

**LEARNING MATERIAL  
OF  
ENGINEERING PHYSICS**



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# SEMESTER I

## UNIT-I

### SI UNITS AND STATICS

#### UNITS AND MEASUREMENT

##### **Introduction:**

The word physics comes from the Greek word meaning "nature". Today physics is treated as the most fundamental branch of science and finds numerous applications in all walks of life. Physics deals with matter in relation to energy and the accurate measurement of the same. Thus physics is inherently a science of measurement. The fundamentals of physics form the basis for the study and the development of engineering and technology.

Measurement consists of the comparison of an unknown quantity with a known fixed quantity. The quantity used as the standard of measurement is called 'unit'. For example, a vegetable vendor weighs the vegetables in terms of units like kilogram.

##### **Fundamental physical quantities**

Fundamental quantities are the quantities which cannot be expressed in terms of any other physical quantity.

(eg) length, mass and time.

##### **Derived quantities**

Quantities that can be expressed in terms of fundamental quantities are called derived quantities.

(eg) area, volume, density.

##### **Unit**

Unit of a physical quantity is defined as the accepted standard used for comparison of given physical quantity.

The unit in which the fundamental quantities are measured are called fundamental unit and the units used to measure derived quantities are called derived units.

## SI Units

SI unit is the abbreviation for **Systeme International d'units** and is the modern form of metric system finally agreed upon at the eleventh International conference of weights and measures, 1960. This system of units is now being adopted throughout the world and will remain the primary system of units of measurement. SI system possesses features that make it logically superior to any other system.

There are seven fundamental units (base units) and two supplementary units.

### SI system of units

Physical quantity	Unit	Symbol
<b>Fundamental quantities</b>		
1. Length	metre	m
2. Mass	kilogram	kg
3. Time	second	s
4. Electric current	ampere	A
5. Temperature	kelvin	K
6. Luminous Intensity	candela	cd
7. Amount of substance	mole	mol
<b>Supplementary quantities</b>		
1. Plane angle	radian	rad
2. Solid angle	steradian	sr

### Derived quantities and their units

SI.No	Physical quantity	Formula	Unit	Symbol
1	Area of the square	side × side	metre <sup>2</sup> or square metre	m <sup>2</sup>
2	Volume of the cube	side × side × side	metre <sup>3</sup> or cubic metre	m <sup>3</sup>
3	Density	Mass/volume	kilogrammetre <sup>-3</sup>	kgm <sup>-3</sup>
4	Velocity	Displacement/ time	metresecond <sup>-1</sup>	ms <sup>-1</sup>
5	Acceleration	velocity/time	metresecond <sup>-2</sup>	ms <sup>-2</sup>
6	Momentum	mass × velocity	Kilogrammetre second <sup>-1</sup>	kgms <sup>-1</sup>

Sl.No	Physical quantity	Formula	Unit	Symbol
7	Force	mass× acceleration	newton	N
8	Impulse	force×time	newtonsecond	Ns
9	Work(or) Energy	force× displacement	newtonmetreor joule	J
10	Power	Work/time	joulesecond <sup>-1</sup> or watt	W

### Dimensions

The fundamental physical quantities namely length, mass and time are symbolically represented by the capital letters L, M and T respectively.

Dimensional formula is the formula in which the given physical quantity is expressed in terms of the fundamental quantities raised to suitable powers.

### Dimensional formula for derived physical quantities.

1. Area of the square = side × side

Applying dimensions

$$\text{Area of the square} = L \times L = L^2$$

$$\text{Dimensional formula for Area} = L^2$$

2. Volume of the cube = side × side × side

Applying dimensions

$$\text{Volume of the cube} = L \times L \times L = L^3$$

$$\text{Dimensional formula for Volume} = L^3$$

3. Density is the mass per unit Volume.

$$\text{Density} = \frac{\text{mass}}{\text{volume}}$$

Applying dimensions

$$\text{Density} = \frac{M}{L^3} = ML^{-3}$$

$$\text{Dimensional formula for Density} = ML^{-3}$$

4. Velocity is the rate of change of displacement

$$\text{Velocity} = \frac{\text{displacement}}{\text{time}}$$

Applying dimensions,

$$\text{velocity} = \frac{L}{T} = LT^{-1}$$

Dimensional formula for Velocity =  $LT^{-1}$

5. Acceleration is the rate of change of velocity

$$\begin{aligned} \text{Acceleration} &= \frac{\text{velocity}}{\text{time}} \\ &= \frac{\text{displacement}}{\text{time}} \times \frac{1}{\text{time}} \end{aligned}$$

$$\begin{aligned} \text{Applying dimensions} \\ &= \frac{L}{T \times T} = \frac{L}{T^2} = LT^{-2} \end{aligned}$$

Dimensional formula for Acceleration =  $LT^{-2}$

6. Momentum is the product of mass and velocity.

Momentum = mass  $\times$  velocity

$$= \text{mass} \times \frac{\text{displacement}}{\text{time}}$$

Applying dimensions

$$\text{Momentum} = M \times \frac{L}{T} = MLT^{-1}$$

Dimensional formula for Momentum =  $MLT^{-1}$

7. Force is the product of mass and acceleration.

Force = mass  $\times$  acceleration

$$\begin{aligned} &= \text{mass} \times \frac{\text{velocity}}{\text{time}} \\ &= \frac{\text{mass}}{e} \times \frac{\text{displacement}}{\text{time}^2} \end{aligned}$$

Applying dimensions,

$$\text{Force} = \frac{M}{T} \times \frac{L}{T} = MLT^{-2}$$

Dimensional formula for force =  $MLT^{-2}$

8. Impulse is the product of force and time

$$\text{Impulse} = \text{force} \times \text{time}$$

$$= \text{mass} \times \text{acceleration} \times \text{time}$$

$$= \text{mass} \times \frac{\text{velocity}}{\text{time}} \times \text{time}$$

$$= \text{mass} \times \text{velocity}$$

$$= \text{mass} \times (\text{Displacement} / \text{time})$$

Applying dimensions,

$$\text{Impulse} = \frac{M \times L}{T} = MLT^{-1}$$

$$\text{Dimensional formula for Impulse} = MLT^{-1}$$

9. Work (or) Energy is the product of force and displacement

$$\text{Work or Energy} = \text{force} \times \text{displacement}$$

$$= \text{mass} \times \text{acceleration} \times \text{displacement}$$

$$= \text{mass} \times \frac{\text{velocity}}{\text{time}} \times \text{displacement}$$

$$= \frac{\text{mass} \times \text{displacement}}{\text{time}} \times \text{displacement}$$

Applying dimensions

$$\text{Work or energy} = M \times \frac{L}{T} \times L = ML^2T^{-2}$$

$$\text{Dimensional formula for Work or Energy} = ML^2T^{-2}$$

10. Power is the rate of doing work.

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

=

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

Applying dimensions,

$$\text{Power} = MLT^{-2} \times \frac{L}{T} = ML^2T^{-3}$$

$$\text{Dimensional formula for Power} = ML^2T^{-3}$$

## CONVENTIONS TO BE FOLLOWED IN SI UNITS

1. When we write a unit in full, the first letters should not be in capital letter.  
(eg) metre and not as Metre  
kilogram and not as Kilogram
2. The symbols of unit in the name of scientist should be in upper case (capital) letter  
(eg) N for newton, J for joule.
3. Only the singular form of the unit is to be used.  
(eg) km and not as kms  
Kg and not as kgs
4. There should be no full stop at the end of the abbreviations.  
(eg) mm and not as mm.  
kg and not as kg.
5. When temperature is expressed in Kelvin, the degree sign is omitted. (eg) 273K not as 273°K
6. Only accepted symbols should be used.  
(eg) ampere is represented as 'A' and not 'amp'  
second is represented as 's' and not 'sec'
7. Use of the solidus or slash is to be avoided, but when used, not more than one solidus be employed.  
(eg) J/k.mol or  $\text{J.k.mol}^{-1}$  but not  $\text{J/k/mol}$   
 $\text{J.kg}^{-1}\text{K}^{-1}$  but not  $\text{J/kg/K}$
8. One letter space is always to be left between the number and the symbol of the unit.  
(eg) 2.3m and not as 2.3m  
 $\text{Kgms}^{-2}$  and not  $\text{kgms}^{-2}$
9. The numerical value of any physical quantity should be expressed as  $p = a \times 10^m$ . Here 'a' is a number between 1 and 10 and 'm' is the appropriate power of 10.  
(eg) velocity of light  $c = 2.997 \times 10^8 \text{ms}^{-1}$ .

### Multiples and Submultiples of units

In SI, some units are not of convenient size to measure certain quantities. Hence multiples and submultiples of the base units are used in measurements.

Multiplication factor	Prefix	Symbol
$1000000000000=10^{12}$	Tera	T
$1000000\ 000=10^9$	Giga	G
$1000\ 000=10^6$	Mega	M
$1\ 000=10^3$	Kilo	k
$100=10^2$	hecto*	h
$10=10^1$	deca*	da
$0.1=10^{-1}$	deci	d
$0.01=10^{-2}$	centi*	c
$0.001=10^{-3}$	milli	m
$0.000001=10^{-6}$	micro	$\mu$
$0.000000001=10^{-9}$	nano	n
$0.000000000001=10^{-12}$	pico	p
$0.0000000000000001=10^{-15}$	femto	f
$0.000000000000000001=10^{-18}$	atto	a

\*These are not preferred ones.They are used where the other prefixes areinconvenient

## STATICS

Statics is the part of mechanics which deals with forces acting on bodiesatrest.Theweightofthebody,thetensionofastring,theloadand thereactionarethedifferentnamesofforcesthatareusedinstatics.

Any force can be represented in magnitude and direction by a straightlinewithanarrowhead.Thebeginningofthelinerepresentsthe pointofapplicationoftheforce.Thearrowheadrepresentsthedirection oftheforce.Themagnitudeoftheforceisgivenbythelengthoftheline drawntoscale.



## Scalar and Vector Quantities

Physical quantities which have magnitude only are called scalar quantities.

(e.g) Mass, Volume, Speed

Physical quantities which have both magnitude and direction are called vector quantities.

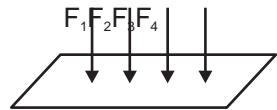
(e.g) Velocity, Momentum, Force

## Concurrent forces

Two or more than two forces acting at a point are called concurrent forces.

## Coplanar forces

Two or more than two forces lie on same plane are called coplanar forces.



## Resultant

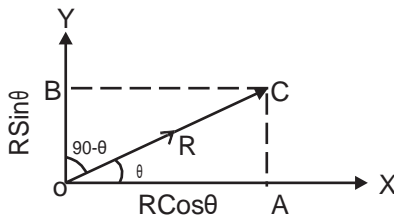
Resultant is that single force which produces the same effect as that produced by two or more forces acting on the body.

## Equilibrant

Equilibrant is the single force which acting along with the other forces keeps the body in equilibrium.

## Resolution of a vector into two perpendicular components

The process of splitting up a vector (force) into two perpendicular component parts is known as resolution of a vector. As the two component forces are mutually perpendicular, they are called rectangular components.



Let a force  $R$  act at a point  $O$  at an angle  $\theta$  with  $x$ -axis. This force can be resolved into two rectangular components along  $x$  and  $y$  axis. Two lines  $CA$  and  $CB$  are drawn perpendicular to  $x$  and  $y$  axis as shown in the figure.

In the right angled triangle  $OAC$ ,

$$\begin{aligned} \cos\theta &= \frac{OA}{OC} \\ OA &= OC \cos\theta \\ &= R \cos\theta \end{aligned}$$

Therefore the horizontal component of the force  $R$  is  $R \cos\theta$ . In the right angled triangle  $OBC$ ,

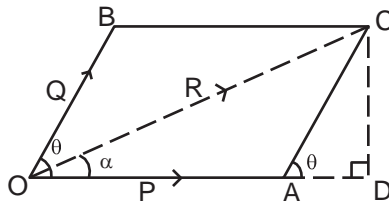
$$\begin{aligned} \cos(90-\theta) &= \frac{OB}{OC} \\ \sin\theta &= \frac{OB}{OC} \\ OB &= OC \sin\theta = R \sin\theta \end{aligned}$$

The vertical component of the force  $R$  is  $R \sin\theta$ .

