability constant of vacuum and is equal to $4\pi \times 10^{-7}$ T m A^{-1}.

If there are 'n' turns in the circular coil, then each turn will contribute to the magnetic field at the centre. So the total magnetic field at the centre due to 'n' turns will be equal to $\frac{n\mu_0 i}{2r}$

Electromagnetism

7. Properties of magnetic lines of force around a circular coil

1. At the points where current enters or leaves coil is like a straight wire. Hence the lines of force around these points are almost circular.
2. The lines of force inside coil are in same direction.
3. Magnetic field near the centre of coil is uniform as the lines of force are parallel to each other.
4. At the centre, the plane of magnetic field is perpendicular to the coil. The magnetic field strength increases with the increase in current.
5. Looking at a face of coil, if current is flowing in clockwise direction, the face develops south polarity. If it is in anti clockwise direction, the face develops north polarity.

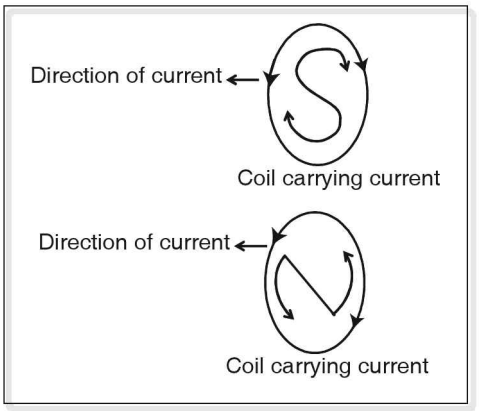


Figure 8.14

☛ *Example*

Find the magnetic induction field B at a distance of 10 cm from a long straight conductor carrying a current of 10 A.

Solution

The magnetic induction field near a straight conductor is given by

$$B = \frac{\mu_0 i}{2\pi r} \quad \text{- - - - - (1)}$$

In the given problem,

R = 10 cm = 0.1 m, i = 10 A, $\mu_0 = 4\pi \times 10^{-7}$ H m^{-1}

Substituting above values in equation (1),

$$B = \frac{4\pi \times 10^{-7} \times 10}{2\pi \times 0.1}$$

$$= 2 \times 10^{-7} \times 10 \times 10$$

$$= 20 \times 10^{-6}$$

$$= 20\ \mu T.$$

☛ *Example*

Find the ratio of the magnetic fields, at distances 5 cm and 50 cm from a long straight current carrying conductor.

Solution

Let the fields be B_1 and B_2 at distances 5 cm and 50 cm, respectively. As current remains same, we can write

$$B \propto \frac{1}{r}$$

$$\Rightarrow \frac{B_1}{B_2} = \frac{r_2}{r_1} \quad ----- (1)$$

Substituting the value of r_1 and r_2 in equation (1), we get

$$\frac{B_1}{B_2} = \frac{50}{5} = 10$$

$\therefore$ The ratio of fields is 10 : 1.

☛ *Example*

A circular coil of radius π cm and 100 turns is carrying a current of 10 A. Find the magnitude of the magnetic induction field at the centre of the coil.

Solution

The induction field at the centre of the coil is given by

$$B = \frac{n\mu_0 i}{2r} \quad ----- (1)$$

In the problem, the values given are, $r = \pi$ cm $= \pi \times 10^{-2}$ m, $n = 100$, $i = 10$ A

$B = ?$

Substituting the above values in equation (1), we get

$$B = \frac{100 \times 4\pi \times 10^{-7} \times 10}{2 \times \pi \times 10^{-2}} = 2 \times 10^{-2}\ T = 0.02\ T$$

III. Magnetic field due to a solenoid carrying current

An insulated copper coil wound around some cylindrical cardboard or some other core such that its length is greater than its diameter and behaves like a magnet when electric current flows through it, is called a solenoid or a helix.

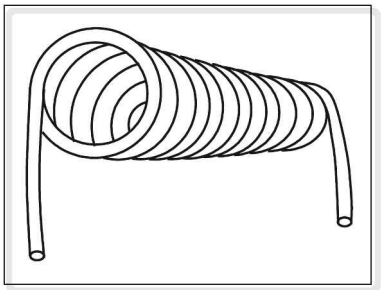


Figure 8.15 A Helix of Solenoid

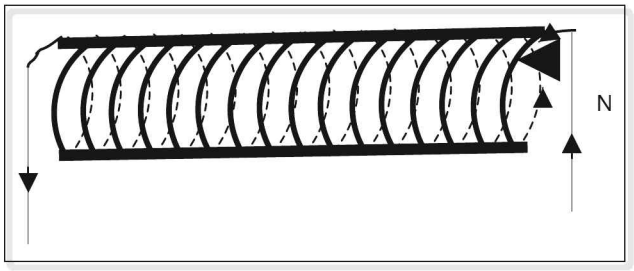


Figure 8.16

When an electric current flows in the solenoid then each turn of the coil behaves like an independent magnet. All these magnets are arranged in order and therefore the total magnetic strength of the solenoid depends upon the number of turns. Thus the solenoid acts like a bar magnet.

The end in which current flows in an anti-clockwise direction becomes North Pole and the other end where the current flows in clockwise direction, becomes south pole.

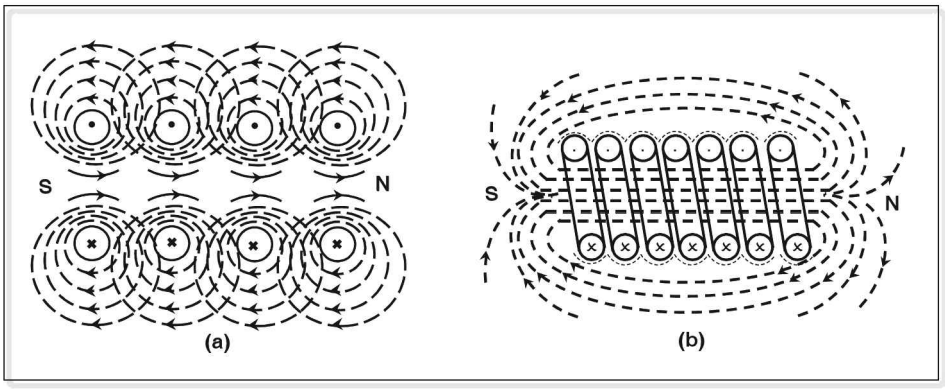


Figure 8.17

Properties of a solenoid

1. The magnetic field due to solenoid is directly proportional to the number of turns per unit length of the solenoid.
2. The magnetic field due to the solenoid is directly proportional to the current passing through the solenoid.
3. The field depends upon the nature of material on which the coil is wound.
4. If an iron core is kept inside the solenoid, the field increases.
5. Laminated soft iron core increases the intensity of the magnetic field inside a solenoid.
6. The magnetic lines of force within the solenoid are almost parallel to each other and axis of solenoid.

Force on a current carrying conductor kept in a magnetic field

It is observed that when a current carrying conductor is kept in a magnetic field perpendicular to the direction of current, it experiences a force.

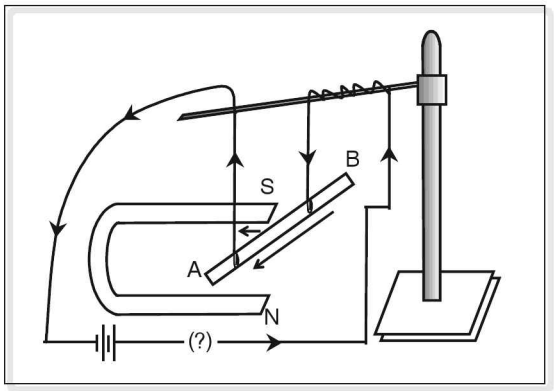


Figure 8.18 A current carrying rod, AB experiences a force perpendicular to the length and the magnetic field

Explanation

The lines of force representing the magnetic field between the poles of a horseshoe magnet would be straight and parallel to each other as shown in figure 1(a) below.

The magnetic field lines around a current carrying conductor would be as shown in figure 1(b). When this conductor is placed in between the poles of the permanent magnet, the two fields would superimpose and there would be a tendency for the lines to crowd on the right side. These lines acting like stretched elastic strings would tend to straighten themselves, and thus push the conductor towards the left. If the current in the conductor is reversed, the effect would be opposite.

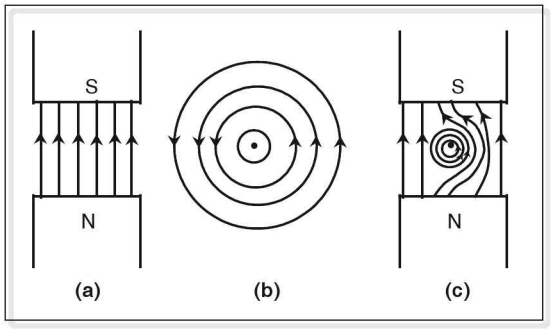


Figure 8.19

The force acting on the current carrying conductor depends upon the following factors:

1. The force on the conductor is directly proportional to the current in the wire.
   $F \propto i$ .....(i)
2. The force on the conductor is directly proportional to the strength of the magnetic field.
   $F \propto B$. .....(ii)
   3. The force on the conductor is proportional to the length of the conductor.
   $F \propto \ell$.....(iii)
   From (i), (ii) and (iii), we have $F \propto Bi\ell$
   $\Rightarrow F = kBi\ell$, where 'k' is the constant of proportionality.
4. Let the force on a conductor of unit length, carrying 1 ampere current and placed in magnetic field of induction 1 T be 1, then
   $1\ N = K \times \frac{1\ N}{A\ m} \times 1\ A \times 1\ m$
   $\therefore K = 1$
   $\therefore F = Bil$

The equation is applicable only when the current carrying conductor is placed at a right angle to the direction of the magnetic field B.

To find the direction of the force acting on the current carrying conductor **Fleming's Left Hand Rule** is used which is stated as follows:

Stretch the forefinger, middle finger and thumb of your left hand in three mutually perpendicular directions, then if the forefinger points the direction of magnetic field (B) and the middle finger points the direction of current (i), then the thumb will point the direction of motion of the conductor(M).

If the magnetic field is not perpendicular to the current carrying wire, then the magnitude of force is given by $F = Bi\ell\sin\theta$ where θ is the angle which the current carrying wire makes with the direction of the magnetic field.

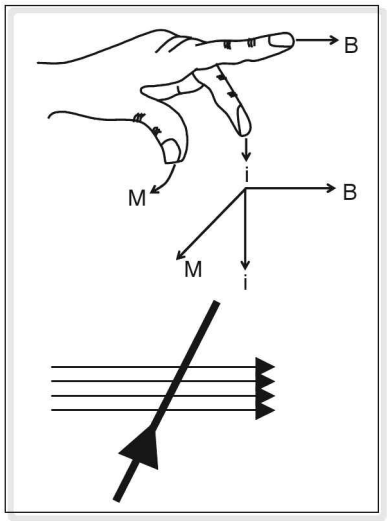


Figure 8.20

Moving coil galvanometer (suspension type)

Principle

A current carrying coil kept in a magnetic field experiences a couple which is proportional to the current. It is used to detect and final the direction of the flow of current.Voltmeter and ammeter are modified forms of a galvanometer.

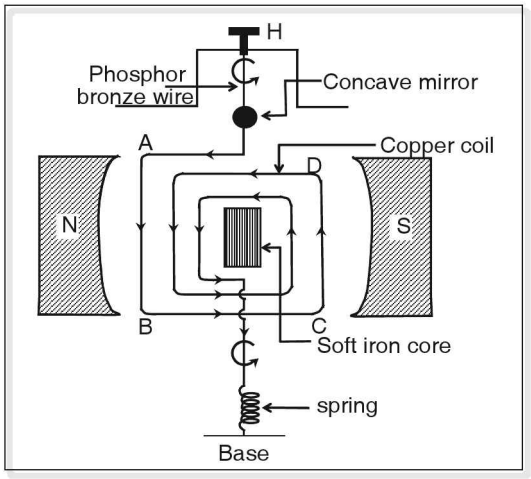


Figure 8.21

Construction

A moving coil galvanometer consists of a rectangular coil of thin, insulated copper wire suspended between the poles of a strong horseshoe magnet NS by means of a strip of phosphor bronze. It is attached to a torsion head 'H'. There is a small mirror attached to the phosphor bronze strip.This is to observe the deflection produced by the coil, when a current is passed through it. Current enters through the torsion head 'H' and leaves through the spring at the phosphor bronze bottom.

The magnet is made concave in shape so that its magnetic field is always parallel to the plane of the coil when the coil rotates and the magnitude of the magnetic field is the same at every point of time. This type of magnetic field is called radial magnetic field. Furthermore, soft iron core is kept suspended within the coil to increase the magnetic field.

Working

The current carrying coil ABCD is kept in a magnetic field such that the direction of the current and the field are perpendicular to each other.The direction of the current in the arms AB and CD are equal and opposite. Therefore the arms AB and CD experience equal and opposite force, forming a couple. Which tries to rotate the coil. This is called the deflecting couple. The direction of the rotation of the coil is given by Fleming's left hand rule.

As the coil rotates, the suspension wire and the loose spring of phosphor bronze get twisted, thus opposing the motion of the coil. This gives rise to an opposing couple called restoring couple. As the phosphor bronze wire is highly tensile at doesn't break when the spring and wire are twisted.

When the moment of force produced by the deflecting couple is equal to the moment of force due to restoring couple, the coil comes to rest, at a new position.

The angle through which the coil has rotated can be measured by lamp and scale arrangement. It is found that the tangent of the angle of deflection is directly proportional to the strength of the current.

$\tan \theta \propto i$.

$\therefore i = k \tan \theta$, where k is a constant and its value depends on

1. the number of turns in the coil,
2. the area of cross section of the coil and
3. the strength of the magnetic field

The value of k is given by the manufacturers of galvanometers and knowing the deflection, we can calculate the current. Most often, using the given k the scale is directly calibrated in terms of current for a given galvanometer.

DC motor

Motor is a device which converts electrical energy into mechanical energy.

Principle: It works on the principle that a current carrying coil experiences a couple when placed in a magnetic field, which sets it into continuous rotations.

The main parts of a motor are,

1. Permanent magnet N, S
2. Armature coil A, B, C, D
3. Commutator C_1, C_2
4. Brush B_1, B_2
5. Battery B

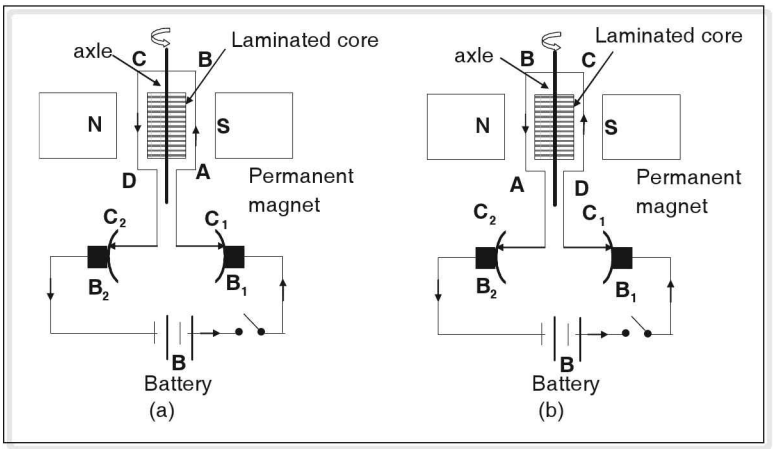


Figure 8.22

The permanent magnet gives the necessary magnetic field. This is made concave in shape so as to give radial field. In big motors, in order to have strong magnetic field electromagnets are used.

Armature coil consists of a large number of insulated copper coils wound on a laminated soft iron core. The function of the coil is to set up magnetic field when a current is passed through it. The core is laminated to avoid any eddy current loss, and to provide strong magnetic field when current flows through the coil.

Commutator

The ends of the wire are connected to the two split rings C_1 and C_2 of the commutator. The function of the commutator is to change the direction of the current after half a rotation of the coil.

Brushes

The function of the two brushes B_1 and B_2 is to give electrical connection from the battery wire to the rotating coil, i.e., between the stationary parts and the moving parts of the electric circuit.

Battery

The function of the battery is to give the motor the necessary electrical energy.

Working of the motor

The current from the positive terminal of the battery passes through the brush B_1, split ring C_1, and the armature coil ABCD returns via split ring C_2, brush B_2, to the negative terminal of the battery. The coil is placed in such a way that the magnetic field created by it is perpendicular to field due to magnets.

The current carrying coil is kept in the magnetic field of the permanent magnet. Therefore it will experience a force.

As per Fleming's left hand rule, the arm AB of the coil experiences a force perpendicular to the plane of the paper and into the paper. The arm CD experiences a force perpendicular to the plane of paper and outwards (opposite to that of end AB) as shown in fig 8.22 (a). These two forces constitute a couple and the motor rotates in *clockwise direction*.

When the coil turns through 90°, the magnetic field due to the coil and the field due to the permanent magnet become parallel to each other. Therefore rotation should stop as there will not be any couple when the two fields are parallel. But due to the inertia of motion, it continues to rotate and it turns through 180°.

Now the commutator reverses the direction of the current which flows along DCBA. (as shown in fig 8.22 (b). The arm CD of the wire experiences a force perpendicular to the plane of the paper and into the paper, and the arm AB experiences a force perpendicular to the plane of the paper and outwards (opposite to that of end AB). Thus the motor rotates continuously in one direction.

The revolutions per minute of a motor coil depends on the following:

1. The number of the turns of the coil
2. The area of the coil
3. The magnitude of the current
4. The strength of the magnetic field.

Electromagnetic induction

We have seen that when current flows through a conductor placed in a magnetic field, it experiences a force which makes it move.

Micheal Faraday, through series of experiments showed that current can be induced in a conductor if it is moved in a magnetic field or if magnet is moved relative to conductor. Due to the relative motion between them, the magnetic flux (ϕ_0) linked with the conductor i.e., the number of lines of force passing normally through it, changes. This sets up an emf at the ends of conductor which causes current to flow. Faraday found that the current flows as long as there is continuous change in the magnetic flux passing through the conducting coil. Thus even if both coil and magnet are stationary, current can still be induced in coil, if the magnetic field is changed continuously which is possible by using electromagnets. This phenomenon of inducing current by changing the magnetic field associated with an electric circuit containing no source is called electromagnetic induction.



Experiment 1

Faraday's experiments

When a magnet is moved towards a coil with its north pole towards the coil as shown in the figure above, Faraday observed a deflection in the galvanometer. The amount of deflection is found to be more if the speed of magnet is more.

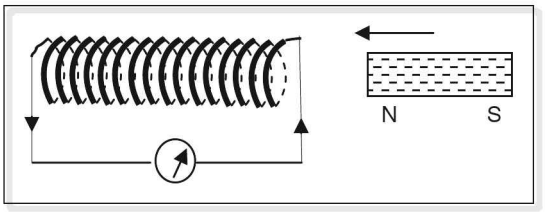


Figure 8.23

When the magnet is moved away from the coil, then it produces deflection in the opposite direction as compared to the case, when the magnet is moved towards the coil.

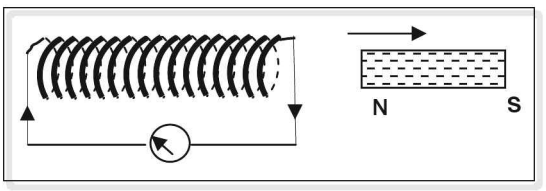


Figure 8.24

When the magnet is moved towards the coil with its south pole pointing towards the coil, then it produced a deflection in the opposite direction as compared to the case, when the north pole was pointing towards the coil.

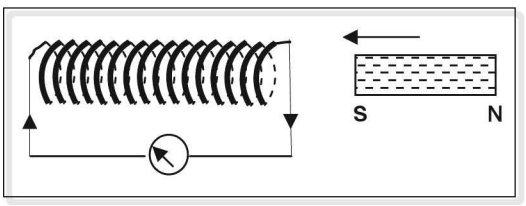


Figure 8.25

Once again, if the magnet is moved away from the coil with the south pole pointing towards the coil, then it produces deflection in the opposite direction as compared to the case when the magnet is moved towards the coil.

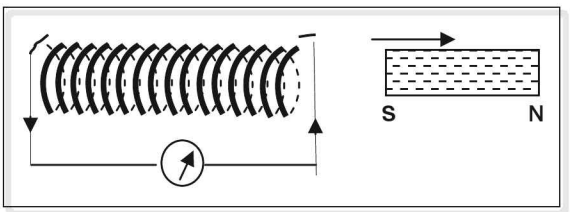


Figure 8.26

Similar observations were made when the coil was moved towards or away from a stationary magnet. The deflections in the galvanometer were governed by the orientation of the magnet and the direction of the motion of the coil.

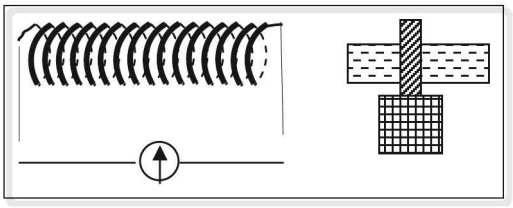


Figure 8.27

Now if we fix both the magnet and the coil such that there is no relative motion, then no deflection is observed in the galvanometer.

On the basis of the observations made in his experiments, Faraday formulated the following laws.

Faraday's Laws of Electromagnetic Induction

1. Whenever the magnetic flux linking a coil changes, then an emf is induced in the coil.
2. The magnitude of the induced emf is directly proportional to the rate of the change of flux.
3. Furthermore, the induced emf depends upon the number of turns of the coil and its area.

From the above observations, we can conclude that if there is a *relative motion between the magnet and the coil* an emf is induced in the coil.

1. This emf lasts so long as there is a relative motion between the magnet and the coil.
2. If there is no relative motion, then no emf is induced.
3. The magnitude of emf depends upon the relative speed.
4. The magnitude of emf depends upon the number of turns in the coil.

Following the above conclusions, Faraday formulated the law for induced emf in a coil as, the induced emf in a coil is equal to the negative rate of the change of magnetic flux linked with the coil. Thus

$$E = -N\frac{d\varphi}{dt}$$

The observations of Faraday's experiments on electromagnetic induction and his laws can also be studied by replacing the permanent magnet with an electromagnet.

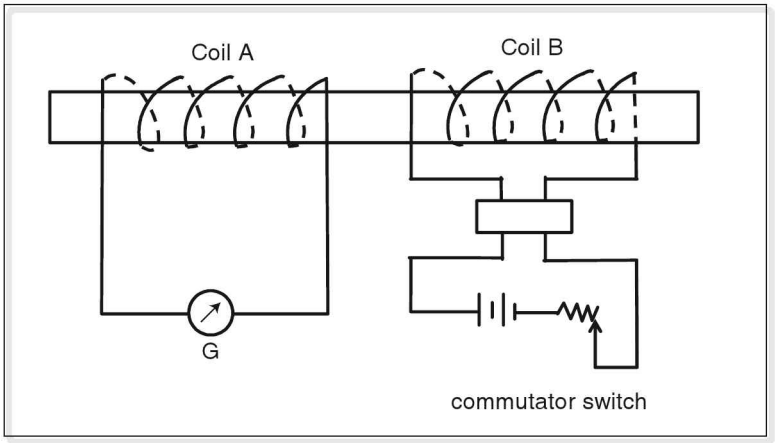


Figure 8.28

Wind two coils on a soft iron core separated by some distance, connecting the ends of coil A to a galvanometer and the ends of B to a battery through a rheostat (to vary the current) and a commutator switch (to reverse the current). It is observed that the needle of the galvanometer instantly jumps

to one side and returns to zero quickly whenever the key is closed. When the key is opened, thus disconnecting the coil from the battery, similar deflections are observed in the galvanometer, but in the opposite directions.

Repeating the experiment after reversing the direction of current in coil B, similar observations are recorded.

Lenz's Law

According to Lenz's law, the direction of induced current is such that it opposes the very cause that produces it.

Consider a bar magnet moving towards a coil with its north pole facing the coil. Due to the movement of magnet, current is induced in anti-clockwise direction in the coil when it is viewed from magnet. The face pointing towards magnet becomes north pole and opposes its movement. When magnet is pulled away such that its north-pole leaves the coil, current is induced in clockwise direction giving the face of the coil pointing towards magnet a south polarity. Thus the movement of magnet which is responsible for induced current, is always opposed by the current itself.

Let $d\phi_B$ be the change in the magnetic flux in a short interval of time dt. The induced emf (E), according to Faraday's law is given by

$$E = \frac{d\varphi_B}{dt} \quad \text{—— (1)}$$

According to Lenz's law, the induced emf is such that it opposes the cause producing it. Hence equation (1) on application of Lenz's law becomes

$$E = -\frac{d\varphi_B}{dt} \quad \text{——} \qquad (2)$$

If the coil contains N turns, equation (2) becomes

$$E = -N\frac{d\varphi_B}{dt}$$

Fleming's right-hand rule

To determine the direction of induced current in a conductor, when it is moved across a magnetic field, Fleming proposed the 'Right-hand rule'.

Stretch the forefinger, middle finger and thumb of your right hand in three mutually perpendicular directions, such that the forefinger points the direction of magnetic field(B), the thumb indicates the direction of motion(M), then the middle finger represents the direction of induced current (i) in the conductor.

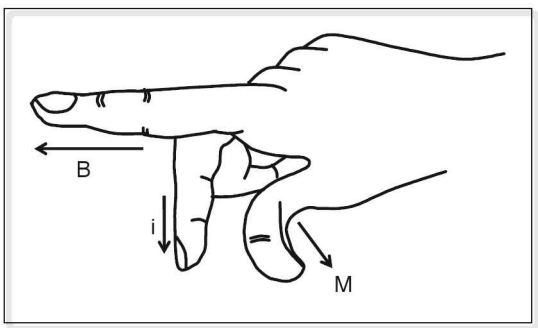


Figure 8.29

Moving coil microphone

It is a device which converts sound energy into electric energy. It works on the principle of electromagnetic induction.

It consists of a diaphragm, a moving coil and a permanent magnet. The coil is kept in the magnetic field. As a person speaks into microphone, diaphragm is set into vibrations. The moving coil connected to it begins to vibrate accordingly. It being under magnetic field, a electric current is induced in it whose variations represent the vibrations of original sound. This electrical signal is amplified and fed to loud speaker.

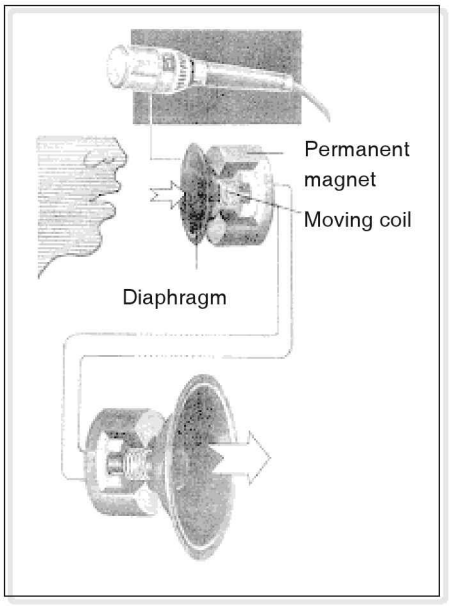


Figure 8.30 Microphone and loudspeaker

Loud speaker

It converts electric energy into sound energy. It consists of a permanent magnet, moving coil and a diaphragm in the form of paper cone as shown in the figure.

The electrical signal corresponding to the variations of original sound from the microphone is fed to the coil placed between the poles of permanent magnet. The coil begins to vibrate and sets the paper cone into vibrations. The air around the loudspeaker vibrates reproducing the original sound more loudly.

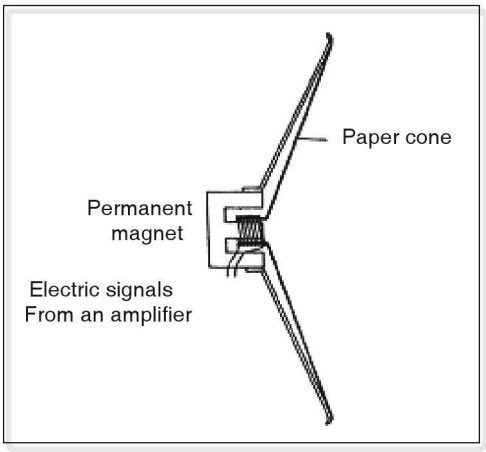


Figure 8.31 Parts of a loud speaker

Electric generator

An electric generator is a device which converts mechanical energy into electrical energy.

When a coil is rotated in a magnetic field, an emf is induced in it. The mechanical energy required to rotate the conductor is converted into an electrical energy. Hence, an electric generator converts mechanical energy into electrical energy.

There are two types of generators—Alternating Current (AC) generator and Direct Current (DC) generator. In the former, the current changes direction after every half-rotation. In the latter, the current is unidirectional throughout.

The alternating current generator

The main parts of an AC generator are the horseshoe magnet, armature (coil), slip rings and carbon brushes.

Construction

The construction of an electric generator is similar to that of an electric motor. An AC generator consists of a rectangular coil ABCD known as armature. This armature is rotated rapidly between the poles of

a permanent horseshoe magnet. The ends of the rectangular coil A and D are connected to two circular metal pieces called slip rings A_1 and A_2. The slip rings rotate along with the coil. The two pieces of carbon brushes B_1 and B_2 are kept in contact with the slip rings. The carbon brushes are capable of tapping the current produced in the rotating coil.

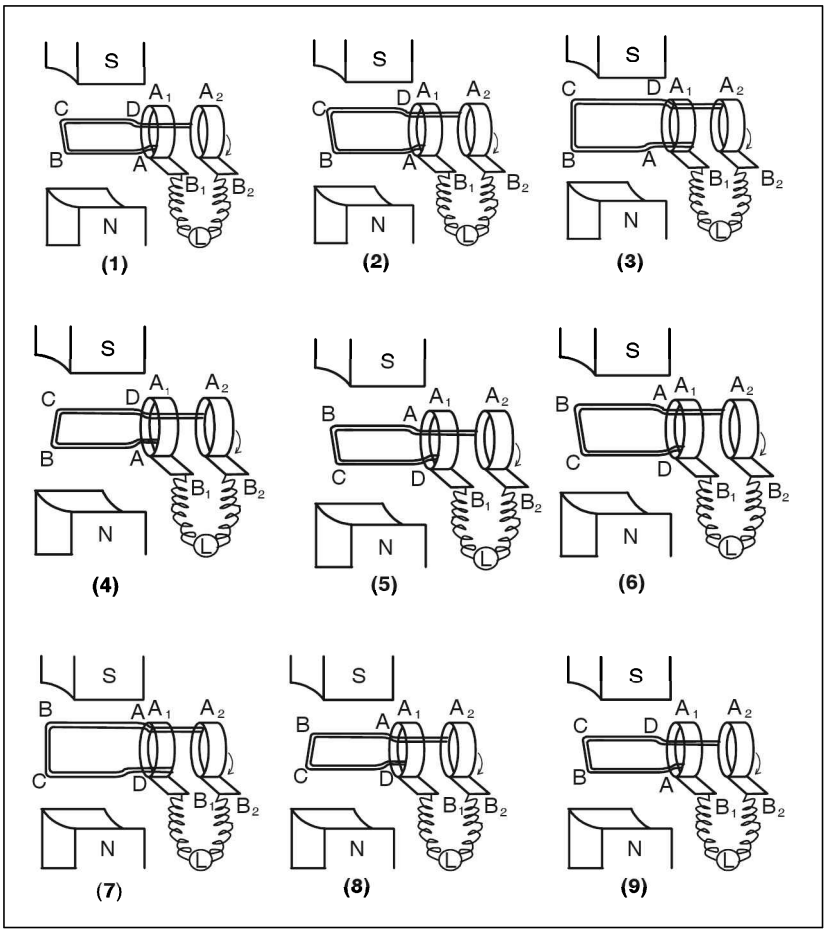


Figure 8.32

Working

1. Assume that the rectangular coil ABCD is in horizontal position initially.
2. Let the rectangular coil ABCD be rotated from 0° to 90°, such that during its rotation, the side AB of the coil moves downwards and the side CD of the coil upwards. Hence the flux linked decreases, but the rate of change in the flux increases from zero to maximum. Hence, emf induced increases from zero to maximum, (according to Faraday's laws of electromagnetic induction). The current flows in the direction ABCD according to Lenz's Law.
3. On further rotation from 90° to 180°, the side AB of the coil starts moving upwards while the side CD of the coil moves down. Hence, flux linked increases from zero to maximum. But the rate of change in flux decreases from maximum to zero. The emf induced in the coil decreases from maximum to zero. The current continues to flow in the direction ABCD, but the magnitude of the current decreases.
4. On further rotation of the coil from 180° to 270°, the side AB of the coil which is now once again horizontal moves upwards while the side CD of the coil moves downwards. The flux linked by the coil decreases from maximum to zero. But the rate of change in flux increases from zero to maximum in the opposite direction. Hence the emf induced is maximum, but in opposite direction, and the current starts flowing in the direction DCBA, with the magnitude increasing, in accordance with the Lenz's Law.
5. On further rotation of coil from 270° to 360°, the side AB of the coil moves downwards while the side CD of the coil moves upwards. The flux linked by the coil increases from zero to maximum and the rate of change in flux decreases from maxiumum to zero in opposite direction. Therefore,

the emf induced also decreases from maximum to zero in opposite direction, with decreasing current continuing to flow in the direction DCBA.

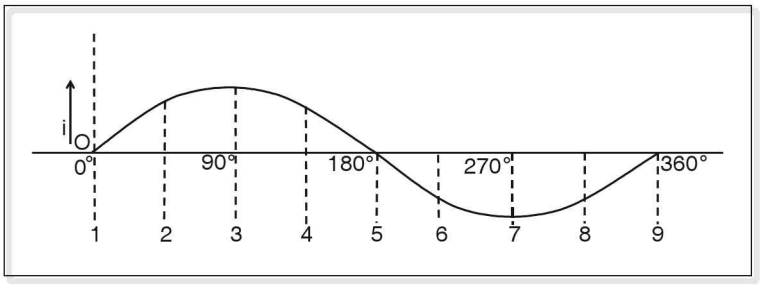


Figure 8.33

In the above, the direction of induced emf at any instant can be found using Fleming's right hand rule or Lenz's Law for one complete rotation of the coil i.e., from 0° to 360°, the coil produces an alternating voltage (AC). The variation in the current produced for various positions of the coil is shown in the graph.

Alternating current (AC) dynamo

Principle

When a conducting coil in a closed circuit in a constant uniform magnetic field rotates, the magnetic flux passing through it changes continuously producing an induced emf or current in the circuit.

Construction

It consists of four major parts.

1. **Armature:** It consists of insulated copper coil wound over a rectangular frame and soft iron core which a laminated. In the figure ABCD represents armature.
2. **Slip rings:** S_1 and S_2 are two slips rings connected to the ends of rectangular copper coil. They rotate along with the coil about the same axis.
3. **Carbon brushes:** B_1 and B_2 are two carbon brushes used as electrical contact between moving parts (S_1 and S_2) and stationary part (load R).
4. **Permanent magnet:** N-S is powerful horse shoe magnet having concave cylindrical poles. It produces a uniform radial magnetic field.

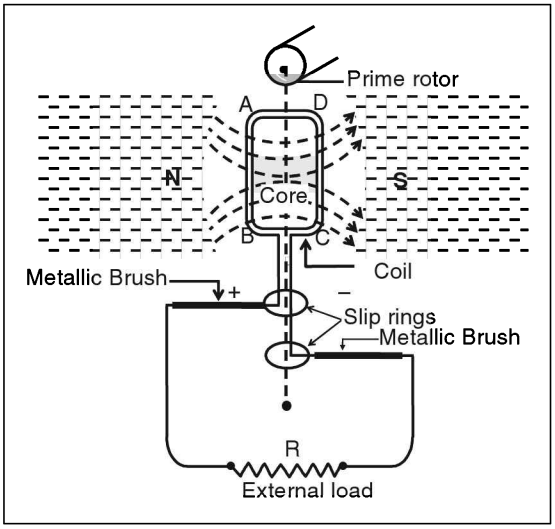


Figure 8.34 An electric generator

Working

As the amature ABCD is rotated between the poles of magnet, the magnetic flux linked with it changes continuously resulting in generation of induced emf which causes current to flow through the coil and load resistance R. As the position of amature coil changes due to its rotation, the magnitude of induced current also changes as shown in following diagram.

In its initial position where angle of rotation is zero, the plane of the coil is perpendicular to the magnetic lines of force or magnetic flux (ϕ_B). Maximum number of lines of force pass through coil but

the induced current is zero as coil is stationary. Let the coil is rotated in counter clockwise direction such that arm ABgoes down and CD comes up. As the coil rotates by $90°\left(\frac{\pi}{2}\right)$, its plane becomes parallel to lines of forces. The magnetic flux linked with it is zero but the change in magnetic flux ($d\phi_B$)with respect to original position is maximum. Hence maximum induced current flows. The direction of this current can be found by applying fleming's right hand rule. As the arm AB moves in downward direction and lines of force are from right to left, current flows from B to A in the arm AB. Arm CD has moved in upward direction, hence current through it flows in opposite direction i.e., from D to C. Through the load R, it flows from B_1 to B_2. As the coil rotates further by 90°, it again becomes perpendicular to magnetic flux with AB below CD. The magnetic flux passing through the coil is same as in initial position. Hence the change $d\phi_B$ is zero causing the induced current to decrease from maximum to zero in the same direction. From this position as the coil rotates by 90°, AB moves up and CD down such that its plane is parallel to magnetic flux. As the change in the flux ($d\phi_B$) is maximum with respect to 'O' the induced current increases from zero to maximum in opposite direction which can be verified by applying flemings right hand rule. Now the current flows from A to B, in AB, C to D in CD and B_2 to B_1 in R. As the coils comes back to its original position by rotating more by 90°, current decreases from maximum to zero in the same direction. This cycle is repeated as coil continues its rotational motion.

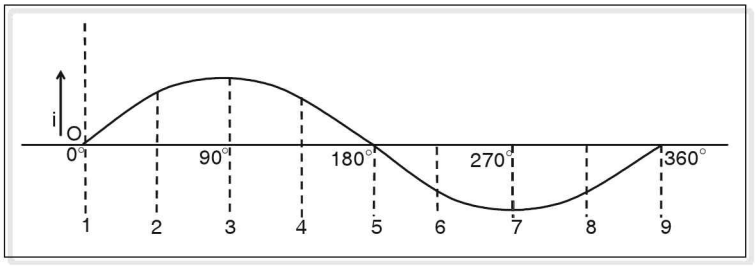


Figure 8.35

It can be seen that during 1st half of rotation i.e., from 0 to π, current flows from B_1 to B_2 in R, while during 2nd half from π to 2π, it flows from B_2 to B_1. Thus current changes its direction after regular intervals of time. Hence alternating current is generated. This reversal in current direction is brought about by slip rings S_1 and S_2 which alternately connect the two ends of coil to B_1 and B_2.

The magnitude of induced current can be increased by increasing:

1. the number of turns in coil.
2. the area of cross section of coil.
3. the intensity of magnetic field.
4. the speed of rotation of coil.

RMS voltage and peak voltage

In this type of a generator, we observe that the induced voltage across the output terminals alternates between $+V_0$ and $-V_0$ as shown in the figure above. The alternation from $+V_0$ to $-V_0$ and again back to $+V_0$ (or $-V_0$ to $+V_0$ to $-V_0$) is referred to as one cycle, and if the frequency is n Hz, it implies that n cycles are completed in one second. The voltage V_0 is called peak value.

If the terminals of this generator are connected across a resistor the current flowing through the circuit would also alternate at the same frequency and is referred to as alternating current. Thus, such a source of electrical energy is referred to as AC source or Ac supply. The symbol for an AC source is. ~

If a heater coil is connected to an AC supply of peak voltage V_0, the heater coil gets continuously heated even though the direction of current alternates. It is found that the heating effect produced by such a source would be the same as that caused by a constant voltage source such as a cell of terminal p.d. equal to $V_0/\sqrt{2}$. It can be shown that this value is the square root of the mean of the squares of the instantaneous voltages across the terminals of the AC supply. Hence it is called the Root Mean Square (RMS) value.

$$\therefore V_{RMS} = \frac{V_0}{\sqrt{2}} \text{ and } I_{RMS} = \frac{I_0}{\sqrt{2}}$$

☛ *Example*

Find the RMS value of +4, +2, 0, –2 and –4

Solution

Let $x_1 = 4, x_2 = +2, x_3 = 0, x_4 = -2$ and $x_5 = -4$.

Then, $x_1^2 = x_5^2 = 16$

$x_2^2 = x_4^2 = 4$

$x_3^2 = 0$

$\therefore\ x_1^2 + x_2^2 + x_3^2 + x_4^2 + x_5^2 = 16 + 4 + 0 + 4 + 16 = 40$

Mean of the squares $= \frac{40}{5} = 8$

Square root of the mean $= \sqrt{8} = 2\sqrt{2}$

$\therefore$ The RMS of the given numbers $= 2\sqrt{2}$

☛ *Example*

If an AC supply is rated as 220 V, find the peak voltage.

Solution

The rating of an AC supply refers to its RMS value

Thus $V_{RMS} = 220\,V$

i.e., $\frac{V_0}{\sqrt{2}} = 220\,V$

$\therefore V_0 = 220\sqrt{2}\,V$

Therefore the peak voltage is $220\sqrt{2}$ V.

The direct current generator

When the current flows continuously in one direction, the current is referred to as direct current. To obtain

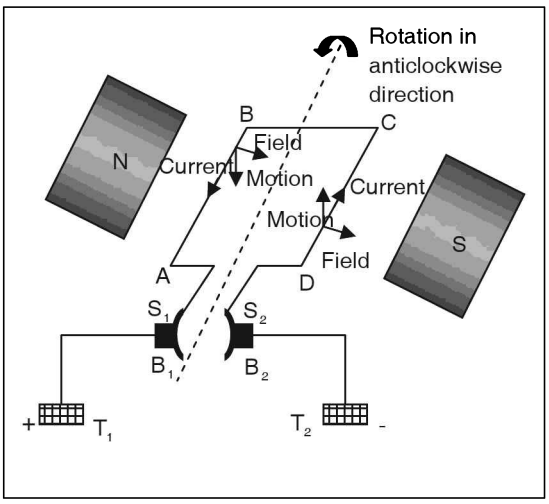


Figure 8.36 DC Generator

such a DC current, the AC generator is modified by replacing the slip rings with a commutator (half-slip rings or split rings) S_1S_2.

Working

Let the generator coil ABCD be initially in the horizontal position. Let the coil be rotated in an anticlockwise direction as shown in the figure above, side AB downwards and side CD upwards. According to Fleming's right-hand rule, the induced current flows in the direction BA in the side AB and in the direction DC in the side CD. Thus induced current flows in the direction DCBA during the first half of rotation.