### Parallelogram law of forces

If two forces acting at a point are represented in magnitude and direction by the two adjacent sides of a parallelogram drawn from that point, their resultant is given in magnitude and direction by the diagonal of the parallelogram drawn through that point.

### Expressions for the magnitude and direction of the resultant of two forces acting at a point, with an acute angle between them.



Consider two forces  $P$  and  $Q$  acting at a point  $O$  with an angle  $\theta$  between them as shown in the figure. The forces  $P$  and  $Q$  are represented in magnitude and direction by the sides  $OA$  and  $OB$  of a parallelogram

OACB. The diagonal OC gives the resultant of two forces P and Q in magnitude and direction. From C draw a perpendicular to meet OA extended at D.

In the right angled triangle OCD

$$\begin{aligned} OC^2 &= OD^2 + CD^2 \\ &= (OA + AD)^2 + CD^2 \end{aligned}$$

$$OC^2 = OA^2 + AD^2 + 2OA \cdot AD + CD^2 \quad \dots(1)$$

In the right angled triangle ACD

$$\cos \theta = \frac{AD}{AC} = \frac{AD}{Q}$$

$$AD = Q \cos \theta$$

$$\sin \theta = \frac{CD}{AC} = \frac{CD}{Q}$$

$$CD = Q \sin \theta.$$

Substituting AD and CD values in eqn. (1)

$$R^2 = P^2 + Q^2 \cos^2 \theta + 2PQ \cos \theta + Q^2 \sin^2 \theta$$

$$R^2 = P^2 + Q^2 (\cos^2 \theta + \sin^2 \theta) + 2PQ \cos \theta$$

$$R^2 = P^2 + Q^2 + 2PQ \cos \theta \text{ where } \cos^2 \theta + \sin^2 \theta = 1 \text{ Resultant,}$$

$$R = \sqrt{P^2 + Q^2 + 2PQ \cos \theta}$$

This is the expression for the magnitude of the resultant.

Let  $\alpha$  be the angle between the resultant R and the force P.

$$\tan \alpha = \frac{CD}{OD} = \frac{CD}{OA + AD} = \frac{Q \sin \theta}{P + Q \cos \theta}$$

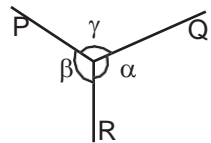
$$\text{Direction, } \alpha = \tan^{-1} \left[ \frac{(Q \sin \theta)}{(P + Q \cos \theta)} \right]$$

In the right angled triangle COD,

This is the expression for the direction of the resultant.

### Lami's theorem

If three forces acting at a point are in equilibrium, then each force is directly proportional to the sine of the angle between the other two forces.

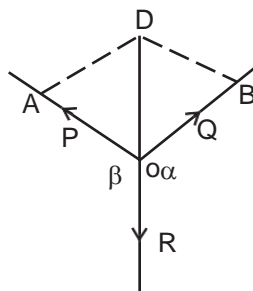
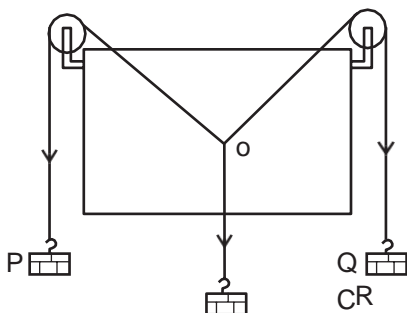


Let P, Q and R be three forces acting at a point O. Under the action of three forces, the point is at rest. Let  $\alpha$ ,  $\beta$  and  $\gamma$  be the angles opposite to the forces P, Q and R respectively.

By Lami's theorem

$$\begin{aligned}
 P \sin \alpha &\Rightarrow P = K \sin \alpha \Rightarrow \frac{P}{\sin \alpha} = K \\
 Q \sin \beta &\Rightarrow Q = K \sin \beta \Rightarrow \frac{Q}{\sin \beta} = K \\
 R \sin \gamma &\Rightarrow R = K \sin \gamma \Rightarrow \frac{R}{\sin \gamma} = K \\
 \therefore \frac{P}{\sin \alpha} &= \frac{Q}{\sin \beta} = \frac{R}{\sin \gamma} = \text{constant}
 \end{aligned}$$

### Experimental verification of parallelogram law of forces



A drawing board is fixed vertically on the wall. Two frictionless pulleys are fixed at the top corners of the drawing board. A light and inextensible string is passed over these pulleys. Another short string is tied to the first string at O. Weighthangers P, Q and R are tied at the ends of the strings as shown in the figure. The weights P, Q and R are adjusted suitably such that the point O is at rest.

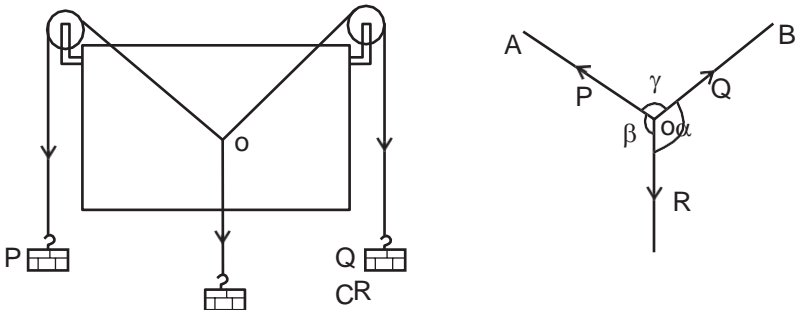
A white paper is held just behind the string without touching them. The point O and the directions of the strings are marked on the paper. Then the paper is removed from the board. With suitable scale the points A, B and C are marked. OA, OB and OC represent the forces P, Q and R respectively.

The parallelogram  $OADB$  is drawn. Then the diagonal  $OD$  is drawn. The length of the diagonal  $OD$  and  $\angle COD$  are measured and tabulated.

Sl.No	P	Q	R	OA	OB	OC	OD	$\angle COD$

The experiment is repeated for different values of  $P, Q$  and  $R$ . It is found that  $OC = OD$  and  $\angle COD = 180^\circ$ . Thus the parallelogram law of forces is verified experimentally.

### Experimental verification of Lami's theorem



A drawing board is fixed vertically on the wall. Two frictionless pulleys are fixed at the top corners of the drawing board. A light and inextensible string is passed over these pulleys. Another short string is tied to the first string at  $O$ . Weighthangers  $P, Q$  and  $R$  are tied at the ends of the strings as shown in the figure. The weights  $P, Q$  and  $R$  are readjusted suitably such that the point  $O$  is at rest.

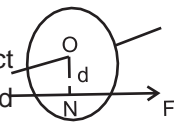
A white paper is held just behind the string without touching them. The point  $O$  and the directions of the forces  $P, Q$  and  $R$  are marked on the paper. Then the paper is removed from the board. Let  $\alpha, \beta$  and  $\gamma$  are the angles opposite to the forces  $P, Q$  and  $R$  respectively. The angles are measured with the help of protractor and tabulated.

Sl.No	P	Q	R	$\alpha$	$\beta$	$\gamma$	$\frac{P}{\sin\alpha}$	$\frac{Q}{\sin\beta}$	$\frac{R}{\sin\gamma}$

The experiment is repeated for different values of P, Q and R. The ratios  $(P/\sin\alpha)$ ,  $(Q/\sin\beta)$  and  $(R/\sin\gamma)$  are calculated. For each set of readings, the ratios are found to be equal. Thus Lami's theorem is verified experimentally.

### Moment of a force

Consider a body which is fixed at a point, about which it can rotate freely. Let a force  $F$  is acting on the body. The effect of the force is to rotate the body about the fixed point, unless the line of action of the force passes through that fixed point  $O$ . This rotating tendency or the turning effect of the force about that point is called moment of force. i.e. the turning effect of the force acting on a body about an axis or point is called moment of force.



It is common experience that in opening or closing a door, the force we apply rotates the door about its hinges. This rotating effect is known as the moment of force.

Moment of force is also equal to the product of the magnitude of the force and the perpendicular distance of the line of action of the force from the axis of rotation.

$$\begin{aligned} \text{Moment of force about the point } O &= \text{Force} \times \text{perpendicular distance} \\ &\quad \text{of the force from the point } O \\ &= F \times ON \end{aligned}$$

The unit of moment of force is Nm and the dimensional formula is  $ML^2T^{-2}$ .

### Clockwise and anti-clockwise moments

If the moment of a force turns or rotates the body in clockwise direction, then it is called as clockwise moment.

If the moment of a force turns or rotates the body in anti-clockwise direction, then it is called anti-clockwise moment.

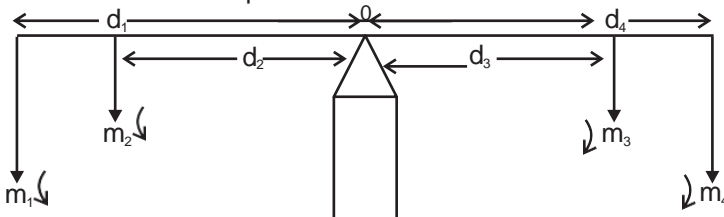


$$\text{Clockwise moment} = m_2 \times d_2$$

$$\text{Anti-clockwise moment} = m_1 \times d_1$$

## Principle of moments

The principle of moment states that if a body is in equilibrium under the action of a number of parallel forces, the sum of the clockwise moments about any point must be equal to the sum of anti-clockwise moments about the same point.



According to the principle of moments,

Sum of anti-clockwise moments = sum of clockwise moments

From figure,

$$m_1 d_1 + m_2 d_2 = m_3 d_3 + m_4 d_4.$$

## Couple:

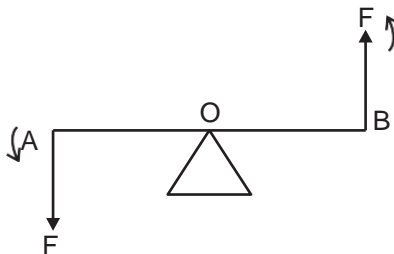
Two equal and opposite parallel forces acting at different points in a body constitute a couple. A body acted upon by the couple will rotate the body in clockwise direction or anti-clockwise direction.

Steering wheel and pedals of bicycles are the examples for couple, where the two forces are equal but acting in opposite direction.

## Torque acting due to couple (or) Moment of couple

Torque is calculated by the product of either of forces forming the couple and the arm of the couple.

i.e.) Torque = one of the force  $\times$  perpendicular distance between the forces.

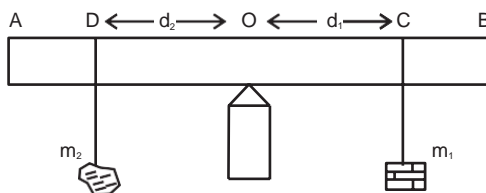


Consider two equal forces  $F$  and  $F$  acting on the arm  $AB$ . Let  $O$  be the midpoint of the arm. The forces  $F$  and  $F$  are acting in opposite directions as shown in the figure, they constitute a couple. Then

$$\text{Moment of couple or torque} = F \times AB.$$

If the forces acting on the body have the same line of action then the moment becomes zero. The torque is maximum when the forces are at right angles to the arm i.e.  $\theta = 90^\circ$ .

### Determination of mass of the given body using principle of moments.



A meter scale  $AB$  is supported on a knife edge at its centre such that it remains in equilibrium state. Let  $O$  be the centre of the scale. On the right side suspend the weight hanger of mass  $m_1$ , and on the left side suspend the given body  $m_2$  whose mass is to be calculated. Adjust the positions of  $m_1$  and  $m_2$  until the scale comes to the exact equilibrium position. Let  $d_1(OC)$  and  $d_2(OD)$  be the distances of  $m_1$  and  $m_2$ .

According to the principle of moments

$$m_2 \times d_2 = m_1 \times d_1$$

$$\text{The mass of the given body } m_2 = (m_1 \times d_1) / d_2$$

By changing the values of  $m_1$ , the experiment is repeated and the readings are tabulated.

Sl.No	$m_1$ kg	$d_1$ m	$d_2$ m	$m_2 = (m_1 \times d_1) / d_2$ kg

Mean =

The average value of the last column gives the mass of the body.



## WORKED EXAMPLES

1. Find the magnitude and direction of the two forces 20N and 25N acting at an angle  $60^\circ$  to each other.

Given

$$P=20\text{N}$$

$$Q=25\text{N}$$

$$\theta=60^\circ$$

$$\begin{aligned} \text{Resultant } R &= \sqrt{P^2+Q^2+2PQ\cos\theta} \\ &= \sqrt{20^2+25^2+2\times 20\times 25\cos 60} \\ &= \sqrt{1525} \end{aligned}$$

$$R=39.05\text{N}$$

$$\text{Direction } \alpha = \tan^{-1} \left[ \frac{Q\sin\theta}{P+Q\cos\theta} \right]$$

$$\alpha = \tan^{-1} \left[ \frac{25\sin 60^\circ}{20+25\cos 60^\circ} \right]$$

$$\alpha = \tan^{-1}(0.6661)$$

$$\alpha = 33^\circ 40'$$

2. Find the magnitude and direction of the resultant of two forces 30N and 40N acting at right angle to each other.

Given

$$P=30\text{N}$$

$$Q=40\text{N} \quad \theta=90^\circ$$

$$\begin{aligned} \text{Resultant } R &= \sqrt{P^2+Q^2+2PQ\cos\theta} \\ &= \sqrt{30^2+40^2+2\times 30\times 40\times \cos 90} \\ &= \sqrt{2500} \end{aligned}$$

$$R=50\text{N}$$

$$\text{Direction } \alpha = \tan^{-1} \left[ \frac{(40\times \sin 90)}{(30+40\cos 90)} \right]$$

$$= \tan^{-1}(1.333)$$

$$\alpha = 53^\circ 7'$$

3. If the resultant of two equal force is  $\sqrt{3}$  times a single force, find the angle between the forces.

Given

Two forces are equal

$$P=Q=P$$

$$R=\sqrt{3}P$$

$$R^2=P^2+Q^2+2PQ\cos\theta$$

$$(3P)^2=P^2+P^2+2.P.P.\cos\theta$$

$$3P^2-2P^2=2P^2\cos\theta$$

$$P^2=2P^2\cos\theta$$

$$\cos\theta=\frac{1}{2}$$

$$\therefore\theta=60^\circ$$

Angle between the forces is  $60^\circ$

4. If the resultant of two equal forces inclined to each other at  $60^\circ$  is  $8\sqrt{3}$  N, find the component forces.

Given

Two forces are equal.

$$P=Q=P$$

$$R=8\sqrt{3}\text{ N}$$

$$\theta=60^\circ$$

$$R^2=P^2+Q^2+2PQ\cos\theta$$

$$(8\sqrt{3})^2=P^2+P^2+2.P.P.\cos60^\circ$$

$$192=2P^2+2P^2.\frac{1}{2}$$

$$192=3P^2$$

$$64=P^2$$

$$=8\text{N}$$

Component force is 8N

5. The sum of two forces is 8N and the magnitude of the resultant is at right angle to the smaller force 4N. Find the forces and the inclination between them.

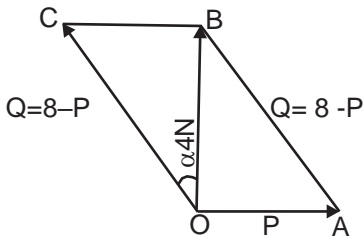
Given

$$\text{Sum of the two forces} = 8\text{N}$$

Resultant of 2 forces = 4N

Let the smaller force be P

Larger force Q = (8 - P)



In the right-angled triangle OAB,

$$AB^2 = OA^2 + OB^2$$

$$(8 - P)^2 = P^2 + 4^2$$

$$64 - 16P + P^2 = P^2 + 16$$

$$16P = 48$$

$$P = 3\text{N}$$

Therefore the larger force Q = 8 - 3

$$Q = 5\text{N}$$

In triangle OCB,

$$\tan \alpha = \frac{BC}{OB} = \frac{P}{4} = \frac{3}{4} = 0.75$$

$$\alpha = \tan^{-1}(0.75)$$

$$\alpha = 36^\circ 52'$$

Angle between the forces P and Q

$$90 + \alpha = 90 + 36^\circ 52'$$

$$= 126^\circ 52'$$

## QUESTIONS

### Part A

1. Define unit of a physical quantity.
2. What are the fundamental quantities?
3. Give two examples for fundamental quantities.
4. What are the derived quantities?
5. Give two examples for derived quantities.
6. Expand SI.
7. What are the supplementary quantities and their SI units?
8. Write the SI units for plane and solid angles.
9. What are the units for momentum and force?
10. What are the units for impulse and work?
11. What are the units for energy and power?
12. Derive the dimensional formula for impulse.
13. Derive the dimensional formula for work or energy.
14. Derive the dimensional formula for power.
15. Derive the dimensional formula for force.
16. Define scalar quantity.
17. Define vector quantity.
18. Give two examples for scalar quantities.
19. Give two examples for vector quantities.
20. What is resultant?
21. What is equilibrant?
22. Write the parallelogram law of forces.
23. Write the Lami's theorem.
24. Write any two conventions followed in SI.
25. Define concurrent forces.
26. Define coplanar forces.
27. Define moment of a force.
28. Write the principle of moments.
29. Define couple.
30. Define moment of a couple.
31. Define clockwise moment.
32. Define anticlockwise moment.

## Part-B

1. Derive dimensional formula for impulse, work and power.
2. Give the rules and conventions followed while writing SI units.
3. Explain how a vector can be resolved into two rectangular components.
4. Derive expressions for the magnitude and direction of the resultant of two forces acting at a point with an acute angle between them.
5. Describe an experiment to verify the parallelogram law of forces.
6. Describe an experiment to verify Lami's theorem.
7. Describe an experiment to determine the mass of the given body using principle of moments.

## Exercise Problems

1. Find the magnitude and direction of the resultant of two forces of 5 N and 3 N at an angle of  $60^\circ$ .  
(Ans:  $R=7\text{N}, \alpha=21^\circ 47'$ )
2. Two forces of magnitude 4 N and 2.5 N acting at a point inclined at an angle of  $40^\circ$  to each other. Find their resultant.  
(Ans:  $R=6.149\text{N}, \alpha=15^\circ 8'$ )
3. Find the resultant of two forces 3 N and 4 N acting on a particle in direction inclined at  $30^\circ$ .  
(Ans:  $R=6.767\text{N}, \alpha=17^\circ 19'$ )
4. Two forces of magnitude 4 N and 3 N respectively, act on a particle at right angle to each other. Find the magnitude and direction of the resultant of two forces.  
(Ans:  $R=5\text{N}, \alpha=36^\circ 52'$ )
5. If the resultant of two equal forces is 2 times the single force, find the angle between them.  
(Ans:  $\theta=90^\circ$ )
6. If the resultant of two equal forces inclined to each other at  $60^\circ$  is 8.660 N, find the component force.  
(Ans: 5N)
7. If the resultant of two forces 6 N and 8 N is 12 N, find the angle between them.  
(Ans:  $62^\circ 43'$ )
8. The sum of two forces inclined to each other at an angle is 18 N and their resultant which is perpendicular to the smaller force is 12 N. Find the forces and angle between them.  
(Ans:  $P=5\text{N}, Q=13\text{N}, \theta=112^\circ 36'$ )

# UNIT - II

## PROPERTIES OF MATTER

### ELASTICITY

#### Introduction

When an external force is applied on a body, which is not free to move, the shape and size of the body change. The force applied is called deforming force. When the deforming forces are removed, the body tends to regain its original shape and size due to a force developed within the body. The force developed within the body, which is equal and opposite to deforming force is called restoring force.

Bodies, which completely regain their original size and shape after the removal of the deforming force, are called elastic bodies. Bodies which change the shape and size on the application of force and which do not regain their original condition on removal of the deforming force are said to be plastic bodies. Bodies which do not change their shape and size on application of force are called rigid bodies.

The property by virtue of which a body tends to regain its original shape and size after removal of the deforming force is known as elasticity.

#### Stress:

When an external force is acting on an elastic body, it causes deformation (change in shape or in size or both). At the same time, due to elastic property, a force is developed within the material, which is equal and opposite to the applied force, to bring the body to its original shape and size. This force is 'restoring force'.

$$\text{Stress} = \frac{\text{Force}}{\text{Area}} = \frac{F}{A}$$

The stress is defined as the restoring force acting on unit area.

Since the applied force and the restoring force are equal in magnitude, the 'stress' is measured as the applied force acting per unit area.

The unit for stress is newton metre<sup>-2</sup> with symbol N m<sup>-2</sup> or 'pascal' with symbol 'Pa'.

When the applied force tends to compress the body, the stress is compressive. When it tends to increase the length in the direction of the force, it is tensile and when it acts parallel to the surface of a body, the stress is tangential stress.

**Hooke's law:**

Within the elastic limits, the strain produced in a body is directly proportional to the stress which causes it.

$$\begin{aligned} \text{i.e., strain} &\propto \text{stress} \\ \text{or stress} &\propto \text{strain} \\ \text{stress} &= (\text{constant}) \text{strain} \end{aligned}$$

This constant is called 'modulus of elasticity'.

**Types of strains:**

$$\therefore \frac{\text{stress}}{\text{strain}} = \text{a constant}$$

**(1) Linear Strain:**

When a wire or bar is subjected to two equal and opposite forces, namely pulls, at its ends, there is an increase in the length. If the forces are tensile, the body is elongated. If the forces are compressive, the length is shortened in the direction of the forces. This is called the 'linear strain'.

The linear strain is defined as the ratio of change in length to the original length. If the change (increase or decrease) in length is 'l' in a wire or bar of original length L,

As the linear strain is ratio of lengths, it has no unit.

$$\text{linear strain} = \frac{\text{Change in length } l}{\text{original length } L}$$

**(2) Bulk (or) Volume Strain:**

When a force is applied uniformly and normally to the entire surface of the body, there is a change in volume of the body, without any change in its shape. This strain is called 'bulk or volume strain'.

Volume strain is defined as the ratio of change in volume to the original volume. It has also no unit. If 'v' is the change in volume produced in a body of original volume 'V',

$$\text{bulk or volume strain} = \frac{\text{Change in volume}}{\text{original volume}} = \frac{\Delta V}{V}$$

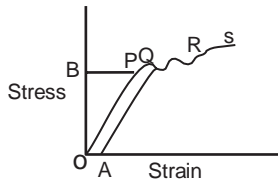
### (3) Shearing (or) Rigidity strain:

When a force is applied parallel to one face of a body, the opposite side being fixed, there is a change in shape but not in size of the body. This strain is called the shearing strain.

Solids alone can have a shearing strain. It is measured by the angle of the shear ' $\theta$ ' in radian.

### Elastic behaviour of a material-stress strain curve:

Let a wire be suspended from a rigid support. At the free end, a weight hanger is provided on which weights could be added to study the behaviour of the wire under different load conditions. The extension of the wire is measured for various load and stress – strain graph is drawn as shown in the diagram below.



The region  $OP$  is linear. Within a normal stress, strain is proportional to the applied stress. This is Hooke's law. Up to  $P$  when the load is removed the wire regains its original length along  $PO$ . The point  $P$  represents the elastic limit,  $PO$  represents the elastic range of the material and  $OB$  is the elastic strength.

Beyond  $P$ , the graph is not linear. In the region  $PQ$  the material is partly elastic and partly plastic. From  $Q$ , if we start decreasing the load, the graph does not come to  $O$  via  $P$ , but traces a straight line  $QA$ . Thus a permanent strain  $OA$  is caused in the wire. This is called permanent set.

Beyond  $Q$  addition of even a very small load causes enormous strain. This point  $Q$  is called the yield point. The region  $QR$  is the plastic range.

Beyond  $R$ , the wire loses its shape and becomes thinner and thinner in diameter and ultimately breaks, say at  $S$ . Therefore  $S$  is the breaking point. The stress corresponding to  $S$  is called breaking stress.



### Three moduli of elasticity

There are three types of moduli depending upon the three kinds of strain.

#### (a) Young's modulus (E):

It is defined as the ratio of linear stress to linear strain.

Let a wire of initial length  $L$  and cross-sectional area  $A$ , undergo an extension  $l$ , when a stretching force  $F$ , is applied in the direction of its length.