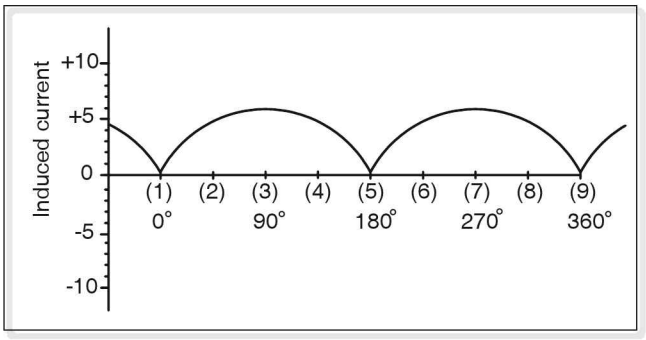


Figure 8.37 Graph of Direct Current

After half rotation, arms AB and CD interchange their positions. When the coils interchange their positions, the commutator split rings S_1 and S_2 automatically change their contact with the carbon brushes. Due to this, the current flowing in the load remains in one direction. Thus, the brush B_1 remains always positive and the brush B_2 remains always negative. Thus, the current would always be flowing out from the terminal T_1.

Inductance of a coil

The relative motion of a coil and a magnet induces emf in the coil. This emf is referred to as dynamically induced emf, since the emf is induced due to the actual motion of either the magnet or coil. If a change in the current in one coil induces an emf in another coil (due to changing magnetic field linked with each other), then the emf induced in the second coil is referred to as statically induced emf.

Inductance is the property of a coil to induce emf in the same coil or another coil placed near it due to the changes in the current in the coil.

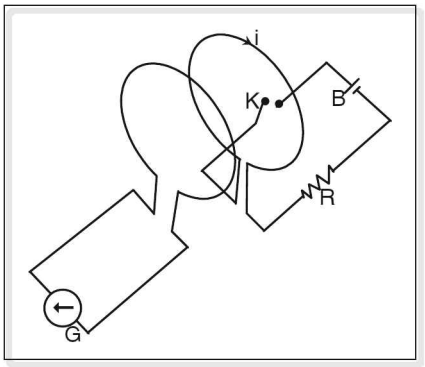


Figure 8.38

Self-induction

Every current carrying coil produces a magnetic field which is linked to the same coil. When the current in the coil changes, the magnetic flux produced also changes and is linked to the coil itself. According to Faraday's law of induction, this change in magnetic field produces an induced emf which opposes the change.

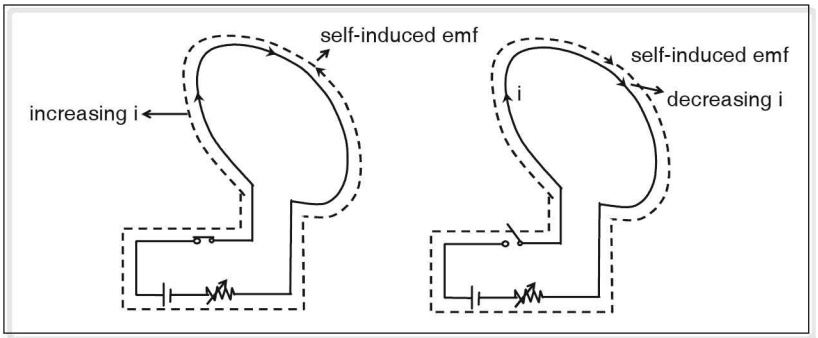


Figure 8.39

This phenomenon in which an induced emf is produced in a coil due to a change in the current in the same coil is called self-induction, and the induced emf is generally referred to as self-induced emf.

The induced emf is equal to the rate of change of flux.

$$E = -\frac{d\varphi}{dt}$$

The negative sign indicates the induced emf opposes the applied emf.

$$\text{But } \frac{d\varphi}{dt} \propto \frac{di}{dt} \text{ and so } \frac{d\varphi}{dt} = L\frac{di}{dt} \text{ and}$$

$$\therefore E = -L\frac{di}{dt}$$ where 'L' is the constant of proportionality and is called self-inductance of the coil.

$$L = \frac{-E}{di/dt}$$

The self inductance 'L' of a coil is numerically equal to the ratio of the induced emf to the rate of change of current in the coil.

Mutual induction

The phenomenon due to which a change in the current in a coil induces an emf in another coil held close to it is called 'mutual induction'.

Here, the current is changing in coil 'B'. Therefore the flux associated with the coil B is changing. If we place one more coil 'A' near it, the changing magnetic field in coil 'B' will produce an emf in the second coil 'A'. This is known as mutual induction.

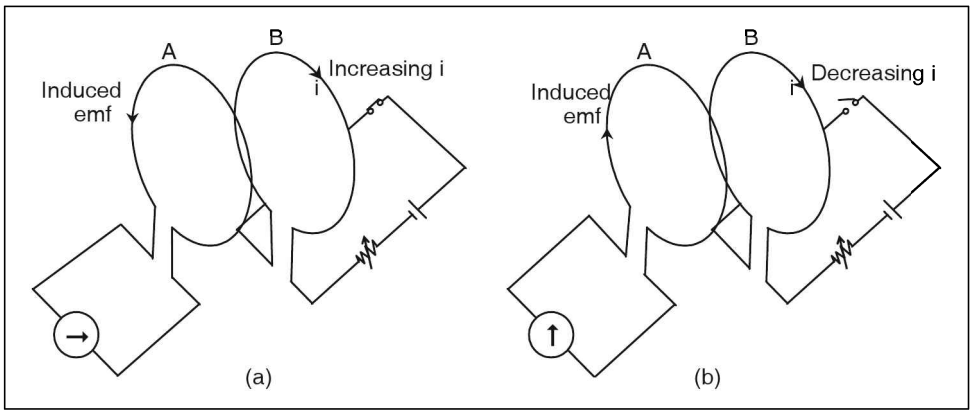


Figure 8.40

$$E_A \propto -\frac{di_B}{dt}$$

$$E_A = -M_{AB}\frac{di_B}{dt}$$

Here, the constant of proportionality 'M' is called the 'mutual inductance'.

The negative sign indicates that the induced emf opposes the change producing it. The mutual inductance of a coil is numerically equal to the ratio of the induced emf in the coil to the rate of change of current in the another coil, placed near it.

$$M_{AB} = -\frac{E_A}{di_B/dt}$$

The unit of inductance (self-inductance and mutual inductance) is henry (H) and 1 H = 1 V s A^{-1} (volt second per ampere)

Transformer

It is a static electrical device used to step-up a low voltage to a high voltage or to step-down a high voltage into a low voltage and is used on AC circuits. Transformer works on the principle of electromagnetic induction. The main parts of a transformer are,

1. a soft iron laminated core,
2. a primary coil and
3. a secondary coil.

Core

The core is made up of a rectangular frame made up of silicon steel. Each sheet is insulated from neighbouring sheet by varnishing it.

1. It is made up of soft iron to increase the magnetic permeability.
2. It is made up of laminations to avoid eddy current loss.

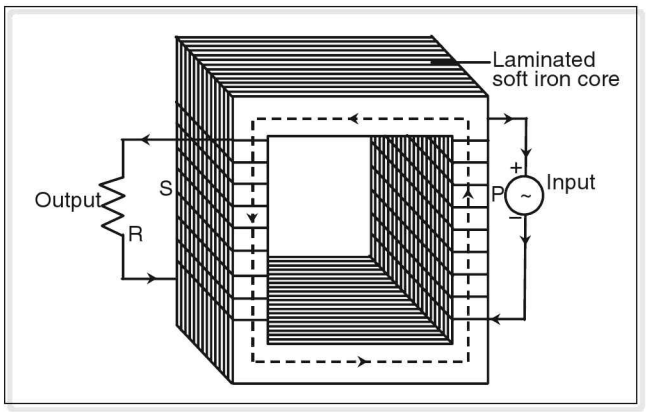


Figure 8.41

Primary coil

The input of the alternating voltage is connected to this coil. Since the voltage changes with respect to time, changing magnetic flux is introduced by this coil. This changing flux is coupled to the secondary coil through the core.

Secondary coil

The output from the transformer is taken from this coil. The changing flux due to changing voltage in the primary coil is coupled to the secondary coil. This will produce an alternating voltage.

The symbolic representation of a transformer in electric circuits is as shown below.

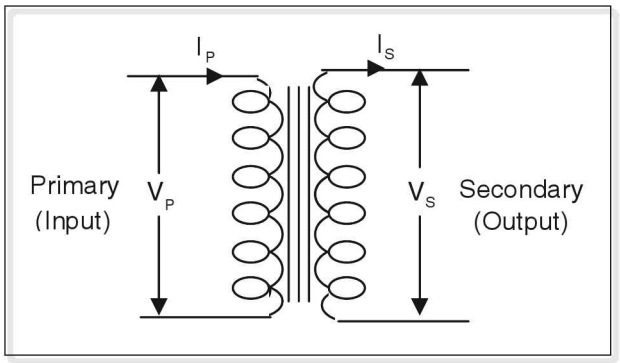


Figure 8.42

Step-up transformer

A step-up transformer steps the primary (incoming) voltage up to a higher value. If the number of turns of the secondary coil is more than the number of turns of the primary coil, it is called a step-up transformer. In this case, the secondary voltage will be more than the primary voltage. The secondary current is less than the primary current.

Step-down transformer

A step-down transformer steps the primary voltage down to a lower value. If the number of turns of the secondary coil is less than that of the primary coil, it is called a step-down transformer. In this case, the secondary voltage will be less than the primary voltage. The secondary current is more than the primary current in a step-down transformer.

Working

If an AC voltage V_1 is applied to the primary coil of N_1 turns, an alternating current flows in the primary coil and produces a changing magnetic flux. This magnetic flux is linked to the N_2 turns of the secondary coil, thereby, an emf is induced in the secondary coil. Let it be V_2.

According to Faraday's law of induction,

$$E = -N\frac{d\varphi_B}{dt}$$

i.e., emf per turn, $E_{turn} = \frac{E}{N} = -\frac{d\varphi_B}{dt} = \text{constant}$

Thus emf per turn is the same for both primary and secondary coils.

$$\therefore \frac{V_1}{N_1} = \frac{V_2}{N_2}$$

Thus, the relation between the number of turns and the voltage in a transformer is given by

$$\frac{V_1}{V_2} = \frac{N_1}{N_2} \quad \text{---------- (1)}$$

Since in an ideal transformer, there is no loss of power,
power in the primary = power in the secondary coils. If I_1 and I_2 are the currents in the primary and secondary coils respectively,

$$P = V_1I_1 = V_2I_2$$

$$\frac{V_1}{V_2} = \frac{I_2}{I_1} \quad \text{---------- (2)}$$

From (1) and (2),

$$\frac{N_1}{N_2} = \frac{V_1}{V_2} = \frac{I_2}{I_1}$$

$$V_2 = V_1\frac{N_2}{N_1} \text{ (transformation of voltage)}$$

$I_2 = I_1 \frac{N_1}{N_2}$ (current transformation)

$\therefore N_2 > N_1 \Rightarrow$ step-up transformer

$\therefore N_2 < N_1 \Rightarrow$ step-down transformer.

☛ *Example*

In a transformer, the ratio of the number of turns of a primary to a secondary coil is 2 : 5. If the current through the primary coil is 10 A and input power is 300 W, find the current and voltage in the secondary coil.

Solution

The turn-ratio and currents are related as

$\frac{N_1}{N_2} = \frac{I_2}{I_1}$ - - - - - (1)

In the problem, values given are $\frac{N_1}{N_2} = \frac{2}{5}$, $I_1 = 10$ A

Substituting the above values in (1) and solving for I_2, we get

$\frac{2}{5} = \frac{I_2}{10}$

$\Rightarrow I_1 = \frac{2}{5} \times 10 \Rightarrow I_2 = 4$ A

In an ideal transformer, the input and the output powers are equal. Therefore, we can write

$V_1 I_1 = V_2 I_2$ - - - - - - (2)

But the input power is given as 300 W. Substituting this in equation (2), we get,

$V_2 I_2 = 300$

Substituting $I_2 = 4$ A in the above equation, we get

$V_2 \times 4 = 300 \Rightarrow V_2 = \frac{300}{4}$

$\Rightarrow V_2 = 75$ V.

Energy loss in transformers

1. Due to the finite resistance of the primary and secondary coils, there will be a power loss in the primary and secondary coils. This is known as copper loss of transformer.

   Copper loss $= I^2R$
2. A part of the energy is wasted on account of the magnetic flux in the soft iron core. These are known as magnetic losses or hysteris losses.
3. A part of energy is wasted on account of the eddy current. This is known as eddy current loss. This is minimized by using laminated core.

Uses of Transformer

1. At the generating station, a transformer is used to step-up the voltage before transmitting it for long distances to avoid losses. At the receiving end, it is used to step-down to the required level to the supply to the domestic users. It is used to change the voltage to the required level.
2. In a radio, TV, X-ray tube etc, either the step-up or step-down transformers are used.
3. In stabilizers, step-up–cum–step-down transformers are used to control the fluctuations in voltage.

Eddy currents

Induced currents produced in a solid core placed in changing magnetic field is called an eddy current. It has advantages and disadvantages.

Let us see why eddy currents are undesirable. The eddy currents produced heat up the metallic solid core. The heating of core by eddy currents is undesirable because

1. it results in the loss of useful energy,
2. it increases the risk of the breakdown of the insulation of the winding.

Eddy currents can be eliminated by the following methods.

1. The solid metal is cut in such a manner so as to increase the path for the eddy current. Hence the resistance increases and eddy current reduces.
2. Eddy currents can be reduced by using a laminated core. Instead of using a solid core, if we use thin sheets packed together and insulated between two layers, the resistance increases and the eddy current is reduced. This is the reason why the coil of a transformer is wound on a laminated core and not on the solid core.

Uses of eddy current

We can make use of the eddy current for some useful purpose. Eddy currents produce heat in metals. Hence metals can be melted using the eddy current. This is the principle in induction melting and heating furnaces.

test your concepts

Very short answer type questions

1. State the expression for magnetic induction field on axial and equatorial line. Give its direction.
2. Define retentivity.
3. What is meant by an electromagnetic induction?
4. State Fleming's right hand rule.

5. Define a neutral point. Give the location of neutral points. When the North Pole of a bar magnet faces geographic north and geographic south.
6. Give the relation between V, N and I with respect to the primary and secondary coils of a transformer.
7. State Lenz's law.
8. Explain the principle of an electric generator.
9. Define: (1) magnetic field (2) magnetic lines of force
10. Define intensity of magnetisation magnetic susceptibility and retentivity. Give their SI units.
11. Which observation made in Oersted's experiment with a current carrying conductor concludes that the magnetic induction is proportional to the strength of the current in the conductor?
12. What is the difference between an AC dynamo and a DC dynamo?
13. What is magnetic saturation?
14. Define dia, para and ferromagnetic substances. Give examples.
15. Why the magnetic lines of force away from a straight conductor carrying current are elliptical?
16. Define (a) inductance (b) self induction (c) mutual induction
17. Define the pole strength of a magnet. What is the S.I. unit of pole strength?
18. What is a domain in ferromagnetic substances?
19. What type of magnetic field is observed at the centre of a circular coil carrying current?
20. Give the principle of a transformer? State the types of transformer
21. What is meant by electromagnetism?
22. Why is Ewing's theory called molecular theory of magnetism?
23. What is meant by a solenoid?
24. Why do magnetic lines of force pass through iron more freely than through air?
25. According to Ewing's molecular theory, why is steel used to make permanent magnets?
26. Which rule is used to find the force on a current carrying wire placed in a magnetic field?
27. Mention the energy conversions taking place in loud speaker and microphone.
28. Define magnetic induction field B. Intensity of magnetic field (H) and state their unit.
29. What happens to the length of a iron bar when it is magnetised?
30. State the principle of a moving coil galvanometer?

Short answer type questions

31. State the properties of
    (a) diamagnetic (b) paramagnetic (c) ferromagnetic substance
32. Explain how will you calculate magnetic moment (M) of a bar magnet using neutral points?
33. Explain domain theory.

33. Explain domain theory.
34. Give the properties of a solenoid.
35. Calculate magnetic field induction B due to a conductor carrying current 5 A at a distance of 10 cm from it.
36. Write a note on important magnetic properties?
37. Give various characteristics of magnetic lines of force.
38. Describe Ewing molecular theory.
39. Calculate magnetic field induction 'B' at the centre of a circular coil having 500 turns, radius π cm and carrying 5 A of current.
40. Explain Oersted's experiment.
41. Explain why isolated magnetic poles do not exist?
42. What is modern electron theory?
43. Explain Faraday's experiment using a magnet and a coil and state Faraday's laws of electromagnetism
44. Explain magnetic induction 'B' and intensity of magnetizing field.
45. Distinguish between absolute and relative permeability of a medium.

Essay type questions

46. Describe moving coil galvanometer.
47. Explain in detail:
    (a) Electric motor (b) AC generator
    (c) DC generator (d) Transformer
48. Write a note on terrestrial magnetism.
49. Write a note on: (1) self-inductance (2) mutual inductance
50. How do you locate the neutral points when
    (1) the North Pole of a bar magnet faces geographical north and
    (2) when the south pole of bar magnet faces geographical north?

CONCEPT APPLICATION



Concept Application Level—1

Direction for questions 1 to 7: State whether the following statements are true or false.

1. In a DC electric motor a pair of split rings is used as commutator.
2. According to Ewing's molecular theory, every molecule of a given substance is a magnet by itself.

3. A transformer is an electrical device that works on the principle of self-induction.
4. Outside the magnet, the magnetic lines of forces pass from north pole to the south pole and inside the magnet, they pass from south to the north pole.
5. Lenz's law is used to find out the magnitude of the induced emf.
6. Relative permeability and absolute permeability have the same units.
7. The angle between geographic and magnetic meridian is called angle of dip.

Direction for questions 8 to 14: Fill in the blanks.

8. The phenomenon of production of back emf in a coil due to flow of varying current through it is called _____.
9. The unit of self-inductance in SI system is _____.
10. The point where resultant magnetic field strength is zero is called _____.
11. Fleming's left-hand rule is used to find the direction of _____ acting on the current carrying conductor placed in a/an _____ field.
12. _____ is the measure of the ability of a given magnetic pole to attract or repel another magnetic pole.
13. The product of pole strength of a magnet and its length is called its _____.
14. An emf is induced in a coil when _____ linked with it changes.

Direction for question 15: Match the entries in column A with appropriate ones from column B.

15.

A.	μ	( )	a. Faraday's law of electromagnetic induction
B.	Diamagnetic substance	( )	b. Magnetic equator
C.	H	( )	c. $\mu_0\mu_r$
D.	Angle of dip = 0	( )	d. Deflection of coil directly proportional to the current flowing through the coil
E.	Tan $\theta \propto I$	( )	e. $\frac{B}{\mu_o}$
F.	$E = -N\frac{d\varphi}{dt}$	( )	f. Thick copper wire in the secondary winding.

(Continued on following page)

G.	Moving coil galvanometer	( )	g.	force on a current carrying conductor placed perpendicular to field.
H.	Step down transformer	( )	h.	Tangent galvanometer
I.	$F = Bi\ell$	( )	i.	Susceptibility small and negative.
J.	Solenoid	( )	j.	Uniform magnetic field.

Direction for questions 16 to 30: For each of the questions, four choices have been provided. Select the correct alternative.

16. A transformer
(1) converts AC to DC
(2) converts DC to AC
(3) increases or decreases (step up or step down) AC voltage
(4) increases or decreases (step up or step down) DC voltage.

17. In an AC generator, maximum number of lines of force pass through the coil when the angle between the plane of coil and lines of force is _____.
(1) 0° (2) 60° (3) 30° (4) 90°

18. Magnetic induction field due to a short bar magnet on its equatorial line at a certain distance is B. Its value at the same distance on the axial line is _____.
(1) 2B (2) $\frac{B}{2}$ (3) 3B (4) $\frac{B}{3}$

19. Magnetic field inside the solenoid is_____.
(1) radial (2) uniform (3) Both (1) and (2) (4) circular

20. A step down transformer steps up _____ and steps down _____.
(1) current, voltage (2) voltage, current (3) power, energy (4) voltage, power

21. The power of a DC motor can be increased by
(1) increasing the area of the cross section of the coil.
(2) increasing the current flowing through the coil.
(3) laminating the soft iron core.
(4) All of the above

22. Keeping the distance between two magnetic poles constant, if the two poles are connected by an iron rod, the force between the two poles _____.
(1) increases (2) decreases (3) remains same (4) None of these

23. In a suspension type of moving coil galvanometer, the couple developed in the suspension wire and the loose spring is called a_____.
(1) deflecting couple (2) restoring couple (3) twisting couple (4) None of these

24. The right hand thumb rule is used to find
(1) force on a charged particle passing through the magnetic field.
(2) force on a current carrying conductor placed in a magnetic field.

(3) direction of induced current.
(4) direction of B around a current carrying straight conductor.

25. Looking from one side towards the face of a coil, the current in it flows in clockwise direction. On looking from the opposite side towards the second face of the coil, the current flows in
(1) anti-clockwise direction and this face represents the south pole.
(2) clockwise direction and this face represents the north pole.
(3) anticlockwise direction and this face represents the north pole.
(4) clockwise direction and this face represents the south pole.

26. The relative permeability of a paramagnetic substance is _____.
(1) much more than one (2) slightly more than one
(3) less than one but greater than zero (4) zero.

27. Ferromagnetic substances achieve a high degree of magnetic alignment due to the effect known as ____.
(1) pole swapping (2) exchange coupling
(3) magnetic poling (4) None of these

28. An AC generator can be converted into DC generator by replacing _____.
(1) armature with coil (2) concave magnets with horse shoe magnet
(3) slip rings with split rings (4) All of the above

29. Magnetic permeability is the ratio of
(1) magnetic induction to susceptibility. (2) magnetic induction to magnetizing field.
(3) magnetizing field to magnetic induction. (4) magnetizing field to susceptibility.

30. When a piece of iron is placed in a changing magnetic field, it gets heated due to flow of _____.
(1) Eddy current (2) Faraday's current
(3) Fleming's current (4) None of the above

Concept Application Level—2

31. When a short magnet is placed with its south pole pointing the geographic north of the earth, the neutral point is formed at a distance 10 cm. At what distance the neutral point will be formed, when the magnet is rotated through 180^0?

$$\left[\text{Take } \frac{1}{3\sqrt{2}} = 0.8(\text{app})\right]$$

32. Calculate the magnetic force per unit length on a conductor carrying a current of 10 A and making an angle of 30° with the direction of a uniform magnetic field of 0.3 T.

33. In a moving coil galvanometer, why do we not use a permanent magnet with flat poles? Explain.

34. A circular coil of area 1.5 m^2 is placed normal to a uniform magnetic field of induction 2 Wb m^{-2}. The field is increased uniformly to 4 Wb m^{-2} in 20 s. Calculate the induced emf.

35. At a given point in a magnetic field is it possible to have different magnetic induction fields? At two different points in a magnetic field, is it possible to have same magnetic induction field? Explain.

36. Why does a solenoid contract when a current passes through it?

37. The current through the coil of inductance 0.15 H increases at a rate of 10 A s^{-1}. Find the induced emf across the coil.

38. Two identical circular coils are placed side by side such that they do not touch each other, in the same plane. An increasing current is flowing through one of the coils in the clockwise direction. Do the two coils attract or repel each other? Explain.

39. A bar magnet of magnetic moment M is cut into 'n' parts along the axial line and 'm' parts perpendicular to the axial line. Find the magnetic moment of each piece.

40. Two infinitely long wires P and Q carry 2 A and 8 A currents respectively. Find at what distance from P the resultant induction field is zero, if the distance between them is 10 cm, and the flow of current in both the conductors is in the same direction.

41. How will you light a bulb rated 100 W, 120 V using 240 V AC supply?

42. Calculate the pole strength of a short bar magnet of length 5 cm which produces a neutral point 15 cm away from its centre on the equatorial line. The earth's horizontal component of magnetic intensity is 50 μT at the given place.

43. A magnetized rod of moment M and length 'L' is bent in the form of arc of circle of radius r. If the arc length is $\frac{1}{4}$th the circumference of the circle then find the magnetic moment of the arc, in terms of M.

44. Explain why two magnetic field lines can never intersect?

45. A cell is connected across AD of a square loop ABCD, made of a conductor of uniform area of cross section, as shown in the figure. Find the ratio of magnetic-fields at the centre produced by
(a) side AB to that of side CD and
(b) side AD to that of side BC.

Concept Application Level—3

46. Lenz's law is in accordance with the law of conservation of energy. Explain.

47. The dipole moment of semicircular magnet is 28 A m^2. It is straightened to form a bar magnet, which is then cut along its axial line to give two magnets. Find the magnetic induction at a point on the equatorial line 0.2 m away from the centre of one of the magnets. If the point is along the axis same distance away from the centre of the magnet then find the magnetic induction.

48. Explain why a current carrying conductor experiences a force when placed in a magnetic field but not when placed in an electric field?

49. Two magnetic poles of strengths 'm_1' and 'm_2' when placed in air at a distance of 'd' from each other, a force 'F' acts in between them. If the pole strength of the poles are doubled, the distance between them is increased by three times and the poles are placed in some other medium, the force between them reduces by 25%. Find the relative magnetic permeability of the medium.

50. A bar magnet is suspended in a uniform magnetic field. Find the ratio of torque when it makes an angle of 30° and 60° respectively with the field.

Very short answer type questions

1. Axial line

$B = \frac{\mu}{4\pi}\frac{2M}{d^3}\, N\, A^{-1}\, m^{-1}$

Direction:- along the axis parallel to $\vec{SN}$ vector

Equatorial line: $B = \frac{\mu_0}{4\pi}\frac{M}{d^3} NA^{-1}m^{-1}$

Direction: Parallel to the axis the magnet and is in the $\vec{NS}$ direction.

2. The property of magnetic material by virtue of which it retains the magnetic property even after removal of the applied magnetic field.

3. Phenomenon of inducing current by changing the magnetic field associated with an electric current containing no source is called electromagnetic induction.

4. M, B, I are represented by thumb, forefinger and middle finger of the right hand.

5. (i) The point where the resultant magnetic field is zero.

   (ii) Lies on the equatorial line axial line of the bar magnet respectively.

6. $\frac{V_p}{V_s} = \frac{I_s}{I_p} = \frac{N_p}{N_s}$

7. The direction of the induced current is such that it opposes the very cause that produces it.

8. Mechanical energy is converted into electrical energy.

9. (i) Space or region around a magnet, in which its effect can be felt.
   (ii) The imaginary line along which a unit north pole moves.

10. (i) magnetic moment acquired per unit volume.

    $I = \frac{m}{v}$

    Unit: $A\ m^{-1}$

    (ii) It is the ratio of intensity of magnetization (I) to the intensity of applied magnetic field H'.

    $\chi = \frac{I}{H}$

    No unit.

    (iii) Property of magnetic substance by virtue of which it retains the magnetic property even after removal of the applied magnetic field.
    No unit.

11. Deflection of the magnetic compass.

12. Produces AC and DC currents respectively.

13. A point beyond which further magnetization is not possible.

14. The distances in which the net magnetic moment of any atom in them is zero are called diamagnetic substances.

    Example air, water, copper, bismuth.

    The substances in which the orientation of electron orbits and axes of their spins are such that the net magnetic moment of any atom is not zero, are called paramagnetic substances.

    Example aluminium, platinum, chromium.

    Like paramagnetic substances, have resultant magnetic moment, but can achieve high degree of magnetic alignment due to a special effect called "Exchange coupling" are called Ferro magnetic substances.

    Example Fe, Co, Ni, Gd and Dy.

15. Due to the influence of earth's magnetic field.

16. (i) Inductance is the property of a coil to induce emf in the same coil or another coil placed near it due to the changes in the current in the coil.

    (ii) emf is produced in a coil due to a change in the current in the same coil.

    (iii) A change in the current in a coil induces an emf in another coil held close to it.

17. The ability of the pole of a magnet to attract or repel another magnetic pole is called pole strength (m).

S.I. unit is A.m.

18. A large number of small local region.

19. The magnetic field is perpendicular to the plane of the coil.

20. Electromagnetic induction.

(i) step-up

(ii) step-down transformers.

21. Magnetic effect of electricity.

22. Each molecule of a magnetic substance is an independent magnet.

23. An insulated copper coil wound around some insulated core such that its length is greater then its diameter.

24. Iron has high magnetic permibility than air.

25. The retencity is high.

26. Flemings left handed rule.

27. (i) Electric energy is converted into sound in loud speaker.

(ii) Sound energy is converted into electrical pulses in microphone.

28. Magnetic flux per unit area is magnetic induction (B).

Unit:- $N A^{-1} m^{-1}$.

Intensity of magnetic field (H) can be defined on the force acting on unit north pole independent of the medium.

Unit: $A\ m^{-1}$

29. Increases.

30. Flemings left handed rule.

Short answer type questions

31. (i) Ferromagnetic substance is divided into a number of domains.

(ii) Magnetic moments of domains cancel each other.

32. (i) $B = \frac{\mu_0}{4\pi}\frac{M}{d^3}$ at neutral points on equational line.

(ii) $B = \frac{\mu_0}{4\pi}\frac{2M}{d^3}$ at axial neutral points.

(iii) At neutral points $B = B_H$ i.e., horizontal component of earth's magnitude field.
$\therefore$ Put $B = B_H$ in above equation.

(iv) Calculate M as B_H at a given place is constant and d distance of material point from magnet can be found.

33. (i) Domain theory used for ferromagnetic substances.

(ii) Define domain.

(iii) On magnetization, size of domains increase.

(iv) On demagnetization, some magnetism is retained.

34. (i) The magnetic field due to solonoid is directly proportional to
the number of tunes per unit length of the solenoid,
the current passing through the solonoid.

(ii) The magnetic field depends upon the nature of material on which the coil is wound.

(iii) The magnetic lines of forces within the solonoid are almost parallel to each other and axis of solenoid.

35. (i) Use $B = \frac{\mu_0 I}{2\pi r}$

Ans: $10\ \mu T$

(ii) $\mu_0 = 4\pi \times 10^{-7}\ H\ m^{-1}$

36. (1) Important magnetic properties i.e.,

(i) Intensity of magnetization

(ii) magnetic susceptibility

(iii) Retentivity

37. Curved, closed, continuous lines.

38. (i) Each molecule is a magnet.
(ii) Molecules are randomly oriented.
(iii) On magnetization, they align in one direction.
(iv) Explain saturation.
(v) Mention drawbacks.

39. (i) use $B = \frac{n\mu_o I}{2r}$
Ans: 0.05 T
(ii) $\mu_O = 4\pi \times 10^{-7}\ H\ m^{-1}$

40. (i) Magnetic needle placed near current carrying conductor is deflected.
(ii) deflection is reversed on reversing the direction of current.
(iii) deflection is increased on increasing the strength of current.
(iv) Thus magnetic field is created around a current carrying conductor.

41. The smallest part belonging to a magnet also possess two poles.

42. (i) Electrons have orbital and spin motion.
(ii) due to movement of electrons, current is produced creating a magnetic field.
(iii) In magnetic substances, net moment of electrons is not zero.
(iv) In non-magnetic substances, net moment of electrons is zero.

43. (i) When a magnet is moved in or out of a stationary coil, galvanometer connected to coil shows deflections, in opposite directions.
(ii) The deflection is also reversed if the polarity of magnet is reversed.
(iii) This current is called induced current and the phenomenon is called electromagnetic induction.

44. $B = \mu_0 H$

45. $\mu = \mu_0 \mu_r$

Essay type questions

46. (i) Mention working principle.
(ii) Find the deflecting couple, restoring couple.
(iii) make use Fleming left hand rule.
(iv) Mention important parts and their use.
(v) The tangent of the angle of deflection (Q) is directly proportional to the strength of the current (i).
$\tan\theta \propto i$
$i = k \tan\theta$
(vi) Mention on which the value of k depends.

47. (1) Mention working principle.
(2) Mention energy conversion, if any
(3) Mention important parts and their use.
(4) Mention working of the device and its uses.

48. (i) Explain origin of earth's magnetism.
(ii) Angle between magnetic and geographic is 17°.
(iii) Explain following elements of earth's magnetism.
(a) Geographic and magnetic meridian.
(b) Angle of drop or inclination.
(c) angle of declination.

49. (i) Define inductance.
(ii) Define self and mutual inductance.
(iii) Write expression for them and their units.

50. (i) Keep bar magnet with as north pole towards geographic north and draw its lines of force.
(ii) Earth's magnetic lines of force are from geographic south to north.
(iii) Locate the two points in the magnetic field where the lines of force due to earth and bar magnet are equal and opposite.
(iv) Repeat same process with south pole of bar magnet towards geographic north.
(v) In 1st case nuetral points are along equatorial line while in 2nd case along axial line of bar magnet.

KEY

Concept Application Level—1

True or false

1. True
2. False
3. False
4. True
5. False
6. False
7. False

Fill in the blanks

8. self-induction
9. henry
10. null point
11. force, magnetic
12. pole strength
13. Magnetic moment
14. the magnetic flux

Match the following

15. A : c
    B : i
    C : e
    D : b
    E : h
    F : a
    G : d
    H : f
    I : g
    J : j

Multiple Choices questions

16. Choice (3)
17. Choice (4)
18. Choice (1)
19. Choice (2)
20. Choice (1)
21. Choice (4)
22. Choice (1)
23. Choice (2)
24. Choice (4)
25. Choice (3)
26. Choice (2)
27. Choice (2)
28. Choice (3)
29. Choice (2)
30. Choice (1)

Concept Application Level—2,3

Key points

31. (i) When south pole faces the geographic north, the null point is formed on the axial line. At null point, the axial field and horizontal component of earth's magnetic field are equal.

$$B_H = B_{axial} = \frac{\mu_o}{4\pi} \times \frac{2M}{d_A^3} \quad \text{-----} \quad (1)$$

Where $d_A = 10$ cm $= 0.1$ m.

When the magnetic is turned through 180°, the null point is formed on the equatorial line.

$$B_H = B_e = \frac{\mu_o}{4\pi} \times \frac{M}{d_e^3} \quad \text{-----} \quad (2)$$

Equating (1) and (2) find d_e

(ii) 8 cm

32. (i) Use $F = B\ell \sin\theta$.

Get the values of B, i and θ from the given data.

Then find $\frac{F}{\ell}$.

(ii) 1.5 N m^{-1}

33. The magnet is made concave in shape so that its magnetic field is always perpendicular to the plane of the coil when it rotates.

34. (i) Get the values of area (A) and initial magnetic induction (B_i). Find magnetic induction (B_f) and time (t) from the given data.

Use $E = \frac{\Delta\varphi}{t}$ and

$\phi = BA$.

Find ϕ_{final}, $\phi_{initial}$ and from them, find $\Delta\phi$.

(ii) 0.15 V

35. (i) Can two magnetic lines of force intersect at a point in a magnetic field?

(ii) Magnetic field can be uniform or non - uniform. In uniform magnetic field it is possible to have same magnetic induction field at two different points.

36. Analyse the type of magnetic poles formed on either sides of two adjacent rings of a solenoid. Are the poles formed like or unlike?

37. (i) The emf induced in the coil is given by

$$E = -L\frac{di}{dt}$$

When L is the self inductance of the coil.

(ii) 1.5 V

38. Recall the principles of mutual induction and Faraday's experiments.

39. (i) When the magnet is cut along axial line, its area of pole decreases. If the area of poles decreases 'n' times, pole strength also decreases 'n' times.

The magnetic moment of each piece is $\frac{M}{n}$

When each piece (obtained by cutting the magnet along the axial line) is divided into 'm' parts by cutting it perpendicular to the axial line, the length and the magnetic moment reduces to $\frac{1}{m}$ th the original value.

(ii) $\frac{M}{mn}$

40. (i) Use $B = \frac{\mu_o i}{2r}$, for both the wires.

If the resultant field at a point is zero, then at the point B due to P it is equal in magnitude and opposite in direction to B due to Q.

(ii) 2 cm

41. Find the resistor of the bulb by using

$$P = \frac{V^2}{R}$$

Use the resistance of the same value as resistance of the bulb and connect in series with the bulb, Another way is to connect a coil of suitable inductance in series with the bulb.

42. (i) At a neutral point we have $B = B_H$. Use the formula $B_{equatorial} = \frac{\mu_o}{4\pi}\frac{M}{d^3}$ and $M = (2l)(m)$.

(ii) 33.75 A m

43. (i) The pole strength of the magnet (m) is

$$m = \frac{M}{L} \quad \text{-----} \quad (1)$$

When the rod is bent in the form of arc, it subtends 90^0 at the centre.

If 'd' is the distance between the poles, then

$$d = \sqrt{r^2 + r^2}$$

$$d = r\sqrt{2} \quad \text{-----} \quad (2)$$

where r is the radius of the arc.

The r is related to ℓ as

$\ell = \pi r$ ----- (3)

Use (1), (2) and (3) and find the new magnetic moment (M_1)

M_1 = pole strength × distance between poles.

(ii) $\frac{2\sqrt{2}}{\pi}M$

44. Recall the definition of a magnetic field line. Is it possible to have two directions for a given magnetic field at a given point in the field?

45. (i) Side AB and side CD carry equal currents. The distance of the centre from both the sides are equal. Therefore, fields produced by both sides are equal at the centre. Side AD and BC carry different currents. If R is the resistance of each side, then current through AD (I_D) is

$$I_D = \frac{3R}{4R} \times V \quad \text{-----} \quad (1)$$

Where V is the terminal PD of the cell.
The current through BC is

$$I_{BC} = \frac{R}{4R} \times V \quad \text{-----} \quad (2)$$

The ratio of currents gives the ratio of fields.

(ii) 1:1, 3:1

46. (i) What happens if the emf is induced in such a way that it support the relative change in magnetic field associated with the coil?

(ii) Work done by the external agent to bring the change in magnetic flux is converted into electrical energy.

47. (i) What is the expression for a magnetic moment of a semicircular magnet?
When the semicircular magnet is straightened, relate the distance between the two poles in the two cases to find the new length of the magnet.
Does the pole strength of the magnet change on straightening?
How does the pole strength change when the straight magnet is cut axially?
What is the magnetic moment of the newly formed (straightened) magnet?

Use $B_{axial} = \frac{\mu_0}{4\pi} \frac{2M}{d^3}$ and $B_{equatorial} \frac{\mu_0}{4\pi} \frac{M}{d^3}$

(ii) B equatorial = 0.0275 mT
B axial = 0.550 mT

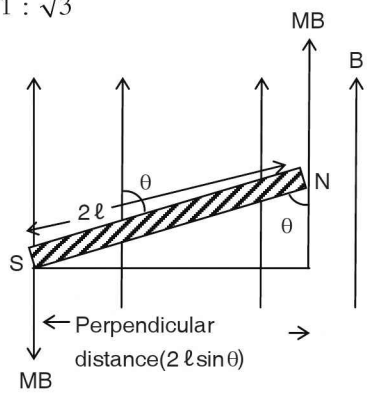
48. How a current carrying conductor behaves? Like a magnet or like a charged body?

49. (i) Use $F = \frac{\mu_0}{4\pi} \frac{m_1 m_2}{d^2}$, in both the cases.
Relate the quantities 'm_1', 'm_2', 'd' and 'F' in the second case with that in the first case using the given data. Then relate μ of the medium with μ_0 and find the value of μ_r of the medium.

(ii) 3

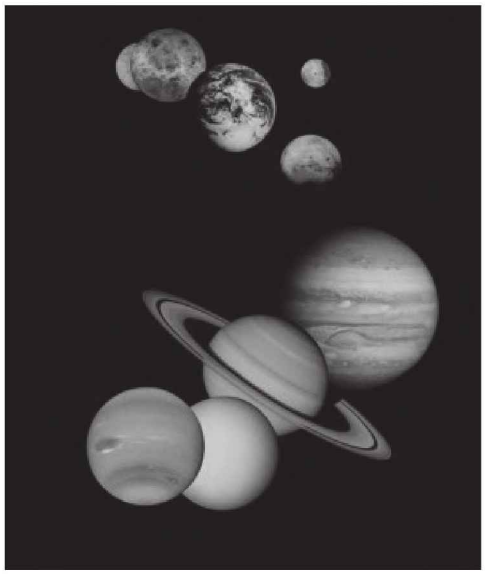
50. (i) The force on the poles of the bar magnet is mB. If the angle made by the bar magnet with the field is θ, then the perpendicular distance between the forces is $2\ell \sin\theta$ where '2ℓ' is the magnetic length, as shown in the figure below. The torque is given by Torque = force X perpendicular distance.
Torque = $mB \times 2\ell \sin\theta$
$\Rightarrow$ Torque $\propto \sin\theta$

(ii) $1 : \sqrt{3}$



9

Modern Physics



INTRODUCTION

Just as the universe at the macro level has intrigued the human mind and intellect for ages, so at the micro level man has been continuously in search of the smallest particle and has discovered a fathomless new world of sub-atomic particles. In the process of all the research having been done in this direction, a whole new branch of physics unfolded leading to the revolutionary discoveries and inventions—radio, transistor, X-ray, photoelectric effect, computers, etc., being only a part of the endless list which opened the doors to a new modern world. This branch of physics came to be referred to as modern physics. Nuclear physics, electronics, etc., are all off-shoots of the developments in modern physics.

Cathode rays

Gases do not conduct electricity at ordinary pressures. However, if very high P.D. (more than 30 kV) is applied between two electrodes in the presence of air, then electric charges pass through air. As electricity passes, intense light in the form of spark is produced. The flash of lightning during rainy days is an example of conduction of electricity through gases at normal atmospheric pressure.

When the gas pressure is lowered, it is found that gases conduct electricity at sufficiently high potential which is less than the potential required at ordinary pressure. The conduction of electricity through gases is also called discharge of electricity or simply electric discharge. During electric discharge, different gases emit light of different colours. To study the electric discharge phenomenon at low pressures, a given gas is enclosed in a hard glass tube called discharge tube, as shown in the figure (9.1).

Usually, the discharge tube is 30 cm to 40 cm long and 2 cm to 3 cm in diameter. Two aluminium dishes, called electrodes are sealed in the tube on its ends. A high potential difference can be applied

across the two electrodes.The electrode connected to positive terminal (high potential) of the DC source is called anode (A) and the one connected to negative terminal (low potential) is called cathode (C). To reduce the pressure inside the discharge tube, it is provided with a side tube which is connected to the vacuum pump.

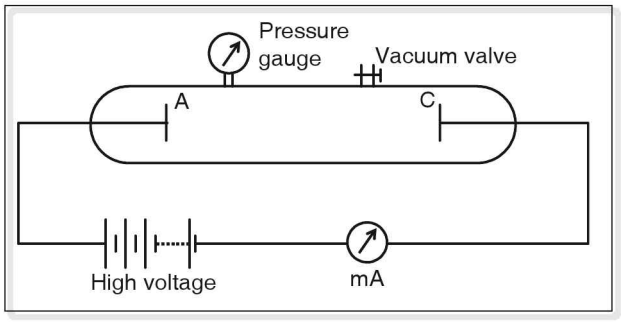


Figure 9.1 Discharge tube

To study the electric discharge, the pressure of the gas inside the tube is gradually reduced. The electric discharge passes through various stages, as the gas pressure is gradually reduced. Finally, at a pressure of about 0.01 mm of Hg, the tube becomes dark. The wall of the tube opposite to the cathode glows with greenish fluorescence which is due to the invisible radiation given out from the cathode surface called cathode rays.

Properties of cathode rays

1. Cathode rays are always emitted along the normal to the surface of cathode, irrespective of the position of the anode.
2. Cathode rays travel in straight line path.
3. Cathode rays exert mechanical force on the objects placed in their path.
4. They are deflected in electric and magnetic fields.
5. They ionize the gas through which they pass.
6. Cathode rays produce fluorescence in many substances.
7. They can pass through thin metal foils.
8. They affect photographic plates.
9. When cathode rays are stopped by certain objects like metals, X-rays are produced.

Nature of cathode rays

Cathode rays are deflected in electric and magnetic fields. This shows that they consist of charged particles. From the direction of deflection in electric and magnetic field, it is concluded that they are negatively charged particles. In 1897, J.J. Thomson measured e/m ratio of those particles and proved that they are the common constituents of all matter. G.J Stroney named these particles electrons.

Dalton, in his atomic theory, proposed that the atom is the smallest indivisible particle of matter. However, discovery of electrons proved that atom itself is made up of still smaller particles and that the atom is divisible.

After the determination of e/m of electrons, the charge on it was discovered to be -1.6×10^{-19} C. Using $\frac{e}{m}$ ratio and charge, the mass of an electron was found to be 9.11×10^{-31} kg.

Uses of cathode rays

1. X-rays are produced when fast moving cathode rays are stopped suddenly.
2. The screen of a cathode ray oscilloscope, a TV picture tube, the monitor of a computer, etc. are coated with fluorescent substances. When the cathode rays fall on the screen, pictures are visible on the screen.

X-rays

X rays were discovered by Wilhelm Roentgen of Germany. While studying electric discharge phenomena, he noticed that a screen (coated with barium salt) placed near a cathode ray tube was glowing whenever the tube was working. He further observed that the screen glowed even after the tube was covered with a black paper. On further investigation, he found that the photographic plates wrapped in a black paper were fogged when kept near the tube. When he placed his hand between the tube and photographic plate, he obtained a photograph of the bones. Roentgen concluded that some unknown, highly penetrating radiations were responsible for his observations. He named these unknown radiations as X-rays.

Production of X-rays

X rays are produced when fast moving cathode rays are stopped suddenly by a metal called target.

The modern device used in the production of X-ray is called the Coolidge tube, which consists of a highly evacuated glass tube as shown in the figure (9.2).

A filament F emits electrons when it is heated by passing a low current through it. It also acts as a cathode. It is surrounded by a molybdenum cylinder M. The cylinder M focusses the electrons on a fine beam. The target is tungsten which is embedded in copper anode A. The anode A is inclined at 45° to the horizontal. It is cooled by circulating water around it. To produce X rays, high speed electrons are required. For this, the electrons emitted from the filament are accelerated by applying a high potential difference between the anode and the cathode. The fast moving electrons produce X-rays when they strike the target.

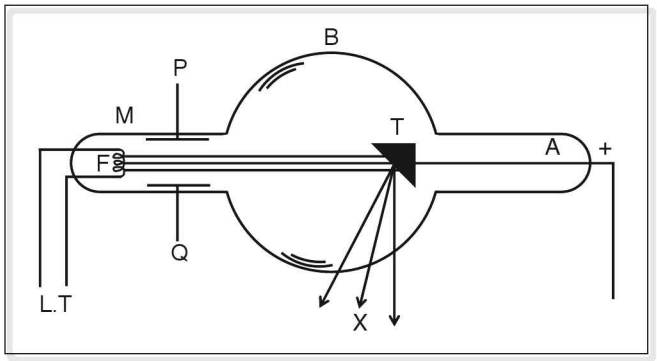


Figure 9.2 Coolidge tube

Properties of X-rays

1. X-rays are electromagnetic radiations. Their wavelength is very short, of the order of 1 A°. They are invisible.
2. X-rays, being electromagnetic radiations, travel in a straight line with a velocity of light (3×10^8 m s^{-1}).

3. They affect photographic plates.
4. They produce fluorescence in certain materials.
5. They are not affected by electric and magnetic fields.
6. X-rays have a high penetrating power. However, they are stopped by bones, dead cells, etc.
7. They ionize the gases through which they pass.
8. They bring about chemical and biological changes in the substances and living beings through which they pass.
9. They are diffracted by crystals.

Uses of X-rays

X-rays are used extensively in different fields.

1. Medicine: In medicine, X-rays are employed to detect fracture and dislocation of the bones, presence of foreign bodies and early detection of certain diseases like tuberculosis.
2. Industry and engineering: X-rays are used to detect flaws in machinery parts, metal castings, metal structures, etc.
3. X-rays are used to detect goods hidden in wooden boxes and leather cases. They are also used in testing.
4. Radiotherapy: X-rays are used to destroy malignant tumours without surgery, especially when malignant tumours are situated deep inside. The tumours are exposed to hard X-rays which possess high penetrating power.
5. X-rays are used to study the arrangement of atoms and molecules in crystals and other matter.

Radioactivity

Natural radioactivity

The phenomenon of radioactivity was first discovered by A.H.Becquerel and further investigations were made by Madame Marie Curie and Prof. Pierre Curie.

In 1896 Henri Becquerel, the French scientist, quite accidentally discovered radioactivity while studying the phenomenon of fluorescence. He found that the photographic plates get fogged when placed near pitchblende, an ore of uranium. It was discovered that certain substances like uranium emit some rays which were initially referred to as 'Becquerel rays'. Research work carried out by the Curies led to the discovery that these rays are emitted due to a phenomenon called radioactivity which is a nuclear phenomenon of self-disintegration.

Thus, radioactivity is the phenomenon of self-disintegration and spontaneous emission of radiation by some unstable nuclei with atomic number greater than 82. This type of spontaneous radiation is referred to as 'natural radioactivity'. Substances which give such spontaneous emission of radiations are known as radioactive substances.

α, β and γ– radiations

The radiations of radioactive substances were found to consist of three distinct constituents_ alpha, beta and gamma radiations. When the radiations were allowed to pass through an electric field, the constituent which deflected towards the negative plate was called the α-rays.The constituent which deflected towards the positively charged plate was called the β- rays. The constituent which emerged undeflected was called γ-radiations.

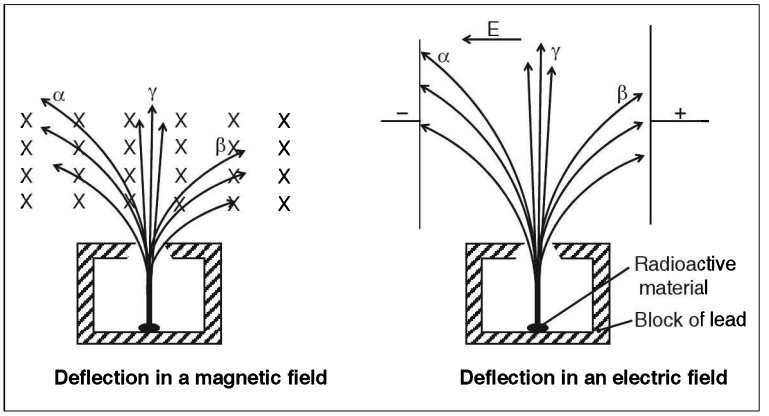


Figure 9.3

Properties of α, β and γ radiations

1. Nature	Positively charged	Negatively charged	Electrically neutral
2. Constituents	He^{++} (2 protons and 2 neutrons)	High energy electrons	Electromagnetic radiations
3. Speed	Nearly 10^7 m s^{-1}	About 2.7×10^8 m s^{-1}	3×10^8m s^{-1}
4. Rest mass	4 times the mass of a proton, i.e., 6.64×10^{-27} kg	Equal to the mass of an electron i.e., 9.1×10^{-31}kg	No mass
5. Charge	Positive charge $= +3.2 \times 10^{-19}$C	Negative charge $= -1.6 \times 10^{-19}$C	No charge
6. Specific charge (q/m)	4.83×10^7 C kg^{-1}	1.76×10^{11}C kg^{-1}	-
7. Wavelength	-	-	10^{-13}m (less than the wavelength of X-rays)
8. Effect of electric field	Deflects towards the negative plate	Deflects towards the positive plate, deflection more than that for α rays	No effect
9. Effect of magnetic field	Less deflected.	More deflected, direction opposite to that of α-rays	No effect
10. Ionizing power	Intense ionization (10000 times that of γ-rays)	High ionization power (100 times that of γ rays)	Less
11. Penetrating power	Small (3-8 cm in air)	Large (up to few metres in air)	Very large (up to a few hundred metres in air)

(Continued on following page)

12. Stopping substance	Thin paper, human skin	About 1 mm of lead or about 3 mm of aluminium	About 30 cm of iron or few metres of concrete
13. Effect on photographic plate	-	Affect photographic plates and cause scintillations on a fluorescent screen	Affect photographic plates and cause scintillations on a fluorescent screen
14. Biological damage	Cause some damage	Can cause radiation damage	Passes through human body and destroy human cells. Used to kill cancer cells

Radioactive transformation

The disintegration of an unstable nucleus resulting in the formation of another nucleus is called radioactive transformation or radioactive decay.

Types of transformation

1. α-decay or α-emission

"When an α-particle is emitted from a radioactive nucleus, the atomic number (Z) decreases by 2 and the mass number (A) decreases by 4.

$$_{Z}P^{A} \rightarrow {}_{Z-2}D^{A-4} + {}_{2}He^{4} \ (\alpha - \text{Particle})$$

'P' is called the parent nucleus.
'D' is called the daughter nucleus.

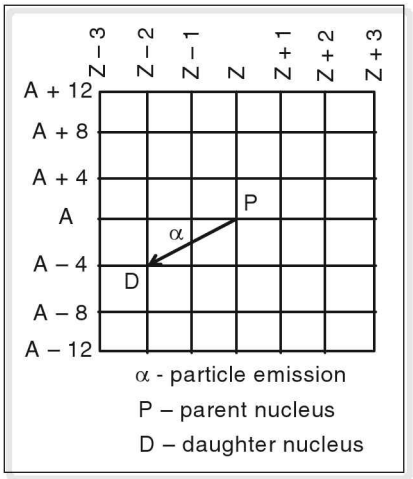


Figure 9.4

☛ *Example* $_{92}U^{238} \rightarrow {}_{90}Th^{234} + {}_{2}He^{4}$

Here, uranium is the parent nucleus and thorium is the daughter nucleus.

This is summarized as Rutherford and Soddy's law of α-emission, which states that "whenever an α- particle is emitted from a radioactive nucleus, the atomic number decreases by 2, the mass number decreases by 4 and the position of the element shifts by two places to the left in the periodic table".

2. β-decay or β-emission

The presence of excess neutrons in an unstable nucleus lead to a neutron changing into a proton by emitting an electron (for charge conservation).

$$_{0}n^{1} \rightarrow {}_{1}P^{1} + {}_{-1}e^{0}$$

(neutron) (proton) (β-particle)

In general, ${}_{Z}P^{A} \rightarrow {}_{Z+1}D^{A} + {}_{-1}\beta^{0}$

(parent nucleus) (daughter nucleus) (β-particle)

☛ *Example*

${}_{90}Th^{234} \rightarrow {}_{91}Pa^{234} + {}_{-1}\beta^{0}$

(thorium) (protactinium) (β-particle)

Rutherford and Soddy's law of β-decay summarizes the above as "Whenever a β-particle is emitted from a radioactive nucleus, the atomic number increases by 1 whereas the mass number remains the same and the position of the element shifts one place to the right in the periodic table".

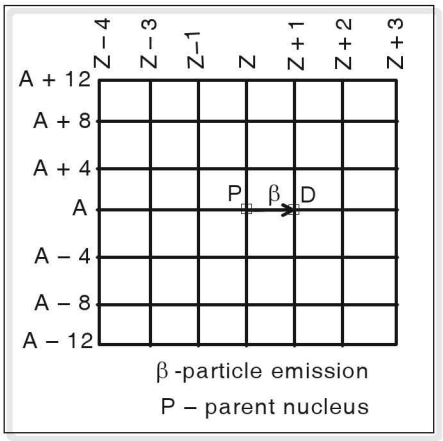


Figure 9.5

3. γ-decay or γ-emission

There is no change in the atomic number (Z) or in the mass number (A) of the atom which undergoes gamma decay. However, the emission of γ-rays results in a change in the energy state of the nucleus.

In a single radioactive decay, both α-decay and β-decay never take place together. There will be either an α-decay or a β-decay, which may be accompanied by a γ–emission.

Radioactive series

An unstable parent nucleus undergoes disintegration several times producing daughter nuclei, which would also be unstable. The radioactive daughter nucleus would undergo further disintegration giving rise to a new daughter nucleus. This sequential decay of a given radioactive nucleus resulting in the production of new nuclei at each stage and giving rise to a series of elements is called radioactive decay series.

There are three radioactive series depending on the starting element naturally available on the earth.

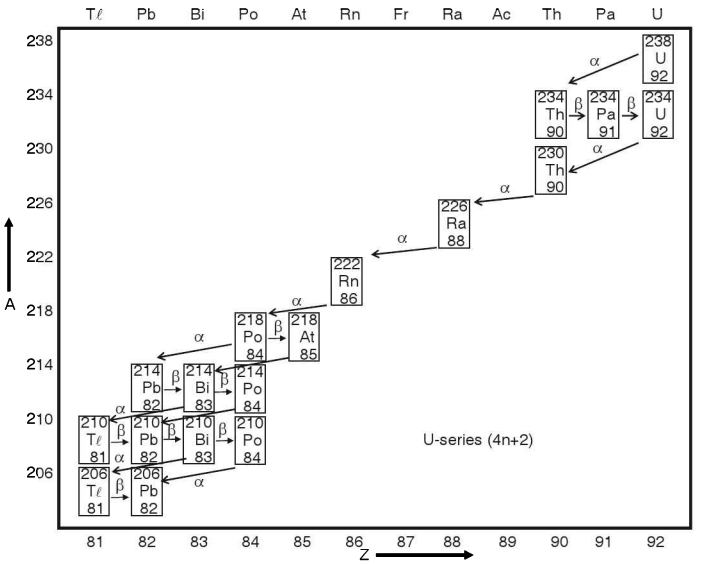


Figure 9.6 Uranium Series

They are

(a) the uranium series (${}_{92}U^{238}$ series),

(b) the actino-uranium series or Actinium series (Ac-U series) and

(c) the thorium series (${}_{90}Th^{232}$ series).

The important features of the radioactive series are as follows:

1. The stable end-product of all the three series is an isotope of lead having an atomic number 82, but the mass number is 206 in U-series, 207 in Ac U-series and 208 in Th-series.
2. In each of the series, parent–daughter sequences exist.
3. More than one kind of decay is possible for an intermediate radioactive species.
4. In each of the three series, the element Radon (Rn) with atomic number 86 is present.
5. In all the three series, a change in the mass number A, if any (due to α-decay), is always equal to 4.
6. The thorium series is referred to as 4n series, since the mass numbers of all its members are multiples of 4.
7. The uranium series is referred to as (4n+2) series, since (A–2) is always a multiple of 4 for each one of the elements of the series.
8. The actinium series is identified as (4n+3) series, since (A – 3) is always a multiple of 4 for each one of the elements of the series.
9. The neptunium series ($_{93}Np^{237}$ series) is obtained from artificially produced radioactive material. This series is also known as (4n + 1) series.

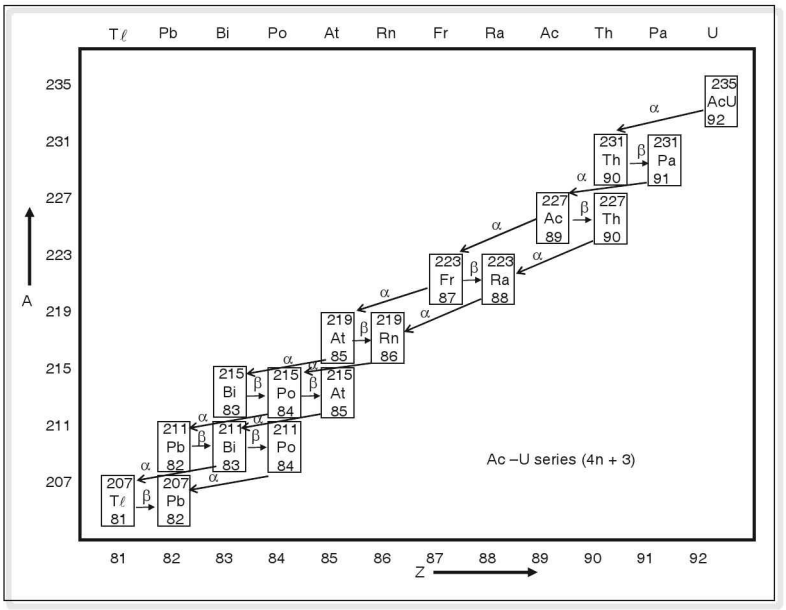


Figure 9.7 Actino–Uranium Series

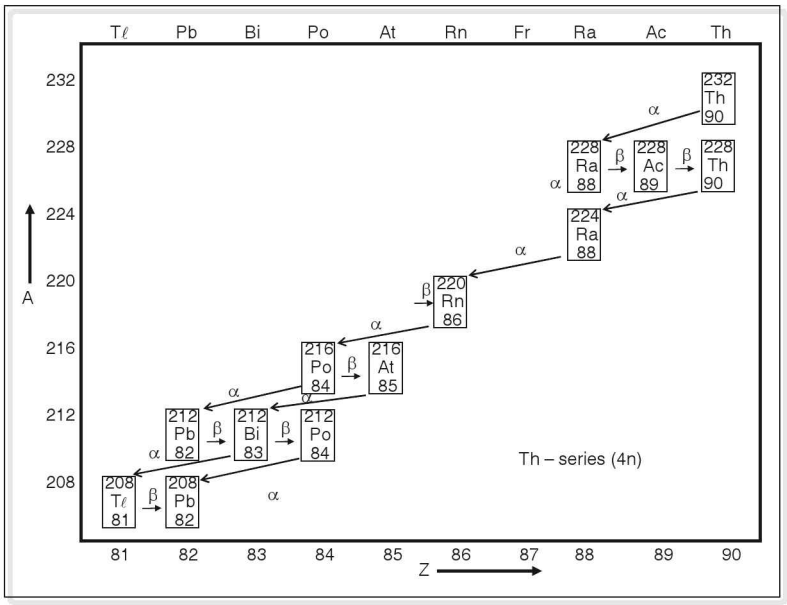


Figure 9.8 Thorium Series

It is found that during radioactive transformation due to either α-emission or β-emission, not all the nuclei undergo disintegration simultaneously. Also the rate at which the decay takes place depends on the number of parent atoms available at a given instant.

This is summarized in the law of radioactive disintegration.

Law of Radioactive Disintegration

The law of radioactive disintegration states that "in any radioactive substance, the number of atoms disintegrating per second is directly proportional to the number of atoms present".

$$\frac{\Delta N}{\Delta t} \propto N$$

$$\frac{\Delta N}{\Delta t} = -\lambda N$$

where λ, the proportionality constant, is called the decay constant. The negative sign indicates that the number of atoms decreases with time.

Using calculus, we can show that $N = N_0 e^{-\lambda t}$ where N_o and N are the number of atoms of the substance at instants when time is zero and 't' s respectively and 'e' is an irrational number.

The half-life period

"The half-life period (T) of a radioactive substance is defined as the time during which half the amount of the substance disintegrates."

The half-life period (T) is different for different substances. It is impossible to say exactly when a particular atom of a radioactive element would disintegrate. The fraction of the substance decaying during a given period will be the same as that during any other equal period. Smaller value of T (half-life period) indicates faster decay. Decay constant (λ) of a radioactive substance is defined as $\lambda = \frac{0.693}{T}$.

The higher value of λ indicates faster decay. Thus, the decay constant is a characteristic property of a radioactive element.

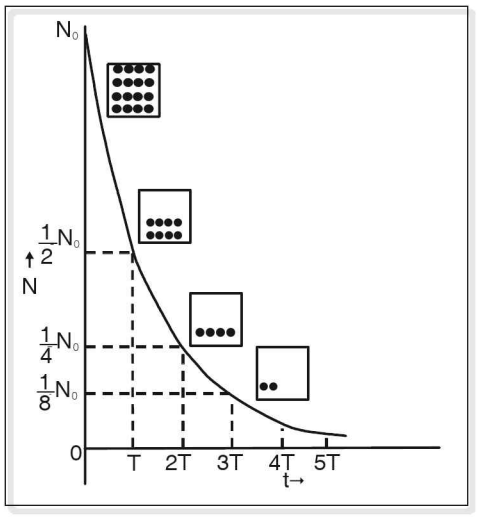


Figure 9.9

☛ *Example*

If $\frac{7}{8}$ th of the initial mass of the radioactive substance decays in 15 days, then what is the half-life period of the radioactive substance?

Solution

Let the initial number of radioactive nuclei be N_0.

Then, after a half-life period, the number of nuclei present would be $N_0 \times \frac{1}{2}$.

After two half-lives, the number of nuclei present would be

$$\frac{1}{2} \times \left(N_0 \times \left(\frac{1}{2} \right) \right) = N_0 \left(\frac{1}{2} \right)^2.$$

After 'n' half-lives, the number of nuclei present would be $N = N_0\left(\frac{1}{2}\right)^n$ ....(1)

If 't' is the given time period and T is the half-life period of the radioactive substance, then

$n = \frac{t}{T}$.....(2)

Also, if 'm_o' is the initial mass and 'm' the mass of undecayed nuclei of the substance, then $\frac{N}{N_0} = \frac{m}{m_o}$.....(3)

∴ Equation (1) can be written as

$m = m_0\left(\frac{1}{2}\right)^n$ .....(4)

In the given problem, undecayed mass $m = m_0 - \frac{7}{8}m_0 = \frac{1}{8}m_0$

$t = 15$ days, $T = ?$

Substituting the values of 'm' and 'm_0' in equation (4), we get

$\frac{1}{8}m_0 = m_0\left(\frac{1}{2}\right)^{\frac{15}{T}} \quad \Rightarrow \frac{1}{8} = \left(\frac{1}{2}\right)^{\frac{15}{T}}$

$\Rightarrow \frac{1}{8} = \left(\frac{1}{2}\right)^{\frac{15}{T}} \qquad \Rightarrow \frac{15}{T} = 3$

$\Rightarrow T = \frac{15}{3}$ days $= 5$ days

☛ *Example*

20 g of radioactive substance decays to 10 g in 2 days. How much of the substance will remain undecayed after 10 days?

Solution

As 20 g decays to 10 g, i.e., the initial mass reduces to half in 2 days, the half life period must be 2 days. By using the formula,

$m = m_0\left(\frac{1}{2}\right)^n$ and substituting $m_0 = 20$ g, $n = \frac{t}{T} = \frac{10}{2} = 5$, we get

$m = 20\left(\frac{1}{2}\right)^5 \Rightarrow m = \frac{20}{32}$

$\Rightarrow m = 0.625$ g.

Isotopes, isobars and isotones

The atoms of some elements exist in different forms depending on the variations in their atomic numbers and mass numbers and are referred to as isotopes, isobars or isotones.

Isotopes

"Isotopes are the atoms of the same element having the same atomic number, but different mass number due to difference in the number of neutrons in their nuclei".

☛ *Examples of isotopes*

(i) $_1H^1$, $_1H^2$, $_1H^3$	isotopes of hydrogen
(ii) $_{17}Cl^{35}$, $_{17}Cl^{37}$	isotopes of chlorine
(iii) $_8O^{16}$, $_8O^{17}$, $_8O^{18}$	isotopes of oxygen
(iv) $_{92}U^{235}$, $_{92}U^{238}$	isotopes of uranium
(v) $_{19}K^{39}$, $_{19}K^{40}$	isotopes of potassium

Isobars

"Atoms of different elements, having the same mass number but different atomic number are called isobars".

☛ *Examples of isobars*

(i) $_6C^{13}$, $_7N^{13}$

(ii) $_{19}K^{40}$, $_{20}Ca^{40}$

(iii) $_{18}Ar^{40}$, $_{20}Ca^{40}$

(iv) $_{12}Mg^{24}$, $_{11}Na^{24}$

Isotones

"Atoms with their nuclei having the same number of neutrons but different number of protons are called isotones".

☛ *Examples of isotones*

The atomic number and the mass number of $_{14}Si^{31}$ are 14 and 31 respectively, which give the number of neutrons as $31 - 14 = 17$. The atomic number and the mass number of $_{15}P^{32}$ are 15 and 32 respectively, which give the number of neutrons as $32 - 15 = 17$.

Thus, $_{14}Si^{31}$ and $_{15}P^{32}$ are isotones.

Artificial transmutation

Unstability in the nucleus leads to a natural and spontaneous radioactive disintegration only in certain heavy elements. However, in 1919, Rutherford succeeded in producing nuclear disintegration even in nuclei of lighter elements by bombarding them with high energy α-particles.

Definition

"Artificial transmutation is the process of transformation of one element into another by artificial means of bombarding it with high energy particles".

$$_2He^4 + {}_7N^{14} \rightarrow {}_8O^{17} + {}_1H^1$$

Incident α-particle — nitrogen (target) — oxygen — proton

Such artificial transmutation can be achieved by employing α-particles, protons or other particles as the bombarding particles. These particles, which are used for the bombardment of target nuclei, are referred to as 'projectiles'.

Artificial transmutation by α-particles

Artificial transmutation of fluorine and aluminum is possible by using the α-particles as projectiles.

$$_2He^4 + {}_9F^{19} \rightarrow {}_{10}Ne^{22} + {}_1H^1$$

$$_2He^4 + {}_{13}Al^{27} \rightarrow {}_{14}Si^{30} + {}_1H^1$$

Artificial transmutation by protons

$$_1H^1 + {}_3Li^7 \rightarrow {}_2He^4 + {}_2He^4$$

$$_1H^1 + {}_{11}Na^{23} \rightarrow {}_{10}Ne^{20} + {}_2He^4$$

In these transmutations, a proton (nucleus of hydrogen atom) is used as a projectile to transmute the lighter elements lithium and sodium. In this type of artificial transmutation which was conducted by Cockroft and Walton in 1932, the α-particles are obtained as one of the products.

In some of the artificial transmutations, the product nucleus is found to be radioactive. It undergoes further disintegration by emitting particles like electrons and protons. It was first discovered in 1934 by Madam Irene Curie Joliot and M. Frederic Joliot. This process came to be referred to as "artificial radioactivity."

Artificial or induced radioactivity is a nuclear process in which a stable element, which is not radioactive, is converted into another element, which is radioactive in nature.

The Curies found that when the atoms of stable aluminium were bombarded by α-particles, an unstable isotope of phosphorous, $_{15}P^{30}$ was produced which, being radioactive, disintegrated producing stable silicon ($_{14}Si^{30}$) and a particle whose mass was the same as that of an electron, but with a positive charge. This particle was referred to as a 'positron'.

$$_{13}Al^{27} + {}_2He^4 \longrightarrow {}_{15}P^{30} + {}_0n^1 + \text{energy}$$

$$_{15}P^{30} \longrightarrow {}_{14}Si^{30} + {}_1e^0 \text{ (positron)}$$

Thus, we see that while β-particles are emitted in natural radioactivity, positrons are emitted in the bombardment of stable aluminum using α-particles as projectiles.

Instead of α-particles, energized protons, neutrons or other heavy elements can also be used to induce radioactivity in stable atoms.

These isotopes of elements which are obtained in artificial transmutation and are radioactive in nature are called 'radio-isotopes'.

About 500 radioactive isotopes have been artificially produced by the bombardment of stable elements with α- particles.

Uses of radio-isotopes

Some of the applications of radio-isotopes are as follows:

1. ${}_{11}Na^{24}$ is used to detect a clot in the blood in the human body.
2. ${}_{27}Co^{60}$ is used to kill cancer cells since it is an intense source of γ-radiations.
3. ${}_{53}I^{131}$ is used to test the functioning of the thyroid gland.
4. Radio-isotopes like ${}_{27}Co^{60}$ are used to detect the internal flaws in a cast-material.
5. Radio-isotopes are also used in agriculture in the selection of fertilizers.
6. Unnecessary mutations of plants can be prevented by irradiating seeds using radio-isotopes.
7. Perishable food items can be kept fresh for longer periods by exposing them to mild γ radiations from a radio-isotope.
8. Radioactive gases are injected into the atmosphere and by suitable detectors, the course of monsoon winds and the changes they undergo can be studied.

Carbon dating or radioactive dating

Carbon in the atmosphere is found to exist as three isotopes ${}_6C^{12}$, ${}_6C^{13}$ and ${}_6C^{14}$. ${}_6C^{14}$ (or C-14) is an unstable isotope with half-life period of 5730 years, and is continuously produced in the atmosphere by the action of cosmic rays on atmospheric nitrogen.

$$\underset{\text{atmospheric nitrogen}}{{}_7N^{14}} + \underset{\text{neutron of cosmic radiation}}{{}_0n^{1}} \rightarrow \underset{\text{Radioactive carbon}}{{}_6C^{14}} + \underset{\text{Proton}}{{}_1H^{1}}$$

This C-14 is continuously absorbed by the cells of living beings. In living beings, the ratio C-14:C-12 is constant. In dead plants or animals, C-14 undergoes radioactive decay without replenishment. By determining the ratio C–14/C-12 in fossils, their age is determined. This process of estimating the age by measuring the disintegration of C-14 is known as 'Carbon dating'.

For estimating the age of earth or rocks, U-235 isotope is used. This is done by determining the present ratio of U-235 to stable lead isotope.

The ill effects of radioactive radiations

The α, β and γ rays emitted during radioactive disintegration of an unstable substance cause several biological effects (kill the living cells and cause radiation burns) on human beings.

The biological effects are of the following types:

(i) Short term recoverable effects,

(ii) Long term irrecoverable effects and

(iii) Genetic effects

The first two types are limited to the affected person, whereas the third type affects the later generations.

Safety precautions for the people regularly working with radio-active substances:

1. Special lead-lined aprons, gloves, etc. should be worn and long tongs made of lead should be used while handling radioactive substances.
2. Special film badges should be worn to determine periodically the amount of radiation to which a person is exposed, and to ascertain whether any particular person is exposed to more than the permissible levels.
3. The radio-active substances should be handled with the utmost care and stored in lead containers with narrow openings to stop radiations in all directions.

Background radiations

Background radiations are those radio-active radiations that everyone is exposed to due to the presence of natural radio-active substances on the earth as well as the cosmic radiations, and also the radio-active substances present in the body.

Internal background radiations are from the radioactive substances K-40, and C-14 present in the entire body and radium present in the bones. External background radiations are from the cosmic rays and radiations from the radioactive rocks on the earth.

Geiger counter

Geiger counter is an instrument used for recording the rate of emission from a radioactive substance.

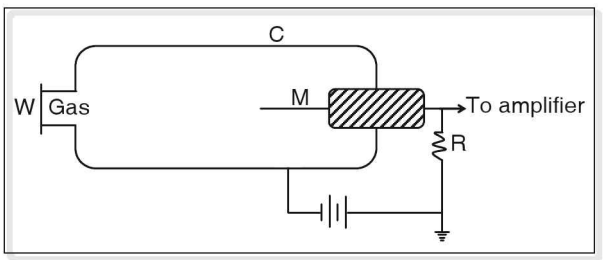


Figure 9.10 Geiger counter

It consists of a metallic cylinder 'C' and a metallic wire 'M' fixed along the axis of the cylinder. The cylinder is filled with argon gas with traces of chlorine and the pressure inside it is maintained between 50 mm of Hg and 100 mm of Hg. A potential difference that is slightly less than the necessary potential to produce discharge in the gas is maintained between the wire and the cylinder. α-particles from a given source enter the cylinder through the thin mica window 'W' and ionize the gas molecules in the cylinder. This leads to a rapid build-up of current and a momentary change of potential across 'R'. This electric surge is amplified and given to an electronic counter.

Nuclear energy

The experiments and the studies conducted by Rutherford in the early twentieth century, in understanding the atomic structure led to further research and the discovery of the process of artificial transmutation. This along with the discovery of neutron in 1932 by Chadwick gave a new direction to the research

being carried out by the physicists in Europe. In 1939, the German scientists Otto Hahn and Fritz Strassmann conducted several experiments and showed that heavy uranium atoms could be split into lighter atoms of smaller atomic number. But what excited the scientists most was the discovery that these reactions resulted in a substantial difference in the masses of the products and reactants. And when Einstein proposed the mass-energy equivalence with his famous $E = mc^2$ equation, the foundation for the fission and fusion reactions was laid down, shifting the focus to the (nuclear) energy released in these nuclear reactions.

Before we venture into the study of the nuclear reactions in detail, we need to understand some basic concepts like mass defect, binding energy, etc.

Atomic mass

The mass of an atom is measured in units called 'atomic mass unit'(u); defined as 1/12th of the mass of a neutral carbon atom ${}_6C^{12}$ (Carbon – 12 atom).

$1\ u = 1.6605402 \times 10^{-27}$ kg.

Thus, the mass of a carbon atom is 12 u.

☛ *Example*

Estimation of mass of helium nucleus.

Consider a helium atom ${}_2He^4$.
It has 2 electrons, 2 protons and 2 neutrons
The mass of a proton = 1.0078 u
The mass of a neutron = 1.0087 u
$\therefore$ The total mass of the nucleus $= 2 \times 1.0078 + 2 \times 1.0087$
$= 4.0330$ u

But the actual mass of a helium nucleus determined experimentally is 4.0026 u.
Then, what happened to the remaining mass of 0.0304 u?

Mass defect

The difference between the sum of the individual masses of constituents in a nucleus and the mass of the nucleus itself is called the 'mass defect'.

In the above example, the mass defect is found to be 0.0304 u.

Einstein's mass energy equivalence

According to Einstein, mass and energy are not independent, but are mutually convertible. Any body which changes its energy content 'E' undergoes a change in its mass, 'Δm' so that the ratio $\frac{E}{Dm}$ is always constant, i.e. if the energy of a particle is increased, its mass also increases.

Thus, if m_0 is the 'rest mass' (mass of the particle at rest), then the 'moving mass' m of the particle having velocity 'v' is given by

$$m = \frac{m_0}{\sqrt{1 - \frac{v^2}{c^2}}}.$$

Einstein's mass-energy equivalence is given as $E = (\Delta m)c^2$, where c is the velocity of light in vacuum. Thus, the mass of 1 u = 931.5×10^6 eV = 931.5 MeV. And 1 kg of mass, if completely changed into energy, would be equivalent to 9×10^{10} MJ.

What had puzzled the scientists the most was how the protons having positive charge are packed together within a distance of less than 10^{-15} m without being repelled due to the electrostatic forces. Since the gravitational forces of attraction between the protons is far less than the electrostatic forces of repulsion between them, there must be some other stronger forces binding the protons together.

Similarly, the role of the neutrons was also puzzling. It was thought that a neutron is formed when the electron of a hydrogen atom collapses into the nucleus under high and extreme pressure, bringing the electron very close to the proton and holding it there, by some form of nuclear energy. Thus a neutron (a collapsed hydrogen atom) is electrically neutral.

It must be noted that these nuclear forces are much stronger than the electrostatic or electromagnetic forces. Thus, the nucleons – protons and neutrons are bound together by the nuclear energy arising out of the mass defect and the mass-energy equivalence.

Thus, here we define two quantities, binding energy and binding energy per nucleon, which play an important role in the stability of a nucleus.

Binding energy

The nucleons of a nucleus are bound together by the energy that results from the loss in the total mass of the nucleons when they are held in the nucleus, i.e. the mass defect; this energy is called the binding energy and is measured in MeV.

Binding energy of a nucleus = $\Delta m \times 931.5$ MeV,

where Δm is the mass defect of the nucleus in atomic mass units 'u'.

Thus, for the nucleus of a helium atom, the binding energy = 0.0304 u × 931.5 MeV
= 28.32 MeV

Binding energy per nucleon

It is the average energy required to extract one nucleon from the nucleus.

$$\text{Binding energy per nucleon} = \frac{\text{binding energy}}{A}$$

Binding energy per nucleon is the measure of the stability of the atom. The higher is the binding energy per nucleon, the greater is the stability of the atom.

☛ *Example*

What is the energy required to increase the mass of a system by one atomic mass unit?

Solution

$\Delta m = 1$ u
$E = 1u \times c^2 = 931.5$ MeV

☛ *Example*

What is the mass defect and binding energy of $_{27}Co^{59}$ which has a nucleus of mass of 58.033 u? ($m_p = 1.0078$ u, $m_n = 1.0087$ u)

Solution

In $_{27}Co^{59}$ number of protons = 27
Number of neutrons = 59 – 27 = 32
∴ The total mass of the nucleus = 27 × 1.0078 u + 32 × 1.0087 u
= 27. 2106 u + 32.2784 u = 59.489 u

Given, The actual mass of the nucleus = 58.933 u
∴ The mass defect = 59.489 u – 58.933 u
= 0.556 u
∴ The binding energy = $\Delta m \times 931.5$ MeV = 0.556 u × 931.5 MeV = 517.914 MeV

Nuclear reactions

Any reaction that tends to change the configuration of the nucleus of an atom can be termed as a nuclear reaction.

There are two types of nuclear reactions. They are

(i) nuclear fission reactions and
(ii) nuclear fusion reactions.

The energy released in a nuclear reaction is generally expressed in electron volt (eV).

1 eV = 1.602×10^{-19} J.

Fission reactions

As already studied in natural radioactivity, U^{235} and U^{238} are highly unstable and emit powerful radiations. In this process, uranium isotope ultimately ends up as 'stable lead', releasing energy of 1 MeV to 5 MeV.

Thus, if it is possible to break an unstable nucleus into two or more smaller ones, this would lead to the release of a great amount of energy. This is the principle of a nuclear fission reaction.

"Nuclear fission is a process in which the heavy nucleus of a radioactive substance like uranium is split into lighter nuclei by the bombardment of a low energy (slow moving) neutron, the reaction being accompanied by the release of energy and two or three or more neutrons".

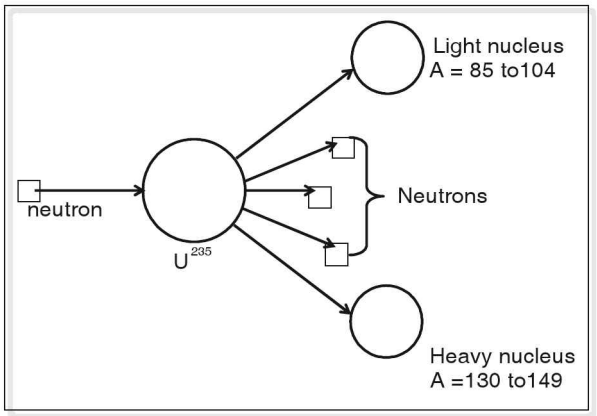


Figure 9.11 Fission reaction

The low energy neutrons used for the bombardment of the nuclei of heavier atoms of uranium are called 'thermal neutrons'.

The salient features of a fission reaction are as follows:

(a) A fission reaction takes place when the nucleus of a heavy atom such as U^{235} is bombarded with a slow moving, low energy thermal neutron.

(b) Fast moving neutrons cannot produce a fission.

(c) The products of a fission reaction are generally a light nucleus, a heavy nucleus and 2 or 3 neutrons. The atomic mass number of the lighter nucleus is in the range of 85 to 104, whereas the atomic mass number of the heavier nucleus is between 130 and 149.

(d) On an average, about 2.5 neutrons are produced in one fission reaction.

(e) There are three types of fission reactions– spontaneous, prompt and delayed reactions. In spontaneous fission reactions, fission takes place continuously and spontaneously in certain (radioactive) substances without bombardment by any projectile. In a prompt reaction, fission takes place as soon as the heavy nucleus is bombarded with a neutron, whereas in a delayed fission, the neutron enters a heavy nucleus causing instability and eventually leading to the fission.

(f) Every fission reaction is accompanied by the release of enormous energy in the form of light and heat. This is due to the mass defect i.e., the difference in the masses of the reactants and the products.

(g) Another by-product of a fission reaction is γ-radiations. These radiations of very short wavelength have harmful effects on living organisms.

Liquid drop model

After Otto Hahn and Fritz Strassmann discovered the fission reaction in 1939, Niels Bohr, Frenkel and Wheeler suggested the liquid drop model to explain the nuclear fission. According to this model, the various stages of a nuclear fission are as follows.

1. A heavy nucleus can be treated as the drop of a liquid with positive charge uniformly distributed.
2. Due to the surface tension, the surface of the liquid drop would act as a stretched rubber membrane.
3. The forces acting between the nucleons inside the nucleus are different from those acting between nucleons close to the surface, just as the cohesive forces between the molecules inside the liquid are different as compared to the adhesive forces at the surface.

4. During the bombardment of a nucleus by neutrous, a neutron could be absorbed by the 'drop-like' nucleus leading to the disturbance of the internal forces and setting up oscillations.
5. These oscillations lead to the development of a thick neck forming two parts, giving the nucleus an elongated shape.
6. From this stage the repulsive forces take over causing the fission (breaking up) of the nucleus.

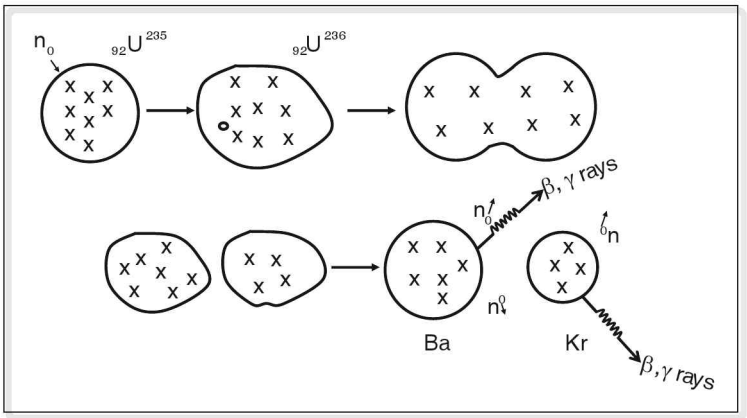


Figure 9.12 Pictorial representation of nuclear fission (liquid drop model)

Determination of energy liberated in a nuclear reaction

In the fission of $_{92}U^{235}$, the energy liberated can be calculated as follows:

$$_{92}U^{235} + {}_0n^1 \longrightarrow {}_{36}Kr^{92} + {}_{56}Ba^{141} + 3\,{}_0n^1 + \text{energy}$$

The mass of $_{92}U^{235}$ = 235.124 u

The mass of $_{36}Kr^{92}$ = 91.944 u

The mass of $_{56}Ba^{141}$ = 140.948 u

The mass of a neutron = 1.009 u

$\therefore$ The total mass of reactants = 235.124 u + 1.009 u = 236.133 u

$\therefore$ The total mass of products = 91.944 u + 140.948 u + 3 (1.009 u)

= 235.919 u

The mass defect = 236.133 u – 235. 919 u = 0.214 u

The mass of the products is less than that of the reactants. This loss in mass can be accounted for by the liberation of heat.