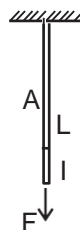
The modulus of elasticity, in this case, is called Young's modulus and is given by

$$\text{i.e., Young's modulus (E)} = \frac{\text{Linear stress}}{\text{Linear strain}}$$

$$\text{Then, longitudinal or linear stress} = \frac{F}{A}$$

$$\text{and longitudinal strain} = \frac{l}{L}$$

$$E = \frac{\text{Linear stress}}{\text{Linear strain}} = \frac{F/A}{l/L} = \frac{F.L}{A.l}$$



The unit for Young's modulus is newton metre<sup>-2</sup> with symbol  $\text{Nm}^{-2}$ . The single term unit which is widely used for Young's modulus is 'pascal' with symbol 'Pa'.

#### (b) Bulk (or) Volume modulus (k):

It is defined as the ratio of bulk stress to bulk strain.

When a body is subjected to a uniform compressive force, its volume decreases and the strain produced is a bulk or volume strain.

If  $v$  is the change in volume and  $V$  is the original volume, then

If  $F$  is the total compressive force acting on a total area  $A$ , then bulk stress =  $F/A = P$

$$\text{Bulk modulus (k)} = \frac{\text{Bulk stress}}{\text{Bulk strain}}$$

If  $P$  is the stress applied i.e. (force/unit area) then,

$$\text{bulk strain} = \frac{\Delta V}{V} = \frac{\Delta P \cdot V}{P \cdot V}$$

$$\text{bulk modulus } k = \frac{P}{\frac{\Delta V}{V}}$$

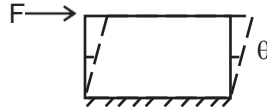
The unit for bulk modulus is 'newton metre<sup>-2</sup>' with symbol 'Nm<sup>-2</sup>'. The single term unit which is widely used for bulk modulus is 'pascal' with symbol 'Pa'.

**(c) Shearing (or) Rigidity modulus 'n':**

The ratio of the shearing stress applied to the body to the shearing strain produced is called the rigidity modulus and denoted by the letter 'n'.

If T is the tangential force/unit area and if θ is the angle of shear measured in radian, then

$$\text{rigidity modulus } n = \frac{T}{\theta} = \frac{F}{A\theta}$$



The unit for rigidity modulus is 'newton metre<sup>-2</sup> radian<sup>-1</sup>' with symbol 'N m<sup>-2</sup> rad<sup>-1</sup>'. The other unit which is widely used for rigidity modulus is 'pascal radian<sup>-1</sup>' with symbol 'Pa rad<sup>-1</sup>'.

**Poisson's ratio: σ**

When a tensile stress is applied to a wire, the wire undergoes not only an extension of length in the direction of the force but also a contraction in its thickness. The ratio of decrease in thickness to the original thickness in lateral direction is known as lateral contraction.

The ratio of lateral contraction to linear elongation is called Poisson's ratio.

$$\text{Poisson's ratio } \sigma = \frac{\text{Lateral contraction}}{\text{Linear elongation}}$$

**Bending of beams:**

A beam is a body having uniform cross section, the length of which is very large compared to its thickness. When such a beam is fixed at one end and loaded at the other end, it is called a cantilever.

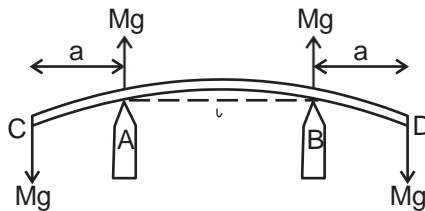
A beam is supported symmetrically on two knife edges A and B at the same level. It is loaded symmetrically with two equal weights W, W, beyond the knife edges, at C and D, so that AC = BD = a. Now the beam is

elevated at the centre. This type of bending is called uniform bending of beam.

A beam is supported symmetrically on two knife edges A and B near its ends in a horizontal level. A weight hanger  $W$  is suspended by means of a loop of thread from a point which is exactly midway between the two knife edges. Now the beam is depressed at the centre. This type of bending is called non-uniform bending of beam.

**Experiment to determine the Young's modulus of the material of a beam by uniform bending method:**

The given bar is supported symmetrically on two knife edges in a horizontal level, so that a quarter of the bar projects beyond each knife edge. Two weight hangers are suspended from two loops of string tightly attached at distances about 5 cm from each end so that their distances from the knife edges are equal. A pin is fixed vertically at the midpoint of the bar. A traveling microscope is focused on the pin such that the horizontal cross wire coincides with the tip of the pin. The reading in the



vertical scale of the microscope is noted. Weights are added in steps of  $M$  kg and the corresponding readings of the image of the pin in the microscope are noted. Similarly readings are taken while unloading and tabulated.

The mean elevation  $y$  is found for a load of  $M$  kg. The distance ' $l$ ' between the knife edges is measured. The distance ' $a$ ' between the point of suspension of the weight and the nearest knife edge is also measured. The breadth ' $b$ ' of the bar is found accurately with the vernier calipers and its thickness  $d$  is found with a screw gauge. Young's modulus is calculated using the formula.

$$E = \frac{3Mgal^2}{2bd^3y} \quad \text{Nm}^{-2}$$

Load in Kg	Microscopereading			Elevationy For M kg
	Load increasing	Load decreasing	Mean	
x				
x+M				
x+2M				
x+3M				

### Worked examples

1. A load of 5 kg is attached to the free end of a wire of length 2 m and diameter 0.6 mm. If the extension of the wire is 0.2 mm, calculate the Young's modulus of the material of the wire. ( $g = 9.8 \text{ ms}^{-2}$ )

Length of the wire  $L = 2 \text{ m}$

Extension of the wire  $e = 0.2 \text{ mm} = 0.2 \times 10^{-3} \text{ m}$

Force acting on the wire  $F = 5 \text{ kgf} = 5 \times 9.8 \text{ N}$

Diameter of the wire  $d = 0.6 \text{ mm} = 0.6 \times 10^{-3} \text{ m}$

Radius of the wire  $r = 0.3 \times 10^{-3} \text{ m}$

Area of cross section of the wire  $a = \pi r^2 = \pi (3 \times 10^{-4})^2 \text{ m}^2$   
 Young's modulus = Linear stress / Linear strain

$$\text{Linear stress} = \frac{F}{A} = \frac{5 \times 9.8}{\pi (3 \times 10^{-4})^2} = 1.734 \times 10^8 \text{ Nm}^{-2}$$

$$\text{Linear strain} = \frac{e}{L} = \frac{2 \times 10^{-4}}{2} = 1 \times 10^{-4}$$

$$= \frac{1.734 \times 10^8}{1 \times 10^{-4}} = 1.734 \times 10^{12} \text{ Nm}^{-2}$$

2. A copper wire of 3 m length and 1 mm diameter is subjected to a tension of 5 kg wt. Calculate the elongation produced in the wire, if Young's modulus of elasticity of copper is 120 GPa.

$$\begin{aligned} \text{Young's modulus } E &= \frac{F/A}{l/L} \\ E &= \frac{FL}{Al} \\ \text{Elongation produced } l &= \frac{FL}{AE} \\ &= \frac{MgL}{\pi r^2 \times E} = \frac{5 \times 9.8 \times 3}{\pi (0.5 \times 10^{-3})^2 \times 120 \times 10^9} \\ &= 1.562 \times 10^{-3} \text{ m} \end{aligned}$$

3. What is the force required to stretch a steel wire to double its length when its area of cross-section is  $0.5 \text{ cm}^2$  and Young's modulus is  $200 \text{ GPa}$ . As the length of the wire is doubled, the change in length is equal to its original length.

$$\begin{aligned} \text{The young's modulus } E &= \frac{FL}{Al} \\ F &= \frac{EAl}{L} \\ &= \frac{EAL}{L} \end{aligned}$$

If  $L$  is the original length, then  $l = 2L$

$$\begin{aligned} F &= EA \\ &= 200 \times 10^9 \times 0.5 \times 10^{-4} \\ &= 2 \times 10^7 \text{ N} \end{aligned}$$

## VISCOSITY

Let us consider a liquid flowing over a horizontal surface. The layer in contact with the surface is at rest. The top most layer has the maximum velocity. The intermediate layers have intermediate velocity. To maintain this relative motion of the layers, an external force must be acting on the liquid. Otherwise the liquid will come to rest due to internal frictional forces acting between the layers of the liquid. These internal frictional forces that bring the liquid to rest are known as viscous forces and this property is known as viscosity.

The property by virtue of which the relative motion between the layers of a liquid is maintained is called viscosity. We can also say viscosity is the resistance to flow.

### Coefficient of viscosity of a liquid

Let  $F$  be the viscous force acting between two layers of a liquid separated by a distance  $dx$ . Let  $dv$  be the difference in velocity between the two layers. Let  $A$  be the area of the layers. The velocity gradient  $dv/dx$  acts perpendicular to the direction of flow of the liquid.

The viscous force  $F$  is found to be directly proportional to

i. the area of the layers ' $A$ '

ii. the velocity gradient  $\frac{dv}{dx}$

$$\propto \frac{A dv}{dx}$$

$$F = \frac{\eta A dv}{dx}$$

If  $A=1$  and  $\frac{dv}{dx}=1$ , then  $F=\eta$

Where  $\eta$  is constant, called Coefficient of viscosity of the liquid

Coefficient of viscosity of a liquid is defined as the viscous force acting between two layers of a liquid having unit area of layers and unit velocity gradient normal to the direction of flow of the liquid.



SI unit of  $\eta$ :

$$\eta = \frac{F}{A \frac{dv}{dx}}$$

Substituting the respective units of the quantities

on the R.H.S,

$$\text{The unit of coefficient of viscosity } \eta = \frac{\text{Nm}^2 \cdot \text{ms}^{-1}}{1/\text{m Nsm}^{-2}}$$

=

Dimensional formula of  $\eta$ :

$$\eta = \frac{F}{A \, dv/dx}$$

Substituting the respective dimensional formula of the quantities on the R.H.S,

The dimensional formula of coefficient of viscosity

$$\eta = \frac{MLT^{-2}}{L^2 \cdot LT^{-1}/L} = ML^{-1}T^{-1}$$

### **Streamline flow (laminar flow) and turbulent flow:**

To maintain the flow of a liquid through a tube, some pressure should be given at one end of the tube to overcome the viscous drag. For a given external pressure, the velocity of flow depends on the coefficient of viscosity. At low pressure the velocity is less than a certain velocity called the critical velocity and the liquid flows in an orderly manner and the flow is called steady flow or streamline flow.

In streamline flow,

1. The liquid particles flow in an orderly manner. i.e., the liquid particles flow along straight lines, each line is called the line of flow.
2. The lines of flow are parallel to the axis of the tube.
3. The velocity of all particles in a line of flow is a constant.
4. The velocity of flow is always less than the critical velocity.
5. Streamline flow will happen only when the driving pressure is small.
6. The flow of water through a capillary tube at the rate of 5 to 6 drops per minute is an example for streamline flow.

When the external pressure causing the flow of the liquid is high, the flow of the liquid is said to be turbulent.

In turbulent flow,

1. There is no orderliness in the flow of liquid.
2. The liquid particles will not flow along straight lines, but will flow in curved paths in a zig-zag manner.
3. The different particles of the liquid will flow in different directions with different velocities.
4. The velocity of flow is always greater than the critical velocity.

5. Turbulent flow will happen when the driving pressure is high.
6. The flow of water from a tap is an example of turbulent flow.

**Distinction between stream line motion and turbulent motion**  
**Critical velocity:**

When the external pressure driving the flow of liquid increases slowly, the velocity of flow of liquid will also increase gradually. At a

Sl. No.	Streamline motion	Turbulent motion
1	Flow of liquid is orderly.	Flow of liquid is zigzag and random.
2	At any point along a straight line parallel to the axis of the tube, velocity is constant	Velocity varies along the straight line parallel to the axis of the tube.
3	Velocity is proportional to the pressure	Velocity is proportional to the square root of pressure
4	Velocity will always be less than critical velocity	Velocity will always be greater than critical velocity

particular velocity the flow of liquid will change from laminar flow to turbulent flow. This velocity is called critical velocity.

The critical velocity is defined as the velocity at which the streamline flow ceases and the turbulent flow sets in.

**Reynolds number:**

Reynolds number is a pure number, which determines the type of flow of a liquid, whether a streamline flow or a turbulent flow through a cylindrical pipe. It has no units and dimensions. It is given by the expression

where  $V$  is velocity of the liquid,  $\rho$  is the density of the liquid,  $\eta$  is the coefficient of viscosity of the liquid and  $D$  is the diameter of the pipe. This

$$R = \frac{V\rho D}{\eta}$$

equation is applicable only for liquids flowing through cylindrical pipe.