1 u = 931.5 MeV

$\therefore$ 0.214 u = 931.5 × 0.214 MeV = 199.3 MeV

$\approx$ 200 MeV

= $200 \times 1.6 \times 10^{-13}$ J = 3.2×10^{-11} J per fission

Considering the fact that 235 g of uranium contains 6.023×10^{23} atoms of uranium, if all the atoms in 1 kg of U^{235} undergo fission, it would liberate energy equal to 8.2×10^{13} J.

Taking 30% efficiency in conversion from nuclear energy to electrical energy, 2.46×10^{13} J of electrical energy can be generated if 1 kg of U^{235} undergoes fission. If coal were to be used for the generation of this amount of energy, 2500 kg of coal would be required.

Similarly, the energy liberated by 1 atom of U^{235} is two million times the energy obtained in the combustion of one molecule of methane.

Thus, harnessing nuclear energy for the generation of electrical energy has become important. Bhabha Atomic Research Centre (BARC) in Mumbai is a major research centre in India working on the atomic research. The major research reactors employed in India are APSARA (1 MW), Cirus (100 MW), Dhruva (100 MW), Purnima (zero energy fast breeder) and Zerlina (zero energy experiment research reactor).

The four nuclear power plants in India used for the generation of electricity are Tarapur Atomic Power Plant (420 MW), Rajasthan Atomic Power Plant (440 MW), Narora Atomic Power Plant (470 MW) and Chennai Power Plant (IGCAR) (470 MW). These power plants have been set up by Nuclear Power Corporation.

☛ *Example*

Find the energy released during the following nuclear reaction.

$${}_1H^1 + {}_3Li^7 \longrightarrow {}_2He^4 + {}_2He^4$$

The mass of ${}_3Li^7$ is 7.0160 u, ${}_2He^4$ is 4.0026 u and proton is 1.0078 u.

Solution

The mass of the reactant nuclei = 7.0160 + 1.0078 = 8.0238 u

The mass of the product nuclei = 4.0026 + 4.0026 = 8.0052 u

Mass defect = Δm = 8.0238 – 8.0052 = 0.0186 u

Energy released = 0.0186 u × 931.5 MeV = 17.326 MeV

Kinds of nuclear fission reactions

There are two kinds of nuclear fission reactions. They are uncontrolled fission reactions and controlled fission reactions.

The main difference between these reactions is that in the former, the amount of energy released is enormous in a short span of time and causes explosion, whereas in the latter case, there is a slow and controlled release of energy, which can be effectively utilized without causing any destructive explosion.

Uncontrolled fission reaction or chain reaction or explosive fission reaction

When uranium (U-235) is bombarded by neutrons, each uranium nucleus is broken into two nearly equal fragments and along with it, huge energy is released and two or three fresh neutrons are liberated.

Under favourable conditions, these neutrons cause fission reaction in other uranium nuclei and the process continues till a stable product is obtained. However, a minimum mass of U^{235} is required to start a chain reaction. This minimum mass is known as critical mass. The probability of fission is high in U^{235} isotope.

The composition of U-235 isotope in natural uranium is only ~ 0.7%. The other isotopes are U-234 and U-238. The concentration of U-235 is increased by a process called 'enrichment' and the uranium thus produced is referred to as 'enriched uranium'.

Chain reaction

"A chain reaction is a series of nuclear fissions whereby the neutrons produced in each fission cause additional fissions, releasing enormous amount of energy".

It is estimated that in one minute 10^{23} uranium atoms undergo fission in an uncontrolled chain reaction and as a consequence, tremendous amount of energy is released in a very short time causing destructive explosion. The atom bombs dropped on Hiroshima and Nagasaki by the Americans in 1945 resulting in the death of about 1.5 lakh people within minutes are examples of this type of fission reaction.

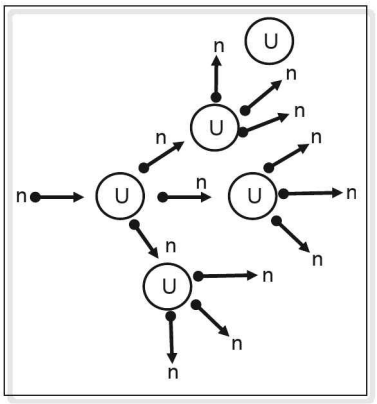


Figure 9.13

Pictorial representation of chain reaction

Controlled chain reaction or critical fission reaction

The energy released in chain reaction is enormous and if this reaction is controlled, the energy can be used effectively.

The control and harnessing of the nuclear energy becomes possible if the number of neutrons in the immediate environment of uranium nuclei is limited by some means.

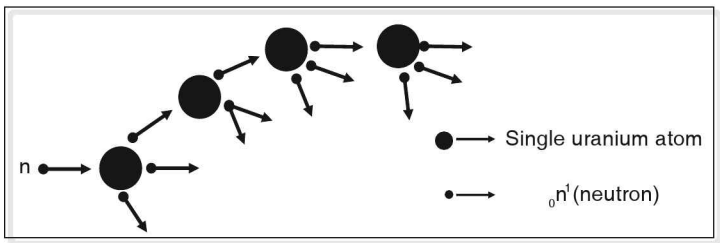


Figure 9.14

Pictorial representation of a controlled chain reaction

Thus, a fission reaction in which the number of neutrons participating in the fission in unit time is controlled by some means is referred to as "controlled fission reaction". In such a reaction, the rate of release of nuclear energy is also controlled.

A fission reaction can be controlled if

1. the neutron absorbed by a nucl eus of uranium atom does not cause fission,
2. the released neutrons are stopped from being captured by other uranium nuclei or
3. the neutrons released in a fission reaction escape from the system.

Nuclear reactor

A nuclear reactor, or a nuclear file, is a device in which a self sustaining controlled chain reaction is produced in a fissionable material. It is, thus, a source of controlled energy which is utilized for many useful purposes. The first nuclear reactor was constructed by Fermi in 1942. It used U^{235} as fuel.

Generation of electricity

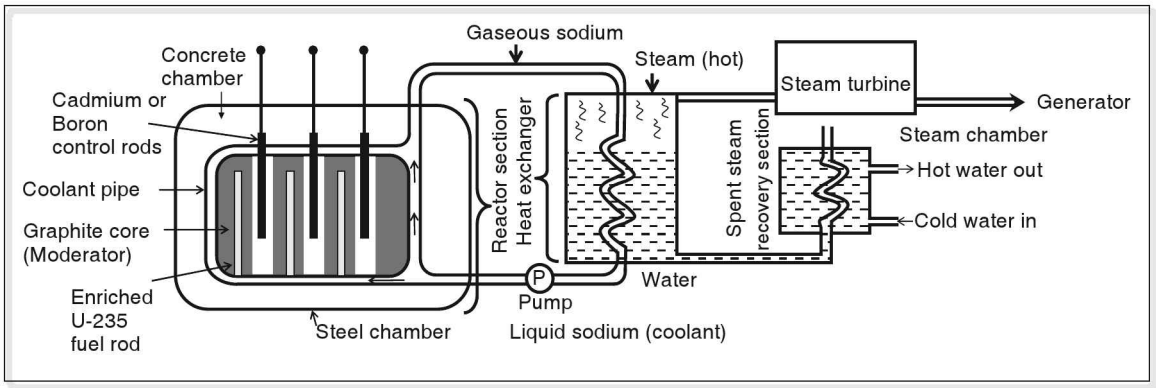


Figure 9.15 Nuclear Power Point

A nuclear reactor used for the generation of electricity has five fundamental parts.

1. **Nuclear fuel**
   It is the most important part of the reactor and is the substance used for fission. The fissionable materials like U^{235} and Pu^{239} used for the generation of heat energy in a nuclear reactor are called 'nuclear fuels'. Enriched uranium containing about 2% to 3% of U^{235} is often used as nuclear fuel.
2. **Moderator**
   It is used to slow down the neutrons. Heavy water, graphite or beryllium oxide is used for this purpose. Heavy water is the best moderator available.
   The neutrons produced in a fission reaction have high energy (1 MeV) and the probability of these neutrons producing fission is 500 times less than that of slow moving thermal neutrons. Thus the moderator converts fast neutrons to thermal neutrons.
3. **Coolant**
   The energy released inside the reactor in the form of heat is removed by a coolant. For this purpose, molten sodium, ice cold water or CO_2 is used in the reactor. This heat removed from the reactor is utilized for producing steam which is, then, used to drive turbines for generating electricity.
   Molten sodium is the most efficient and commonly used coolant.
4. **Shield**
   Various types of harmful intense rays are emitted from the reactor during fission reactions. To protect the people from these rays, the reactor is surrounded by 7 to 8 feet thick concrete and lead walls forming a shield.

5. **Controlling rods**
   For controlled fission reaction to take place, only one of the three neutrons produced in a fission reaction should be allowed to participate in the next fission, the other two being absorbed by some other material. These materials are called 'control rods'. Cadmium rods are used to control the rate of fission in the reactor. These rods are fixed in the walls of the reactor.

During the operation of a nuclear reaction two stages occur - shut down position and critical stage. In the shut down position, the control rods are completely inserted into the reactor so that all the neutrons emitted are absorbed, and no fission reactions take place.

In the critical stage, the control rods are drawn out of the reactor appropriately such that the right number of neutrons is available to sustain the controlled fission reaction at the required rate.

Nuclear power plant

A nuclear power plant essentially consists of four units,

1. Nuclear reactor
2. Heat exchanger
3. Steam turbine
4. Electric generator or dynamo

Working

1. In order to start the nuclear reactor, the control rods are slowly lifted till the reactor goes critical.
2. The heat energy produced in a reactor vaporizes the molten sodium. The vapors of sodium pass on the heat to the water in the heat exchanger, converting water to steam.
3. The steam turns the turbine converting heat energy into mechanical energy.
4. The turbine is coupled to a generator, which, in turn, converts mechanical energy into electric energy.
5. The condensed steam in the turbine and liquefied sodium in the heat exchanger are circulated through pumps.

Nuclear fusion

Fusion means joining together or combining together. "A nuclear reaction combining nuclei of lighter atoms together to form heavier nuclei resulting in the release of tremendous amount of energy is called a nuclear fusion reaction".

It is found that this heavier nucleus has a mass less than that of the sum of the masses of the reacting nuclei. Thus, such a process results in the conversion of mass into energy.

Nuclear fusion reactions were discovered in the 1930s. In 1939, Hans Bethe suggested that the enormous amount of heat energy produced in sun and other stars are due to such fusion reactions

taking place in their core. These fusion reactions which take place at high temperatures are referred to as thermo-nuclear reactions. The main advantage of a fusion reaction is that they are not accompanied by the release of harmful radiations.

Fusion Theory

In a typical nuclear fusion reaction, four protons (nucleus of hydrogen) combine forming a helium nucleus. The other products are two positrons and two neutrinos (neutral particles of negligible mass) and the reaction is accompanied by the release of about 27.6 MeV energy. This reaction takes place in three stages.

I **stage reaction**

$$\underset{\text{(Proton)}}{{}_1H^1} + \underset{\text{(Proton)}}{{}_1H^1} \longrightarrow \underset{\text{(deuteron)}}{{}_1H^2} + \underset{\text{(positron)}}{{}_{+1}e^0} + \underset{\text{(neutrino)}}{{}_0\upsilon^0}$$

II **stage reaction**

$$\underset{\text{(deuteron)}}{{}_1H^2} + \underset{\text{(Proton)}}{{}_1H^1} \longrightarrow \underset{\text{(Helium isotope)}}{{}_2He^3}$$

III stage reaction

$$\underset{\text{(helium isotope)}}{2\ {}_2He^3} \longrightarrow \underset{\text{(helium)}}{{}_2He^4} + 2\ \underset{\text{(proton)}}{{}_1H^1}$$

Combining the three equations, we have

$$\underset{\text{(proton)}}{4\ {}_1H^1} \longrightarrow \underset{\text{(helium)}}{{}_2He^4} + 2\ \underset{\text{(proton)}}{{}_{+1}e^0} + 2\ \underset{\text{(neutrino)}}{{}_0\nu^0} + \underset{\text{(energy)}}{Q}$$

This group of reactions of the three stages are called proton–proton cycle since two protons are released in the III stage reaction.

The mass–energy equivalence in a fusion reaction:

The mass of proton = 1.0078 u

The mass of helium nucleus = 4.001506 u

The masses of positrons and neutrinos being negligible, we have,

The mass of reactants = 4 × 1.0078 u = 4.0312 u and

The mass of products = 4.001506 u

∴ The mass defect = 4.0312 u – 4.001506 u = 0.0297 u

The equivalent energy = 0.0297 u × 931 MeV/u = 27.67 MeV

All this energy is converted into heat.

A fusion reaction can be initiated when a high energy proton (KE = 0.01 MeV) accelerated in a particle accelerator, is made to strike on a target and fuse with the nuclei of the target. The target consists of nuclei of low atomic number.

Alternately, the kinetic energy of the light nuclei (protons) can be increased by raising the temperature to as high as 4×10^6 K in an enclosure. At these high temperatures, the atoms are completely ionised and the electrons move at high speeds. Matter in such a state is called plasma in which the fusing nuclei easily overcome the electrostatic repulsion between them.

Hydrogen bomb

Hydrogen bomb, like atom bomb, is a weapon of mass destruction, employing the thermo-nuclear reactions, which can release enormous energy within a few minutes. The high temperature required for the uncontrolled fusion reactions in a hydrogen bomb is produced by a fission bomb containing uranium (U^{235}) or plutonium (Pu^{239}). Thus the hydrogen bomb is a fission-fusion bomb.

Nuclear fusion reactor

The energy released in a fusion reaction is much higher than the energy released in a fission reaction. Thus, there is a considerable interest in harnessing this energy for useful applications such as generation of electricity. However, fusion reactors could not be built so far for the following reasons.

1. Unlike fission reactions, fusion reactions are not chain reactions. Hence it is not possible to sustain the reaction.
2. The order of the magnitude of temperature enabling a fusion reaction to take place is 10^6 K. Producing such high temperature is very difficult. Even if the temperature is produced using a fission bomb, there is no material which can withstand such a high temperature.
3. It is difficult to harness and convert the high amount of heat energy produced in a fusion reaction into mechanical energy.

Nuclear hazards and safety measures

Nuclear radiations affect all living beings as well as materials including buildings. The effects due to nuclear radiations given out during the explosion of a nuclear bomb cause death of lakhs of people, plants and animals. The materials and buildings are damaged extensively. In humans exposure to nuclear radiations causes somatic and genetic effects. Due to somatic disorders, deadly diseases like cancer develop whereas genetic effects are observed generation after generation. In genetic effect, nuclear radiations change the genes of the exposed person for the worst and pass them on to the next generation. The offsprings born to such persons have deformities. Nuclear reactors are another source of nuclear radiations. Continuous nuclear reactions in them produce a large quantity of radioactive materials. The fuel rods, after use, become highly radioactive and contain toxic elements like plutonium. The spent nuclear fuel rods should be handled carefully so that humans are not exposed to the radiations given out by them. For this, they are first stored under water in a nuclear plant and then sent to a processing plant so as to retrieve usable materials. After that, they are stored in safe places like unused deep mines. Care should be taken to ensure

that there is no leakage present in moderators and cooling water used in a nuclear plant. Such a leakage, if present, can cause contamination of water bodies and soil.

Nuclear research in India

In 1945, under the leadership of Dr Homi Jehangir Bhabha, the atomic research in India started. The atomic energy commission is established and Dr Bhabha was made its first Chairman.

Under his leadership, atomic reactors such as Apsara, Cirus, Zerleina and a plant to enrich Uranium were started. Indian scientists have developed the capability to prepare nuclear fuel, design and fabrication of nuclear reactors and also explosive devices like atom bomb and hydrogen bomb.

The major centre of nuclear related research and development is Bhabha Atomic Research Centre (BARC) situated at Bombay in Maharashtra. The major research nuclear reactors are Apsara, Cirus, Dhruva, Purnima and Zerlina.

The nuclear power plants are Tarapur atomic power plant (Maharashtra), Rajasthan atomic power plant, Narora atomic power plant (Uttar Pradesh), and Chennai Power Plant (Tamil-Nadu).

In India, Uranium occurs at Jaduguda in Jharkhand. The natural ore is processed and enriched at Nuclear Fuel Complex in Andhra Pradesh.

☛ *Numericals*

1. In a given radio-active sample, there are 10^{24} nuclei present. If its half life period is 20 years, how many nuclei will be present after 10 years?

   (Take $\frac{1}{\sqrt{2}} = 0.707$)

Solution

The number of nuclei (N) present after 'n' half lives is given by

$$N = N_o\left(\frac{1}{2}\right)^n \quad \ldots\ldots\ldots(1)$$

Where $n = \frac{t}{T}$

In the question, we are given that $N_o = 10^{24}$

$n = \frac{10}{20} = \frac{1}{2}$, $N = ?$

Substituting the above values in equation (1), we get

$$N = 10^{24}\left(\frac{1}{2}\right)^{\frac{1}{2}}$$

$= 10^{24} \times 0.707$

$= 7.07 \times 10^{23}$

2. Calculate the binding energy per nucleon for Beryllium $^{9}_{4}Be$, its mass being 9.012 u. The masses of proton and neutron are 1.008 u and 1.009 u.

   (Take 1 u = 931.5 MeV)

Solution

First, let us find the mass defect. The Beryllium nucleus contains 4 protons and 5 neutrons.

Mass of 4 protons = $4 \times 1.008 = 4.032$ u

Mass of 5 neutrons = 5×1.009 u = 5.045 u

Total mass of protons (4) and neutrons (5) = 4.032 + 5.045 = 9.077 u

Mass defect = 9.077 – 9.012 = 0.065 u

The mass defect converted into equivalent energy gives binding energy.

1 u = 931.5 MeV

$\therefore$ 0.065 u = 0.065×931.5 MeV = 60.5475 MeV

Binding energy per nucleon

$$= \frac{\text{binding energy}}{\text{No. of nucleons}} = \frac{60.5475}{9} = 6.7275 \text{ MeV}$$

3. Assuming 200 MeV energy is released per fission of $^{235}_{92}U$, calculate the energy released when 1 kg of $^{235}_{92}U$ undergoes complete fission. What percentage of mass of $^{235}_{92}U$ gets converted into energy?

Solution

235 g of $^{235}_{92}U$ contains Avogadro number (6.023×10^{23}) of atoms.

No. of atoms of $^{235}_{92}U$ in 1 kg of uranium

$$= 1000 \times \frac{6.023 \times 10^{23}}{235} = 2.563 \times 10^{24}$$

Energy released = No. of atoms × energy released per fission.

$= 2.563 \times 10^{24} \times 200$ MeV

$= 5.126 \times 10^{26}$ MeV

The mass that liberates the above energy can be calculated using $E = mc^2$ .....(2)

where $E = 5.126 \times 10^{26} \times 1.6 \times 10^{-13}$ J

$= 8.2 \times 10^{13}$ J

$c = 3 \times 10^{8}$ m s^{-1}

Substituting in (2), we get

$$m = \frac{E}{c^2} = \frac{8.2 \times 10^{13}}{9 \times 10^{16}} = 9.11 \times 10^{-4} \text{ kg}$$

$$\% \text{ of mass} = \frac{9.11\times10^{-4}}{1}\times100 = 0.0911\%$$

= 0.1% approximately

4. Two protons and two neutrons combine to form a nucleus of ^4_2He. Find the energy released during the process. What type of nuclear reaction is it? Take masses of proton, neutron and helium as 1.007 u, 1.009 u, 4.002 u respectively.

Solution

The nuclear reaction is $2^1_1P + 2^1_0n \rightarrow {}^4_2He$

First, let us find mass defect and then energy.

The mass of 2p and 2n is $(2 \times 1.007 + 2 \times 1.009) = 2 \times 2.016 = 4.032$ u.

The mass of helium is = 4.002 u.

The mass defect = 4.032 – 4.002 = 0.030 u.

1 u liberates 931.5 MeV of energy. The energy equivalent to 0.030 u = 0.03×931.5

= 27.94 MeV.

The above nuclear reaction is called fusion as lighter nuclei combine together to form a single nuclei.

5. The binding energy of ^7_3Li and ^4_2He are 39.2 MeV and 28.24 MeV respectively. Which one of the two nuclei is more stable?

Solution

The nucleons present in ^7_3Li is 7.

The binding energy per nucleon for lithium is = $\frac{39.2}{7} = 5.6$ MeV.

The binding per nucleon for helium is $\frac{28.24}{4} = 7.06$ MeV.

The binding energy per nucleon is the measure of stability of the nuclei. Therefore, helium is more stable than lithium.

6. A nuclear reactor generates 500 MW electrical energy using the fission of $^{235}_{92}U$. Find the mass of $^{235}_{92}U$ required to produce electricity for 30 days. (0.1% of the mass of $^{235}_{92}U$ gets converted into energy).

Solution

The electrical energy produced in one second is 500 MJ. The electrical energy generated in 30 days is

$E = 500 \times 60 \times 60 \times 24 \times 30 \times 10^6$

$= 1.3 \times 10^{15}$ J. (approximately)

To generate above energy the mass required can be found using the formula

$E = mc^2$

$$\text{or } m = \frac{E}{c^2} = \frac{1.3\times10^{15}}{(3\times10^8)^2} = 0.144\times10^{-1}\text{ kg} = 1.44\times10^{-2}\text{ kg.}$$

Let M be the mass of uranium required to generate energy E. Then 'm' is 0.1% of M.

$$\text{i.e. } m = \frac{0.1}{100}\times M$$

$\Rightarrow M = 1000\ m = 1000 \times 1.44 \times 10^{-2} = 14.4$ kg .

7. In a Coolidge tube, cathode rays emitted from the cathode are accelerated through a potential difference of 20 kV. Find the minimum wavelength of the X – rays produced.
(Take Plank's constant $h = 6.6 \times 10^{-34}$ J s)

Solution

When cathode rays fall through a potential difference, the work done on them equals change in their kinetic energy. When cathode rays are stopped, they lose kinetic energy which is then converted into X-rays.
Thus,

$$\text{Work done on cathode rays (W)} = V \times q = \frac{1}{2}mv^2 \quad \text{———(1)}$$

The energy and frequency of X-rays (any radiation) is related as $E = h\nu$.
Where 'h' is Plank's constant.

$$\text{But } \nu = \frac{c}{\lambda}$$

$$\therefore E = h\nu = \frac{hc}{\lambda} \quad \text{.....(2)}$$

Equating (1) and (2) and solving λ.

$$\frac{hc}{\lambda} = v \times q$$

$$\lambda = \frac{hc}{vq}$$

Substituting $h = 6.6 \times 10^{-34}$ J s, we have

$c = 3 \times 10^8$ m s^{-1}

$V = 20 \times 10^3$ V, $q = 1.6 \times 10^{-19}$ C

$$\lambda = \frac{6.6\times10^{-34}\times3\times10^8}{20\times10^3\times1.6\times10^{-19}} = 0.625\text{ Å(approximately)}$$

test your concepts

Very short answer type questions

1. Define artificial transmutation.
2. Energy released in a nuclear reaction is given by _____.
3. What is a chain reaction?
4. _____ are used to absorb the neutrons in a nuclear reactor.
5. In a discharge tube, at what pressure cathode rays are produced?
6. Thorium series is also called _____.
7. If the half-life period of a radioactive substance is 0.693 years, what is its decay constant?
8. The first radioactive element discovered is _____.
9. What happens when cathode rays are suddenly stopped?
10. In a nuclear reactor, what is the function of heat exchanger?
11. What is the measure of the stability of the atom?
12. ${}_1H^3 +$ _____ $\rightarrow {}_2He^4 + {}_0n^1 +$ Energy
13. How can we say that cathode rays consist of negatively charged particles?
14. Radioactive isotope of carbon is _____.
15. Define binding energy.
16. What is a fusion reaction?
17. How are X-rays produced?
18. In artificial transmutation, the bombarding particles are called _____.
19. What forces are responsible for keeping the nucleons together in a nucleus?
20. Why are nuclear fusion reactions also known as thermo-nuclear reactions?
21. Which rays among α, β and γ are deflected most in a magnetic field? Which rays remain undeflected?
22. The radioactive isotope used in the treatment of cancer is a good source of _____ radiations.
23. What is a moderator? Give one example.
24. What is radioactive transformation or radioactive decay?
25. What is the use of control rods?
26. What is the use of Geiger counter?
27. In all radioactive series, the end element formed is
28. Name the three radioactive series. Which series is referred to as 4n Series?
29. 4n + 2 series is called _____ series.
30. Define isotopes, isobars and isotones.

Short answer type questions

31. State the laws of radioactive emissions.
32. Find the binding energy per nucleon for a lithium nucleus ${}_3Li^7$. Take mass of ${}_3Li^7$ = 7u, mass of proton 1.007825 u and mass of neutron = 1.008665 u. Take 1 u = 931.5 MeV.

33. Explain how the age of fossils is determined.
34. What are prompt fission and delayed fission reactions?
35. Mention the properties of cathode rays.
36. What is nuclear fusion? What are the conditions for a nuclear fusion reaction to take place?
37. Explain nuclear chain reactions.
38. Explain why the energy released during fusion cannot be used to produce electricity.
39. Mention the properties of X-rays.
40. What is a nuclear reactor? Name its different parts.
41. Give any three uses of radio-isotopes.
42. Explain nuclear fission by giving an example.
43. Mention the uses of X-rays.
44. Using $E = mc^2$, find out the energy released, when 1 u of mass disappears.
    Take 1 u = 1.66×10^{-27} kg.
45. What is a radioactive series? How is it formed? Explain.

Essay type questions

46. What is a Liquid Drop Model? How does it help in the understanding of fission reaction?
47. Explain in detail the nuclear hazards and the safety precautions to be taken.
48. Describe how cathode rays are produced in a discharge tube. Write down the properties and uses of cathode rays.
49. Describe in detail the process of β-decay and γ-decay.
50. Compare the properties of α, β and γ radiations.

CONCEPT APPLICATION



Concept Application Level—1

Direction for questions 1 to 7: State whether the following statements are true or false.

1. High energy neutrons are used in a nuclear reactor to initiate the fission reaction.
2. Cadmium rods are used as moderators in a nuclear reactor.
3. Oxygen can be obtained from nitrogen by artificial nuclear transmutation.
4. ${}_0n^1 + {}_7N^{14} \rightarrow {}_1H^0 + {}_6C^{15}$
5. X–rays have higher penetrating power as they travel with very high velocity.

6. Higher the mass defect, higher will be the stability of the nucleus.

7. The e/m ratio of anode rays is constant irrespective of the gas filled inside the discharge tube.

Direction for questions 8 to 14: Fill in the blanks.

8. If the binding energy per nucleon for $_3Li^7$ is 5.6 MeV, the total binding energy of a lithium nucleus is _____ MeV.

9. The conversion of 1 u of mass results in _____eV of energy.

10. The common product formed in the artificial transmutation by a proton is _____.

11. In plants and animals, the ratio of C^{14} to C^{12} is _____.

12. The radioactive radiation which can be stopped easily is _____.

13. When an α-particle is ejected, the atomic number of the atom decreases by _____.

14. The cathode rays emitted from a concave cathode meet at _____.

Direction for question 15: Match the entries in column A with appropriate ones from column B.

15.

A.	Coolidge tube	( )	a.	present in all natural radioactive series
B.	α decay	( )	b.	isobars
C.	Radon	( )	c.	istones
D.	$_{14}S^{31}$, $_{15}P^{32}$	( )	d.	stability of nucleus
E.	β-decay	( )	e.	to slow down neutrons
F.	Binding energy per nucleon	( )	f.	decrease in atomic number by two
G.	Moderator	( )	g.	X-rays
H.	Thermonuclear reaction	( )	h.	$E = mc^2$
I.	Mass energy equivalence	( )	i.	clot of blood in human body
J.	$_{11}Na^{24}$	( )	j.	nuclear fusion

Direction for questions 16 to 30: For each of the questions, four choices have been provided. Select the correct alternative.

16. The specific charge is the highest for _____.
(1) α-rays (2) β-rays (3) protons (4) neutrons

17. Which of the following is/are subatomic particles?
(1) Protons (2) Positron (3) Neutrino (4) All the above

18. Which of the following radiations has the highest penetrating power?
(1) α (2) β (3) γ (4) X-rays

19. The nuclear reaction that takes place in a nuclear reactor is _____.
(1) controlled fission reaction (2) uncontrolled fission reaction
(3) nuclear fusion reaction (4) None of the above

20. _____ is an example of isobars.
(1) $_{14}Si^{31}$, $_{15}P^{32}$ (2) $_{6}C^{13}$, $_{7}N^{13}$ (3) $_{7}N^{14}$, $_{7}N^{15}$ (4) None of the above

21. Higher value of decay constant indicates _____.
(1) faster decay (2) high half-life period
(3) slower decay (4) Both (1) and (2)

22. To convert $_{90}Th^{232}$ into $_{90}Th^{228}$, the least number of α and β particles that have to be emitted are _____ and _____ respectively.
(1) 1,1 (2) 2,1 (3) 1,2 (4) 2,2

23. The mass number of an element in a radioactive series is 223. Then the radioactive series is _____.
(1) 4n (2) 4n + 3 (3) 4n + 2 (4) 4n + 1

24. Carbon dating is used to
(1) assess the age of fossils. (2) assess the amount of radioactivity.
(3) assess the rate of radioactivity. (4) none of the above.

25. The kinds of force(s) that are present between two protons in a nucleus is (are) _____.
(1) electrostatic force (2) nuclear force
(3) magnetic force (4) Both (1) and (2)

26. $_{Z}X^{A}$ and $_{Z+1}Y^{A}$ are two radioactive elements. X emits one α particle whereas Y emits one β particle. The daughter nuclei formed are _____.
(1) isotopes (2) isobars (3) isotones (4) isomorphic

27. The e/m of three singly ionized atoms A, B and C are x, y, z respectively. If $x > y > z$ and m_A, m_B, and m_C are their atomic weights, the correct relation among the following is _____.
(1) $m_A > m_B > m_C$ (2) $m_c > m_B > m_A$ (3) $m_A = m_B = m_C$ (4) $m_A > m_B = m_C$

28. A radioactive substance was reduced to $\frac{1}{4}$ of the total amount in 8 days. The time taken by this radioactive substance to get reduced to $\frac{1}{8}$ th of the remaining amount is _____.
(1) 12 days (2) 16 days (3) 20 days (4) 8 days

29. $_{1}H^{1} + _{3}Li^{7} \rightarrow$ _____ + $_{2}He^{4}$
(1) $_{2}He^{4}$ (2) $_{1}H^{1}$ (3) $_{4}Be^{9}$ (4) $_{4}Be^{8}$

30. The property of cathode rays used in a monitor of a computer is _____.
(1) high velocity (2) high ionization power
(3) cause fluorescence of phosphors (4) rectiliesnear propagation

Concept Application Level—2

31. The binding energy per nucleon for deuterium and helium are 1.1 MeV and 7 MeV respectively. Find how much energy is released by the fusion of 1 g of deuterium.

32. Calculate the energy equivalent of 1 u of mass. Express the energy in MeV.

33. Show that iron $_{26}Fe^{56}$ is stable than carbon $_{6}C^{12}$, given binding energy of $_{26}Fe^{56}$ is 492. 35 MeV and that of $_{6}C^{12}$ is 89.14 MeV.

34. How many electrons, protons and neutrons are there in 12 g of $_{6}C^{12}$?

35. In Ac-U series the decay of the parent nucleus starts with an α-particle. The decay is continued such that an α and β particle are emitted alternately until an unstable nucleus radium is formed. Identify the isotopes, isobars, isotones formed and the change in the number of neutrons and protons of parent and daughter nuclei.

36. How many grams of U^{235} is required to produce 8×10^{11} J of energy, if 200 MeV of energy is released when one U^{235} atom undergoes fission?

37. $\frac{15}{16}$th of a radioactive substance disintegrates in 12 days. How much more time will it take for the remaining amount of radioactive substance to reduce to $\frac{1}{32}$ of the reduced amount? What is the half-life period of the radioactive substance?

38. The half-life period of C^{14} is 5670 years. If 56 g of C^{14} was present initially, how many atoms of C^{14} is left after 22680 years?

39. Is there any force of attraction between two neutrons in a nucleus? Explain.

40. The total binding energies of $_{1}H^{2}$, $_{2}He^{4}$, $_{26}Fe^{56}$ and $_{92}U^{235}$ are 2.22 MeV, 28.3 MeV, 492 MeV and 1786 MeV respectively. Which among the above four is the most stable nucleus? Why?

41. Complete the following nuclear reactions

(1) $_{92}U^{238} + _____ \rightarrow {}_{99}E^{247} + 5\,{}_{0}n^{1}$

(2) $_{96}Cm^{244} + {}_{2}He^{4} \rightarrow _____ + {}_{97}Bk^{245} + {}_{1}H^{1}$

(3) $_{94}Pu^{292} + \ldots\ldots \rightarrow {}_{104}Ku^{260} + 4\,{}_{0}n^{1}$

42. What percentage of initial amount of radioactive substance is approximately left, after 10 half-lives?

43. Write down the sequence of radiations emitted in the following radioactive processes.

(1) $_{Z}P^{A} \rightarrow {}_{Z-1}Q^{A} \rightarrow {}_{Z}R^{A}$

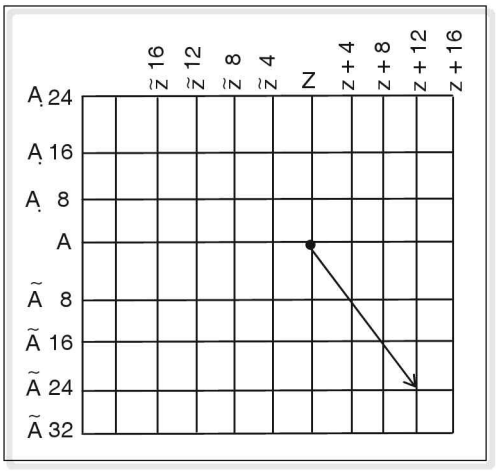
(2) $_{Z}K^{A} \rightarrow {}_{Z-2}L^{A-4} \rightarrow {}_{Z-1}M^{A-4}$

44. How many nuclei of $_{92}U^{235}$ should undergo fission per second, to produce a power of 10MW, if 200 MeV of energy is released per fission of $_{92}U^{235}$?

45. The change in atomic number and mass of a radioactive substance during a radioactive decay process is graphically represented as follows. Determine the number of α and β particles emitted.

Concept Application Level—3

46. Explain how a radioactive nucleus can emit negatively charged β-rays, even though the nucleus itself is positively charged.



47. $_{60}Nd^{144} \rightarrow {}_{58}Ce^{140} + {}_{2}He^{4} + 1.77$ MeV. In the above nuclear reaction if the parent nucleus is at rest determine the velocity of α-particle.

48. A certain radioactive isotope has a half-life of 50 days. Over a period of 50 days, the average number of α- emissions per day is found to be 6.023×10^{13} from a sample of initial mass 1 μg. Assuming that only one emission of α particle takes place per atom, find the atomic weight of the radioactive isotope.

49. When the nuclear reaction started in a nuclear reactor the amount of $_{92}U^{235}$ and $_{36}Kr^{92}$ present after 2000 s is found to be 10 g and 23 g respectively. Determine the output power of the nuclear reactor if its efficiency is 70%. Fission of one nucleus of U–235 releases 200 MeV of energy.

50. How many α and β particles are emitted when uranium $_{92}U^{238}$ disintegrates to form lead $_{82}Pb^{206}$?

key points for selected questions

Very short answer type questions

1. The process of transformation of one element into another by artificial means of bombardment with high energy particles.

2. $E = m c^2$

3. A series of nuclear fissions whereby the neutrons produced in each fission cause additional fissions, releasing enormous amount of energy.

4. cadmium rods

5. 0.01 mm of Hg.

6. 4n series

7. $\lambda = \dfrac{0.693}{T}$

8. uranium

9. X-rays are produced.

10. To convert water into stream.

11. Binding energy per nucleon.

12. $_1H^2$

13. Direction of deflection in electric and magnetic fields.

14. C^{14}

15. Energy that results from the loss of mass when nucleus is formed.

16. Lighter nuclei combine to form a heavier nucleus.

17. When cathode rays are suddenly stopped by a metal target.

18. projectiles

19. Nuclear forces are responsible.

20. Take place at very high temperatures.

21. (i) β-rays

(ii) γ - rays

22. γ rays

23. Used to slow down neutrons. Heavy water.

24. Disintegration of an unstable nucleus resulting in the formation of another nucleus.

25. For controlling the fission reaction.

26. To record the rate of emission of radiations from a radioactive substance.

27. lead

28. (i) Uranium series, actino-uranium series and thorium series.

(ii) Thorium series.

29. Uranium series

30. (i) Atoms of same element with same atomic number but different mass number- Isotopes.

(ii) Atoms of different elements, with same mass number but different atomic number- Isobars.

(iii) Atoms with their nuclei, with same number of neutrons but different number of protons Isotones.

Short answer type questions

31. $N = N_o e^{-\lambda t}$,

Where N_0 and N are the number of atoms of the substance at t = 0 and 't' s respectively

λ- disintegration constant.

32. (i) Mass defect = $[3 \times m_p + 4 \times m_n] - 7u$

(ii) Binding energy = mass defect × 931.5 MeV

(iii) Binding energy per nucleon

$$= \frac{\text{Binding energy}}{7}$$

33. By determining the ratio of C-14 and C-12 in fossils.

34. (i) Fission takes place as soon as the heavy nucleus is bombarded with a neutron.

(ii) Neutron enters a heavy nucleus causing instabili.ty and eventually leading to the fission.

35. (i) Travel in straight line.

(ii) Exert mechanical force on the objects.

(iii) Deflect in presence of electric and magnetic fields.

(iv) Ionize the gas through which they pass.

(v) Affect photographic plates.

36. (i) Two or more lighter nuclei combine to give a heavy nucleus.

(ii) High temperature.

37. (i) A series of nuclear fissions where by the neutrons produced in each fission cause additional fissions, releasing enormous amount of energy.

(ii) Pictorial representation of nuclear chain reaction.

38. (i) Fusion reactions are not chain reactions – Not possible to sustain the reaction.

(ii) Requires a high temperature for nuclear reaction to take place.

(iii) Difficult to harness and convert the heat energy produced into mechanical energy.

39. (i) Electromagnetic radiations.

(ii) Wavelength is of the order of 1 A°.

(iii) Affect photographic plates.
(iv) Travel with velocity of light in straight line.
(v) Produce fluorescence in certain materials.
(vi) Not affected by electric and magnetic fields.

40. (i) A device, in which a self-sustaining controlled chain reaction is produced in a fissionable material.
(ii) Nuclear fuel, moderator, coolant, shield, control rods.

41. (i) $_{11}Na^{24}$ – To detect a clot in the blood in the human body.
(ii) $_{27}Co^{60}$ – To kill cancer cells.
(iii) $_{53}I^{131}$ – To test the functioning of the thyroid gland.

42. (i) A process in which a heavy nucleus of a radioactive substance like uranium is split into lighter nuclei by the bombardment of a low energy neutrons.
(ii) $_{92}U^{235} + {}_{0}n^{1} \longrightarrow {}_{36}Kr^{92} + {}_{56}Ba^{141} + 3\,{}_{0}n^{1}$ + energy

43. (i) Medicine– To detect fracture in the bones.
(ii) Industry and engineering–To detect flaws in machinery parts.
(iii) Radiotherapy–To detect and destroy malignant tumours.

44. $E = \Delta mc^2$

45. (i) Sequential decay of a radioactive nucleus.
(ii) Results in the production of new nuclei at each stage.
(iii) Gives rise to a series of elements called radioactive decay series.

Essay type questions

46. Surface tension, breaks.

47. Genetic and somatic disorders.

48. Two electrodes. High potential difference, low pressure.

49. Shift in the position of daughter nuclei in periodic table.

50. Charge, velocity, penetration, ionization etc.



Concept Application Level—1

True or false

1. False
2. False
3. True
4. False
5. False
6. False
7. False

Fill in the blanks

8. 39.2
9. 931.5×10^6
10. helium
11. a constant, i.e., 10^{-12}
12. α - rays
13. 2
14. centre of curvature

Match the following

15. A : g
B : f
C : a
D : c
E : b
F : d
G : e

H : j
I : h
J : i

Multiple choice questions

16. Choice (2)
17. Choice (4)
18. Choice (3)
19. Choice (1)
20. Choice (2)
21. Choice (1)
22. Choice (3)
23. Choice (2)
24. Choice (1)
25. Choice (4)
26. Choice (3)
27. Choice (2)
28. Choice (3)
29. Choice (1)
30. Choice (3)

Concept Application Level—2,3

Key Points

31. 3.55×10^{24} MeV or 5.68×10^{11} J.

32. (i) $E = mc^2$
Energy equivalence of 1 kg = $1u \times 9 \times 10^{16}$J
$1u = 1.66 \times 10^{-27}$ kg
11eV is the energy acquired by an electron in a potential difference of one volt.
$1eV = 1.6 \times 10^{-19}$J.
(ii) 931.5 MeV.

33. Stability increases with binding energy per nucleon.

34. (i) 12 g of carbon consists of 6.023×10^{23} atoms.
Number of protons in one carbon atom = its atomic number = number of electrons.
Number of neutrons in one carbon atom = A - Z
(ii) Number of neutrons = number of protons = 3.6138×10^{24}.

35. (i) Isobars are atoms with different atomic number and same mass number.
Isotopes are atoms having same atomic number but different mass number.
Isotones are the atoms with same number of neutrons in their nucleus.
$^{A}P_{Z} \rightarrow {}^{A-4}D_{Z-2}$
$^{A}P_{Z} \rightarrow {}^{A}P_{Z+1}$
(ii) $_{90}Th^{231}$, $_{91}Pa^{231}$
$_{90}Th^{231}$, $_{90}Th^{227}$
4, 8

36. (i) $1 \text{ MeV} = 1.6 \times 10^{-13}$J
$200 \text{ MeV} = 200 \times 1.6 \times 10^{-13}$ J = energy released by one U^{235} atom
235 g of uranium contains 6.023×10^{23} atoms
From the given information, we can determine the number of uranium atoms required to produce 8×10^{11} J of energy and the mass of uranium required.
(ii) 9.8 g

37. (i) The fraction of amount of radioactive substance left = 1 – fraction of radioactive substance disintegrated.
Use the definition of half life period to determine the time taken to disintegrate to the given amount.
(ii) 15 days

39. (i) Nuclear forces are charge independent.
(ii) Nuclear attractive forces are between any two nucleons.

38. (i) Number of half lives = $\dfrac{\text{total time}}{\text{half life period}}$
Number of atoms in 14 g of $C^{14} = 6.023 \times 10^{23}$
Determine the amount of radioactive substance left in given number of halflives.
(ii) 1.50575×10^{23}

40. (i) Stability $\propto$ binding energy per nucleon.
(ii) $_{26}Fe^{56}$ is the most stable nucleus as its binding energy per nucleon is the largest among the given four entities.

41. (i) Balance the mass number and atomic number in the reactant side and in the product side.
Find out the difference in atomic and mass number. Identify the missing part with that atomic number and mass number.

(ii) $_{92}U^{238} + {}_{7}N^{14} \rightarrow {}_{99}Es^{247} + 5{}_{0}n^{1}$

(iii) $_{96}Cm^{234} + {}_{2}He^{4} \rightarrow 2{}_{0}n^{1} +$
$_{92}Bk^{245} + {}_{1}H^{1}$

(iv) $_{94}Pu^{242} + {}_{10}Ne^{22} \rightarrow {}_{104}Ku^{260} + 4{}_{0}n^{1}$

42. (i) $N\% = \frac{1}{2^n} N_o \times 100$ where n is the number of half lives, N_o. is the total number of atompresent initially.

(ii) Approximately 0.1%

43. (i) α decay $_{Z}P^{A} \rightarrow {}_{Z-2}D^{A-4}$
β - decay $_{Z}P^{A} \rightarrow {}_{Z+1}D^{A}$
emission of positron
$P \rightarrow n$ + positron

(ii) Positron, β-particle

(iii) α-particle, β-particle

44. (i) Power $= \frac{\text{energy}}{\text{time}}$
Convert 200 MeV into MJ and calculate the number of atoms required to produce 10 MJ of energy in one second = 10 MW.

(ii) 3.125×10^{17}

45. (i) The number of α particles emitted $= \frac{\text{change in mass number}}{4}$
Number of β particles emitted = (new atomic number – atomic number after α decay)

(ii) 6, 24

46. How does the neutron form initially?
What is the change in atomic number and mass number during the emission of a 'β' particle?

47. (i) $m_1v_1 = m_2v_2$
Write the velocity of one of the daughter nuclei in terms of the other.
Law of conservation of energy.
$\frac{1}{2}m_1v_1^2 + \frac{1}{2}m_2v_2^2 = Q$ value.

(ii) 0.92×10^7 m s^{-1}

48. (i) The number of atoms present in gram atom, i.e. atomic weight of the radioactive substance $= 6.023 \times 10^{23}$.
Determine the number of atoms in 1 μg of the radioactive substance. Half of this gives the number of atoms disintegrated in 50 days (half life period). Equate the number of atoms disintegrated in 50 days in both the cases to find the atomic weight.
That is $\frac{6.023 \times 10^{23}}{\text{at.wt} \times 2} \times 10^{-6} = 6.023 \times 10^{13} \times 50$

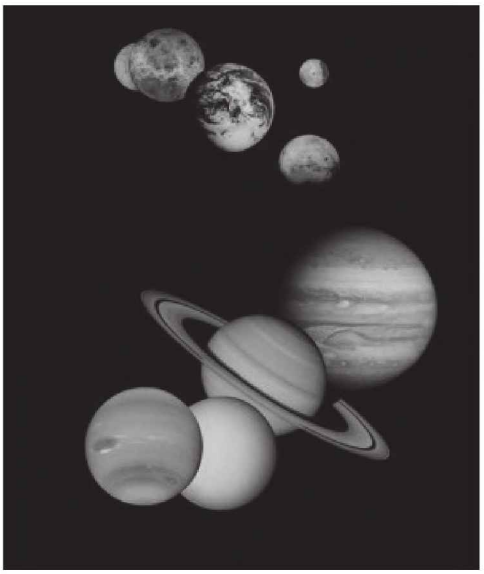
(ii) 100 g

49. (i) The amount of uranium disintegrated when 92 g of krypton is formed = 235 g.
Determine the number of uranium atoms disintegrated in a given time.
The energy released per uranium atom = 200 MeV
Output power $= \frac{\text{output energy}}{\text{time}}$

(ii) 1687 MW

50. (i) The mass number changes only during the emission of α particles.
Therefore the number of α particles emitted $= \frac{\text{change in the mass number}}{4}$
The atomic number of the parent nucleus decreases by two during the emission of an α particle and increases by one during the emission of a 'β' particle.
Therefore the number of β particles emitted = (number of α particles $\times$ 2) – change in atomic number.

(ii) 8 α particles and 6 β particles are emitted.



10

Sources of Energy

INTRODUCTION

Law of conservation of energy states that energy can be neither created nor destroyed. From work-energy theorem we understand how energy spent and work done are related. Whenever certain work is done, some form of energy is always involved, that is, to do some work, energy is required. For example, if water is to be brought to a height, electrical energy supplied to a pump motor gets transformed to mechanical energy which lifts the water. But what is the source of electrical energy? It could be either hydel energy, thermal energy or nuclear energy. Again the source of hydel energy is the water stored in a reservoir, those of thermal energy and nuclear energy are coal and fissionable uranium respectively. Thus, we find that energy is stored in nature in some substances (coal, petrol etc) or in some forms (winds, tides, etc). An indepth study of these sources of energy has become an important branch of science, as we need to replenish continuously those substances or sources, which are used up during conversion to other forms. The current chapter deals with different categories of sources of energy and the related topics.

Sources of energy

We are dependent upon various sources of energy in one way or the other for our energy requirements.

Sources of energy are mainly classified into two types. They are:

(i) Renewable sources of energy and
(ii) Non-renewable sources of energy

Characteristics of sources of energy

Fuels can be classified based on the following:

1. **Net energy output**
   The difference between the output energy and the input energy is termed as net energy output. For example, when we use a fuel, say coal or wood, we have to provide some initial heat to ignite it. This

will be the input. The heat energy given by it will be the output. Now, the net energy will be the difference between the two. A good fuel should have high net energy output.

2. **Availability**

   The energy source should be able to provide energy over a long period of time. Sun, wind, etc are the best examples for this.

   A good fuel should be capable of providing desired quantity of energy at a steady rate over a long period of time.

3. **Safe and convenient to use**

   This is one of the important characteristics of a fuel; a fuel has to be safe and convenient to use.

   For example, nuclear energy cannot be used for cooking. Correct fuel has to be selected for the required purpose.

4. **Easy for storage and transportation**

   Most of the sources of energy like petroleum, LPG, etc. are not found in all the places. So, they have to be transported over long distances.

   Moreover, these fuels have to be safely stored so that they can be helpful during crisis of energy sources. A good fuel should be safe to store and transport.

5. **Economical**

   A fuel has to be economical. A common household or an industry should not feel burdened because they have to use it.

A good fuel is one which has all the given characteristics.

Non-renewable sources of energy

"Energy sources which, when once used up, cannot be replenished are called non-renewable sources of energy".

Some non-renewable sources of energy are coal, petroleum and natural gas. These sources are related to solar energy, either directly or indirectly. The formation of these fuels have occurred over millions of years due to slow compression of plant bodies under high pressures inside the earth.

Renewable sources of energy

"The energy sources that can be replenished, once they are used up, are called renewable sources of energy". Some examples of renewable sources of energy are the sun, water, wind, earth, wood and biomass.

Solar energy

Sunlight consists of waves called electromagnetic waves. Visible light has a wavelength of 400 nm to 700 nm. Our eyes are insensitive to radiations having wavelength beyond these values. Nearly one-third of sunlight consists of infrared rays. These rays heat up the objects that lie in their path. The heat that we

feel in the sunlight is due to infrared radiation present in it. All bodies that are hot radiate energy in the form of infrared rays.

Nuclear fusion reactions in the interior of the sun continuously release enormous amounts of energy. It is estimated that the earth directly receives only 0·000000045792% of the sun's total energy output.

The sun is the ultimate source of energy. All forms of energy which are available on earth owe their origin to the sun.

The solar energy reaching the periphery of the earth's atmosphere is considered to be constant for all practical purposes, and is known as the solar constant.

Solar constant = 1.4 kJ m^{-2} s^{-1} or 1.4 kW m^{-2}

Only 47% of the solar energy that strikes the periphery of the earth reaches the surface. The remaining part of the solar energy gets reflected to space and some portion of solar energy is absorbed by the atmosphere. During this process u-v rays get eliminated and only visible light and infrared radiation reaches the earth's surface.

One third of sunlight consists of infrared radiation. Infrared radiation heats the substance through which they pass. Hot bodies transmit or radiate heat in the form of infrared radiation. The infrared radiation emitted by hot bodies have longer wavelength than the infrared radiation emitted by the sun.

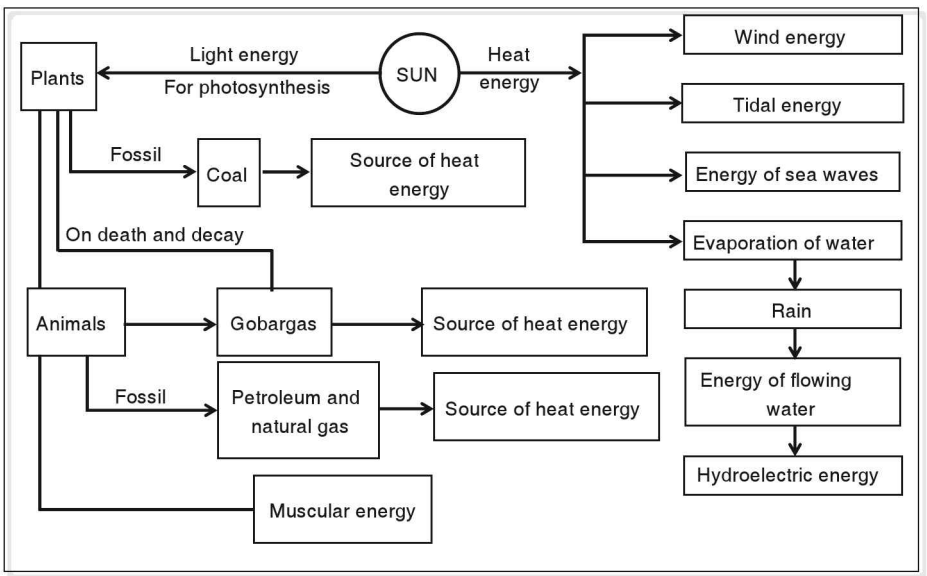


Figure 10.1 Functions performed by solar energy

The landmass and water bodies absorb a part of the solar energy that reaches the surface of the earth. This energy, then, manifests itself in a variety of forms through many natural processes, both biological and physical.

The various functions performed by solar energy on earth are illustrated by a flow chart given below.

1. Solar energy provides wind energy, tidal energy and the energy of sea waves.

2. Solar energy causes large-scale evaporation of water. Water vapours condense to form clouds, cause rain and ultimately change to flowing water. The energy of flowing water is utilised in water-mills and in the generation of.hydroelectric power.
3. Light energy present in the form of solar energy is stored in plants during photosynthesis in the form of chemical energy. The chemical energy of plants remains intact when they form fossil fuels such as coal. The chemical energy of coal is, then, used as a source of heat energy from domestic fuels to thermal power plants.

If plants do not become fossil after death, they can be converted to biogas (gobar gas). In biogas, the chemical energy stored in plants changes into the heat energy on burning. Similarly, the remains of dead animals can be changed into biogas.

When animals consume plants the chemical energy of plants changes into heat energy during respiration. It is the heat energy which helps to keep animals warm and perform various activities using their muscular energy.

If animals die in large numbers and get buried under the surface of the earth, then their bodies under intense pressure and temperature for long durations change to petroleum and natural gas. These products are a source of heat energy for most of the industries.

Solar energy provides light energy which helps us to see the things around.

Direct harnessing of solar energy

In this type, solar energy is directly converted into heat energy (as in the case of solar cookers, water heaters, solar concentrators, etc.,) or electrical energy (as in case of solar cells).

Indirect harnessing of solar energy

In nature, plants convert solar energy into chemical energy. The chemical energy of plants is used as food by animals or used as fuels to produce heat energy. Solar energy also changes into wind energy, the energy of flowing water, the energy of tides and the sea waves. This energy can be harnessed by water mills either in the form of mechanical energy or electrical energy.

Methods of enhancing the collection of solar energy

Devices working with solar energy as fuel need to enhance their solar energy collection so as to work quickly and effectively. The following are the methods to enhance the collection of solar energy.

Use of black painted surface

Black bodies are good absorbers as well as good radiators of heat. Thus, the surfaces which are to collect radiant heat are painted black so that they can absorb (or radiate) more energy.

A perfect black body is one which absorbs (or radiates) all the light incident on it.

Use of insulated box with glass cover

The wavelength of infrared radiation emitted by hot bodies is shorter than that of the radiation emitted by bodies of comparatively lower temperature.

It is observed that an insulated box with a glass cover prevents heat losses that occur due to conduction, convection and radiation. The glass cover allows solar energy to pass through it. It has the property of allowing radiation of shorter wavelength to pass through it. When the object with blackened surface gets hot within the box and radiates radiant heat with longer wavelength, the glass will not allow radiations to pass through, as they are of longer wavelength. Thus, the heat radiation is trapped within the insulated box, thereby raising the temperature of the objects with blackened surface.

Solar cooker

Solar cooker is a device that cooks food using heat from solar energy.

Box type solar cooker

This type of cooker consists of a rectangular wooden box, which is lined from inside with some insulating material such as thermocol or glass wool. The insulating material prevents heat losses that occur due to conduction, convection and radiation. The inner walls are painted in black. A plane mirror is hinged on the top of the box. The mirror is adjusted in such a manner that it reflects sunlight into the box.

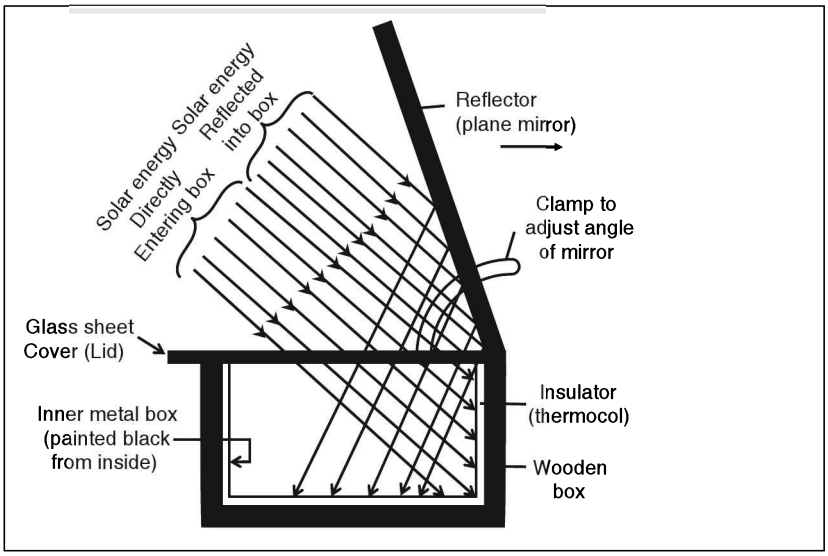


Figure 10.2 Solar cooker: A solar heating device for collecting maximum solar energy

Green house

Some crops require sufficient heat to grow. In cold countries, the heat required for the growth of plants is achieved by growing them in a green house.

The principle used in this is the same as in box type solar cookers.

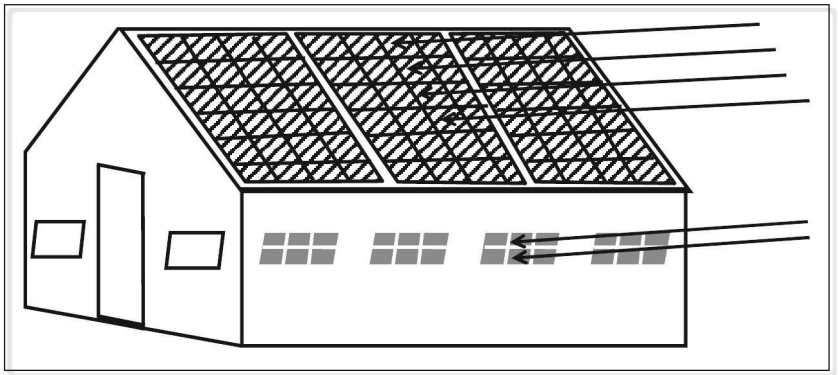


Figure 10.3

The glass surface on the top receives light and the walls which are coated in black retain the heat inside the room, thus maintaining a temperature higher than that outside of the house.

Reflector type solar cooker

In this type of cooker, the solar energy received over an area is concentrated at a particular point where the cooker is kept. The surface area of the disk-like structure is coated with a reflecting type material. This type of cooker can generate temperatures upto 180 °C at the focus (point where it is concentrated). Thus, this is more effective than a box type cooker.

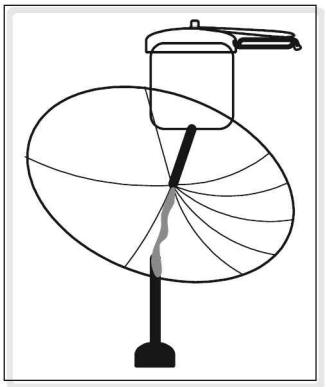


Figure 10.4 A reflector type solar cooker concentrates solar energy on a small area

Solar concentrators

The principle behind reflector type solar cooker is concentration of all the received solar energy at a particular point. Thus, it can be called a solar concentrator.

Solar furnace is of reflector type design where thousands of reflectors are arranged spherically so that each and every reflector reflects the incident light in a direction so that the reflected light focuses at a particular point. In this type of design, a temperature up to 3000°C can be produced.

Solar water heater

This is the most commonly used one among solar heating devices.

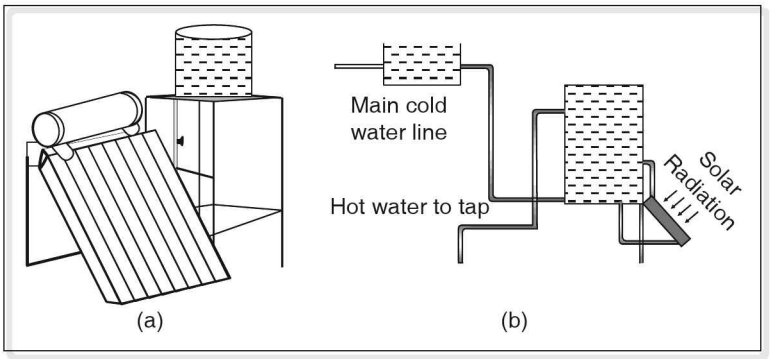


Figure 10.5 Schematic arrangement of a solar water heater

Here, a copper, pipe with its outer surface painted in black, is fixed in the form of a coil in a box. Bending copper pipe as a coil helps to increase the surface area for heating. A reservoir kept at a higher level is used to store cold water, which is connected to a small tank kept slightly above the water heater. One end of the copper pipe is connected to the bottom of the small tank of water, while the other end is joined near its middle. Water in the tank continuously circulates through the copper pipe due to a small difference in pressure between its two ends.

As the water slowly moves through the pipe, it absorbs solar energy and gets heated up. The hot water again goes to the tank while the cooler water from the lower part of the tank replaces it.

Solar cells

"A device which can directly convert solar energy into electrical energy is called solar cell". Solar cell is also called 'photovoltaic cell'.

Photoelectric effect

The phenomenon due to which light energy directly changes into electrical energy, when light is incident on certain sensitive materials is called photoelectric effect.

First photovoltaic cell that was invented in 1877 was only 0·6% efficient. It used selenium metal. In 1954, Bell laboratories of USA produced the first silicon solar cell, which had an efficiency of 5%.

Solar cells having multi layer silicon solar cells have been designed, which convert both visible and infrared radiation into electricity. This increases the efficiency of solar cells to a very great extent.

A large number of solar cells combined together produce large amount of current. The arrangement of large number of solar cells together is called solar panel.

Advantages of solar cells

1. It can be used as a source of generating electricity at any place.
2. It provides pollution free environment.

Disadvantages

1. High cost of installation. This is due to the technology to obtain pure silicon is expensive and due to use of silver as conducting material which is used to connect solar cells. Silver is good conductor of electricity. If other metals are used in place of silver, a fraction of electricity generated by solar cells is used to overcome the resistance of the conductor or the connectors.
2. Electricity generated by solar cells is stored in batteries which can give DC. In order to operate devices working on AC, the DC has to be converted to AC. This decreases the efficiency and further adds up to the cost.

Uses of solar cells

1. Solar cells are the main source of electricity in satellites and space probes.
2. They can be used for street lighting, power source to devices like TV, radio, etc.
3. Solar cells are used to provide electricity to off shore oil rigs and light houses.
4. Certain electronic calculators and wrist watches are powered by solar cells.

Wind energy

Moving air is called wind. Sun's energy (solar energy) is one of the main factors responsible for the motion of air in the atmosphere. Windmills are devices that convert wind energy into mechanical or electrical energy. They were used for grinding grains in many parts of the world, until a 100 years ago. Modern windmills are designed to convert wind energy into mechanical or electrical energy, in a large scale. A windmill essentially consists of a structure similar to that of a large electric fan that is erected at some height on a rigid support.

The principle of a windmill is that the blades of a windmill are designed to create a pressure difference between its different regions when wind strikes them. This pressure difference produces a turning effect to make the blades rotate.

In a water-lifting pump, the rotational motion of windmill is utilised to do mechanical work and draw underground water to the surface of the earth.

Production of electricity: wind energy farm

The rotatory motion of the windmills is utilised to rotate the armature of an electric generator. However, the electricity produced by a single windmill is quite small and cannot be used for commercial purposes.

A number of windmills are erected over a large area, which is known as wind energy farm. The energy output of each windmill in a farm is coupled together to generate electricity on a commercial scale.

Limitations in harnessing of wind energy

Wind energy farms can be established only at those places where wind blows for large duration of the year. The wind should be strong, steady, and should have a velocity of atleast 15 kmph.

Hydel energy

Energy that comes from flowing of water is called hydel energy. Water flowing through rivers has both kinetic energy and potential energy. The energy of flowing water can be used to transport heavy logs of wood from inaccessible areas of forests in hilly regions.

Generation of electricity: Hydroelectric power plant

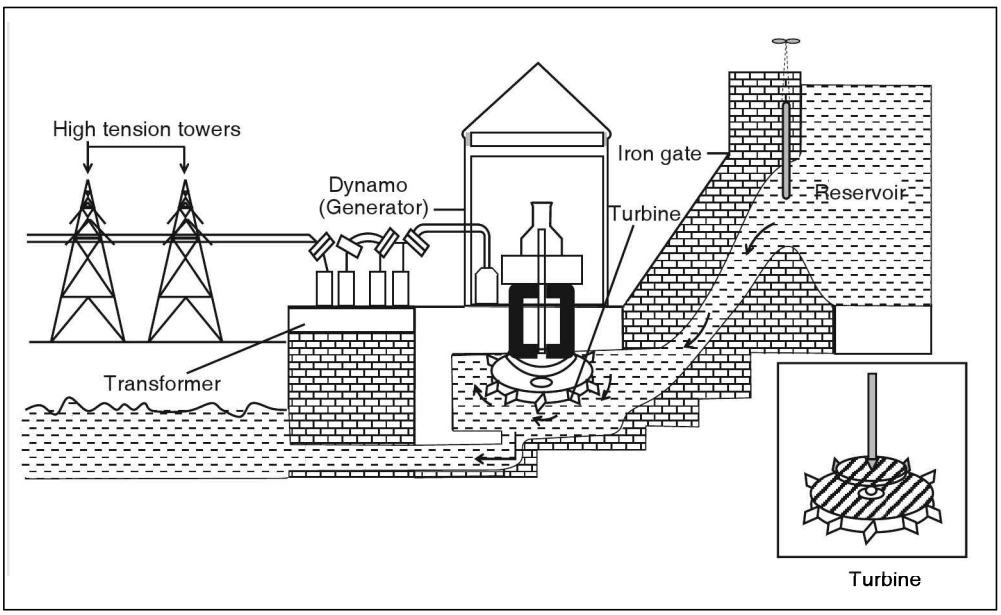


Figure 10.6 Hydel Power Plant

The energy of flowing water is utilised to produce electricity on a large scale at hydroelectric power plants. The water stored in dams has potential energy. When this water is released, it possesses high kinetic energy.

When the released water is made to fall on the blades of a turbine, they rotate and in turn rotate the armature of the electric generators to produce electricity.

Limitations of hydroelectric power plant

1. The initial cost of the construction of a dam is very high.
2. For generation of electricity, abundant water should be available in the dam.
3. These hydroelectric power plants can be constructed only at certain places.

Advantage of hydroelectric power plants

Water is a renewable source of energy and, further, these power plants do not pollute the environment.

Tidal energy

Oceans are huge reservoirs of water and energy. The oceans store energy in many forms. Tidal energy caused by the ocean tides is a renewable source of energy. The daily rise and fall of the tides can be used to drive turbines for generating electricity.

A small dam is constructed across a narrow opening of the sea. During high tide water flows to reservoir and during the low tide the water flows from reservoir through a pipe and this flowing water is used to rotate a turbine which in turn generates electric current.

However, these plants can be set up at certain places only where the amount of power released by waves striking the shoreline is at least 40 MW km^{-1}.

The three most common devices that are utilised for harnessing the energy of the ocean waves are:

(i) Surface-followers, using floats or pitching devices,
(ii) Oscillating water column and
(iii) Surge or focusing devices.

Since the rise and fall of water during tides is not so high, this is not likely to be a major source of energy.

The waves that continuously hit the ocean coast possess large amount of kinetic energy. The total amount of power released by the waves striking the short lines is about 2 to 3 million mega watts, and this energy can be harnessed only in limited sites.

Ocean thermal energy conversion (OTEC)

The difference in temperature between warm surface waters heated by the sun and colder waters found at ocean depths is another form in which solar energy becomes available from the oceans. This is known as ocean thermal energy.

The power plants used to harness this form of energy are called OTEC power plants.

A temperature difference of 20°C or more between surface waters and water at depths upto 1000 m is required for operating OTEC systems.

In this, the warm water is used to boil a liquid like ammonia or CFC. The vapour of the liquid is used to drive the turbine of a generator. The cold water from ocean depths is pumped up to convert the vapour again into liquid.

The advantage of OTEC plant is that it can be operated for 24 hours a day through out the year.

Geothermal energy

Geothermal energy is the energy stored as heat in the earth.

The temperature inside the earth is high enough to liquify the rocks, etc. (called magma) and heat the substances in it.

In certain areas below the surface of the earth, hot magma swells up into the crust, but remains trapped below the surface of the earth. Such areas in the earth's crust are called hot spots. When these hot spots are nearer to the surface of the earth, the water which seeps through gets converted to steam and comes out of the rocks in the upper layer (crust) of the earth with great force. These are called geysers.

Generation of electricity

This steam from geothermal sources can be directly utilised to turn turbines of the electrical generator to produce electricity. Such electricity is called geothermal electricity.

Advantages of geothermal energy

(i) Geothermal power plants can operate round the clock.
(ii) It is pollution free.
(iii) Operating cost involved is less (but initial cost in setting up this plant will be high).
(iv) This type of source is free and renewable.

Nuclear energy

Huge amount of energy can be generated from nuclear reactions.

Nuclear reactions are of two types: nuclear fission and nuclear fusion reactions.

Nuclear fission is a process in which a heavy nucleus splits into lighter nuclei along with the release of huge amounts of energy.

Nuclear fission reaction is the principle behind a nuclear power plant.

Nuclear fusion is a reaction in which lighter nuclei (like hydrogen) combine to become a heavier nucleus at very high temperatures with evolution of enormous amounts of energy. The Sun's heat is due to the nuclear fusion reactions occurring in the interior of the sun.

Nuclear fusion reactions occur generally in stars which exert high temperatures (nearly 4×10^7 K). However, the energy released in fusion reactions cannot be harnessed into productive applications as these reactions cannot be controlled.

Bioenergy

A small fraction of the solar energy radiation on earth's surface is converted into chemical energy by green plants through the process of photosynthesis. This energy becomes available in the form of biomass stored in different parts of a plant. Animals, including humans, utilise a part of this energy for their food.

The excreta of living organisms and their bodies after death also contribute to biomass.

Biogas

Animal and plant wastes are easily degraded by anaerobic micro organisms in the presence of water, producing gases such as methane, CO_2, H_2 and H_2S. This mixture of gases is called 'biogas'.

Figure (a) and (b) show the two designs of commonly used biogas plants.

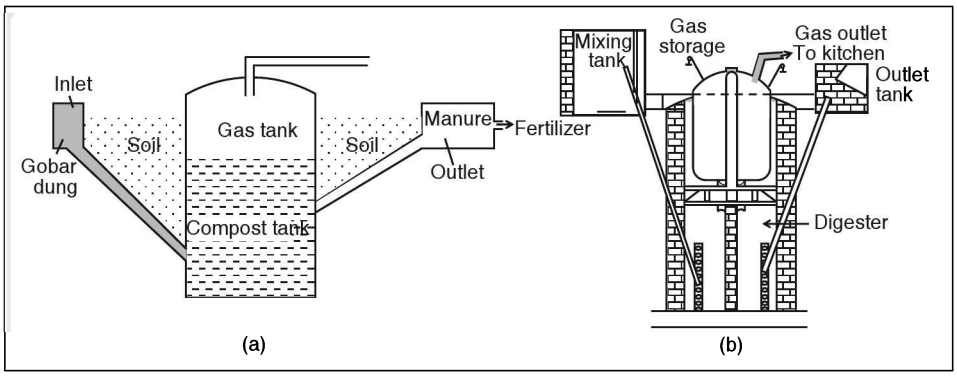


Figure 10.7 Schematic view of two designs of biogas plants: (a) Fixed dome type (b) Floating gas-holder type

Floating gas-holder type biogas plant

Floating gas-holder type biogas plant consists of a dome-like structure, usually made of steel, that floats up as the pressure of biogas evolved increases. This type of biogas plant is more frequently used where cow dung is used to obtain biogas.

Fixed dome type biogas plant

This has an advantage that the dome can be constructed by laying bricks.

Fixed dome has a longer life as compared to that of a floating dome, which reduces the cost of the biogas.

Formation of biogas

Slurry of animal dung and water is made in the mixing tank from where it is fed into the digester tank. In the digester, the action of micro organisms breaks down or decomposes the complex compounds of the biomass in the slurry. Anaerobic micro organisms, in the presence of water, do the degradation of biomass in the digester.

Since anaerobic micro organisms do not require oxygen, the digesters are designed like a sealed chamber. This process takes a few days to complete during which gases like methane, CO_2, H_2 and H_2S are produced. Biogas, which is a mixture of these gases, is supplied to consumers through pipes.

Advantages of a biogas plant

Combustion of biogas produces no smoke but a large amount of heat as its calorific value is high. The use of biogas in villages improves the sanitary conditions and checks environmental pollution. It is convenient for domestic use. The slurry obtained as a by-product, left in the plant after the gas is used, forms a good and enriched manure, containing compounds of nitrogen and phosphorous. The use of biogas reduces the dependence on the fossil fuels and wood.

Hydrogen and alcohol as fuels

The reaction between hydrogen and oxygen to form water is an exothermic reaction which evolves a lot of heat. Hydrogen is the substance that releases heat more than any fuel. Hydrogen is easily produced by the electrolysis of water and is also obtained as a by product in many industrial processes. Combustion of hydrogen does not pollute the environment.

Hydrogen is used in spaceships and high temperature flames as source of energy. However, it is difficult for the common man to work with hydrogen as the amount of heat emitted is very high. Moreover, storing and transportation of hydrogen is also a difficult task. The main reason for not using hydrogen as a fuel is due to its explosive nature.

Alcohol is another fuel that is gaining significance as a fuel. It reacts with oxygen and produces carbon dioxide and water, releasing heat energy. A mixture of alcohol and petrol is being used in some countries as a fuel for automobiles.

Example: Spirit lamp used in laboratories.

Fuels

Any substance that gives out energy on burning is called a fuel.

Fuels are classified into three types.

(a) Solid fuels,
(b) Liquid fuels and
(c) Gaseous fuels.

Calorific value

Any fuel, on burning, gives heat energy. But the heat energy produced is not equal for all fuels. The heat energy produced by a fuel is measured in joules. And it depends on the mass and the nature of the fuel.

Thus heat energy liberated on burning unit mass of a fuel is considered for comparison and this is called calorific value of the fuel.

Definition: "Calorific value is the energy liberated in the form of heat by a fuel when its unit mass is burnt completely under ideal conditions".

The SI unit of calorific value is J kg^{-1}. Calorific value is generally high for gases. Hydrogen has the highest calorific value.

Calorific values of some of the fuels are given below.

Solid fuel	Cow-dung cake	6000-8000
	Wood	17000-22000
	Coal	25000-33000
Liquid fuel	Ethanol	30000
	Kerosene	45000
	Petrol	47000
Gaseous fuel	Biogas	35000-40000
	Butane (LPG)	55000
	Methane	55000

Fuels containing oxygen atoms have low calorific value than those of hydrocarbons but they burn readily.

Solid fuels

Wood, charcoal, cow-dung cakes, coal, etc., are some examples of the solid fuels.

Wood is the fuel that has been used over a very long time as the main source of fuel and it is becoming scarce due to uncontrolled felling of trees.

Wood is a source of heat. When wood is burnt in a limited supply of oxygen, water and other volatile substance are removed from it and the residue left is known charcoal. The process of burning wood in limited supply of oxygen is called destructive distillation.

Hence, many nations have restricted felling of trees and have taken up afforestation as a major project.

In rural areas, cow-dung cakes, firewood and agricultural wastes are being used as a fuel for cooking. However, this increases pollution of atmosphere.

Coal is another important fuel used in industries, cities and power stations. Coal can be called as a fossil fuel.

Fossil fuels

Fossil fuels are the fuels formed by the decay of plants and marine animals that were buried in the earth millions of years ago under extreme conditions of heat and pressure. Since these conditions do not exist any more, there is no chance of new reserves of these fuels. So, these fuels are termed as non-renewable sources of energy.

Coal is the fossiled remains of the plants that have been buried deep into the earth's crust for millions of years while petroleum and natural gas are the fossiled remains of dead bodies of marine animals and plants.

Coal

Coal is a non-renewable source of energy. It is found below the surface of the earth. The main constituent of coal is carbon. The other constituents are hydrogen and oxygen. The percentage of carbon and other constituents in a sample of coal varies from place to place. The quality of coal depends on the percentage of carbon present in it.

The following table gives the information of different types of coal and the percentage of carbon in them.

Peat	27%
Lignite	28 - 30%
Bituminous	78 - 98%
Anthracite	94 – 98%

Anthracite is found to be the best quality of coal as the percentage of carbon in it is maximum.

Coke is made from bituminous coal. High grade coke is a perfect fuel for blast furnaces.

Burning of coal creates serious air pollution. When coal is burnt, the carbon present in it reacts with oxygen to produce carbondioxide. This is an exothermic reaction which yields a lot of heat energy. Coal, on destructive distillation, gives coke. It is used as reducing agent in the extraction of some metals from their ores and in making steel.

Liquid fuels

Fuels like petrol, diesel, kerosene, fuel oil, etc., come under the category of liquid fuels. All these fuels are liquids under normal state. They are obtained from petroleum, also called crude oil. Petroleum, and coal

are fossil fuels which have been formed under the earth over thousands of years. Since these fuels are formed by natural evolutionary process of earth, their replenishment is not in our control.

Petroleum products are mixtures of hydrogen and carbon and are called hydrocarbons.

Petroleum

The word 'petroleum' is derived from the Greek words 'Petra' which means 'rock' and 'oleum' which means 'oil' because it is an oil derived from rocks.

Petroleum is a complex mixture of many compounds of hydrogen and carbon known as hydrocarbons. It also contains small amounts of other compounds of oxygen, nitrogen and sulphur.

Formation of petroleum

Petroleum was formed from the dead remains of marine plants and animals, which settled in muddy sediments at the bottom of the sea millions of years ago. Due to prolonged sedimentation and accumulation of organic debris, these got buried deep below the surface of the earth. In intense heat and high pressure conditions underneath and in the presence of natural catalysts, the remains were ultimately converted into petroleum. For this reason, petroleum is also a fossil fuel. It is a dark oily liquid, which does not burn, and is insoluble in water and floats on it.

Extraction of petroleum

Large reservoirs of petroleum have been formed in nature for millions of years between porous rocks beneath the earth. It is pumped out by drilling holes and by digging wells. Petroleum in its natural form is a thick black liquid.

Refining of petroleum

The crude oil extracted from deep earth is separated into its constituents by a process called 'fractional distillation'. It is based on the principle that different fractions of petroleum boil at different temperature ranges.

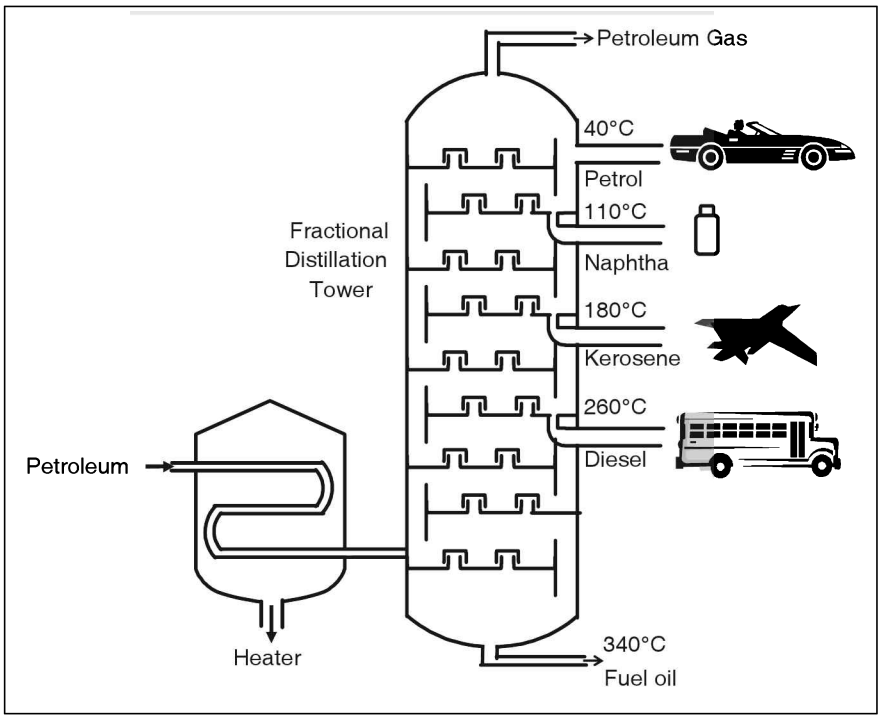


Figure 10.8 Distillation of petroleum

In refining of petroleum, the crude oil is fed into the base of the fractionating column and heated to 650°C. The mixture of vapours of different liquid components so formed rise in the fractionating column also called distillation column. The vapours of higher boiling fraction cool first, get liquefied and finally get collected from the lower outlet of the chamber.

Gaseous	$C_1 - C_4$	Below 298	As a fuel
Petroleum ether	$C_5 - C_7$	298 – 333	As solvent in dry cleaning
Petrol/gasoline	C_7-C_{10}	333 – 473	Motor fuel
Kerosene	$C_{12} - C_{16}$	448 – 550	Fuel in jet engines and for domestic purposes
Fuel oil or diesel oil	$C_{15} - C_{18}$	525 – 670	In diesel engines, in industries
Lubricating oil	$C_{16} - C_{20}$	Above 623	Lubrication, vaseline, greases
Petroleum wax	C_{20} and above	Above 623	Candles, waterproofers, fabrics, fuel, electrodes
Petroleum coke			Fuel

Gaseous fuels

Petroleum gas and natural gas are two common gaseous fuels.

Petroleum gas is obtained as a by-product in the process of refining of petroleum. It consists mainly of butane and small fractions of ethane and propane. Butane can be liquefied easily under high pressure. Cooking gas cylinders contain liquefied petroleum gas LPG. The main constituent of LPG is butane. Small quantities of ethane and propane are also present.

Biogas is another example of gaseous fuel.

Natural gas is an important fossil fuel that is found on the top of petroleum under the earth. However, some oil wells yield only natural gas. Its main component is methane.

It is compressed under high pressure to form compressed natural gas (CNG). CNG is easier for transportation than natural gas and is being used as a fuel in automobiles. Since burning of CNG evolves CO_2 and water, it is environment friendly.

Ignition temperature

Substances begin to burn only after it has attained certain minimum temperature. This minimum temperature at which the substance begins to burn is called ignition temperature. Ignition temperature is a characteristic of a substance. Hence a substance does not catch fire until it attains the ignition temperature.

Condition for combustion of fuels

Combustion is a process of burning of a fuel. During combustion, fuel combines with oxygen of the air to produce heat and light. Combustion is an oxidation and an exothermic reaction.

The conditions required for combustion of fuels are:

1. Attainment of ignition temperature of the fuel.
2. Presence of oxygen and the combustible material or the fuel.
3. Adequate supply of oxygen is necessary to maintain combustion.

☛ *Example*

Burners of cooking gas and kerosene stove are designed such that to maintain adequate supply of oxygen. If stove burns with a blue flame, it indicates proper burning of the fuel. A yellow or smoky flame indicates inadequate supply of oxygen and this results in an incomplete combustion i.e., the fuel is not completely burnt or unburnt. The partial burnt particles begins to glow due to heat which gives yellow colours to flame. Example, flame of the candle. When fuel is burnt completely,

1. fuel is not wasted.
2. smoke will be less.
3. carbon dioxide is produced instead of carbon monoxide.
4. charcoal is not produced.

When fuel is partially burnt.

1. fuel is wasted.
2. large quantity of smoke is produced.
3. poisonous gas, carbon monoxide is produced.
4. carbon particles are expelled with smoke.

Choice of a good fuel

It is very important to have a knowledge of the chemical properties of different fuels so that it can be of a great help in choosing the best suitable fuel. For example, we cannot use nuclear energy for cooking or burning leaves or use of wood for producing electricity.

Some of the qualities that we should consider while choosing the most suitable fuel are:

(i) easy availability, low cost,
(ii) no problems in storage and transport,
(iii) a high calorific value,
(iv) an ignition temperature that is well above the normal room temperature,
(v) a controllable rate of combustion for different purposes,
(vi) low rate of evaporation at room temperature,
(vii) low content of non-combustibles and
(viii) no combustion products that are poisonous or environmental pollutants.

test your concepts

Very short answer type questions

1. What is the source of heat on the surface of the sun?
2. The S.I. unit of calorific value is _____.
3. Several solar cells when connected in series and parallel form a _____.
4. What is the chief source of energy for all natural activities?
5. Define ignition temperature.
6. The fuel used as aviation fuel in jet aeroplanes is a special grade of _____.
7. What is the principle behind a solar power plant?
8. Name any two materials that are used for making solar cells.
9. Nearly 30% of sunlight consists of _____ rays that produce heating effect.
10. What is solar energy?
11. What are combustible and non-combustible materials?
12. A device which converts solar energy into electrical energy is _____.
13. What are the types of sources of energy?
14. What is nuclear fission?
15. _____ is left behind as a residue, when wood is burnt in a limited supply of oxygen.
16. What is combustion?
17. What is petroleum? What is natural gas?
18. The difference between the output energy and the input energy is termed as _____.
19. What is destructive distillation?
20. What are fossil fuels?
21. _____ metal is used in connecting wires that connect solar cells in a solar panel.
22. What are the types of fuels?
23. What is a hydroelectric power plant?
24. Energy sources that can be replenished, once they are used up, are called _____.
25. Between petrol and kerosene, which has a high calorific value?
26. What is a windmill?
27. The type of coal containing the lowest percentage of carbon is _____.
28. Define calorific value.
29. What is a solar cell?
30. The heat that we feel in the sunlight is due to _____.