Experiments show that the flow of a viscous liquid is laminar or streamlined, if Reynolds number is less than 2000. The flow of a viscous

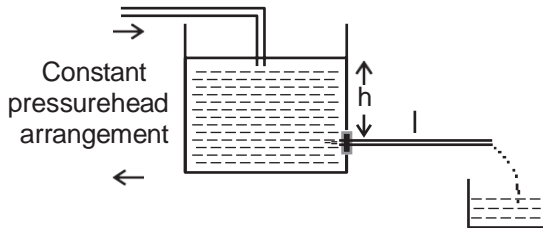


liquid is turbulent if Reynolds number is more than 2800. Between 2000 and 2800 there is a transition region in which the flow may change from streamline to turbulent.

Narrow tubes and highly viscous liquids tend to promote streamline motion, while wide tubes and liquids of low viscosity lead to turbulence.

**Poiseuille's formula:**

Consider a capillary tube of length 'l' and radius 'r' fixed horizontally



to a side of a vessel. Let a liquid be filled in the vessel. The level of the liquid is maintained constant by constant pressure head arrangement. Let 'h' be the height of the liquid and P be the pressure at the capillary tube. Let 'η' be the coefficient of viscosity of the liquid and 'ρ' be the density of the liquid.

When the liquid is in streamline flow, it is found that, the volume of liquid flowing per second 'v'

Is proportional to

- i) some power of pressure gradient (P/l)<sup>x</sup>
- ii) some power of radius of the capillary tube (r)<sup>y</sup>
- iii) some power of coefficient of viscosity of the liquid (η)<sup>z</sup>

i.e.  $v \propto (P/l)^x (r)^y (\eta)^z$

$v = K(P/l)^x (r)^y (\eta)^z$  Where ... (1)

K is a constant.

Substituting the respective dimensional formula of the quantities in the equation 1

$$L^3 T^{-1} = [ML^{-1} T^{-2} / L]^x L^y [ML^{-1} T^{-1}]^z$$

$$L^3 T^{-1} = M^{x+z} L^{-2x+y-z} T^{-2x-z}$$

Equating the powers of like terms, we get

$$x+z=0$$

$$-2x+y-z=3$$

$$-2x-z=-1$$

Solving the above equations,

$$x=1, y=4, z=-1$$

Substituting these values in equation (1)

$$v = KPr^4 / (l \eta)$$

The value of K is found to be  $\pi/8$  by experiments,

$$= \frac{\pi Pr^4 v}{8l\eta} \quad (\text{or}) \quad \eta = \frac{\pi P r^4}{8lv} = \frac{\pi h \rho g r^4}{8lv}$$

### Experiment to determine coefficient of viscosity of a low viscous liquid-Capillary flow method.

#### Aim:

To determine the coefficient of viscosity of water by capillary flow method, using graduated burette without stopper.

#### Formula:

The coefficient of viscosity of the water

$$\eta = \frac{\pi \rho g r^4 H t}{8 l v}$$

#### where

$\rho$  = density of water

$g$  = acceleration due to gravity

$r$  = radius of the capillary tube

$l$  = length of the capillary tube

$v$  = volume of water comes out of the capillary flow (5cc)  $t$  = time

taken for 5 cc flow of water

$H$  = the average pressure head height of the liquid levels from the capillary tube.

$$= \frac{h_1 + h_2}{2} = h_3$$

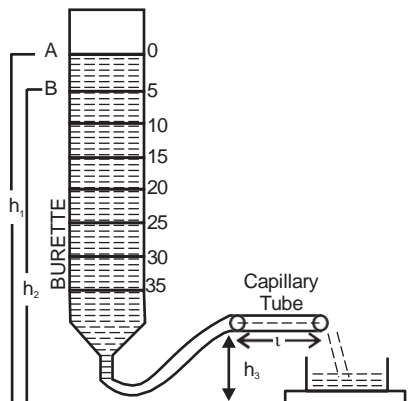
$h_1$  = height of initial level from the experimental table  $h_2$  =

height of final level from the experimental table

$h_3$  = height of the axis of capillary tube from the table

### Description:

The graduated burette without stopper is mounted vertically in the stand. A rubber tube is connected to the bottom of the burette. To the other end of the tube, a capillary tube is inserted and placed in a perfectly horizontal position.



### Procedure:

The pure water is poured into the burette using a funnel. The height of the liquid level in the burette is adjusted such that there is a streamline flow of water through the capillary tube i.e. 5 to 8 drops per minute. This can be achieved by raising or lowering the capillary tube. When the water level in the burette comes to zero mark a stop watch is started and the time for the water level to reach 5cc, 10cc, 15cc, 20cc and 25cc mark is noted and entered in the tabular column.

From the above readings, the time of flow of water ( $t$ ) 5cc, at various ranges like 0cc to 5cc, 5cc to 10cc, 10cc to 15cc, 15cc to 20cc and 20cc to 25cc are tabulated. In each case the heights of the respective initial levels and final levels from the experimental table are noted. The average pressure head height is calculated and entered in the tabular column. Substituting the average value of  $Ht$  values in the formula given, the coefficient of viscosity of water is calculated.

**Tabular column:**

Height of the axis of cap. tube from the table  $h_3 =$  m

Sl. No	Initial water level	Final Water level	Height of Initial level $h_1$	Height of final Level $h_2$	Ave. Pr. Head height $H = \frac{h_1 + h_2 - h_3}{2}$	Time Of Flow t	H.t
	cc	cc	m	m	m	s	ms

Mean

**Terminal velocity**

When a small metallic sphere is gently placed on the surface of a highly viscous liquid it descends down with an acceleration. As its velocity increases, the viscous force opposing the motion of the sphere also increases.

At some stage the apparent weight of the sphere becomes equal to the force due to viscous drag. So the resultant force on the sphere is zero. Therefore the sphere moves down with a constant velocity known as terminal velocity.

Terminal velocity is that velocity at which the apparent weight of the sphere is equal to the viscous force acting on the sphere moving in a highly viscous liquid.

**Experimental determination of the coefficient of viscosity of a highly viscous liquid by Stokes' method:**

**Aim:**

To determine the coefficient of viscosity of highly viscous liquid (Castor oil) by Stokes' method.

**Formula:**

The coefficient of viscosity of the highly viscous liquid

$$\eta = \frac{2(\rho - \sigma)g(r^2 t)}{9h}$$

where

$\rho$  = the density of the material of the solid sphere  
(glass beads)

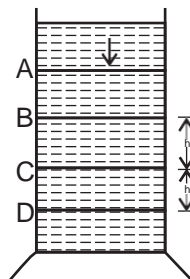
$\sigma$  = the density of the liquid (castor oil)

$g$  = the acceleration due to gravity

$r$  = the radius of the glass beads

$t$  = the time taken by the solid sphere to travel the marked distance  $h$ , inside the highly viscous liquid.

$h$  = the distance between the marks B and C

**Description:**

The pure, transparent, highly viscous liquid (castor oil) is taken in the tall glass jar. On the outer surface of the jar, two markings B and C or C and D are made at a distance of say 50 cm. The marking B should be well below the free surface of the liquid. Then only the sphere can attain terminal velocity when travelling from B to C or C to D.

**Procedure:**

First the least count, zero error and zero correction of screw gauge are determined. Then using the screw gauge, the diameters of the given glass beads are measured. The radii of the beads are calculated. Values of  $\rho$  and  $\sigma$  are taken from standard tables.

The sphere is placed gently on the surface of the highly viscous liquid and dropped. When it crosses the marking B, a stop clock is started and when it crosses the marking C, the clock is stopped. The time taken ( $t$ ) to travel by the sphere from B to C is recorded in the tabular column. The distance between B and C is noted as 'h'. The same procedure is repeated for all the given spheres. Then the coefficient of viscosity of highly viscous liquid (castor oil) is calculated, using the formula.

**Tabular column:**

To find  $r^2t$

Glass beads	Radius(r)	Radius <sup>2</sup> (r <sup>2</sup> )	Time taken(t)	r <sup>2</sup> t
Unit	mm	mm <sup>2</sup>	s	mm <sup>2</sup> s
Bead1				
Bead2				
Bead3				

$$\begin{aligned} \text{The average of } r^2t &= \text{mm}^2\text{s} \\ &= \text{ } \times 10^{-6}\text{m}^2\text{s} \end{aligned}$$

**Applications of viscosity**

To reduce friction between the two moving surfaces which are in contact, a drop or two of some oil is introduced between the two contact surfaces. The oil enters into the depressions in the contact surfaces and forms thin layers so that direct contact between the surfaces is prevented. Now, the viscous force of the oil in between the surfaces (lubricant) reduces the friction between the two moving surfaces because the fluid friction is always less than the friction between the solid surfaces. However a good viscous lubricant can only reduce friction, but cannot completely eliminate the friction.

Generally, for high speeds and low pressures, highly viscous oils such as Mobil oil, grease, etc. are used. For low speeds, oils of low viscosity are used. It may be interesting to note that air also acts as a sort of lubricant in reducing friction. The friction between two metallic surfaces in vacuum is found to be greater than that in air.

**SURFACE TENSION**

**Introduction**

The free surface of a liquid at rest behaves like a stretched elastic membrane with a tendency to contract in area. The following simple experiments will illustrate this property.

1. When a camel hair brush is dipped into water, the hairs remain spread out. When the brush is taken out, the hairs cling together on account of the films of water between them contracting.

2. A film of the soap solution is formed in a wire ring. If a wetted loop of thread is gently placed on the film, the film is unaffected and the thread remains in any form in which it is placed on the film. But if the film is pricked inside the loop, the film outside contracts on all sides and hence the circular loop is formed. The area of the film is reduced to a minimum.

The above examples show that a force acts on the surface of liquid to reduce the surface area to a minimum. This force is known as surface tension.

A water droplet takes the spherical shape because for a given volume, a sphere has the minimum surface area. This is due to surface tension.

### **Surface tension**

Surface tension is the force acting on unit length of an imaginary line drawn on the free surface of a liquid, the force acting normal to the line and parallel to the surface. The unit of surface tension is  $\text{Nm}^{-1}$  and its dimension is  $\text{MT}^{-2}$ .

### **Angle of contact**

The angle of contact is the angle by which the tangent drawn to the liquid surface at the point of contact with the solid, makes with the surface of the solid, the angle being measured within the liquid.

For liquids like water and kerosene which spread on glass, the angle of contact  $\theta = 0^\circ$ . In the case of mercury and clean glass, the angle of contact  $\theta = 137^\circ$ . For ordinary water and glass, the angle of contact  $\theta = 8^\circ$ .

When a glass rod is dipped vertically in water, which wets the glass, the water is drawn up around the glass rod at all points of contact. Here, the angle of contact is acute. When a glass rod is dipped vertically in mercury, which does not wet the glass, the liquid is depressed around the glass rod at all points of contact. Here, the angle of contact is obtuse.

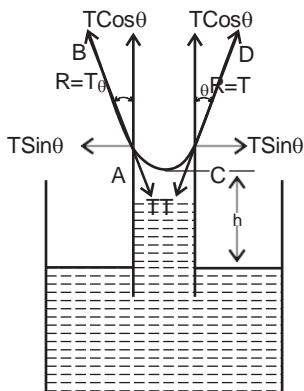
### **Explanation for capillary rise and capillary dip**

When a capillary tube is dipped vertically into a liquid, it is found that the liquid rises or falls in the tube. This phenomenon is known as capillarity and it is due to the surface tension of the liquid.

When a capillary tube is dipped vertically into a liquid like water, the liquid rises through a height in the capillary tube. This is called capillary rise.

When a capillary tube is dipped vertically in mercury, taken in a beaker, the level of mercury inside is found to be lower than the level of mercury outside the capillary tube, in the beaker. This is called 'capillary dip'.

**Expression for the surface tension by capillary rise:**



Let a capillary tube of uniform radius ( $r$ ) be dipped vertically into a liquid of density  $\rho$  taken in a beaker. Let the liquid rise through a height ( $h$ ) in the capillary tube. Let  $\theta$  be the angle of contact with glass and  $T$  be the surface tension of the liquid. The surface tension is acting tangential to the liquid surface. The reaction  $R=T$ , offered by the walls of tube acts at an angle  $\theta$  to the verticals. This reaction  $T$  along  $AB$  can be resolved vertically as  $T \cos \theta$  and horizontally as  $T \sin \theta$ . Similarly the reaction  $T$  along  $CD$  can also be resolved along the horizontal and vertical directions. The horizontal components along the circumference are equal and opposite in direction. So they get cancelled. So the total force due to surface tension is only due to vertical components.

The total upward force acting through the length of the meniscus  $2\pi r$   

$$= 2\pi r \cdot T \cos \theta$$

This force balances the weight of the liquid column of height  $h$  and radius  $r$ .

The weight of the liquid column  $= \pi r^2 h \rho g$



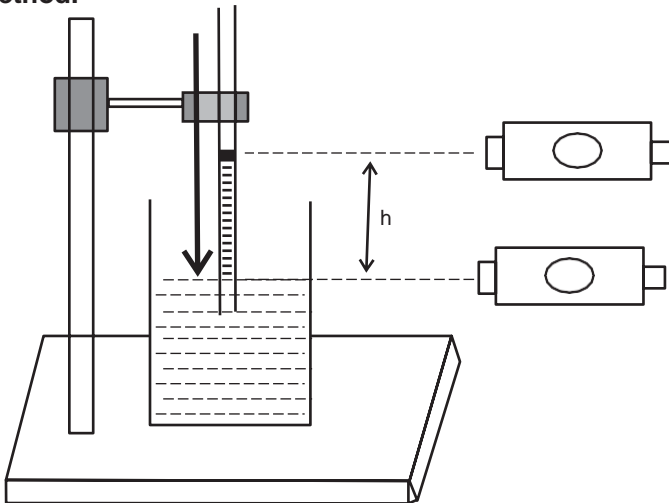
$$2\pi r T \cos\theta = \pi r^2 h \rho g$$

$$T = \frac{hr\rho g}{2\cos\theta}$$

In the case of water  $\theta = 0^\circ$  and hence  $\cos\theta = 1$ . Hence for water, Surface tension

$$T = \frac{hr\rho g}{2}$$

**Experimental determination of the surface tension of water by capillary rise method.**



A capillary tube of uniform cross sectional area is first cleaned with dilute acid and washed with distilled water and dried. It is then clamped vertically with one end inside the water taken in a beaker. Due to surface tension, water rises to a definite height in the tube. A pointer is attached to the capillary tube such that it just touches the surface of the water in the beaker.

Using a travelling microscope the reading in the vertical scale corresponding to the lower meniscus of the water in the tube is taken. The beaker is removed and the microscope is brought down and the reading corresponding to the tip of the pointer is taken. The difference between the two readings gives the height of the capillary rise of water 'h' in the

capillary tube. The diameter and hence the radius 'r' of the capillary tube is also determined with the help of microscope. If  $\rho$  is the density of water and  $g$ , the acceleration due to gravity, then the surface tension of water,

$$T = \frac{hr\rho g}{2}$$

### Applications of capillarity:

1. Lubricating oils spread easily on all parts because of their low surface tension.
  2. Cotton dresses are preferred in summer because cotton dresses have fine pores which act as capillaries for sweat.
  3. Dirt gets removed when detergents are added while washing clothes because surface tension of water is reduced.
  4. The absorption of ink by blotting paper is due to capillary action, as the blotting paper is porous. When it is placed over the ink, the ink rises into the pores. Also rise of oil in the wick of a lamp is due to capillary action.
  5. If one end of a towel is dipped into a bucket of water and the other end hangs over the bucket, the entire towel soon becomes wet due to capillary action.
  6. The supply of water to the leaves at the top of even a tall tree is through capillary rise.
7. A fabric can be made waterproof by adding suitable water proofing materials to the fabric. This addition increases the angle of contact, thereby making the fabric waterproof.

### Worked examples

1. A capillary tube of diameter 0.5 mm is dipped into a liquid vertically and the liquid rises to a height of 6 cm. If the density of the liquid is  $1000 \text{ kg m}^{-3}$ , calculate the value of surface tension of the liquid. ( $g = 9.81 \text{ ms}^{-2}$ )
  - Diameter of the capillary tube  $2r = 0.5 \text{ mm}$
  - Radius of the capillary tube  $r = 0.25 \times 10^{-3} \text{ m}$
  - Rise of liquid in the tube  $h = 6 \text{ cm} = 6 \times 10^{-2} \text{ m}$
  - Density of the liquid  $\rho = 1000 \text{ kg m}^{-3}$
  - Liquid being water, angle of contact  $\theta = 0^\circ$

$$\text{Surface tension of the liquid } T = \frac{h \rho g r \cos \theta}{2 \cos \theta}$$

$$= \frac{6 \times 10^{-2} \times 1000 \times 9.81 \times 0.25 \times 10^{-3}}{2 \cos 0}$$

$$T = 72.5 \times 10^{-3} \text{ Nm}^{-1}$$

2. Calculate the diameter of a capillary tube in which a liquid rises  $2.34 \times 10^{-2} \text{ m}$ . Surface tension of the liquid is  $25 \times 10^{-3} \text{ Nm}^{-1}$ . Relative density of the liquid is 0.79. Angle of contact =  $11^\circ 12'$ . ( $g = 9.81 \text{ ms}^{-2}$ )

Relative density of the liquid = 0.79

Density of the liquid  $\rho = 0.79 \times 10^3 \text{ kgm}^{-3}$

Diameter of the tube =  $2r = 5.41 \times 10^{-4} \text{ m}$

$$\text{We know } T = \frac{h \rho g r \cos \theta}{2 \cos \theta}$$

$$r = \frac{2T \cos \theta}{h \rho g}$$

$$= \frac{2 \times 25 \times 10^{-3} \cos 11^\circ 12'}{2.34 \times 10^{-2} \times 0.79 \times 10^3 \times 9.81} = 2.705 \times 10^{-4} \text{ m}$$

3. A liquid of density  $1000 \text{ kgm}^{-3}$  is taken in a beaker. A capillary tube of diameter 1.0 mm is dipped vertically in it. Calculate the rise of liquid in the tube. Surface tension of the liquid is  $72 \times 10^{-3} \text{ Nm}^{-1}$ .

For water, angle of contact  $\theta = 0^\circ$

Surface tension  $T = 72 \times 10^{-3} \text{ Nm}^{-1}$

Diameter of the capillary tube  $2r = 1 \text{ mm}$

Radius of the capillary tube  $r = 0.5 \times 10^{-3} \text{ m}$

$$h = 2.939 \times 10^{-2} \text{ m}$$

Rise of liquid in the tube  $h = \frac{2T \cos \theta}{r \rho g}$

$$h = \frac{2 \times 72 \times 10^{-3} \cos 0^\circ}{0.5 \times 10^{-3} \times 1000 \times 9.8} = 2.939 \times 10^{-2} \text{ m}$$

4. A capillary tube of internal diameter 0.5 mm is dipped into a liquid of density  $1140 \text{ kgm}^{-3}$ . The liquid rises to a height of 2.6 cm in the capillary tube. If the angle of contact of liquid with glass is  $55^\circ$ , calculate the surface tension of the liquid.

Diameter of the capillary tube = 0.6 mm

Radius of the capillary tube =  $0.3 \times 10^{-3}$  m

Density of the liquid,  $\rho = 1140 \text{ kg m}^{-3}$

Height through which liquid rises =  $2.6 \times 10^{-2}$  m

Angle of contact,  $\theta = 55^\circ$

$$\begin{aligned} \text{Surface tension of the liquid } T &= \frac{h\rho gr}{2\cos\theta} \\ T &= \frac{2.6 \times 10^{-2} \times 0.3 \times 10^{-3} \times 1140 \times 9.8}{2\cos 55^\circ} \\ T &= 75.96 \times 10^{-3} \text{ Nm}^{-1} \end{aligned}$$

### Questions

#### Part– A

1. Define stress.
2. Define strain.
3. Which is more elastic, rubber or steel and why?
4. What is elastic limit and plastic limit?
5. Define elastic body and plastic body.
6. What are the three moduli of elasticity?
7. Define (i) linear strain (ii) bulk strain and (iii) shearing strain.
8. Write the statement of Hooke's law.
9. Define Modulus of elasticity
10. Define (i) Young's modulus (ii) Bulk modulus and (iii) Rigidity modulus.
11. Define Poisson's ratio.
12. What is uniform bending of a beam?
13. What is nonuniform bending of a beam?
14. Define coefficient of viscosity.
15. Derive the dimensional formula and the SI unit for the coefficient of viscosity.
16. What is stream line motion?
17. What is turbulent motion?
18. What is Reynolds number?

19. What is critical velocity?
20. What is terminal velocity?
21. Write the application of viscosity.
22. Derive the S.I. unit and the dimensional formula of surface tension.
23. What is the effect of surface tension on the surface area of liquids?
24. The droplet of rain is spherical. Why?
25. How do insects run on the surface of water?
26. Define surface tension of a liquid.
27. Define angle of contact.
28. Define capillarity.
29. Write any two applications of capillarity.

### **Part-B**

1. Explain the elastic behavior of a material with the help of stress-strain curve.
2. Explain uniform and non-uniform bending of beams.
3. Describe an experiment to determine the Young's modulus of the material of a bar by bending it uniformly.
4. Distinguish between streamline and turbulent motion.
5. Derive the Poiseuille's formula for the coefficient of viscosity of a liquid by dimensional method.
6. Describe an experiment to determine the coefficient of viscosity of a low viscous liquid by capillary flow method.
7. Describe Stokes' method of determining the coefficient of viscosity of a transparent, high viscous liquid.
8. Explain the phenomenon of surface tension, capillary rise and capillary dip.
9. Derive an expression for the surface tension of a liquid in the case of the capillary rise.
10. Describe an experiment to determine the surface tension of water by capillary rise method.
11. List the various applications of capillarity.

## Exercise Problems

- The length of a wire increases from 1.25 m to 1.2508 m when a load of 12 kg is suspended. The radius of the wire is 0.5 mm. Find the stress, strain and young's modulus of material of the wire. ( $149.7 \times 10^6 \text{ Nm}^{-2}$ ,  $6.4 \times 10^{-4}$ ,  $23.4 \times 10^{10} \text{ Nm}^{-2}$ )
- What load in kilogram must be applied to a steel wire of length 1 m and diameter 0.05 cm to produce an extension of 1 cm. The Young's modulus of steel =  $2 \times 10^{11} \text{ Pa}$ . (40 kg)
- A wire 10 m long of area of cross section  $1.22 \text{ cm}^2$  elongates by 1.5 cm when 4.5 kg is suspended from it. Find the (i) stress (ii) strain and (iii) young's modulus of the wire. ( $g = 9.8 \text{ ms}^{-2}$ ) [(i)  $3.528 \times 10^5 \text{ Nm}^{-2}$  (ii) 0.0015 (iii)  $2.352 \times 10^8 \text{ Nm}^{-2}$ ]
- A body of mass 5 kg is suspended by means of a steel wire of length 1 m and diameter 1 mm. The Young's modulus of steel is  $21 \times 10^{10} \text{ Nm}^{-2}$ . Calculate the elongation of the wire. ( $l = 2.97 \times 10^{-4} \text{ m}$ )
- A uniform metal bar, one metre long, is placed symmetrically on two knife-edges separated by a distance of 64.2 cm. When two equal weights of 0.5 kg each are suspended from points 5 cm from the two ends respectively, the mid-point of the bar is elevated by 0.76 mm. Calculate the Young's modulus of the material of the bar if the width of the bar is 2.2 cm and its thickness 0.62 cm. (Ans:  $9.81 \times 10^{10} \text{ Nm}^{-2}$ )
- Calculate the surface tension of water if it rises to a height of 4.2 cm in a capillary tube dipped vertically in it. Radius of the capillary tube is  $3.5 \times 10^{-4} \text{ m}$ . ( $72.03 \times 10^{-3} \text{ Nm}^{-1}$ )
- A capillary tube of diameter 0.5 mm is dipped into a liquid vertically and the liquid rises to a height of 4.4 cm. If the density of the liquid is  $1000 \text{ kg m}^{-3}$ , calculate the value of the surface tension of the liquid.
- A capillary tube of bore 0.84 mm is dipped into a liquid of density  $840 \text{ kg m}^{-3}$  and surface tension  $0.049 \text{ Nm}^{-1}$ . Find the rise of the liquid in the tube.