Short answer type questions

31. How can wind energy be converted to electrical energy?

32. Write a short note on various types of fuels (with reference to their physical presence).

33. Give the characteristics of sources of energy.

34. Explain the production of geothermal electricity.

35. Write a short note on natural gas.

36. What are the different types of sources of energy? Give two examples for each.

37. What is a calorific value? Which gaseous substance has the highest calorific value?

38. Explain the construction of a fixed dome type biogas plant.

39. What are the types of fuels? Give two examples for each.

40. Give the differences between box type and reflector type solar cookers.

41. What are the limitations of a hydroelectric power plant?

42. How is petroleum formed? How is it extracted?

43. What are the various factors to be considered in choosing a fuel?

44. What are the limitations of harnessing wind energy?

45. Explain the construction of reflector type solar cooker.

Essay type questions

46. What is a biogas plant? Explain the construction and working of a biogas plant.

47. Explain how various petroleum products are obtained from crude oil.

48. "Solar energy is the chief source for all other forms of energy". Explain.

49. How is coal formed? What are the different types of it?

50. Explain the principle, construction and working of a windmill.

CONCEPT APPLICATION



Concept Application Level—1

Direction for questions 1 to 7: State whether the following statements are true or false.

1. Tidal energy is a major source of energy due to the presence of large water bodies on the earth.

2. The amount of heat energy released when a fuel burns completely is called calorific value.

3. Sunlight contains ultraviolet radiations which make it hot.
4. Anaerobic microorganisms decompose animal and plant wastes in the presence of water and produces gases such as methane, CO_2, H_2 and H_2S. The mixture of these gases is called biomass.
5. The calorific value of hydrogen is highest among fuels.
6. The process of burning wood in limited supply of oxygen to obtain charcoal as a resultant product is called destructive distillation.
7. Coal is the fossiled remains of dead bodies of marine animals that are buried in the earth's crust for millions of years.

Direction for questions 8 to 14: Fill in the blanks.

8. A substance started burning when it is exposed to air. Then, the ignition temperature of the substance is less than the _____.
9. The quality of coal depends on the percentage of _____ present in it.
10. Burning of a substance in the insufficient amount of oxygen is called _____.
11. Petroleum is a complex mixture of many compounds of _____ and _____ known as _____.
12. Solar cooker is commonly used for harnessing _____.
13. Energy harnessed from flowing water is called _____.
14. Energy sources which cannot be replenished are called _____.

Direction for question 15: Match the entries in column A with appropriate ones from column B.

15.

A.	Peat	( )	a.	liquid fuel
B.	Alcohol	( )	b.	27% of carbon
C.	Decay of biomass	( )	c.	difference in temperature between warm surface waters and colder waters.
D.	Rise and fall of water levels in oceans	( )	d.	biogas
E.	OTEC	( )	e.	solar cell
F.	Ultimate source of energy	( )	f.	tidal energy
G.	Solar energy into electrical energy	( )	g.	sun
H.	Stored in food grains	( )	h.	wind
I.	Moving air	( )	i.	bioenergy

Direction for questions 16 to 30: For each of the questions, four choices have been provided. Select the correct alternative.

16. Which of the following is a false statement?
    (1) To overcome the energy crisis the use of solar cooker must be increased.
    (2) To overcome the energy crisis more amount of non-renewable sources of energy must be used.
    (3) The re-usage of waste material as a source of energy can be done to overcome the energy crisis.
    (4) To overcome the energy crisis water has to be saved.

17. The main constituent of LPG is butane.Then
(A) butane can be liquefied easily under high pressure.
(B) butane is liquefied by chemically reacting with ethane and propane.
(1) Only A is true (2) Only B is true
(3) Both A and B are true (4) Both A and B are false

18. Which element contained in a fuel contributes to its high calorific value?
(1) Carbon (2) Hydrogen (3) Oxygen (4) Nitrogen

19. Combustion, the process of burning a fuel, is _____.
(A) an oxidation and an exothermic reaction.
(B) a reduction and an endothermic reaction.
(1) Only A is true (2) Only B is true
(3) Both A and B are true (4) Both A and B are false

20. _____ is used as a fuel in space ships.
(1) Hydrogen (2) Alcohol (3) Petrol (4) Diesel

21. In solar water heater, a copper pipe with its outer surface painted in black is fixed in the form of a coil in a box.
(1) The only purpose of bending copper pipe is to increase the capacity of water storage.
(2) Bending copper pipe as a coil helps to increase the surface area for heating.
(3) Both (1) and (2) are true
(4) Both (1) and (2) are false

22. In the extraction of some metals from their ores, coke can be used as a/an _____.
(1) oxidizing agent (2) reducing agent (3) catalyst (4) flux

23. Find the false statement from the following statements given below:-
(1) Geothermal power plants cannot operate round the clock.
(2) The initial cost in setting up this plant will be high.
(3) This type of source is free and renewable.
(4) Operating cost involved in a geothermal plant is less.

24. _____are used to produce energy in OTEC.
(1) Tidal energy
(2) Temperature difference between the different layers of water in ocean
(3) Ocean waves
(4) None of the above

25. The crude oil extracted from the earth is separated into its constituents by a process called _____.
(1) disintegration distillation (2) compound distillation
(3) destructive distillation (4) fractional distillation

26. The sun's energy is due to
(1) the nuclear fission of hydrogen. (2) the nuclear fusion of hydrogen.
(3) the natural combustion of hydrogen. (4) the nuclear fission of uranium.

27. Among the following the sources of energy for which source sun is not a chief source of energy is _____.
(1) Hydroelectric power plant.
(2) Ocean thermal energy conversion (OTEC).

(3) Tidal energy.
(4) Biomass.

28. What is the value of solar constant if the energy received by 12 m^2 area in 2 minutes is 2016 kJ?
(1) $1{\cdot}4 \times 10^2$ J s^{-1} m^{-2} (2) 1400 J s^{-1} m^{-2} (3) 84 kJ s^{-1} m^{-2} (4) 84 J s^{-1} m^{-2}

29. A good fuel should _____.
(1) be safe to store and transport
(2) be able to provide desired quantity of energy at a steady rate over a long period of time
(3) have low content of non-combustibles and no combustion products that are poisonous or environmental pollutants
(4) All the above

30. Find the false statement from the following statements given below:-
(1) Geothermal power plants cannot operate round the clock.
(2) The initial cost in setting up this plant will be high.
(3) This type of source is free and renewable.
(4) Operating cost involved in a geothermal plant is less.

Concept Application Level—2

31. Suppose a green tree absorbs, on an average, 10^8 J of solar energy per day, incident on it and the tree could convert 1% of the solar energy incident on it to produce wood, how many days would it take to produce 50 kg of wood?
(The calorific value of wood is 15 kJ g^{-1}).

32. Water that falls from a waterfall is used to generate electricity. If the height of water fall is 100 m and 10^2 m^3 of water falls every minute, then find the hydel energy in MW h generated in one day. Assume that 60% of the energy of the flowing water is converted into electricity. (Take density of water as 1000 kg m^{-3}).

33. Among methane (CH_4) and methyl alcohol (CH_3OH) which is a better fuel and why?

34. 10 kg of water at 30°C is heated to its boiling point on a stove that uses LPG as fuel. If the efficiency of the stove is 70%, then find the mass of the fuel that is consumed. Take the calorific value of LPG (Butane) as 55,000 kJ kg^{-1}.

35. The combustion of 20 g of a certain fuel can melt completely 0.5 kg of ice at 0°C to form water at 0°C. Find the calorific value of the fuel.

(The latent heat of fusion of ice is 336 J g^{-1}).

36. How much time would it take for 1 litre of water to boil, initially at 0°C, if a concentrator type solar heater of area 5 m^2 is used, which can reflect 90% of energy incident on it? (Take solar constant as 1.4 kW m^{-2}).

37. Why is wood not preferred as a source of energy?



38. A power plant uses 100 tonnes of coal to produce a certain amount of electric energy. To produce the same amount of electric energy, find how much nuclear fuel U^{235} would be required by a nuclear power plant. The calorific value of coal is 30 kJ g^{-1} and a nucleus of U^{235} on fission releases about 200 MeV energy per fission. (1eV = 1.6×10^{-19} J). Assume that the energy conversion efficiency is 50% for coal and 25% for fission of U-235 nucleus.

39. Find the electrical power generated in 1s by a wind mill that is erected in a location when air flows with an average speed of 54 km h^{-1}. Assume that the moving air is completely stopped by the blades of the wind turbine of area 10 m^2, over which air flows normally and the energy conversion from wind energy to electrical energy is only 50%. (Take density of air = 0.3 kg m^{-3})

40. The solar energy incident normal to the surface of the water in a container raises its temperature from 10°C to 20°C, in one minute. If the mass of water in the container is 1400 g, then find the surface area of the water in the container (solar constant = 1.4 kW m^{-2}) [Neglect the heat absorbed by the container].

41. What is the source of energy in artificial satellites? Why is this source of energy not used for meeting all our domestic electricity needs?

42. A nucleus of U-235 on fission releases about 200 MeV energy per fission. (1 MeV = 1.602×10^{-13} J). Estimate how much coal (calorific value = 33 kJ g^{-1}) has to be burnt to obtain energy equivalent to that released in the fission of 1g of U-235 assuming that energy conversion efficiency of 30% for charcoal and 25% for fission of U-235 nucleus.

43. An engine motor of 1 Hp power consumes 74.6 kg of a variety of biofuel to operate for one hour. If the efficiency is 100%, then find the calorific value of the fuel.

44. 1 kg of water at 0°C is taken in a vessel of water equivalent 100 g. This vessel is kept at the focal plane of a concentrator type solar heater of area 10 m^2, which can reflect 90% of energy incident on it. Find the time taken to raise the temperature of water by 100°C. (Take solar constant as 1.4 kW m^{-2}, specific heat capacity of water as 4.2 kJ kg^{-1} $°C^{-1}$).

45. Why is it not feasible to produce hydroelectricity by spending energy in lifting a huge amount of water? Explain.

Concept Application Level—3

46. Why is geothermal energy better than wind energy for power generation?

47. "A burning match stick can ignite a piece of paper or an incense stick but cannot burn a wooden block". Give reasons.

48. The sun is the ultimate source of energy on earth. Is the given statement true with respect to fossil fuels?

49. Calculate the energy emitted by the sun in one second. Also estimate the mass of the sun that is converted to energy in one second.

(Take solar constant as 1.4×10^3 W m^{-2}, distance between the earth and the sun as 1.5×10^{11} m and the velocity of light as 3×10^8 m s^{-1}).

50. Why is hydrogen not used as a fuel for domestic purposes, even though it has a high calorific value?

Very short answer type questions

1. Nuclear fusion reaction.
2. J kg^{-1}
3. Solar panel
4. Solar energy.
5. The minimum temperature at which the substance begins to burn.
6. kerosene
7. Light energy changes into electrical energy.
8. Selenium, silicon.
9. infrared
10. Energy obtained from sun in the form of light and heat.
11. (i) Materials that can be burnt–Combustible.
    (ii) Materials that cannot be burnt–Non-combustible.
12. photovoltaic cell
13. Renewable and non-renewable sources of energy.
14. A process in which a heavier nucleus splits into lighter nuclei with release of huge amounts of energy.
15. Charcoal
16. Burning a substance in presence of air or oxygen or a supporter of combustion.
17. (i) An oil derived from rocks. A complex mixture of hydrocarbons.
    (ii) Natural gas is a fossil fuel that is found on the top of petroleum under the earth.
18. net energy output
19. Process of burning wood in limited supply of oxygen.
20. Fuels obtained from the decay of plants and marine animals.
21. Silver
22. Solid, liquid and gaseous fuels.
23. Energy of flowing water is utilized to produce electricity on a large scale.
24. renewable sources
25. Petrol – presence of hydrogen.
26. A device that converts wind energy to mechanical or electrical energy.
27. peat
28. Energy liberated when a unit mass of fuel is burnt completely.
29. A device which directly converts solar energy into electrical energy.
30. infrared radiations

Short answer type questions

31. Wind mill, turbine.
32. (i) Solid fuels
    (ii) Liquid fuels
    (iii) Gaseous fuels
33. (i) Net energy output.
    (ii) Easy availability.
    (iii) Safe and convenient to use.
    (iv) Easy for storage and transportation.
    (v) Economical.
34. Steam from water, turbine runs with the help of steam.
35. (i) Main component is methane.
    (ii) Compressed under high pressure to form CNG.
    (iii) Used as fuel in automobiles.
36. (i) Renewable sources of energy – solar energy, wind energy.
    (ii) Non-renewable sources of energy – coal, petrol.
37. (i) Energy liberated when a unit mass of a fuel is burnt completely.
    (ii) Hydrogen.
38. Digester tank, raw material, inlet tank, outlet tank.
39. (i) Solid fuels – Wood, coal
    (ii) Liquid fuels – Petrol, kerosene
    (iii) Gaseous fuels – Natural gas, biogas

40. Box type solar cooker:
- (i) Rectangular wooden box, lined from inside with insulating material which prevents heat losses.
- (ii) Inner walls painted black.
- (iii) A plane mirror hinged at the top, which can be adjusted such that it reflects sunlight into the box.

41. (i) More economical to construct.
(ii) Need of abundant water in the dam.
(iii) Can be constructed only at certain places.

42. Formation:
- (i) Dead remains of plants and animals that settled in muddy sediments at the bottom of the sea millions of years ago.
- (ii) Got buried deep below the surface of the earth.
- (iii) Due to intense heat and high pressure, remains ultimately converted into petroleum.

Extraction:
- (i) Large reservoirs.
- (ii) Pumped out by drilling holes and digging wells.

43. (i) Easy availability, low cost.
(ii) High calorific value.
(iii) Easy storage and transport.
(iv) Ignition temperature above the room temperature.
(v) Low rate of evaporation at room temperature.

44. Minimum speed of air, cost of construction and erection.

45. Concave reflection surface, cooking items at focus.

Essay type questions

46. (i) A plant designed to produce biogas.
(ii) Types of biogas plants—Its construction and working.

47. (i) Fractional distillation.
(ii) Principle:-Different fractions of petroleum boil at different temperature ranges.

48. (i) Every source of energy either directly or indirectly is manifested from solar energy.
(ii) Flow chart for the various functions performed by solar energy.

49. (i) Fossiled remains of the plants.
(ii) Peat, lignite, bituminous, anthracite.

50. Huge stem like body, impellors, connecting levers and turbine.



Concept Application Level—1

True or false

1. False
2. False
3. False
4. False
5. True
6. True
7. False

Fill in the blanks

8. room temperature
9. carbon
10. destructive distillation
11. hydrogen, carbon, hydrocarbons

12. Solar energy

13. hydel energy

14. non-renewable sources of energy

Match the following

15. A : b
B : a
C : d
D : f
E : c
F : g
G : e
H : i
I : h

Multiple choice questions

16. Choice (2)

17. Choice (1)

18. Choice (2)

19. Choice (1)

20. Choice (1)

21. Choice (2)

22. Choice (2)

23. Choice (1)

24. Choice (2)

25. Choice (4)

26. Choice (2)

27. Choice (3)

28. Choice (2)

29. Choice (4)

30. Choice (1)

Concept Application Level—2,3

Key points

31. (i) Find q (i.e.,) 1% of 10^8 J. This is the solar energy that is converted to wood in one day.
Find the amount of heat energy liberated on burning 50 kg of wood by using the formula, Q = mS -- (1) where S is the calorific value of wood.
Then, find the number of days required to produce 50 kg of wood by using formula, Q = nq = mS ------- (2)

(ii) 750 days

32. (i) Find the amount of water (m) flows in one day.
Let the volume of water that flows in every minute = V.

(ii) Density of water (d) = $\frac{m}{V}$

(iii) One day = 24 × 60 minutes

(iv) Mass of the water that flows in one day is m = Vd × (60 × 24)

(v) Potential energy, PE = mgh

(vi) The energy of the flowing water in one day is = mgh

(vii) 60% of the (mgh) is used to generate electricity

(viii) 24 × 10^3 kW h or 24 MW h

33. (i) Which one of the given fuels-methane and methyl alcohol, has got more percentage of hydrogen?
Among the given molecules, which has got simpler molecular structure?

(ii) CH_4 is a better fuel.

34. (i) Find the amount of heat (Q) required to boil given water by using Q = ms(Δ t).

(ii) The above value of Q is only 70% of the amount of heat liberated by combustion of fuel (Q^1)

(iii) Then the mass of the fuel combusted is,
$m_{fuel} = \frac{Q^1}{S}$

(iv) 76 g

35. (i) Find the amount of heat energy required to melt 1/2 kg of ice at 0°C to water at 0°C by using formula, Q = mL.
The amount of heat energy absorbed by ice is supplied by burning 20 g of fuel, then find the amount of heat energy produced by burning the fuel by using formula Q = mS.

Where S is the calorific value of the fuel.
Now, equate (1) and (2), and obtain the value of 'S'.

(ii) 8.4 kJ g^{-1}

36. (i) Find the change in temperature of water of 1 litre.
Then, find the amount of heat energy absorbed by water to raise its temperature using the formula, Q = ms Δt. ------ (1)
Convert (1) in S.I system by taking s= 4200 J $kg^{-1}°C^{-1}$
Find the percentage of incident solar energy that causes rise in temperature of water.
Then, the heat absorbed by water is equal to the heat produced in solar cooker.
Then, find the heat produced in solar heater by using formula,

$$Q = (\text{Solar constant})\ (\text{Area})\ (\text{time}) \times \frac{90}{100}$$

--- (2)

Substitute the appropriate values in (1) and (2).
Equate (1) and (2), and find the value of time.

(ii) 1.11 minute

37. Find the calorific value of wood with respect to the other sources of energy.

38. (i) Heat released by combustion of 100 tonnes of coal is, Q = mS
$\Rightarrow$ Q = 100 × 1000 kg × 30 kJ kg^{-1} → (1)

(ii) Find 50% of the 'Q'.

(iii) Let above value is Q^1.

(iv) The amount of heat released by fission process should be equal to = $\frac{100 \times Q^1}{25}$ J

(v) 1 fission produces → 200 × 10^6 x 1.6 × 10^{-19} J

(vi) Find the number of fissions to produce $\frac{100 \times Q^1}{25}$ J

(vii) 235 grams of U^{235} produces 6.023 × 10^{23} fissions

(viii) Find the amount of U^{235} that produces $\frac{100 \times Q^1}{25}$ J.

(ix) 0.07 g

39. (i) The volume of air that makes blades to rotate = V = ℓ × area of the blade

(ii) Velocity (v) = $\frac{\ell}{t}$

(iii) Mass of the air = m = volume × density of air

(iv) Kinetic energy (KE) of air = $\frac{1}{2}mv^2$

(v) 50% of kinetic energy of air is converted into electrical energy (EE)

(vi) The power of electrical generation is, P = $\frac{EE}{1\ s}$

(vii) 2.53 kW

40. (i) By using formula, Q = ms Δt, find the amount of heat energy absorbed by the water. --- (1)
Is this heat energy absorbed by the water equal to the solar energy that is incident on the water.
Then, find the solar constant and time for which solar energy is incident on water surface.
Find the amount of solar energy incident on water by using formula,
Q = (solar constant) (Area) (time) --- (2)
Then, substitute appropriate values and obtain the value of the area.

(ii) 0.7 m^2

41. What is the source of energy (Renewable) in outerspace?
Is it economical to harness solar energy?
Is there steady availability of the source of energy?
Then, find the efficiency of energy conversion into electrical energy.

42. (i) Find the Avogadro number.
How many atoms are contained in 235 g of uranium?
Then, find the number of atoms present in 1 g of uranium.
How many fissions are produced by 1 atom of uranium?
Find the energy of 1 eV in terms of joules.
Convert 200 MeV into joules by multiplying 200 with $(10^6 \times 1.6 \times 10^{-19})$.
Find the heat liberated by fission of 1 g of U – 235. ----- (1)
The amount of heat liberated by burning coal is = (m as of coal) × (its calorific value) ----- (2)
Find the energy conversion efficiencies of coal and uranium.
Then, equate the heat energies liberated by fission of uranium and combustion of coal.
From this, find mass of coal.

(ii) 2000 kg.

43. (i) The power of the engine = 1 Hp = 746 J $s^{-1} = \frac{Q}{3600\ s} \rightarrow$ (1)

(ii) Find the value of Q from (1)

(iii) We know, $Q = mS \rightarrow$ (2)

(iv) Equate (1) and (2) and find the value of 'S'.

(v) $S = 36{,}000\ J\ kg^{-1}$

44. (i) Solar energy incident in time "t" is $= Q = 1.4 \times 10^3 \times t\ J \rightarrow$ (1)

(ii) $Q = m_w s_w (\Delta t) +$ (water equivalent of vessel) $s_w(\Delta t)_w$; $\Delta t = 100°C$

(iii) $\frac{90}{100} Q = 1\ kg \times 4200\ J\ kg^{-1}\ °C^{-1} \times 100°C + \frac{1}{10}\ kg \times 4200\ J\ kg^{-1}\ °C^{-1} \times 100°C \rightarrow$ (2)

(iv) From (1) and (2), find the value of 't'.

(v) 36.6 s

45. How much energy is needed in lifting water to a certain height?
Then, this energy stored in the water is converted to produce electricity.
What is the efficiency of conversion to electrical energy?
Is it economical, if electricity produced is less than the energy spent in lifting water?

46. From given sources of energy ,which one is available for 365 days in a year and 24 hours in a day?

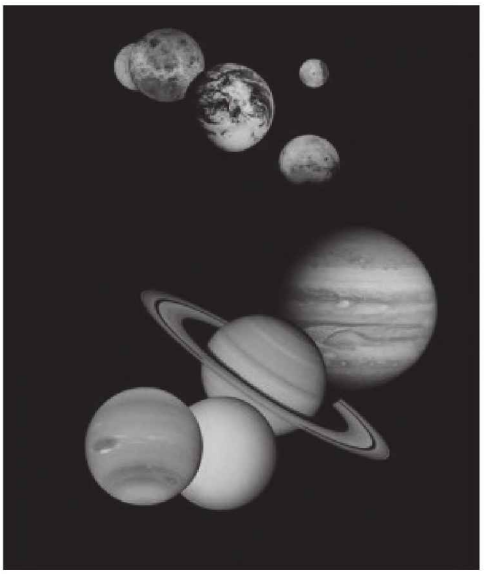
47. Is the cross section of wooden block more?
Will the end of the incense stick completely covered by flame of the burning match stick.
Then, the end exposed to flame reaches ignition temperature. In case of wooden block, is the energy supplied by match stick enough to reach ignition temperature?

48. What are fossils?
Are these impressions (or) remaining of past living creatures?
Do the living beings depend on sun for their survival?
Are these dead creatures converted into fuels?
Then which is the ultimate source of energy for the fossil fuels?

49. (i) Is the energy emitted by the sun in one second equal to the energy reaching an imaginary sphere of radius equal to the distance between the sun and the earth?
Take the distance between the sun and the earth as radius of the sphere.
Then, the total surface area (T.S.A) of the sphere is $4\pi r^2$.
Then, energy received in one second (E) = (solar constant) (T.S.A).
Find the velocity of light (c) from given data.
Find the mass of the sun that is converted into energy in one second by using the formula, $E = mc^2$.
Then, $m = \frac{E}{c^2}$.

(ii) $4.14 \times 10^9\ kg\ s^{-1}$.

50. Is it more difficult to store, handle and transport the hydrogen fuel?



11

Electronics

INTRODUCTION

Today computers play an important part in our everyday life. The use of computers today in many fields is the result of research and development carried out in the field of electronics, the branch of physics that deals with the study of conduction of electrons and the related phenomena. From their primitive stage to the present stage, computers have passed through many generations. In the early stages of development a computer was as big as a huge building. The essential components that were used in the computers in those days were vacuum tubes, which are bulky; and that is the reason for the huge size of computers. These computers were referred to as the first generation computers.

Later, some electronic devices called semiconductor diodes and triodes were invented and used in the circuits of these electronic devices like computers which increased their speed and efficiency. The replacement of vacuum tubes with semiconductor diodes and triodes has reduced the size of computers as well. These computers are referred to as the second generation computers. The development of semiconductor technology led to the improvisation of new devices and a large number of semiconductor diodes and triodes were arranged in a single circuit in a miniature form called integrated circuit (generally referred to as an IC or a chip). The computers that consist of ICs in their circuits are referred to as the third generation computers. The use of ICs in computers increased their efficiency. The fourth generation computers contain very large scale integrated circuits, i.e., large number of chips are integrated in a small circuit; and are referred to as microprocessors. Today microprocessors are used in the electronic circuits of most of the electronic devices.

In the current chapter the fundamentals regarding electronic conduction and some of the basic components that are the building blocks of an electronic circuit are discussed in addition to the fundamentals of computers.

The term electronics comes from the word electron. Electronics is a branch of applied physics which deals with controlled motion of charged particles. Electronics is a vast ocean of knowledge and the applications of electronics are endless.

Radio, T.V. broadcast, air transport, control of rockets and space craft, electron microscope, oscilloscope, remote controlled devices, electro radiograph and computer operated machinery are some of the devices which make use of electronics.

The branch of electronics which deals with vacuum tubes is called tube electronics.

Thermionic Emission—Production of Electron Stream

It is obvious that electrical conduction in solids takes place due to the drift of electrons through them. In liquids electrical conduction takes place through motion of ions; in gases the conduction takes place only at a low pressure. A gas contained in a tube conducts electricity at a low pressure only when electrodes exist in it, similar to that in a discharge tube, and when the electrodes are connected to a source of electricity.

Thus electrons are emitted from one electrode and pass through the gas towards another electrode, thereby conducting electricity through the gas. The electrode from which electrons are emitted is called a 'cathode' and the other electrode is called an 'anode'. Thus, electric conduction in a gas present in a tube takes place at a low pressure from the cathode to the anode.

It is found that the movement of the electrons from the cathode to the anode takes place even in vacuum. Thus electronic conduction is possible in a vacuum tube that contains an anode and a cathode. Such vacuum tubes were used in the circuitry of the first generation computers, mentioned earlier.

The electrons that are emitted from the surface of the cathode in a vacuum tube are called free electrons. The emission of free electrons from the surface of a cathode is an important phenomenon. The electrons, that revolve around the nucleus in the outermost orbit of an atom are valence electrons. These valence electrons are loosely bound to the nucleus and can be made free by supplying a little energy.

Thus on supplying energy to the cathode of the vacuum tube, free electrons are emitted from its surface. The energy to be supplied to the loosely bound electrons of the cathode to make them free, can be in the form of heat, light or high energy ultra violet radiation or X-ray radiation. If the energy supplied is in the form of heat, the phenomenon of electron emission is called thermionic emission.

Thus thermionic emission is defined as "the phenomenon in which free electrons are emitted from a metallic surface, by the absorption of heat energy supplied to them". The electrons thus emitted are called thermions.

The free electrons of the metallic cathode are free in the sense that they can move about from one atom to another within the metal. When heat is supplied to the cathode, the free electrons at the surface of the metal gain enough kinetic energy so that they move away from the cathode and are no more bound to the cathode surface.

The amount of heat energy to be supplied to a metallic surface for it to liberate free electrons (thermions) is different for different metals and depends on their nature. However, for a given metallic surface, there exists a minimum heat energy required for a thermion to be liberated and this minimum heat energy is called **'threshold energy'** or **'work function'**. The threshold energy or work function is usually measured in electron-volts (eV). The corresponding temperature of the metal is known as 'threshold temperature'.

Rate of thermionic emission—factors affecting it

Thermionic emission from a metallic surface is measured by its rate, i.e., the number of thermions emitted per unit time. This rate of thermionic emission depends on the following factors.

1. The rate of emission of thermions depends on the nature of the metallic surface. The work function of each metal is different. If the work function is less, the rate of emission of thermions is more. Thus the rate of emission of thermions is inversely proportional to the work function of metals.
2. The rate of thermionic emission depends on the temperature of the cathode. The higher is the temperature of the surface of the cathode, the more is the rate of emission of thermions. Thus the rate of thermionic emission is directly proportional to the temperature of the cathode.
3. The rate of thermionic emission depends on the surface area of the cathode. The more is the surface area of the cathode, the more is the thermionic emission. Thus, the rate of thermionic emission is directly proportional to the surface area of the cathode.

Keeping in view the above factors on which the rate of thermionic emission depends, the characteristics of a good thermionic emitter can be asserted as follows.

1. A good thermionic emitter should have a low work function. When an emitter has a low work function, less amount of heat energy is required to liberate thermions from its surface. Thus, as more amount of heat is supplied to it, a larger number of thermions are emitted from it, increasing the rate of thermionic emission.
2. A good thermionic emitter should have a high melting point. If a thermionic emitter has a high melting point, it can be heated to a high temperature and thus the rate of thermionic emission can be increased.

Following are some of the thermionic emitters generally used.

(a) Tungsten: It has a high melting point of 3665 K. Its work function is 4.52 electron volt (eV) and starts emitting thermions at a temperature of about 2500 K.

(b) Thoriated tungsten: A tungsten filament on which a mixture of thorium and carbon is coated is known as thoriated tungsten. The advantage in this emitter is that heat is produced by passing electric current through tungsten filament whereas thermions are emitted from thorium. The work function of thorium is 2.6 eV corresponding to a temperature of about 2000 K. If the temperature is increased beyond 2000 K, the rate of thermionic emission also increases.

(c) Alkali metal oxides: The work function of alkali metal oxides is much less than that compared to that of the above-mentioned thermionic emitters. It is less than one electron-volt, and corresponds to a temperature of about 1000 K. Thus a tungsten filament coated with alkali metal oxides like oxides of barium, strontium, cesium, etc., is an efficient thermionic emitter.

Kinds of thermionic emitters

Thermionic emitters can be broadly classified into two categories. They are directly heated thermionic emitters and indirectly heated thermionic emitters.

In the first category of emitters, current is passed through the emitter, thereby its temperature increases. When the temperature exceeds the temperature corresponding to the work function of the emitter, it starts emitting thermions. Tungsten filament is an example of directly heated emitters.

In the second category of emitters, heat is supplied to the emitter by a heating element like tungsten and when the temperature of the emitter crosses the temperature corresponding to its work function, i.e., its threshold temperature, it starts emitting thermions. Thoriated tungsten and alkali metal oxides fall under this category of emitters.

Applications of thermionic emission

The phenomenon of thermionic emission is used widely in some of the basic components used in electric circuits of electronic devices like radio, cathode ray tube (CRT) or cathode ray oscilloscope (CRO) etc. One such basic component is a diode valve.

Diode valve

A diode valve is an electronic device that consists of two electrodes and is used to allow the flow of electrons in a circuit in a particular direction.

There are two types of diode valves, one that contains a directly heated thermionic emitter and the other that contains an indirectly heated thermionic emitter.

Directly heated diode valve

The following figure (11.1) shows the schematic representation of a directly heated diode valve used in an electric circuit. It consists of a vacuum tube in which a tungsten filament is the cathode and an aluminium cylinder that surrounds the cathode, acts like an anode. Anode is also referred to as a plate. Figure (11.2) shows the symbolic representation of a diode valve in an electric circuit. The filament (F) is connected to a low tension battery (LT) via an ammeter and a variable resistance. The current through the filament can be changed by varying the resistance and thereby the emission of thermions from the cathode can be controlled. The high-tension battery (HT) in the circuit provides the potential difference between the cathode and the anode so that the emitted thermions

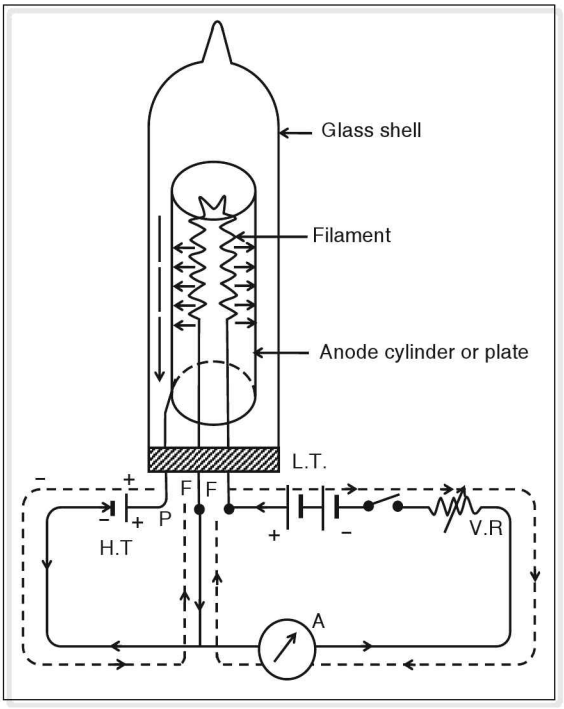


Figure 11.1 A directly heated diode valve

from the cathode move towards the plate (P) and conduct electricity in the circuit. This diode is not efficient as the work function of tungsten is high and even the number of thermions emitted is not large.

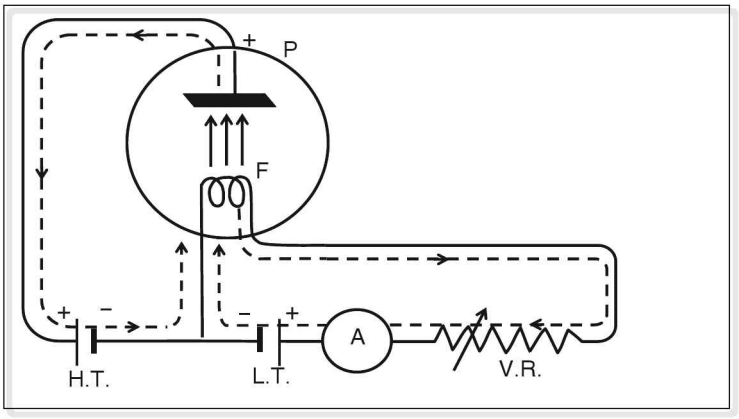


Figure 11.2 Symbolic representation of a directly heated diode valve

Indirectly heated diode valve

In an indirectly heated diode valve, the filament producing heat is surrounded by a cathode that emits thermions, as shown in figure (11.3). The terminals of the filament are connected to a low tension battery via an ammeter and a variable resistance. Adjusting the variable resistance regulates the current through the filament and thus the regulation of thermionic emission from the filament takes place. The plate and the cathode are connected to a high tension battery. The symbolic representation of an indirectly

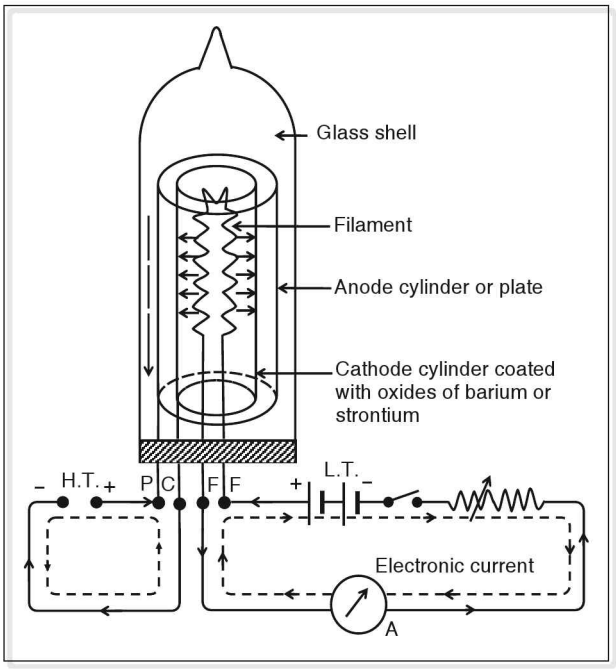


Figure 11.3 An indirectly heated diode valve

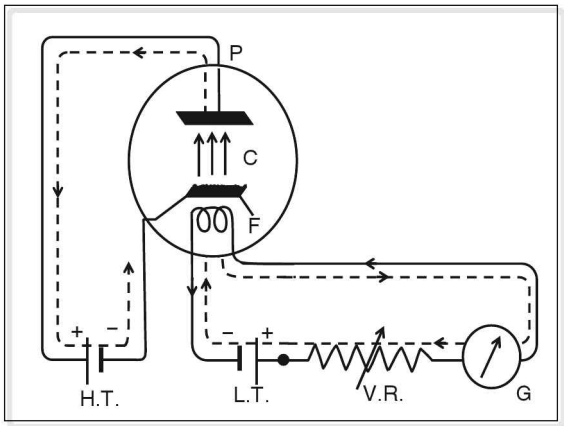


Figure 11.4 Symbolic representation of an indirectly heated diode valve

heated vacuum diode valve is shown in figure (11.4). As the work function and threshold temperature of the cathode is less than that of the filament, thermionic emission takes place at a comparatively lower temperature than that of the filament. As the temperature of the cathode is increased, more thermions are emitted from its surface.

Space charge

In a vacuum diode valve, there are two circuits, one for the filament and the other for the conduction of thermions. Out of these two circuits, if only the filament circuit is switched on, the filament gets heated and produces thermions. The drift of thermions towards the plate depends on the potential difference between the plate and the cathode. If the potential difference mentioned above is zero or very less, there is no drift of thermions emitted from the cathode towards the plate and the emitted thermions are accumulated over the cathode in the form of electron cloud, called space charge. The space charge around the cathode exists as long as the rate of drift of thermions towards the anode is less than the rate of thermion production.

Cathode ray oscilloscope (CRO)

A cathode ray oscilloscope (CRO) or a cathode ray tube (CRT) is a device that converts electrical signals into visual signals. The device is used to detect the nature and features of electric signals, different wave forms and also the changes in the potential difference of alternating currents.

A cathode ray oscilloscope essentially consists of an electron gun, a set of deflecting plates and a fluorescent screen. Figure (11.5) represents a cathode ray oscilloscope.

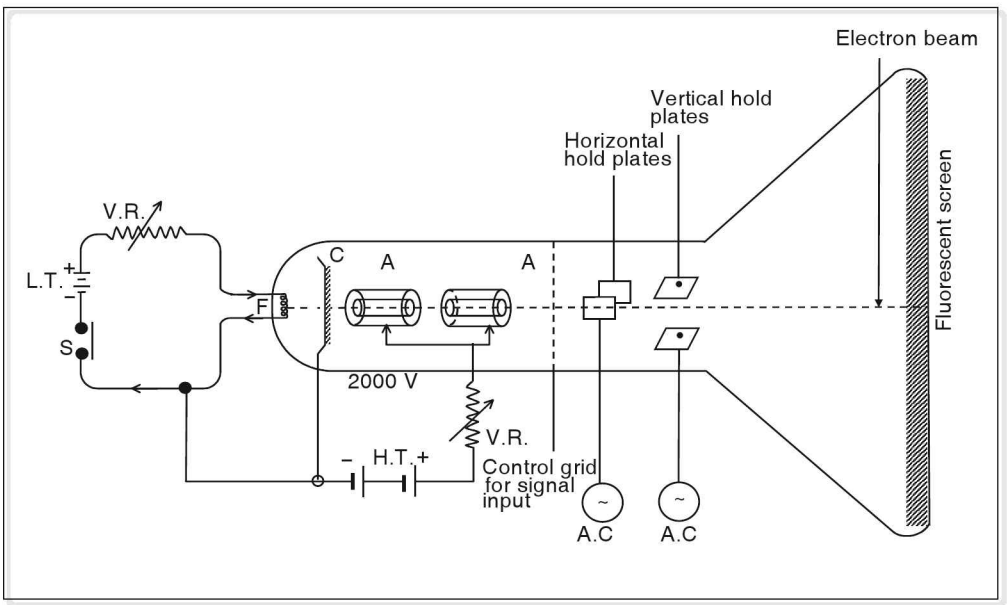


Figure 11.5 Cathode ray oscilloscope

The electron gun consists of a tungsten filament connected to a low tension battery via a switch and a variable resistance. To the front of the filament, a cathode plate coated with barium oxide or strontium oxide is mounted and it generates thermions. To the front of the cathode a double anode of cylindrical

shape is mounted. The anode contains a fine hole in it. This is provided so as to accelerate the electrons (thermions) emitted from the cathode. In front of the anode, a control grid is mounted to which the electric signals are fed.

The fine electron beam emitted from the cathode and accelerated through the anode, after passing through the grid passes through two pairs of deflecting plates that are connected to a high voltage source of alternating current. One pair of deflecting plates is placed in vertical position and produces deflections in the horizontal plane. The other set of deflecting plates is positioned horizontally and causes deflections in the vertical plane. Thus the operation of these pairs of deflecting plates causes the electron beam to scan the whole screen.

The electron beam that passes between the deflecting plates impinges on the fluorescent screen and causes scintillations. Thus the electron beam scans the screen with a bluish white glow. The intensity of the electron beam is controlled by the variable resistance in the low tension circuit. The acceleration of the emitted electrons is controlled by changing the potential difference between the cathode and the anode in the high tension circuit. The electron beam emerging through the anode is repelled after it passes through the control grid depending on the electric signal fed to the grid. This causes variation in the number of electrons striking the screen and hence in the brightness of the light observed on the screen. If the number of electrons striking the screen is more, a white dot is observed at the place or else, a black dot is observed. This is the basic principle adopted in television picture tubes and computer monitors.

But vacuum tubes are outdated and are replaced by semiconductors. The branch of electronics which deals with semiconductors is called solid state electronics.

Semiconductors are preferred to vacuum tubes because of their compact size, low power consumption and economy.

Based on the electrical properties, materials are classified as conductors, insulators and semiconductors.

Semiconductors are devices whose electrical properties (conductivity or resistivity) lie between that of conductors and insulators.

Examples of semiconductors: Germanium, silicon, cuprous oxide (Cu_2O), gallium arsenide etc.

It is obvious that electrons in an atom revolve around the nucleus in elliptical orbits. The energy of electrons in a given orbit is fixed and thus the orbits are referred to as stationary orbits. The energy of the electrons increases as the radius of their orbits increases. The energy associated with electrons in a given orbit is referred to as their energy level. In an isolated atom, the energy levels of electrons are discrete and there exists an energy gap between two energy levels that is equivalent to the difference in the energy of the two levels. In a crystal, the atoms are not isolated. They are fixed to their positions by chemical bonds between them. The interatomic spacing in a crystal is of the order of one angstrom. In the process of forming a crystal, atoms come closer. In this process, the outermost

orbits of the atoms come much closer. They superpose on one another and finally spread into a band of energy levels within a small space and this group of energy levels is called an 'energy band'. The electrons present in an energy band have no fixed energy and their energy varies continuously within the range of the energy band. Thus an electron of an atom in a given energy band is free to move within any of the orbits of atoms present in the band.