## Brain teasers

1. A wire of diameter 2.5 mm is stretched by a force of 980 N. If the Young's modulus of the material of the wire is  $12.5 \times 10^{10}$  percentage  $\text{Nm}^{-2}$ . Find the increase in the length of the wire.
2. Two wires are made of the same material. The length of the first wire is half of the second wire and its diameter is double that of the second wire. If equal loads are applied on both the wires, find the ratio of increase in their lengths.
3. The diameter of a brass rod is 4 mm. Calculate the stress and strain when it is stretched by 0.25% of its length. Find the force exerted ( $E = 9.2 \times 10^{10} \text{Nm}^{-2}$  for brass)

## UNIT - III

# DYNAMICS- I

### Introduction

Dynamics is a branch of mechanics which deals with the motion of the bodies under the action of forces.

A body is said to be at rest if it does not change its place or position with respect to its surroundings. No body in this earth is at absolute rest. Because, they are sharing the motion of the earth. A body can be at rest with respect to another body.

A body is said to be in motion if it changes its place or position with respect to its surroundings. For example, a person in the moving train feels the motion, seeing the passing objects through the windows. Hence rest and motion are relative terms.

The motion of the body is characterised by two quantities, Force and Momentum. If a body is in motion with acceleration, then the body is acted upon by a force ( $F=ma$ ). If a body is in motion with uniform velocity, there is no acceleration, then the motion of the body is only under momentum ( $mv$ ).

**Newton's Laws of motion:** These laws are very important in dynamics.

**I Law:** Every body continues to be in its state of rest or of uniform motion in a straight line unless compelled by an external force.

**II Law:** The rate of change of momentum of a body is directly proportional to the force acting on it and it takes place in the direction of the force.

**III Law:** For every action, there is always an equal and opposite reaction.

In this unit we shall study about the Projectile motion, Circular motion, its application and Simple Harmonic Motion.

### PROJECTILE MOTION

In this chapter we shall consider the motion of the objects projected in any direction from the earth making the following assumptions.



- 1) The air resistance is negligible
- 2) The acceleration due to gravity remains constant.

When an object is projected vertically up from the earth, we regard the upward direction as the positive direction. The force of gravity produces acceleration downwards on the object. Acceleration is therefore taken as negative.

**Equations of motion:** Let  $u$  = initial velocity,  $v$  = final velocity,  
 $a$  = acceleration,  $t$  = time and  $s$  = displacement.

**(i) For Objects in motion**

- (1)  $v = u + at$
- (2)  $s = ut + \frac{1}{2}at^2$
- (3)  $v^2 = u^2 + 2as$

**(ii) For objects falling freely**

Here initial velocity  $u = 0$  and  $a = +g$

- (1)  $v = gt$
- (2)  $s = \frac{1}{2}gt^2$
- (3)  $v^2 = 2gs$

**(iii) For objects thrown vertically upwards**

Here  $a = -g$

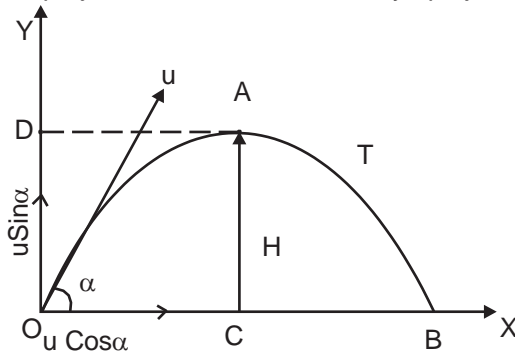
- (1)  $v = u - gt$
- (2)  $s = ut - \frac{1}{2}gt^2$
- (3)  $v^2 = u^2 - 2gs$

**Projectile**

Projectile is any object thrown obliquely into the space. The object which is given an initial velocity and afterwards follows a path determined by the gravitational force acting on it is called projectile. Its motion is called projectile motion.

A stone projected at an angle, a bomb released from an aeroplane, a shot fired from a gun, a shotput or javelin thrown by the athlete are the various examples for the projectile.

Consider a body projected from a point 'O' with velocity 'u'. The point 'o' is called point of projection and 'u' is called velocity of projection.



**Velocity of Projection (u):** The velocity with which the body projected is called velocity of projection.

**Angle of Projection ( $\alpha$ ):** The angle between the direction of projection and the horizontal plane passing through the point of projection is called angle of projection.

**Trajectory (OAB):** The path described by the projectile from the point of projection to the point where the projectile reaches the horizontal plane passing through the point of projection is called trajectory. The trajectory of the projectile is a parabola.

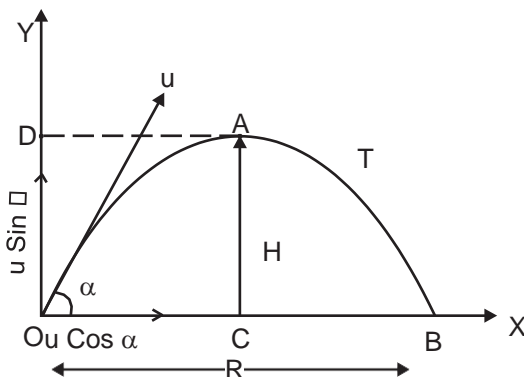
**Maximum height (H):** The maximum vertical displacement of the projectile from the horizontal plane passing through the point of projection is called maximum height.

**Time of flight (T):** The time taken by the projectile to travel from the instant of projection to the instant when it reaches the horizontal plane passing through the point of projection is called time of flight (or) The time taken by the projectile to describe the trajectory.

**Range (R):** The distance between the point of projection and the point where the projectile reaches the horizontal plane passing through the point of projection is called the range of the projectile.

**Note:** The horizontal component of velocity  $u \cos \alpha$  is constant throughout the motion and the vertical component of velocity  $u \sin \alpha$  goes on decreasing and becomes zero at the highest point A, because of the constant retardation '-g' acting on the projectile.

**Expressions for maximum height, time of flight and horizontal range of a projectile.**



Consider a body be projected upwards from the point of projection O with the velocity  $u$ . Let  $\alpha$  be the angle of projection. The velocity  $u$  can be resolved into two components. Let  $u \sin \alpha$  and  $u \cos \alpha$  are the vertical and horizontal components of the velocity.

**Maximum height reached by the projectile:**

Let  $H$  be the maximum height reached by the projectile. At the highest point the vertical component of velocity becomes zero.

Considering the vertical motion of the projectile,

Taking the equation of motion

$$v^2 = u^2 + 2as \tag{1}$$

Final velocity at A,  $v = 0$

Initial velocity at O, along vertical direction  $u = u \sin \alpha$  Acceleration  $a = -g$

Vertical displacement AC,  $s = H$

Substituting these values in equation (1) 0

$$= u^2 \sin^2 \alpha + 2(-g)H$$

$$0 = u^2 \sin^2 \alpha - 2gH$$

$$2gH = u^2 \sin^2 \alpha$$

$$\therefore \text{Maximum height, } H = \frac{u^2 \sin^2 \alpha}{2g}$$

### Time of Flight of the Projectile:

The time taken by the projectile to describe the path is called time of flight. i.e., It is the time taken by the projectile to travel from 'O' to 'B', via A.

Let  $T$  be the time of flight. Here, the vertical displacement at B is zero.

Considering the vertical motion, Taking the equation of motion S

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

Vertical displacement  $S = 0$

Initial velocity at O, along vertical direction  $u = u \sin \alpha$

Acceleration  $a = -g$ ; Time  $t = T$ ; Substituting these values in equation (2)

$$0 = u \sin \alpha T + \frac{1}{2} (-g) T^2$$

$$0 = u \sin \alpha T - \frac{1}{2} g T^2$$

$$\frac{1}{2} g T^2 = u \sin \alpha T$$

$$\therefore \text{Time of flight, } T = \frac{2u \sin \alpha}{g}$$

### Horizontal Range of the Projectile:

The horizontal distance between the point of projection and the point where the projectile meets the horizontal plane passing through the point of projection is called range of the projectile.

Let  $R$  be the range of the projectile.

Range  $R = \text{Horizontal component of velocity} \times \text{Time of flight}$

$$R = u \cos \alpha \times \frac{2u \sin \alpha}{g}$$

$$= \frac{u^2 \sin \alpha \cos \alpha}{g} \quad \therefore \sin 2\alpha = 2 \sin \alpha \cos \alpha$$

$$\therefore \text{Range, } R = \frac{u^2 \sin 2\alpha}{g}$$

**Condition for maximum range ( $R_{\max}$ ):**

$$\text{The expression for Horizontal Range, } R = \frac{u^2 \sin 2\alpha}{g}$$

For a given velocity of projection, range will be maximum only when  $\sin 2\alpha$  is maximum. The maximum value for sine function is 1.

$$\text{i.e. } \sin 2\alpha = 1 \text{ but, } \sin 90^\circ = 1$$

$$\therefore \sin 2\alpha = \sin 90^\circ \therefore 2\alpha = 90^\circ \quad \therefore \alpha = 45^\circ$$

For getting maximum range, the condition is the angle of projection should be  $45^\circ$ .

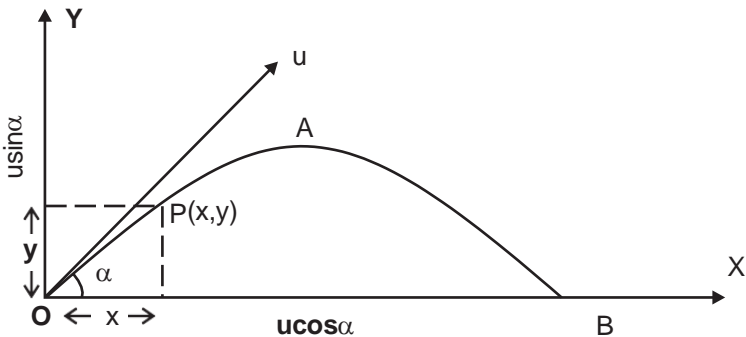
$$\therefore \text{Maximum range, } R_{\max} = \frac{u^2 \sin(2 \times 45^\circ)}{g}$$

$$R_{\max} = \frac{u^2 \sin 90^\circ}{g} = \frac{u^2}{g} \quad (\text{Q } \sin 90^\circ = 1)$$

$$\therefore R_{\max} = \frac{u^2}{g}$$

This is the expression for the maximum range of the projectile.

**Path of the projectile (the trajectory) is a Parabola:**



Consider a body be projected from the point of projection O with the velocity  $u$  as shown in the figure. Let  $\alpha$  be the angle of projection of the body. The velocity  $u$  can be resolved into two components.  $u \sin \alpha$  and  $u \cos \alpha$  are the vertical and horizontal components of velocity  $u$ . Let P(x,y)

Horizontal displacement in  $t$  second,  $x = u \cos \alpha t$

$$t = \frac{x}{u \cos \alpha}$$

Taking the equation of motion  $S = ut + \frac{1}{2}at^2$  (1)

be the position of the body after  $t$  second.

Vertical displacement in  $t$  second,  $S = y$

Time taken to reach P,  $t = \frac{x}{u \cos \alpha}$

Initial velocity at O along vertical direction,  $u = u \sin \alpha$  Acc

eleration  $a = -g$

$$y = u \sin \alpha \frac{x}{u \cos \alpha} + \frac{1}{2}(-g) \frac{x^2}{u^2 \cos^2 \alpha}$$

$$y = x \tan \alpha - \frac{g}{2u^2 \cos^2 \alpha} x^2$$

$$y = x \tan \alpha - \frac{g x^2}{2u^2 \cos^2 \alpha}$$

Substituting these values in equation (1)

The above equation is similar to the equation of the parabola

$y = bx - cx^2$ . Hence the path of the projectile is a parabola.

### SOLVED PROBLEMS:

1. A body is thrown with a velocity of projection  $49 \text{ ms}^{-1}$  at an angle of projection  $45^\circ$ . Find (i) Maximum height (ii) Time of flight (iii) Range ( $g = 9.8 \text{ ms}^{-2}$ ).