Electrons of one atom interact with electrons of the neighbouring atom. As a result each discrete energy level splits into as many number of separate levels as the number of interacting atoms. The split energy levels lie very close to each other and they form a continuum known as the energy band.

Energy bands in a silicon crystal

A silicon crystal consists entirely of silicon atoms. The atomic number of silicon is 14 and its electronic configuration is $1s^2\ 2s^2\ 2p^6\ 3s^2\ 3p^2$. In its outermost orbit, there are 4 valence electrons, whereas it can accommodate a total of 8 electrons. When many silicon atoms are brought together to form a silicon crystal, their outermost energy levels are superposed to form an energy band. As the inter-atomic distance is decreased, the energy band corresponding to 3s level is superposed on the energy bands corresponding to 3p level. In this situation, the electrons present in the 3s energy levels are free to move even in the 3p energy level. When the inter atomic spacing is still reduced and is equal to the normal interatomic spacing in a crystal (shown as 'r' in figure (11.6)), the combined energy band is split into two with a gap called energy gap (E_g) formed between them.

Consider that there are 'n' number of atoms in a crystal, then the 3s and 3p energy levels of all the 'n' number of atoms are superposed to form a '3s' energy band and a '3p' energy band. Since there are 'n' number of atoms, '2n' number of electrons occupy the '2n' number of available '3s' energy states and similarly '2n' number of electrons occupy '6n' number of available '3p' energy states. When all the '3s' and '3p' energy states are superposed, there are '4n' number of electrons which occupy the '8n' number of energy states. At the normal inter-atomic spacing, the total energy bands are split into two, each having '4n' energy states in each energy band. All the '4n' electrons in the crystal occupy the '4n' states at the lower level and the '4n' states in the upper level are left vacant. The energy band at the lower level which is occupied by '4n' electrons is called **valence band** (V.B) as the electrons which occupy those states are valence electrons. When these valence electrons gain energy equivalent to the energy gap (E_g) and transit to the higher energy level, they are free to move within the crystal and conduct electricity. Thus the energy band at the higher level is called **conduction band** (C.B.), since no electrons exist in the energy gap (E_g), it is also called **forbidden gap**.

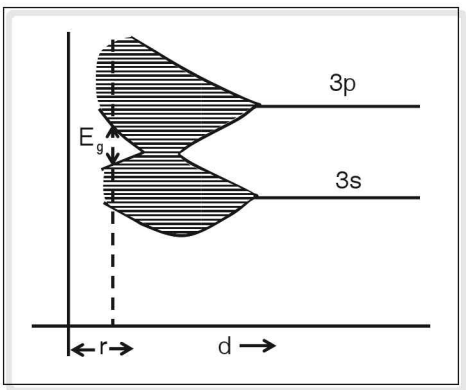


Figure 11.6 Energy bands in a silicon crystal

Classification of solids based on band theory

Solids are classified into conductors, insulators and semiconductors based on their energy bands.

Conductors

Energy band structure of conductors are of two types.

1. The conduction band and valence band are separated by a very small gap. The conduction band is partially filled with electrons at zero kelvin. By application of a very small field, the conductor starts conducting.

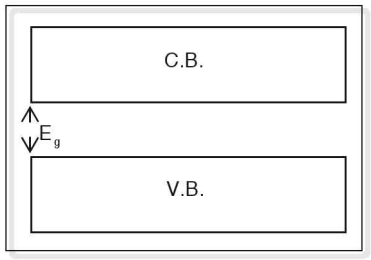


Figure 11.7

2. In the second type, the valence band and the conduction band overlap. Hence there is no forbidden energy gap. Thus the resistance of these conductors is very low. The electrical conduction takes place even when a very small electric field is applied.

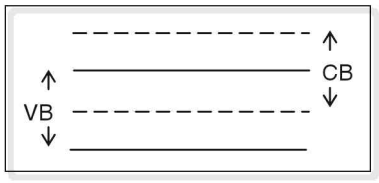


Figure 11.8

Insulators

In insulators, the valence band and the conduction band are separated by a large energy gap (E_g). The valence band is completely filled and the conduction band is completely empty. Since there are no free electrons, electrical conduction does not take place. Further due to large energy gap, electrons are unable to jump to the conduction band even if electric field is applied. Therefore electrical conduction is impossible in insulators.

☛ *Example* E_g of diamond is of the order of 6 eV.

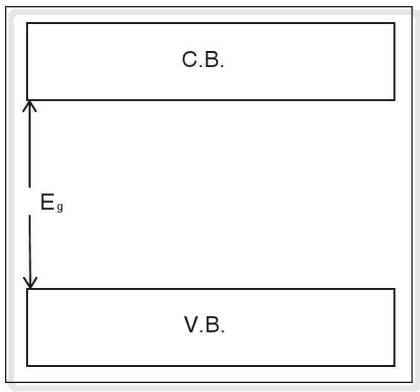


Figure 11.9

Semiconductors

In semiconductors, the valence band and the conduction band are separated by a small forbidden energy gap. At absolute zero, the valence band is filled, the conduction band is empty and no electrons are available for conduction. Hence at zero degree kelvin, semiconductor behaves like a perfect insulator. At room temperature, some of the electrons gain thermal energy and jump to the conduction band. They are free to move in the crystal and are available for conduction.

E_g of silicon = 1.1 eV

E_g of germanium = 0.75 eV

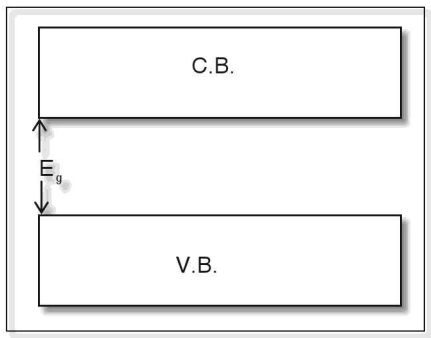


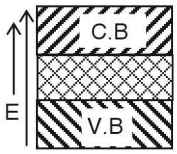
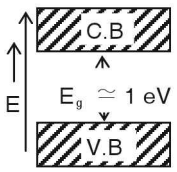
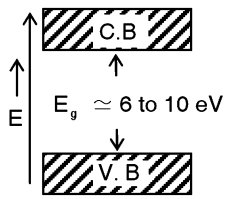
Figure 11.10

C.B → Conduction Band

V.B → Valence band

E_g → Energy gap

The following table compares conductors, semiconductors and insulators.

Number of free electrons	Of the order 10^{28} per cm^3	Of the order 10^{17} per cm^3	Of the order 10^7 per cm^3
Energy gap	Nil. Both the conduction band and valence band overlap.	Of the order 1 eV. For silicon, it is 1.1 eV and for germanium, it is 0.7 eV.	Of the order 6 eV to 10 eV.
Electric conductivity	Highest	Lesser	Nil or negligible
Temperature coefficient of resistance	As the temperature is increased, the collisions of free electrons with positive ions increase and thus the electrical resistance increases. Thus conductors have positive temperature coefficient of resistance.	As the temperature of a semiconductor is increased, more number of free electrons are released and so its conductivity increases. Thus its electrical resistance decreases with increase in temperature. Hence semiconductors have negative temperature coefficient of resistance.	Similar to that of semiconductors, the resistance of an insulator is decreased, when its temperature is increased and thus also has a negative temperature coefficient of resistance.
Band diagrams VB → Valence band CB → Conduction band E_g → Energy gap			
Usage	For electrical conduction.	In electronic devices.	As electric shock proofers.
Examples	All metals and alloys	Silicon (Si) and germanium (Ge)	Plastics, rubber, wood, glass, ebonite, PVC, etc.

Effect of temperature on semiconductor

A semiconductor is a perfect insulator at absolute zero, since the conduction band of a semiconductor is completely empty. But as the temperature increases, the electrons gain energy and jump to the conduction band. Thus the conductivity of a semiconductor increases with temperature.

In general the conductivity of a semiconductor can be increased not only by heating but also by

1. exposing to radiation
2. adding impurities to a semiconductor

Semiconductors are classified into intrinsic semiconductors or pure semiconductors and extrinsic semiconductors or impure semiconductors.

Intrinsic semiconductor

A semiconductor in the purest form is called an intrinsic semiconductor. Examples of intrinsic semiconductors are silicon crystals and germanium crystals i.e., every atom in the crystal is of the same type.

Silicon or germanium atom is a tetravalent element i.e., it has four valence electrons. Each valence electron forms a covalent bond with the neighboring atom. Each silicon atom is tied down in the crystal by four covalent bonds as shown. Hence in the pure silicon crystal all the valence electrons are involved in bonding process and there are no free electrons available for conduction. Therefore pure silicon or pure germanium is a perfect insulator at absolute zero.

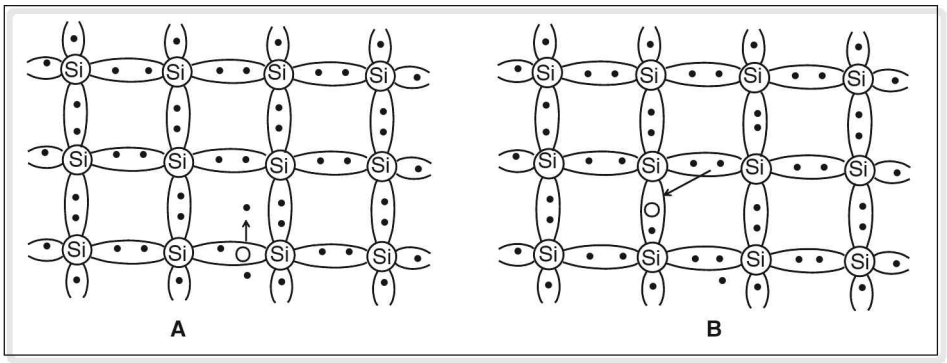


Figure 11.11

As the temperature increases, it causes the atoms to vibrate and this vibrational energy of atoms is sufficient to break the covalent bonds and release electrons. Some of the electrons gain sufficient energy to overcome the forbidden energy gap and these electrons jump from valence band to the conduction band and are available for conduction.

When an electron moves from valence band to conduction band a vacant site is created in the valence band and this vacant site behaves as net positive charge of magnitude +e. This vacant site is called hole. For every electron released, a hole is formed. Thus the number of electrons is equal to the number of holes in an intrinsic semiconductor. A hole can capture or attract an electron from the vicinity and a hole is formed at the site of the captured electron.

When potential difference is applied, the holes move in a direction opposite to that of electrons. Hence there are two types of charge carries in a semiconductor.

Hence both the electrons and holes contribute to the current and thus they are referred to as charge carriers. If 'I_n' indicates the hole current and 'I_e' indicates the electronic current, the total current in the

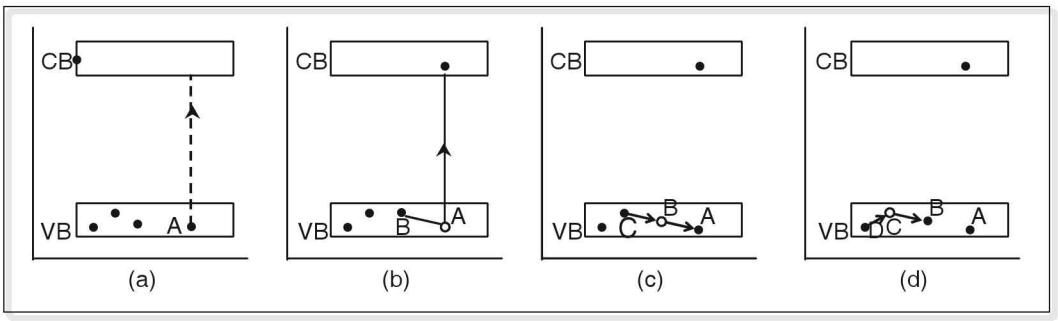


Figure 11.12 Formation of hole and hole movement

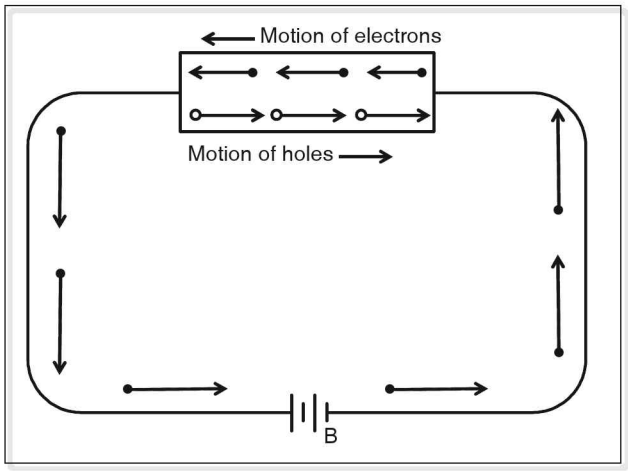


Figure 11.13 Electric conduction in an intrinsic semiconductor

crystal is given by $I = I_n + I_e$. Since the electrons and the holes are equal in number, $I_n = I_e$ and thus $I = 2I_n = 2I_e$.

In an intrinsic semiconductor, the conductivity is very low because the number of charge carriers at room temperature is very small.

In order to increase the conductivity, impurities are added to a pure semiconductor and the impure semiconductor is known as extrinsic semiconductor. The process of adding impurities to a pure semiconductor is called doping and the impurities added to a pure semiconductor is called dopant. The percentage of doping depends upon the amount or concentration of impurity atom added.

Extrinsic semiconductors are classified into two types namely n-type semiconductor and p-type semiconductor.

n-type semiconductor

If a pentavalent impurity like arsenic, antimony, or phosphorus etc., is added to molten silicon or germanium, during the process of crystal growth, an n-type semiconductor is obtained. Four out of the five valence electrons of the dopant form a covalent bond with the neighbouring silicon atoms. The fifth valence electron of the dopant is loosely bound to the parent atom and this electron can be easily set free by giving energy much less than the energy required to rupture a covalent bond. Thus the impurity atom added donates an electron and the impurity atom is known as donor impurity. Each impurity atom added to a semiconductor donates one electron without leaving a hole.

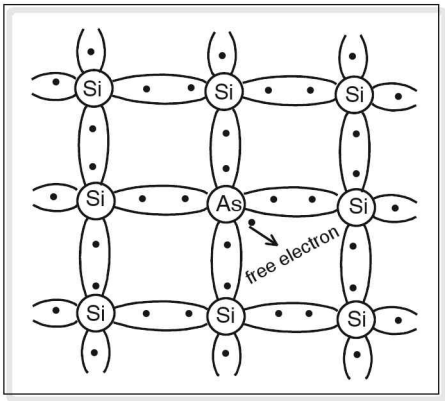


Figure 11.14 Bonding in an n-type semiconductor crystal

Further electron hole pairs are formed due to thermal agitation but still the number of electrons is greater than the number of

holes. This type of semiconductor is called n-type semiconductor. Thus in an n-type semiconductor majority charge carriers are electrons and minority charge carriers are holes.

p-type semiconductor

If a trivalent impurity like indium, boron or gallium is added to molten silicon or germanium during the crystal growth, a p-type semiconductor is obtained.

When a trivalent impurity atom is added to a pure silicon crystal, the three valence electrons of the impurity atom form a covalent bond with the neighbouring silicon atom. The fourth bond is incomplete and this gives rise to a hole. This hole accepts an electron and completes the bond and hence the impurity atom in this case is known as acceptor impurity. A hole is formed at the site from where the electron is captured. Every impurity atom gives rise to a hole. Further electron-hole pairs are formed due to increase in temperature but still the number of holes exceeds the number of electrons. This type of semiconductor is called acceptor semiconductor or p-type semiconductor. In p-type semiconductor majority charge carriers are holes and minority charge carriers are electrons.

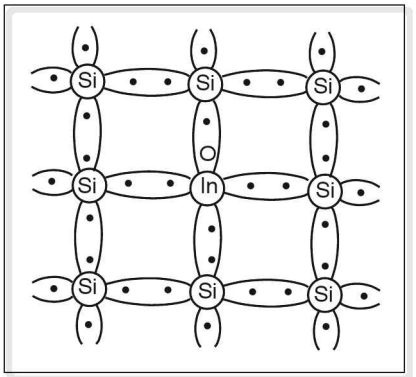


Figure 11.15 Bonding in a p-type semiconductor crystal

p-n junction

p-n junction is a junction between a p-type and n-type semiconductor such that it is continuous at the boundary. p-n junction is the basic unit for all semiconductor devices.

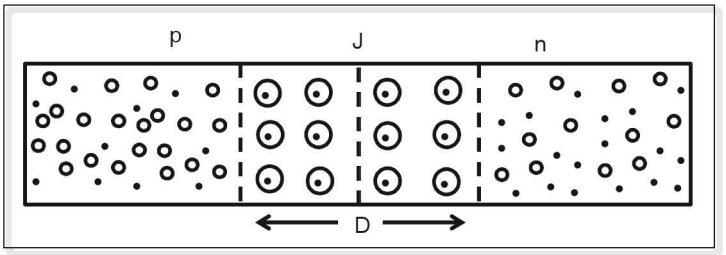


Figure 11.16

There are number of methods of making a p-n junction. In one of the methods, molten liquid of silicon or germanium is doped with pentavalent impurity and the crystal is allowed to grow upto a certain stage. This part of the crystal is n-type semiconductor. After that the crystal is allowed to grow in the presence of a trivalent impurity so that p-type semiconductor is formed on the other side. Thus in a single piece of pure semiconductor, one half is doped with a pentavalent impurity and the other half of the semiconductor is doped with a trivalent impurity.

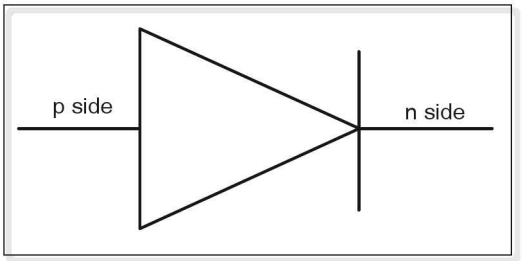


Figure 11.17 Schematic diagram of a p-n junction

In the p-n junction which is just formed, electrons are the majority charge carriers in the n-side and holes are the majority charge carriers in the p-side. Due to the difference in the charge concentration, electrons diffuse from n-side to p-side and holes diffuse from p-side to n-side. These electrons and holes recombine and when an electron leaves the n-side, it makes the impurity atom a positive ion and similarly when holes capture electrons it makes the trivalent

atom a negative ion. Thus the region near the junction contains immobile ions and since this region is depleted of mobile charge carriers, it is called as depletion region. The immobile ions act as a barrier for further diffusion of electrons and holes across the junction. Due to the immobile ions an electric potential difference is developed across the junction. This potential is called the junction potential or the barrier potential.

This junction potential is 0.3 V for germanium and 0.7 V for silicon at 300 K.

p-n junction is called p-n junction diode or semiconducting diode.

When p-n junction is not connected to external source it is said to be unbiased.

p-n junction can be biased or connected to applied voltage in two ways namely, forward bias and reverse bias.

Forward biased p-n junction

When the p-side of a p-n junction is connected to the positive terminal of a battery and n-side is connected to the negative terminal of the battery or the applied source, then the p-n junction is said to be forward biased.

In the forward biased p-n junction, the holes on p-side are repelled by the positive terminal of the battery and the electrons on the n-side repelled by the negative terminal of the battery. Hence the majority charge carriers move towards the junction. The applied potential opposes the junction potential and hence the resistance or the width of the junction decreases. Once the applied voltage (V_B), is greater than the barrier voltage, the majority charge carriers easily diffuse across the

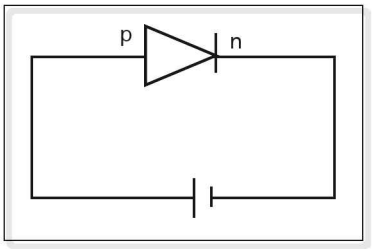


Figure 11.18

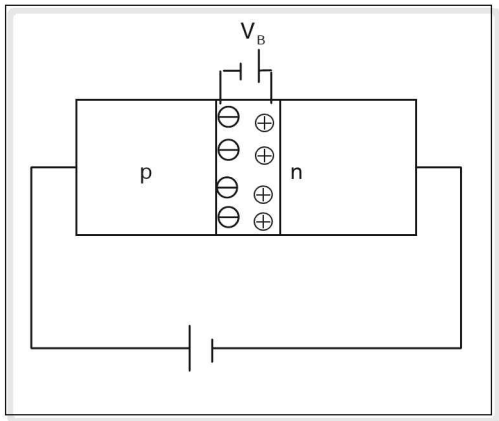


Figure 11.19

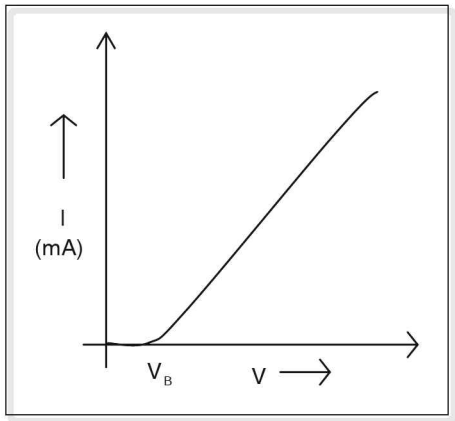


Figure 11.20

junction and the current flows through the diode. When the applied voltage is further increased, the current through the diode also increases.

Reverse biased p-n junction

When p-side of a p-n junction is connected to the negative terminal of a battery and n-side is connected to the positive terminal of the battery, then the p-n junction is said to be reverse biased.

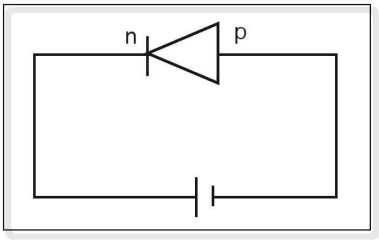


Figure 11.21

In the reverse biased p-n junction, the electrons on the n-side are attracted towards the positive terminal of the battery and the holes on p-side are attracted towards the negative terminal of the battery. Thus the majority charge carriers move away from the junction. The applied voltage supports or acts in the direction of the junction potential. As a result the width of the depletion layer and the resistance of the junction increases. There is no diffusion of majority charge carriers across the junction and hence no current flows through the p-n junction. However there is a very small current (in μA) due to diffusion of minority charge carriers (for an ideal diode no current flows).

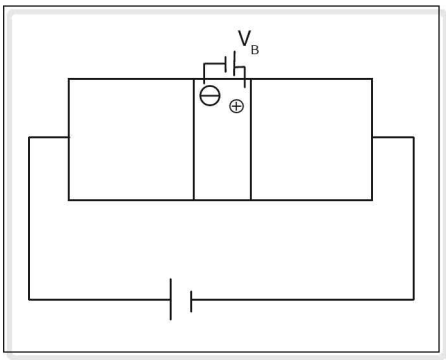


Figure 11.22

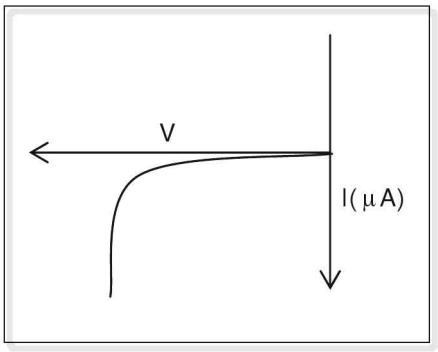


Figure 11.23

Thus the p-n junction conducts current only if it is forward biased and does not conduct when it is reverse biased. This property of unidirectional current flow in a p-n junction is similar to that of a vacuum diode and thus a p-n junction is generally referred to as a p-n junction diode. This unique property of the semiconducting diode makes the diode act as a rectifier.

Rectifier

Rectifier is a device which converts AC into DC. The process of converting AC to DC is called rectification.

p-n junction diode as a rectifier

D – Semiconducting diode.

T – Step down transformer

R_L – Output resistor or load resistor

V_{out} – Output voltage

V_{in} – Input voltage

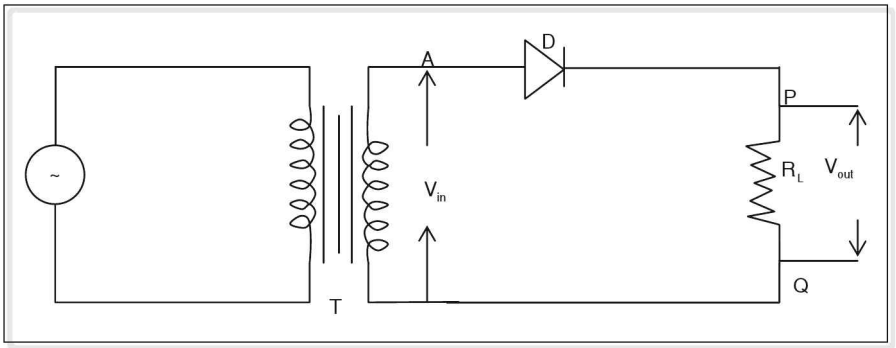


Figure 11.24

The circuit diagram for rectification is as shown in fig(11.24). The AC voltage to be rectified is applied to the free ends of the primary coil of the transformer and the ends A and B of the secondary coil of the transformer are connected to the semiconductor diode and the output resistor R_L.

During the positive half cycle of the AC or the transformer output i.e., when the end A is positive, the diode D is forward biased and current flows through R_L from P to Q. The output voltage is obtained across R_L.

During the negative half cycle of the AC output, the end A is negative. As a result the diode, D is reverse biased and it does not conduct, and hence no current flows through R_L. The cycle repeats and the output voltage is obtained only when diode is forward biased, but the output voltage (V_O) is unidirectional of varying magnitude.

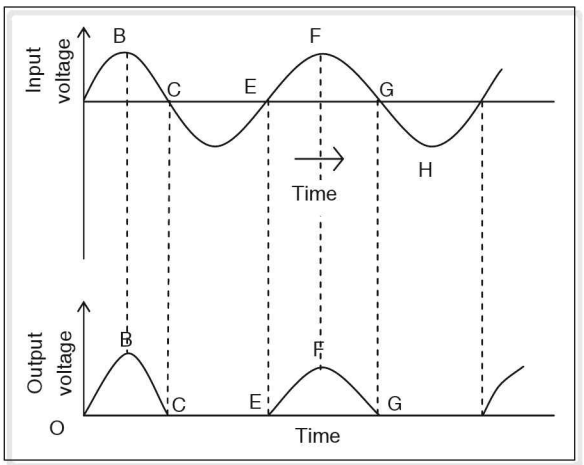


Figure 11.25 Graphical representation of the input and output voltages in a half wave rectifier.

In the above rectification the output voltage is obtained across the load resistor only during positive half cycle of AC input. Such type of rectification is called half-wave rectification.

If the current flows through the output resistor (R_L) in both the cycles of AC input then it is called full wave rectification.

There are many types of semiconducting diodes. LED or light emitting diode is one of the important semiconducting diodes.

LED works on the principle of optics and electronics.

LED's are p-n junction diodes which emit light when the diode is forward biased.

Schematic diagram of LED

When the LED is forward biased, the electrons being majority charge carriers on n-side cross the junction and combine with the holes in the valence band in the p-region. Thus the electrons fall from conduction band (higher energy level) and the difference in the energy is given out as radiation. The colours emitted depend on the nature of semiconductor used in the manufacture of LED. Example-Gallium phosphide radiates red or green colour. Silicon and germanium are not used in the preparation of LED, since they give out energy in the form of heat.

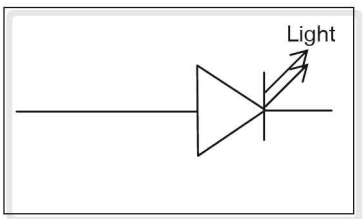


Figure 11.26

Application

1. LEDs are used in display in various systems such as calculators., stereo amplifiers, digital clocks etc.
2. LEDs emitting radiation in infrared region is used in burglar alarms, cd players etc.

Liquid crystal display (LCD)

Liquid crystal is a state of matter which has property of both solid and liquid. In addition it has its own characteristic property.

Liquid crystals are very sensitive to temperature and they respond even to a very weak magnetic and electric fields. By altering or changing the temperature or the electric field, the colour of the liquid crystal can be changed. Liquid crystals are used in liquid crystal thermometer in T.V and computer monitors.

Transistor

Transistor was invented by J.Bardeen and W.H. Barttain and modern transistor was invented by William Shockley.

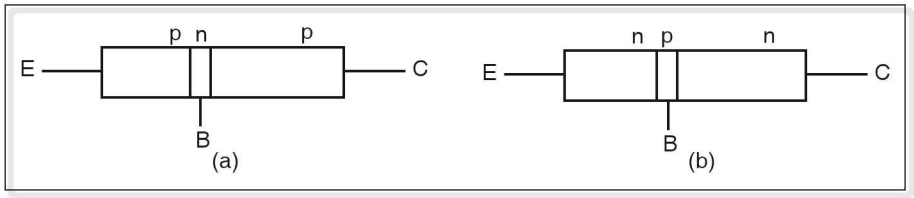


Figure 11.27 (a) p n p transistor and (b) n p n transistor
E – emitter, B – base, C – Collector

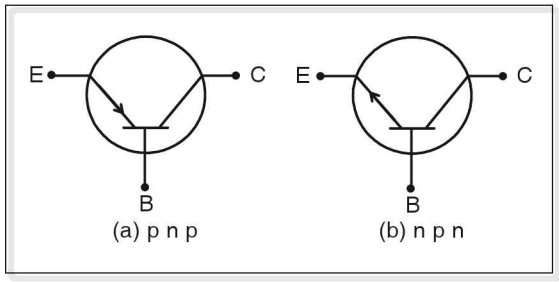


Figure 11.28

Transistor is a three terminal and two p-n junction device. In this device, transfer of resistance takes place, hence this device is called as transfer resistor or transistor. Transistors are of two types namely p-n-p and n-p-n transistor. In p-n-p transistor, n-type is sandwiched between two p-types and in n-p-n transistor p-type is sandwiched between two n-types.

The central region of the transistor which is lightly doped and narrow is called the base. One of the outer regions is heavily doped and moderate in size and this region is called the emitter. Emitter emits majority charge carriers. The other outer region which is moderately doped and larger in size is called the collector. The function of the collector is to collect majority charge carriers.

Action or working of a transistor

In normal mode of operation of a transistor, emitter base junction is forward biased and collector base junction is reverse biased.

V_{EB} → Emitter base voltage

V_{CB} → Collector base voltage

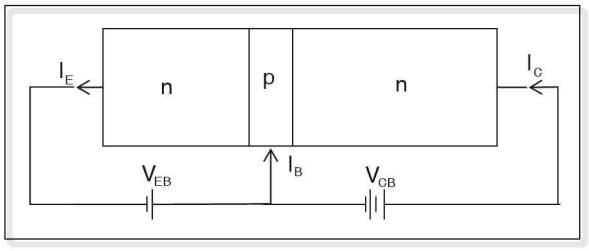


Figure 11.29

The above circuit shows the action of n-p-n transistor with base as a common terminal for both input and output. Since the emitter base junction is forward biased, electrons being the majority charge carriers in the emitter of n-p-n transistor move towards the base (p region). If applied voltage V_{EB} overcomes the barrier potential of the emitter base junction, the electrons move to the base region. This constitutes an emitter current I_E. In the base region, few electrons combine with the holes which give rise to base current I_B. since the base region is narrow and lightly doped and the base – collector junction is reverse biased most of the electrons cross the base collector junction and are collected in a collector. This gives rise to collector current I_C. Thus $I_E = I_B + I_C$. By controlling the base current, the collector current can be controlled. In n-p-n transistor the conduction is mainly due to electrons.

Transistors are used as amplifiers, oscillators and in the switching circuits.

Amplifier

It is a device which strengthens weak AC signals. The process of strengthening weak signal is known as amplification.

Transistor as an amplifier

Circuit diagram for amplification using n-p-n transmission is as shown. The emitter base junction is forward biased with a DC source V_{BE} and the collector base junction is reverse biased with DC source V_{CE}. Emitter is made common to both input and output terminals and it is connected to earth.

The AC signal to be amplified is applied between base and the emitter, the output is obtained across the load resistor R_L.

When the input signal is applied, emitter base voltage is altered. This causes the variation in the emitter current which in turn alters the collector current to a large extent. The voltage across the load resistance R_L varies and large output voltage is obtained across R_L. Thus the weak signal is amplified.

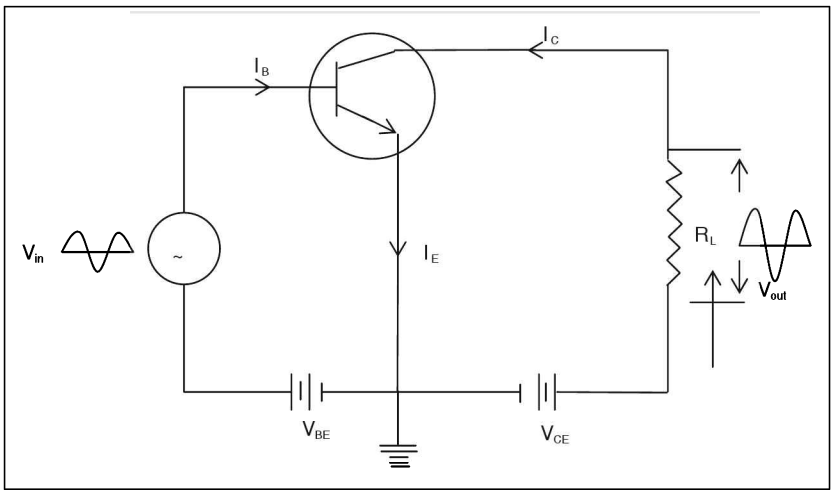


Figure 11.30

Capacitors

It is a device used to store a certain amount of electric charge.

Capacitor consists of two similar parallel metal plates separated by a small distance. One of the plates is positively charged and this induces an equal amount of negative charge on the other plate. The second plate is connected to earth. The space between the two plates is filled with insulator which is also known as dielectric. Terminals are provided to the two plates in order to connect the capacitor in the circuit.

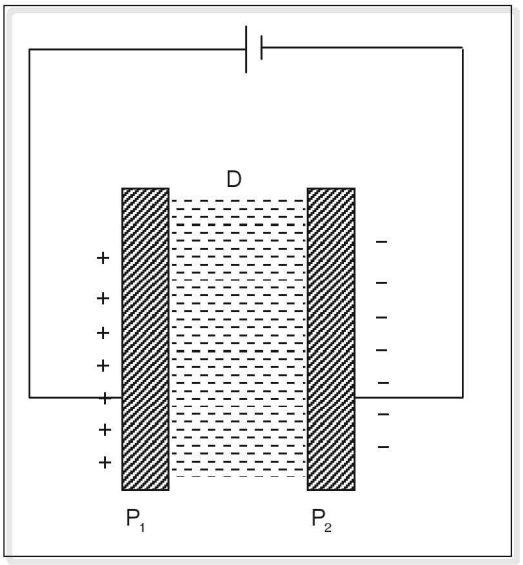


Figure 11.31 A parallel plate capacitor
P_1, P_2 Metal plates D – Dielectric

Whenever charge is added to a conductor its potential increases i.e., $V \propto Q$, where V is the potential and Q is the charge.

$\therefore V \propto Q$

or

$$\frac{Q}{V} = C$$

Where C is called as the capacitance of a capacitor or a conductor.

Capacitance is the ability to store charge. Capacitance is measured in farad (symbol is F). Since farad is a large quantity; micro-farad and pico farad are commonly used to measure capacitance.

$1\mu F = 10^{-6}$ F

1 pF $= 10^{-12}$ F

There are different types of capacitors. Based on the dielectric used in them, some of the types of capacitors are ceramic capacitor, paper capacitor, mica capacitor, polyester capacitor. In these capacitors, the terminals can be either connected to positive or negative, but in case of electrolytic capacitor which works on the property of chemical change of substance. When current is passed, the terminals of this capacitor cannot be connected either to positive or negative as their poles cannot be inter changed. Hence electrolytic capacitor cannot be used in AC circuit.

Radio and television—broadcasting principles

In the broadcasting of radio and television programmes, electromagnetic waves in the frequency range of radio frequency are utilised. The fact that the induction coils detect the radio waves transmitted over long distances was discovered by Marconi in 1895, which led to the invention of radio. The fundamental stages in radio and television communication are the production of messages, the transmission of the produced messages and the reception of the transmitted messages. Initially, the messages to be transmitted are converted into electrical signals using a transducer.

Transducer is a device which converts audio or video information into electrical signal.

Microphone is used as a transducer for sound transmission and photoelectric device is used as a transducer for video transmission.

These message signals are superposed over radio frequency waves, that are capable of propagating over long distances. These radio frequency waves thus are called 'radio frequency carriers'. The values of the radio frequency of the transmitting waves are different for different radio and television stations. For radio transmission, the radio frequency (R.F) is in the range of 300 kHz to 30 MHz and for television it is in the range of 30 MHz to 300 MHz. The process of superposing a message signal onto a radio frequency wave is known as 'modulation'. In modulation, there are two types, namely, 'amplitude modulation' (AM) and 'frequency modulation' (FM). At the receiving end, i.e. in the radio set or a television set, these modulated waves pass through a circuit called filter circuit and the output from the circuit is only a message signal. Thus the message waves are separated from the modulated waves. This process of extracting message signals from a modulated radio frequency carrier waves is known as 'demodulation' or 'detection'. The following points summarize the principles of radio and television broadcasting.

1. The sound or/and the images produced during the making of a program are converted into electrical signals.
2. The electrical signals produced are amplified by using suitable amplifiers. Amplification is the process of increasing the amplitude of a given signal.
3. The amplified signal is modulated with a suitable radio frequency carrier wave, generated in an oscillator.
4. The modulated waves are fed to the transmitting antenna placed on the top of a tower of considerable height.
5. Through the transmitting antenna these modulated waves are transmitted in all possible directions in the space. This procedure is called transmission.
6. The transmitted modulated waves are received at the receiving antenna when struck with it.
7. The received radio frequency modulated waves set up alternating currents in the receiving antenna and through the antenna they are sent to the receiver, i.e. a radio or a television.
8. The tuning circuits in the radio or the television select the desired modulated wave out of the huge number of waves received at the antenna and are amplified by an amplifier.
9. The selected modulated wave is fed to the detector circuit, which consists of an electronic oscillator that separates the message signal (audio or video or both) from the radio frequency carrier wave. (This process is called demodulation or detection.)
10. The separated message signal is passed through an amplifier circuit which is fed to the corresponding devices for reception of audio/video.

Radio broadcasting and reception

The following block diagram shows the basic principles discussed earlier that are involved in radio communication.

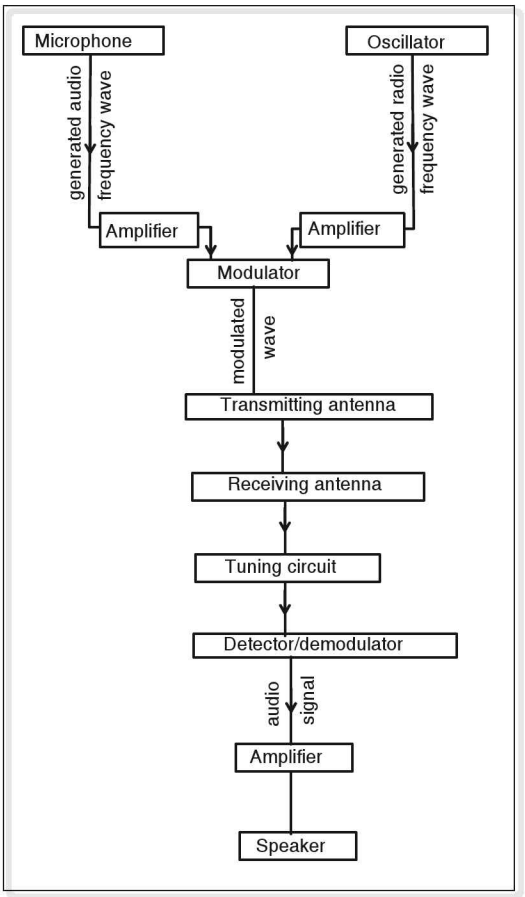


Figure 11.32

Television broadcasting (telecasting) and reception

In radio broadcasting only audio signals are involved whereas in telecasting, video signals are also involved along with audio signals. This makes the difference between radio and television broadcasting. The production of a video signal is done through the scanning of the image of an object. In the early stages of telecasting when there was no technological advancement, scanning was done using a mechanical system called 'nipkow disc'. Now-a-days, an improved system called 'iconoscope' is used for the purpose. Iconoscope is a video camera that consists of a cathode ray beam and photo electric cells. The image of an object is focused onto a screen. The image formed on the screen is divided into a large number of small square portions.

The light incident on each small square portion is converted into an electric signal by photo cells present on the screen. The strength of the electric signal produced depends on the intensity of light incident on it. These electrical (video) signals are amplified further, and then used for modulation. The block diagram in figure (11.33) shows the telecasting and reception. The modulation done for video

signal is amplitude modulation, whereas the audio signals are modulated using frequency modulation. Video signals are modulated in 'AM picture transmitter' and audio signals are modulated in 'FM sound transmitter'. Then the modulated audio and video signals are combined and fed to the transmitting antenna placed on a tower. Different television stations use different radio frequency waves for their transmission and these are referred to as 'channels'. A television antenna receives these transmitted waves and passes them to the tuning circuit of the television set, where a desired channel is selected. Then the selected modulated wave passes through detection circuit where the messages (both audio and video signals) are extracted and they are further amplified. The audio and video signals are separated. The audio signals are amplified and sent to the speaker from which the sound is heard. The video signals are fed to the picture tube where the electric signals are converted into light and the transmitted image is reconstructed. The television set is also referred to as kinescope.

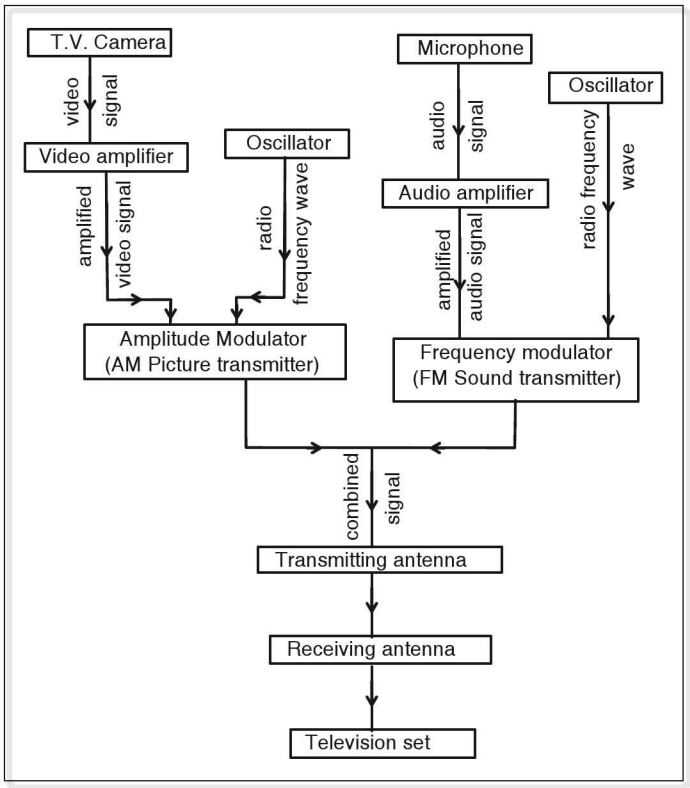


Figure 11.33

The audio frequency and video signals (electrical signals) are very weak and are amplified by using suitable amplifiers.

The audio frequency signals (AF) and video signals cannot be transmitted over long distances since they have low energy. In order to propagate these electrical signals the AF signals or the video signals are mixed with radio waves generated by radio frequency oscillation. These radio frequency waves (RF waves) are superposed over AF waves or the video signal waves in a mixer. The process of superposing electrical message signal over RF wave is called modulation. The RF wave which carries electrical signals is called carrier wave. The carrier wave is amplified by RF amplifier.

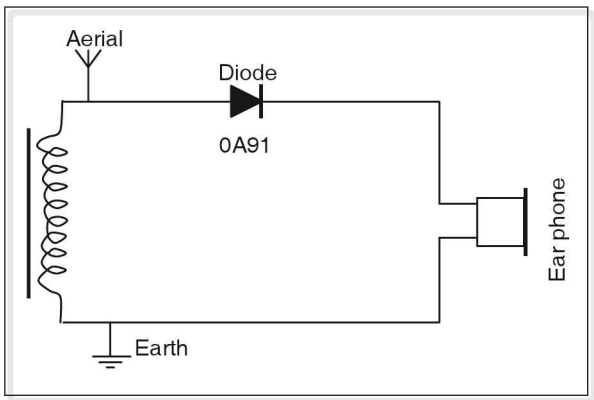


Figure 11.34

Assembly of a radio

Components required

1. OA91 Diode, Ferrite rod, copper wire, earphone, long metal rod (aerial).
2. Variable capacitor (500 PF) and capacitor 0.01μF.
3. Resistor (100 Ω) n-p-n transistor, DC source.

Procedure of building a circuit

Wind a copper wire around a ferrite rod and connect one end of the copper wire to p-type of diode and long metal rod which is the aerial. The n-type of diode is connected to the terminal of an earphone and the other terminal is connected to the other end of the copper wire which is connected to the ground. This circuit is a crude radio.

To increase the efficiency of the crude radio, connect a 500 pF variable capacitor across the copper wire and connect a capacitor of 0.01μF in series with n-type of the diode and the copper wire which is connected to the earphone. The capacitor filters AC from DC.

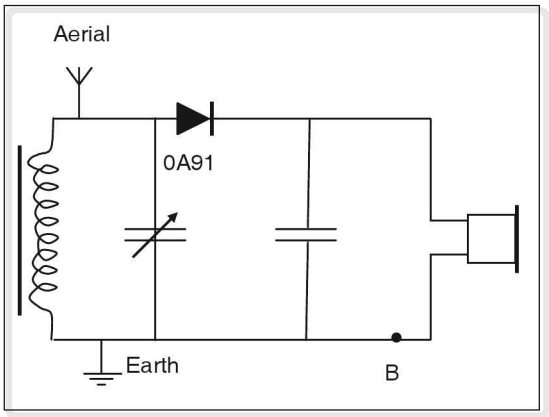


Figure 11.35

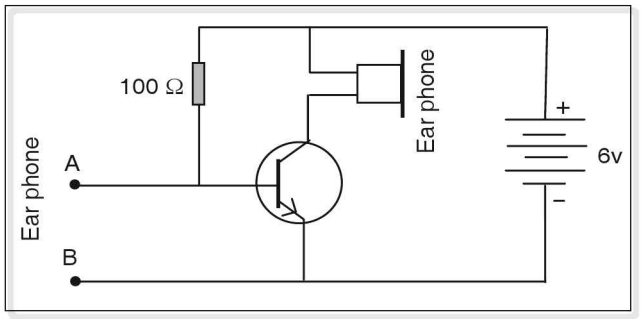


Figure 11.36

In order to increase the clarity and the loudness, a transistor (npn) and a resistor of 100 Ω are introduced.

Base region and the emitter of the transistor are connected across 0.01μF capacitor (as shown). One end of the 100 Ω resistor is connected to the base of the transistor and the other end of the resistor is connected to positive terminal of 6V DC source which is connected to n-type of a transistor. The negative terminal of the DC source is connected to the emitter of the transistor. The n-p-n transistor amplifies the signal. Hence the sound can be heard louder.

The above circuit can work only near AM radio station.

Integrated circuits (IC chips)

Integrated circuit is a single semiconductor crystal or a chip which can perform the operation of resistor, capacitor, diodes and transistors in various combinations. IC chips are compact in size and are very light and very economical.

IC chips are not affected by ordinary temperature and they consume less power and the time taken to activate an IC chip is in order of nano second. Transformer and induction coil cannot be fabricated on IC chips.

Microprocessor (μp)

Combination of many ICs is called microprocessor. A microprocessor can compute data make logical decisions and store data.

The data, either numeric or non-numeric, called instructions are given to a microprocessor through a device known as input. The instruction is processed by a control unit and is sent to a decoder. The Arithmetic Logic Unit (ALU) of the micro processor does all mathematical operations and makes logical decisions. The output data is sent to a device known as output device. The output and the input data are stored in the memory unit of the micro processor.

Thus the microprocessor which can compute data is called computer.

Computer

It is a device that helps in computing. It is a multitasking device that performs arithmetic and logical operations as well. It takes the information (data) given for doing the operations, stores them, processes the information and gives the result. It can also store the result (called output). The block diagram of a computer is as shown in figure (11.37).

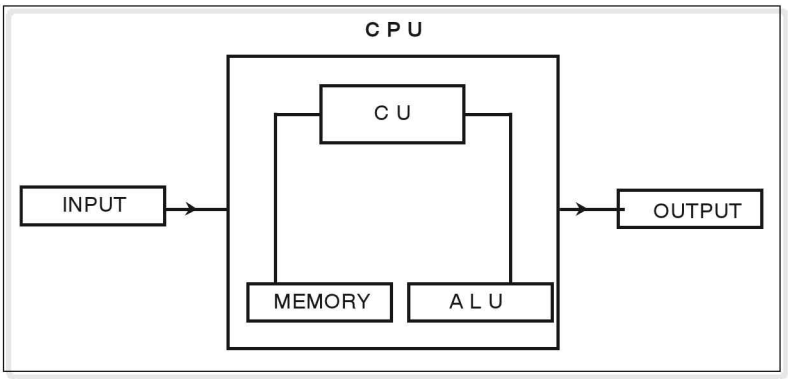


Figure 11.37 Block diagram of a computer

Input

The information that is fed to the computer is known as input. The devices through which information is fed are called input devices. A key board, a mouse, a floppy disc, a compact disc are the different kinds of input devices.

CPU

CPU is the central processing unit. It is the most important part of a computer and thus is generally referred to as the heart of it. This unit comprises three sub units. They are control unit (C.U), memory unit (M.U) and Arithmetic Logic Unit (ALU). Control unit controls the entire system and its related processes. The memory unit stores the information and Arithmetic Logic Unit performs the operations according to the instructions given.

Output

The processed information is the result and is given out to the user. It is called the output. The devices through which output is obtained are called the output devices. Monitor, printer, floppy, etc., are the output devices.

The CPU of a computer consists of a circuit board which has an utmost importance and is called 'mother board'. The mother board consists of the circuits related to all the components of the computer. These circuits consist of compact components called microprocessors. A micro processor consists of integrated circuits (ICs) and an integrated circuit consists of a large number of transistors, junction diodes and logic circuit.

A computer cannot perceive the information in the form that we understand, as it uses a binary system. The binary system consists of only two digits, 0 and 1, '0' refers to 'off' mode for a diode and '1' refers to 'on' mode. In the 'off' mode, the diode does not conduct electricity. In the 'on' mode, it conducts electricity. Thus the data given to a computer is perceived by it in the binary code. The information given to a computer is in the form of either alphabet or numerals or symbols.

These put together are called 'characters'. Each character is understood by the computer in its binary code. A binary digit, i.e. either '0' or '1' is called a 'bit'. A group of eight bits is called a 'byte', and a group of one or more bytes is called a 'word'. Thus the information is stored in the computer in the form of 'bytes' and 'words'. The digits in the decimal system, the alphabet and also the special characters in the information given to a computer are converted into a code that is understood by it before processing. One of the most commonly used codes is Binary Coded Decimal code (referred to as BCD in short). Generally the numbers, letters and symbols are represented by 8 bit codes in the BCD code. The first four bits are called 'zone bits' and the remaining four are called 'numeric bits'. Zone bits help to differentiate between a number, an alphabet and a symbol.

For example, the binary equivalents of the digit 1 and the English alphabet A, both are 0001. Thus to differentiate between the two, zone bits are used. The zone bit of a numeral is '0101' and that of an alphabet is '1010'. Thus the BCD code for '1' is '01010001' and for 'A' is '10100001'.

☛ *Example*

Conversion of 8 into binary system

$$\begin{array}{l|l} 2 & 8 \quad - 0 \\ 2 & 4 \quad - 0 \\ 2 & 2 \quad - 0 \\ & 1 \end{array}$$

Binary equivalent of 8 is 1000 and this is also called as numeric bit zone. Bit of numerals is 0101.

$\therefore$ BCD code for 8 is 0101 1000.

Once the computer perceives the data given to it, the next stage is the processing of the data. This is done based on the instructions given to it. The instructions given to a computer in the form that we understand, however, cannot be perceived by it. So the instructions are also transformed into the BCD code. The set of instructions given to a computer is called a 'program'. The program converted into the BCD code is called 'machine language' and it depends on the CPU. Assembler is an example of a

machine language. The physical components of the computer, i.e., the input and output devices and the CPU are referred to as the 'hardware' of the computer and they work on the machine languages.

But writing the program in a machine language is very tedious. Thus the programs are written in a 'high level language' which consists of English words and symbols, thus making it easy to write a program. There are certain rules to be followed while writing the programs in a high level language, similar to the grammar rules in English. These rules are called 'syntax'.

Once a program is written in a high level language, it has to be translated into a machine language for execution. The computer takes the help of another program called 'compiler' for this translation. Thus the compiler is a program that converts a high level language into a machine language.

The complier and the programs written in high level language are referred to as the 'software' of the computer. BASIC, FORTRAN, COBOL, PASCAL, C, JAVA are some of the high level languages.

Applications of computers

Computers have innumerable applications and today they are part of human life. The following are the various steps to be followed for solving a problem using a computer.

1. Identifying the problem to be solved by a computer and understanding it.
2. Scripting a suitable program for the problem to be solved by the computer.
3. Supplying data to the computer according to the written program.
4. Getting the output from the computer.

test your concepts

Very short answer type questions

1. What is the effect of temperature on semiconductor?

2. What are light emitting diodes?

3. What is a CPU?

4. What is an amplifier?

5. What is a forbidden gap?

6. What is a diode?

7. Define capacitance and write its unit.

8. How are conductors, semiconductors and insulators differentiated on the basis of energy bands?

9. What are thermions?
10. What are dielectrics?
11. What is an extrinsic semiconductor?
12. Define thermionic emission.
13. What is a transducer?
14. What is doping?
15. How many types of thermionic emitters are there and what are they?
16. What is an IC?
17. What is a hole?
18. What is a directly heated thermionic emitter?
19. What is a microprocessor?
20. What is a p-n junction?
21. What is an indirectly heated thermionic emitter?
22. What is the range of radio frequency waves used for radio transmission?
23. What is barrier potential?
24. Define space charge.
25. What is modulation?
26. What is rectifier?
27. What is an energy band?
28. What is a computer?
29. What are liquid crystals?
30. What is a valence band?

Short answer type questions

31. Explain why modulation is necessary.
32. Explain the working of a LED.
33. Explain the reason for keeping very tall TV transmission towers.
34. Explain the functioning of a p-n junction as a forward bias and reverse bias.
35. Explain how the circuit containing diodes can be used to represent '0' and '1' digits which are used in binary system?
36. Explain the formation of an n-type semiconductor with a diagram.
37. In an intrinsic semiconductor, the number of conduction electrons per unit volume is 7×10^{18} m^{-3}. If the volume of a semiconductor is 1 cm^3, find the total number of current carriers.

38. Explain the formation of electron-hole pair with a diagram and illustrate the motion of electrons and holes in an intrinsic semiconductor.

39. "The p-type and n-type semiconductors have large number of electrons and holes respectively, but still they are electrically neutral". Explain.

40. What are the factors that effect thermionic emission?

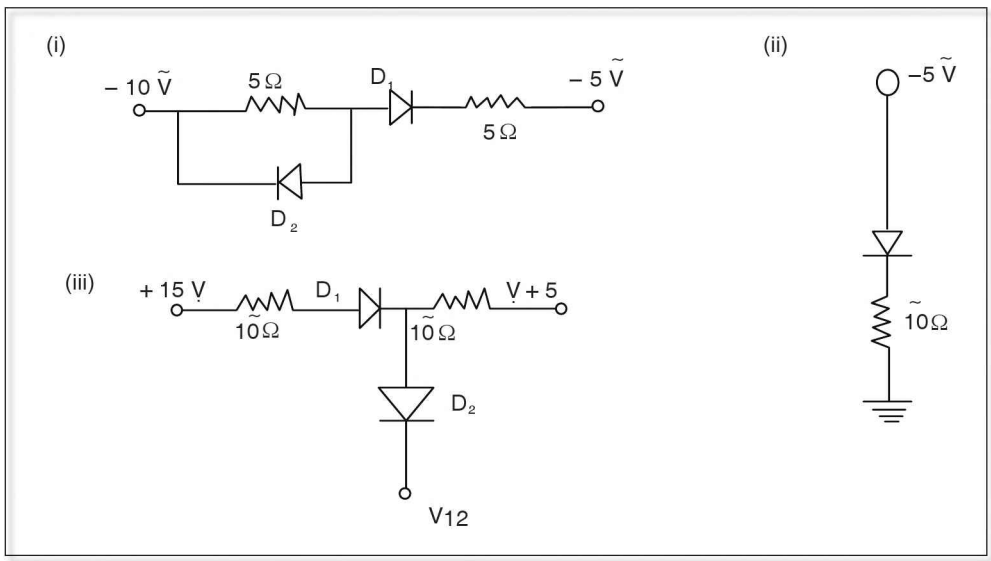
41. The current flowing through a diode valve for various plate voltages are as shown below:

V (in volts)	0	60	80	150
I (in mA)	0	1	4	8

Using the above readings, show that a vacuum diode is non ohmic conductor.

42. Write BCD for 5 and D.

43. In each of the following circuit diagrams, explain whether the diode (diodes) is (are) forward or reverse biased.



44. Does a diode obey Ohm's law? Explain.

45. Are all tetravalent elements semiconductors? Explain.

Essay type questions

46. Explain the working of p-n-p transistor.

47. Describe briefly about television (and radio) broadcasting with the help of neat block diagram.

48. Describe the construction and the working of a cathode ray tube with a neat diagram.

49. Explain with a neat circuit diagram the working of npn transistor as an amplifier.

50. Explain with a neat circuit diagram the action of semiconducting diode as a half wave rectifier.

CONCEPT APPLICATION

Concept Application Level—1

Direction for questions 1 to 7: State whether the following statements are true or false.

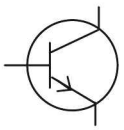
1. A high level language is independent of the hardware of a computer.
2. The process of superimposing an audio or video messages on r.f. carrier waves is called modulation.
3. Addition of either trivalent or pentavalent impurities to an intrinsic semiconductor increases its conductivity.
4. The electric current in an extrinsic semi conductor is the sum of currents due to holes and electrons.
5. The energy gap E_G of a semi conductor decreases with rise in temperature.
6. In an oscilloscope the vertical deflecting plates deflect the electron beam in a vertical plane.
7. The energy bands are present only when the substance is present in the solid state.

Direction for questions 8 to 14: Fill in the blanks.

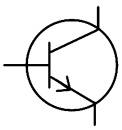
8. The band structure determines the _____ behaviour of a solid.
9. A computer program is the group of _____ given to the computer.
10. The minimum heat energy required to emit an electron from the surface of a metal is called _____.
11. In a reverse biased p-n junction diode, the p-type semiconductor is connected to _____ terminal of the cell and n-type to the _____ terminal of the cell.
12. A group of 8 bits is called a _____.
13. The depletion region in a semiconductor diode is formed at the _____.
14. The process of introducing impurities in small quantities into an intrinsic semiconductor is called _____.

Direction for question 15: Match the entries in column A with appropriate ones from column B.

15.

Column A			Column B
A. Iconoscope	( )	a.	machine language
B. Kinescope	( )	b.	junction transistor
C. Assembler	( )	c.	
D. Microprocessor	( )	d.	space charge
E. Oscillator	( )	e.	T.V. camera
F. Extrinsic semiconductor	( )	f.	alkali metal oxides
G. Symbol of npn transistor	( )	g.	combination of IC's

(Continued on following page)

H.	Symbol of pnp transistor	( )	h.	T.V set
I.	Vacuum diode	( )	i.	
J.	Thermion emitter	( )	j.	p-type semiconductor

Direction for questions 16 to 30: For each of the questions, four choices have been provided. Select the correct alternative.

16. Compiler is a computer program that converts
(1) machine language instructions to high level language instructions.
(2) high level language instructions into machine language instructions.
(3) Basic language to Fortran language.
(4) Fortran language to Basic language.

17. An oscilloscope is basically designed to convert _____.
(1) visual signals to electrical signals
(2) sound signals to electrical signals
(3) electrical signals to visual signals
(4) sound signals to visual signals

18. In an intrinsic semiconductor, the number of electrons in the conduction band is _____ the number of holes in the valence band.
(1) equal to (2) less than (3) greater than (4) None of the above

19. A good thermionic emitter has
(1) high resistance.
(2) high temperature coefficient of resistance.
(3) high work function and high melting point.
(4) low work function and high melting point

20. The characteristics which a material should possess to act as a good thermion emitter is _____.
(1) high work function and high melting point
(2) high work function and low melting point
(3) low work function and high melting point
(4) low work function and low melting point

21. At absolute zero, a semiconductor is an insulator because _____.
(1) no electron is present in the conduction band.
(2) all electrons occupy the the valence band.
(3) the value of E_G is large.
(4) All of the above

22. The heart, or the main component of the computer is the _____.
(1) key board (2) printer
(3) monitor (4) central processing unit (CPU)

23. In a semiconductor holes are present only in the _____.
(1) conduction band (2) valence band
(3) forbidden gap (4) all of the above

24. The forbidden gap for pure silicon at room temperature is _____ eV.
(1) less than one (2) 1.1 (3) 3 (4) 9

25. For television transmission, the range of carrier waves is approximately _____.
(1) 300 kHz to 30 MHz (2) 100 kHz to 30 MHz
(3) 30 MHz to 300 MHz (4) None of the above

26. In an intrinsic semiconductor, if N_e is the number of electrons in the conduction band and N_p is the number of holes in the valence band then
(1) $N_e > N_p$. (2) $N_e = N_p$.
(3) $N_e < N_p$. (4) None of the above.

27. To make a p-type semiconductor, an instrinsic semiconductor is doped with _____.
(1) gallium or indium (2) arsenic or phosphorus
(3) aluminium or boron (4) Both (1) and (3)

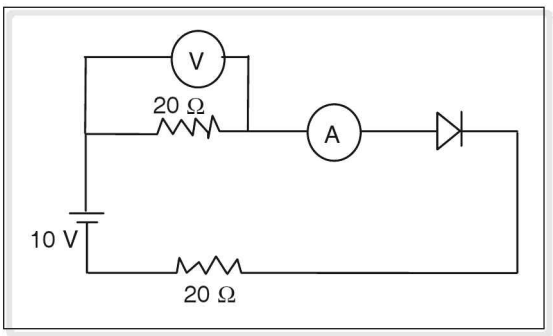
28. The transistors are used in the _____.
(1) electronic circuits called oscillators (2) stabilized power suppliers
(3) integrated circuits (4) All of the above

29. The arrow in the symbol of a transistor indicates _____.
(1) base (2) direction of electronic current
(3) direction of conventional current (4) Both (1) and (3)

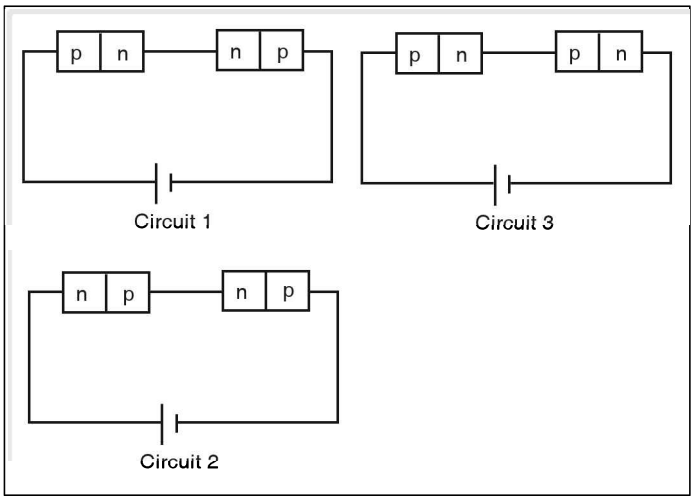
30. A _____ is used as a rectifier.
(1) diode. (2) transistor. (3) resistor. (4) capacitor.

Concept Application Level—2

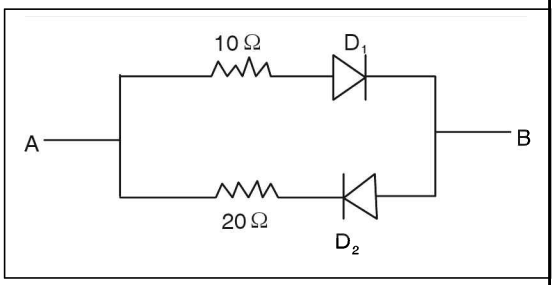
31. In the following circuit diagram, what are the values recorded by the ammeter and the voltmeter if the potential across the diode is 1 V?



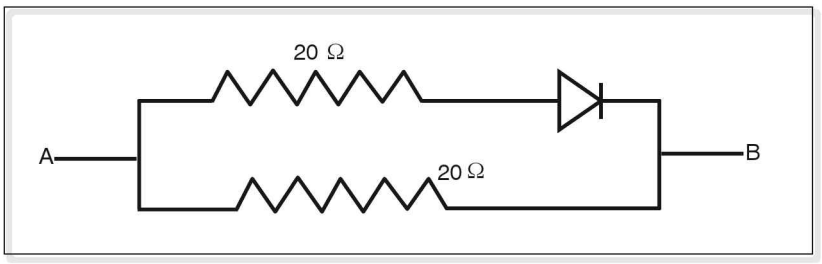
32. Two identical p-n junctions may be connected in series with a battery in three different ways as shown in the circuit diagrams. In which circuit diagram will the potential drop across the p-n junctions be equal?



33. What is the resistance offered by the following circuit if the current flows from
(1) A to B and
(2) B to A?
Assume the diodes to be ideal.
What is the PD across D_1 and D_2, when they were non-conducting, i.e., when they are reverse biased, if 7 V is applied across AB?



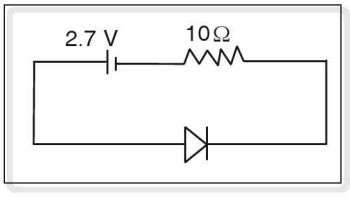
34. In the following circuit diagram, what is the resistance between A and B when the current flows from A to B? (Assume the diode is an ideal diode with no barrier potential and bias resistance)



35. A specimen of tungsten and an alkali oxide, having equal surface areas, emit 5×10^5 electrons per second. Find the temperature of tungsten if the temperature of the alkali oxide is 1000 K. (The work function of tungsten is 4.52 eV and that of alkali oxide is 1eV)

36. Find the energy dissipated in 10 minutes in the diode in the circuit shown below.

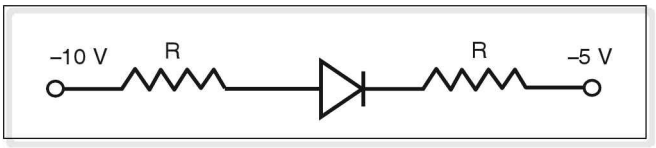
The barrier potential across a diode is 0.7 V which remains constant. (Assume the diode to be ideal)



37. A diode with a barrier potential of 0.7 V is connected to a battery of 4 V in series with a resistor whose resistance is 10 Ω. Find the current through the diode, if the diode is forward biased, and the forward resistance of the diode is 20 Ω.

38. In each of the following circuit diagrams, explain whether the diode is forward biased or reverse biased.

(i)



(ii) and other bits

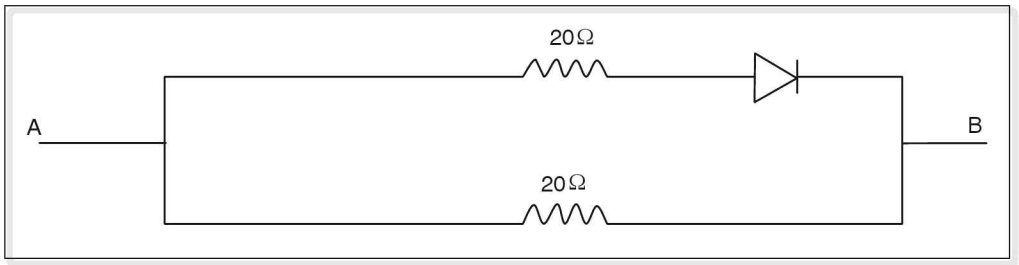
39. 1.6×10^{18} electrons, emitted from cathode, reach the plate of the diode in two seconds, if the power consumed is 1 watt, calculate the potential difference between cathode and the plate.

40. What is the resistance offered by the following circuit if current flows from

(1) A to B and

(2) B to A?

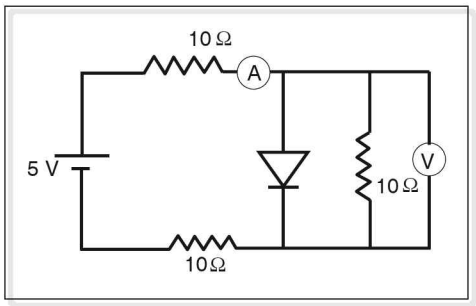
(Assume diode to be an ideal diode)



41. An electron crosses the depletion region to reach p-side from n-side. Find the speed of the electron in p-region if its speed in n-region is 3×10^6 m s^{-1} and potential difference across the depletion region is 0.7 V.

(Take the charge and mass of electron as 1.6×10^{-19} C and 9.1×10^{-31} kg respectively)

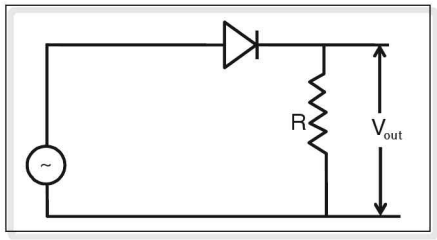
42. A certain radio station transmits the programme at 1 MHz frequency. If we can hear in the frequency range of 20 Hz to 20 kHz, then how can one hear the programme in the radio receiver?



43. In the following circuit diagram, if the diode is ideal, what are the values recorded by the ammeter and the voltmeter?

44. In a vacuum diode, 2×10^{19} electrons emitted from the cathode reach the plate of a vacuum diode in 4 seconds. Find the current and the power consumed if the PD between the cathode and the plate is 200 V.

45. An AC voltage is applied to the diode, as shown in the figure below. Draw the graph of output voltage versus time. Explain the action of the diode.



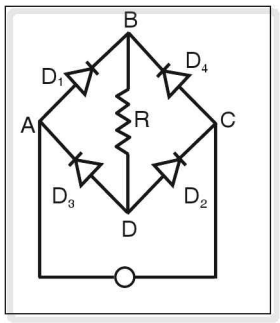
Is the output voltage across R a DC voltage? Explain.

Concept Application Level—3

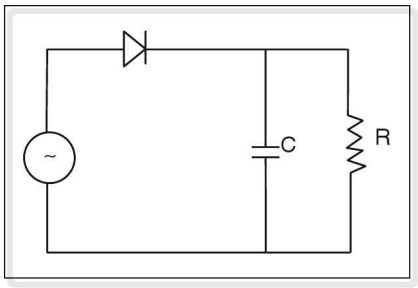
46. In an oscilloscope, the cathode emits the electrons which are accelerated by the anodes. Explain the kind of voltages to be applied to the horizontal and vertical plates so that images formed on the deflecting screen are

(1) horizontal line (2) vertical line

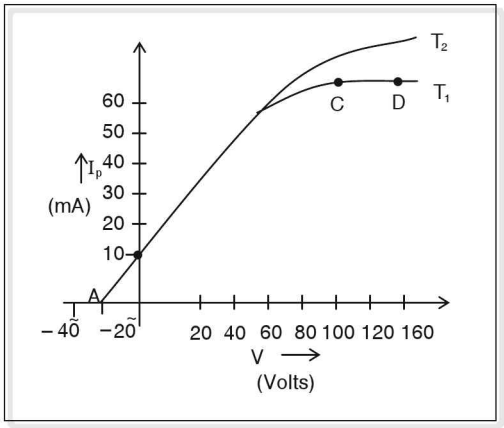
47. In the following circuit, explain the flow of current through resistor R. Can we use the circuit to convert AC to DC? Explain.



48. In the circuit diagram shown below, explain the voltage across R. The sinusoidal voltage has a peak value of 200 V and the capacitor used has long charging time. Assume the diode to be an ideal diode.



49. A graph of vacuum diode current I, is plotted against the anode voltage, V_p for two different temperatures T_1 and T_2 ($T_2 > T_1$) of the cathode, as shown in the figure below.



(1) For the same anode voltage, (called plate voltage) why is the current more when the cathode temperature changes from T_1 to T_2?
(2) In the part CD of the graph, why change in V_p has no effect on I_p?
(3) What is the plate current when plate voltage is zero? At what plate voltage the plate current is zero?

50. Electrons emitted from the surface of a cathode have sufficient kinetic energy to reach the anode even if the anode is at a negative potential. In a vacuum diode, no current flows when the anode potential is less than –5 V. Find the maximum kinetic energy of the electrons in electron volts, emitted from the cathode surface.

key points for selected questions

Very short answer type questions

1. Conductivity increases with increase in temperature.
2. p-n junction diodes, which emit light.
3. Central processing unit – Part of a computer referred to as the heart of it.
4. A circuit which strengthens weak AC signals.
5. Energy states where no electrons exists.
6. (i) A p-n junction.
   (ii) Unidirectional flow of current.
7. (i) $c = \frac{q}{v}$
   (ii) Unit – farad (F)
8. Energy gap between valence band and conduction band.
9. Electrons emitted from a metal surface, by absorption of heat energy.
10. Insulators which are filled between the plates of a capacitor.
11. Impure semiconductor.
12. The phenomena in which free electrons are emitted from a metallic surface, by absorption of heat energy supplied to them.
13. A device which converts audio or video information into electrical signal.
14. Addition of impurities to a pure semiconductor.

15. (i) Two types
(ii) 1. Directly heated thermionic emitter and
2. Indirectly heated thermionic emitter.

16. A single semiconductor crystal or a chip that performs the operation of resistor, capacitor, diode etc.

17. Vacant site created in the VB when electrons jump from VB to CB.

18. (i) Current is passed through the emitter.
(ii) Temperature of the emitter crosses the temperature corresponding to the work function of emitter.
(iii) Emission of thermions.

19. Combination of ICs is a microprocessor.

20. A junction between a p-type and n-type semiconductor such that it is continuous at the boundary.

21. (i) Heat is supplied to the emitter by a heating element.
(ii) Temperature of the emitter crosses the temperature corresponding to the work function of emitter.
(iii) Emission of thermions.

22. 300 kHz to 30 MHz.

23. Electric potential difference developed across the p-n junction due to immobile ions.

24. (i) No drift of thermions emitted from cathode towards the plate.
(ii) Accumulation of emitted thermions over the cathode in the form of electron cloud – space charge.

25. The process of superposing a message signal onto a radio frequency wave.

26. A circuit which converts AC into DC.

27. Superposition of outermost orbits of the atoms and spreading into a band of energy levels within a small space.

28. A microprocessor which can process data.

29. Liquid crystals are very sensitive to temperature and respond even to a very weak magnetic and electric field.

30. Energy band at the lower level, which is occupied by valence electrons.

Short answer type questions

31. (i) Do audio signal and video signal possess sufficient energy to travel a long distance?
(ii) How is the message carried to a long distance?

32. (i) p-n junction, forward biased.
(ii) Electrons fall from CB to VB.
(iii) Difference in the energy given out as radiation.
(iv) Colour emitted depends on the nature of semiconductor used in LED.

33. (i) Is it possible for water ripples in the water tank to move beyond the walls of the water tank?

34. FORWARD BIAS:
(i) p-side connected to positive terminal and n-side connected to negative terminal of a battery.
(ii) Holes on p-side repelled by positive terminal and electrons on n-side repelled by negative terminal of the battery.
(iii) Width of the junction potential decreases.
(iv) Current flows through the diode.
(v) Current through the diode increases with increase in applied voltage.

REVERSE BIAS:
(i) p-side connected to negative terminal and n-side connected to positive terminal of a battery.
(ii) Holes on p-side and electrons on n-side attracted by respective terminals.
(iii) Width of the junction potential increases.
(iv) No current flows through the diode.
(v) Small current (μA) due to diffusion of minority carriers.

35. (i) When a diode conducts it is represented by 1, and refers to 'ON' position.
(ii) When a diode does not conduct, it is represented by 0, and refers to 'OFF' position.

36. Pentavalent impurities.

37. (i) From the given problem find out the number of electrons per unit volume. (n)
(ii) Calculate the number of electrons in a given volume (V), $n_e = n \times v$.
(iii) In an intrinsic semiconductor, the number of electrons = the number of holes.
(iv) Total charge carriers = $2n_e$

38. Place where electron leaves creates a vacant site.

39. (i) A body can be charged either by adding or removing electrons. In the case of semiconductors, do we add or remove electrons?
(ii) What is the total charge of the crystal as a whole?

40. (i) Nature of metallic surface.
(ii) Work function of metals.
(iii) Temperature of the cathode.
(iv) Surface area of the cathode.

42. BCD : 5 – 01010101
D – 1010 0100

43. (i) The anode of D_1 is at low potential withrespect to cathode.
(ii) The anode of the diode is at low potential (0 – 5 V) with respect to the cathode which is at zero potential.
(iii) The anode of D_1 is at higher potential than the cathode. The anode of D_2 is at low potential with respect to the cathode.

44. (i) Is p-n junction diode a conductor?
(ii) How does the current vary with voltage in case of a conductor provided temperature and other physical condition remains constant?
(iii) How does the current vary with voltage in the case of a semiconductor?

45. (i) Platinum, lead, carbon, silicon and germanium are all tetravalent elements.
(ii) How does the nature of platinum and carbon differ from germanium or silicon?

Essay type questions

46. Emitter base junction is forward biased and collector base junction is reverse biased.

47. (i) Transducer and modulation.
(ii) Carrier wave.
(iii) Amplitude modulation and frequency modulation.
(iv) Amplifier, detector.

48. Electron gun, deflecting plates, fluorescent screen.

49. Large output voltage drop across load resistor.

50. Diode conducts when it is forward biased.



Concept Application Level—1

True or false

1. True
2. True
3. True
4. True
5. True
6. True
7. True.

Fill in the blanks

8. electrical
9. instructions
10. work function
11. negative, positive

12. byte
13. junction
14. doping

Match the following

15. A : e
    B : h
    C : a
    D : g
    E : b
    F : j
    G : c
    H : i
    I : d
    J : f

Multiple choice questions

16. Choice (2)
17. Choice (3)
18. Choice (1)
19. Choice (4)
20. Choice (3)
21. Choice (4)
22. Choice (4)
23. Choice (2)
24. Choice (2)
25. Choice (3)
26. Choice (2)
27. Choice (4)
28. Choice (4)
29. Choice (3)
30. Choice (1)

Concept Application Level—2,3

Key points

31. (i) A constant voltage of 1 V appears across the diode.
The remaining voltage (10 – 1 = 9 V) appears across the two resistors connected in series. The total resistance is 20 + 20 = 40 Ω and the PD is 9 V.
Find current using Ohm's law, which is recorded by the ammeter.
Find voltage across 20 Ω resistor, which is recorded by the voltmeter.
(ii) 0.225 A, 4.5 V

32. (i) Find whether the p-n junction diodes in a given circuit is forward or reverse biased.
(ii) If both p-n junction diodes are forward biased or both are reverse biased then the potential drop is the same across the two junctions.
(iii) If one p-n junction diode is forward biased and the other is reverse biased then the potential drop across the two diodes is unequal.
(iv) P.D. is equal in circuits (2) and (3).

33. (i) When current flows from A to B, diode D_1 is conducting and D_2 is non conducting. The resistance offered is 10 Ω.
When current flows from B to A, the diode D_2 is conducting and D_1 is non conducting. The resistance offered equals 20 Ω.
(ii) 10 Ω, 20 Ω, 7 V

34. (i) What is the resistance of an ideal semiconductor diode when it is forward biased?
(ii) Calculate the resistances in series $R_S = R_1 + R_D$ where $R_1 = 20$ Ω and R_D is the resistance of the diode.
(iii) Find the effective resistance using the formula $R = \frac{R_S \times R_2}{R_S + R_2}$ where $R_2 = 20$ Ω.
(iv) R gives the effective resistance between A and B.
(v) 10 Ω

35. (i) Rate of emission is directly proportional to temperature and varies inversely as the work function ($R_0 \propto T/w$) ------ (1)
(ii) Let 'R_1' be the rate of emission of Tungsten where ($R_1 \propto T_1/w_1$) ---- (2)
T_1 and w_1 are temperature and work function of tungsten.
(iii) Similarly find 'R_2' rate of emission of alkali oxide ($R_2 \propto T_2/w_2$) --- (3)

(iv) Since $R_1 = R_2 . T_1/w_1 = T_2/w_2$

(v) Find T_1 by substituting T_2, w_1 and w_2

(vi) 4520 K

36. (i) The PD across the diode remains constant at 0.7 V. The remaining PD (2.7 – 0.7 = 2V) appears across the resistor. Find the current using Ohm's law. Find the energy dissipated in the diode by using the formula.

$Q = V \times I \times t$

(ii) 84 J

37. (i) Find the effective resistance of the circuit using the formula $R = R_D + R_1$, where R_D is the resistance offered by the diode and R is the external resistance.

(ii) Find the current through the diode using the formula, $I = \frac{V - V_B}{R}$,

where V is the potential of the battery and V_B is the barrier potential.

(iii) 0.11 A.

38. (i) 1 Find whether P end of the diode is connected to higher potential and n-end to lower potential.

2. If the above is true, then the diode is forward biased.

(ii) 1. What is the potential of earth?

2. Now examine whether the P type is connected to higher potential and n type to lower potential.

(iii) reverse

(iv) reverse

(v) forward

(vi) forward

(v) forward

39. (i) Power = $I^2 R$ or V. I

Power = $\frac{Vq}{t}$ ----(1) ($I = q/t$)

Where V is the potential difference.

(ii) Find q by using $q = ne$

where n is the number of electrons, e is the charge of an electron.

(iii) Substitute the value of power, q and t (time) in (1) and find V.

(iv) $\frac{125}{16}$ or 7.8125 V

40. (i) What is the resistance (R_2) of an ideal diode when the current passes from A to B?

(ii) Find the effective resistance between A and B using the formula $R = \frac{R_1 \times R_2}{R_1 + R_2}$

(iii) What is the resistance (R_2) of an ideal diode when the current flows in the circuit from B to A?

(iv) Find the effective resistance using the formula $\frac{1}{R} = \frac{1}{R_1} + \frac{1}{R_2}$.

(v) 10 Ω, 20 Ω

41. (i) Potential difference × charge = change in kinetic energy. i.e.

P.D × q = 1/2 m($v_1^2 - v_2^2$)----(1)

where q is the charge of an electron, m is the mass of an electron, v_1 and v_2 are speeds of an electron in the n region and p region of the semiconductor respectively.

(ii) Substituting the value of m, q, v_1 and potential difference in equation (1), calculate v_2.

(iii) 2.95×10^6 m s^{-1}

42. (i) Audio signal is superposed over carrier wave.

(ii) Is the frequency of the audio signals same as the carrier waves?

(iii) How are the carrier wave and the message signal (audio signal) separated at the receiving end?

43. (i) What is the resistance of an ideal diode?

(ii) Find the value of resistance R_S connected in series.

(iii) Current through the diode $I = \frac{V}{R_S}$

(iv) What is the current (I_R) through 10 Ω resistor, which is connected parallel to semiconductor diode?

(v) Find the potential difference across the 10 Ω using the formula $V = I_R R$

(vi) Ammeter reading is 0.25 A and voltmeter reading is zero.

44. (i) The electric current can be found from

$$I = \frac{Q}{t} = \frac{ne}{t}$$

The power can be found from

$P = V \times I$

(ii) 0.8 A, 160 W

45. (i) Will the current flow through the output resistor R during the positive half cycle of AC input?

(ii) If so what is the nature of the current?

(iii) Will the current flow through the output resistor R during the negative half cycle of an AC input?

(iv) Is the current flowing through the resistor R unidirectional?

46. A horizontal line can be formed on the screen if the electron beam is deflected horizontally. To move it horizontally, over the whole screen, an alternating voltage is required.
A vertical line can be formed on the screen, if the electron beam is deflected vertically. To move it vertically over the whole screen, an alternating voltage, is required.

47. (i) Identify the diodes in the circuit which conduct during the positive half cycle of an AC input.

(ii) Does the current flow through the resistor?

(iii) Find out the diodes which conduct during the negative cycle of an AC input.

(iv) Find the direction of current flow through the resistor.

(v) Is the direction of the current through R unidirectional in both positive and negative cycles of an AC input?

48. The capacitor charges to peak AC voltage during positive half cycle and discharges through load (R) during the negative half cycle when the diode is non conducting. This, the voltage across R is maintained nearly at peak AC voltage.

49. (i) As the temperature of the cathode increases, it starts emitting more and more number of electrons which on reaching the anode increases the plate current.

(ii) As the plate potential is increased, the plate current increases but it remains constant when the rate at which electrons are emitted from the cathode equals the rate at which they reach anode.

50. (i) Minimum negative potential required to stop the electrons reaching the anode = kinetic energy of the electrons.