Given

$$u=49\text{ms}^{-1}$$

$$\alpha=45^\circ$$

$$\begin{aligned} \text{Maximum height, } H &= \frac{u^2 \sin^2 \alpha}{2g} \\ &= \frac{49 \times 49 \times \sin 45^\circ \times \sin 45^\circ}{2 \times 9.8} \\ &= 61.25\text{m} \end{aligned}$$

$$\begin{aligned} \text{Time of flight, } T &= \frac{2u \sin \alpha}{g} \\ &= \frac{2 \times 49 \times \sin 45^\circ}{9.8} \\ &= 7.07\text{s} \end{aligned}$$

$$\begin{aligned} \text{Range, } R &= \frac{u^2 \sin 2\alpha}{g} \\ &= \frac{49 \times 49 \times \sin (2 \times 45^\circ)}{9.8} \\ &= \frac{49 \times 49 \times \sin 90^\circ}{9.8} \\ &= 245\text{m} \end{aligned}$$

2. A bullet fired from a gun with a velocity of  $80\text{ms}^{-1}$  strikes the ground at the same level as the gun at a distance of 460 m. Find the angle of inclination with the horizontal at which the gun was fired.

Given

$$u=80\text{ms}^{-1}, R=460\text{m}, g=9.8\text{ms}^{-2}$$

$$\text{Range } R = \frac{u^2 \sin 2\alpha}{g}$$

$$\sin 2\alpha = \frac{g \cdot R}{u^2} = \frac{9.8 \times 460}{(80)^2} = 0.7044$$

$$2\alpha = \sin^{-1}(0.7044) = 44^\circ 46' \quad \therefore \alpha = 22^\circ 23'$$

3. The range of a projectile is equal to double the maximum height attained. Find the angle of projection.

Given

Horizontal range = 2 × maximum height

ie.,  $R = 2H$

$$\frac{u^2 \sin 2\alpha}{g} = 2 \frac{u^2 \sin^2 \alpha}{2g}$$

$$\sin 2\alpha = \sin^2 \alpha$$

$$2 \sin \alpha \cos \alpha = \sin^2 \alpha$$

$$2 \cos \alpha = \sin \alpha$$

$$\therefore \tan \alpha = 2$$

$$\therefore \alpha = \tan^{-1}(2) = 63^\circ 26'$$

4. A projectile is thrown at an angle  $\theta$  and another is thrown at  $(90 - \theta)$  from the same point both with the velocities  $78.4 \text{ ms}^{-1}$ . These two reach 36.4 m higher than the first. Find the individual heights.

Given

Angle of projection of first projectile =  $\theta$

Angle of projection of second projectile =  $(90 - \theta)$

Velocity of projection  $u = 78.4 \text{ ms}^{-1}$  and

$$H_2 = H_1 + 36.4 \text{ m}$$

$$H_2 - H_1 = 36.4 \text{ m}$$

Where  $H_1$  is the maximum height of first stone and  $H_2$  is the maximum height of second one.

$$H_1 = \frac{u^2 \sin^2 \theta}{2g} \quad \text{and} \quad H_2 = \frac{u^2 \sin^2 (90 - \theta)}{2g} = \frac{u^2 \cos^2 \theta}{2g}$$

$$\begin{aligned} \text{Adding,} \quad H_2 - H_1 &= \frac{u^2 \cos^2 \theta}{2g} - \frac{u^2 \sin^2 \theta}{2g} \\ &= \frac{u^2}{2g} = \frac{78.4 \times 78.4}{2 \times 9.8} = 313.6 \end{aligned}$$

$$2H_2 = 36.4 + 313.6 = 350$$

$$H_2 = 175 \text{ m}$$

$$\text{Also, } H_1 = H_2 - 36.4 = 175 - 36.4$$

$$H_1 = 138.6 \text{ m}$$



5. If the range of a projectile is  $256\sqrt{3}$  m and the maximum height reached is 64 m. Calculate the angle of projection.

$$\text{Given } H = 64 \text{ m } \quad R = 256\sqrt{3} \text{ m}$$

$$\text{Maximum height } H = \frac{u^2 \sin^2 \alpha}{2g}$$

$$64 = \frac{u^2 \sin^2 \alpha}{2g} \tag{1}$$

$$\text{Range, } R = \frac{u^2 2 \sin \alpha \cos \alpha}{g}$$

$$256\sqrt{3} = \frac{u^2 2 \sin \alpha \cos \alpha}{g} \tag{2}$$

Dividing eqn.(1) by (2)

$$\frac{64}{256\sqrt{3}} = \frac{u^2 \sin^2 \alpha}{2g} \times \frac{g}{u^2 2 \sin \alpha \cos \alpha} = \frac{\tan \alpha}{4}$$

$$\tan \alpha = \frac{1}{\sqrt{3}} \quad \alpha = \tan^{-1} \left( \frac{1}{\sqrt{3}} \right)$$

$$\alpha = 30^\circ$$

### 3.2 CIRCULAR MOTION

#### Introduction

In this chapter we study the motion of the bodies moving in a circular path. The motion of the body in a circular path is called circular motion. The motion of the body in the circular path with uniform velocity is called uniform circular motion. When it is in uniform velocity, the direction of motion at any point is along the tangent to the circle at that point. If the body is set free, it would move in the direction of the tangent at that point.

#### Angular velocity

Consider a body (or particle) moving along a circular path. If  $\theta$  is the angle turned by the radius vector in  $t$  second, then

$$\text{Angular velocity } \omega = \frac{\theta}{t}$$

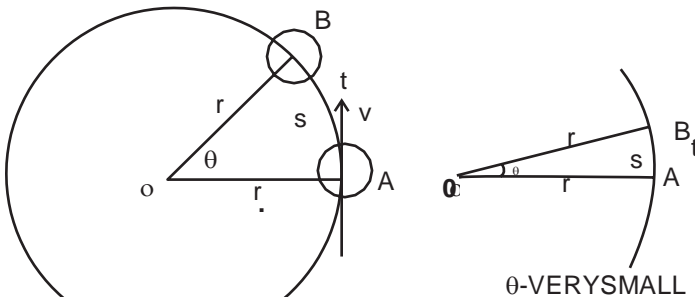
The angle covered or swept or turned by the radius vector in one second is known as angular velocity.

SI Unit of angular velocity is radian/second ( $\text{rads}^{-1}$ )

Radian is the angle made at the centre of the circle by an arc whose length is equal to the radius of the circle.

### Relation between angular velocity and linear velocity

Consider a particle moving along a circular path of radius  $r$ . Initially the particle is at a point  $A$ . After  $t$  seconds the particle reaches the point  $B$ . Let  $\theta$  be the angle swept by the radius vector in  $t$  seconds.



Angular velocity  $\omega = \frac{\theta}{t}$

If  $\theta$  is very small, the distance  $AB$  will be considered as a straight line. Let  $v$  be the linear velocity of the particle.

Linear velocity,  $v = \frac{s}{t}$

In  $\Delta OAB$ ,  $\sin \theta = \frac{s}{r}$

If  $\theta$  is very small,  $\sin \theta = \theta$  in radian

$\therefore \theta = \frac{s}{r}$  (or)  $s = r\theta$

Linear velocity,  $v = \frac{r\theta}{t} = r\omega$

$$\left( \begin{matrix} Q \\ \omega = \frac{\theta}{t} \end{matrix} \right)$$

$\therefore v = r\omega$

Linear velocity = Radius of the circle  $\times$  Angular velocity

**Note:**

If  $\theta = 1^\circ$ ,  $1^\circ = 0.0175$  radian; Also  $\sin 1^\circ = 0.0175$   
 Then  $\sin 0.0175$  (radian)  $= 0.0175$  (natural value)  
 $\theta = 2^\circ = 0.0349$  radian; Also  $\sin 2^\circ = 0.0349$   
 Then  $\sin 0.0349$  (radian)  $= 0.0349$  (natural value)  
 Hence  $\sin \theta = \theta$  for very small values of  $\theta$ .  
 It is practically valid upto  $\theta = 4^\circ = 0.0698$  radian.

**Period and frequency of revolution**

**Period (T):** Period of revolution of a particle is the time taken by the particle to complete one revolution.

**Frequency (n):** Frequency of revolution of a particle is the number of revolutions made in one second.

**Relation between period T and angular velocity  $\omega$ :**

For one complete revolution (in T second),  
 Angle turned by the radius vector in T second  $= 2\pi$

$\therefore$  Angle turned by the radius vector in one second  $= \frac{2\pi}{T}$

i.e., Angular velocity,  $\omega = \frac{2\pi}{T}$  or  $T = \frac{2\pi}{\omega}$

**Relation between period T and frequency n:**

Let the particle complete n rotations in one second.

This is called frequency. So, the time taken complete one rotation which gives period  $= 1/n$ . Therefore,  $T = 1/n$ .

**Relation between frequency n and angular velocity  $\omega$ :**

No. of revolutions in 1 second  $= n$

Angle turned by the radius vector for one revolution  $= 2\pi$

$\therefore$  Angle covered by the radius vector for n revolutions in 1 second  $= 2\pi n$

i.e., Angular velocity,  $\omega = 2\pi n$

**Normal acceleration:**

For a particle in uniform circular motion, the direction of its velocity changes continuously but the magnitude remains constant. For this, a force must act along the radius, towards the centre of the circle, to maintain the circular path. Hence, there is an acceleration acting on the body towards the centre of the circle.

Thus, for a particle executing uniform circular motion, there must always be acceleration directed towards the centre of the circular path along the radius. This acceleration is called normal acceleration and the force responsible for this acceleration is known as centripetal force.

When a particle is moving along a circular path, the particle is accelerated towards the centre of the circular path in a direction normal to the direction of its motion. This acceleration of the particle is known as normal acceleration.

**Centripetal force and Centrifugal force:**

When a particle is moving along a circular path, it is accelerated towards the centre of the circular path. Due to this acceleration there must be force acting on the particle towards the centre and this force is known as centripetal force.

Centripetal force is defined as the force which acts towards the centre along the radius of the circular path on which the particle is moving with uniform velocity.

Let a stone be tied to one end of a string and whirled in a horizontal plane with constant speed. The stone is subjected to a force directed towards the centre of the circle. This force acting towards the centre of the circle is called centripetal force.

The expression for centripetal force is given by

$$\frac{mv^2}{r} \text{ or } mr\omega^2$$

By Newton's third law, there must be a reaction to this force. This reactive force which is directed away from the centre is called centrifugal force.

The centrifugal force is defined as the force acting along the radius of the circle, away from the centre, to balance the centripetal force.

The centripetal force and centrifugal force are equal in magnitude but opposite in direction. As centrifugal force is a reaction force, it ceases once the centripetal force stops.

**Examples for centripetal force:**

The centripetal force is responsible for circular motion. Without this force the body cannot move along a circular path. This force can be produced in different ways.

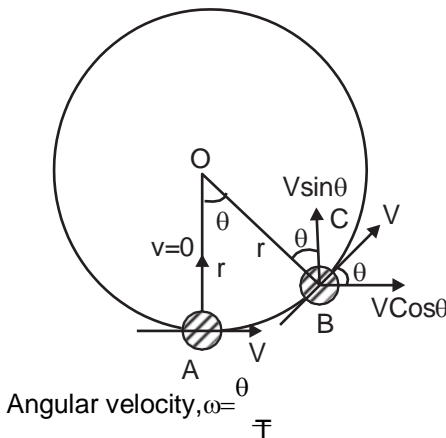
1. A body tied to one end of a string is being rotated in a circle; the centripetal force is supplied by the tension of the string.
2. The moon revolving around the earth, the centripetal force is the gravitational pull of the earth on the moon.
3. A person riding a cycle along a circular path, the necessary centripetal force is supplied by a sidewise pressure exerted by the road on the tyres.

**Examples for centrifugal force:**

1. A stone tied to one end of a string is being rotated in a circle, the stone itself exerts an equal and opposite force on the hand. It is due to the centrifugal force.
2. A centrifugal pump is used to pump well water on the basis of centrifugal force.

**Expressions for Normal acceleration and Centripetal force**

Consider a particle of mass 'm' moving along a circular path of radius  $r$ . Let  $v$  be the velocity of the particle. Initially the particle is at a point A. After  $t$  second the particle reaches the point B. Let  $\theta$  be the angle covered by the radius vector. Let B be very close to A, so that the angle covered  $\theta$  is small.



Initially,  
 The velocity component in AO direction = zero  
 After  $t$  second

The velocity component in BC direction =  $v \sin \theta$

Since  $\theta$  is very small, the direction BC is very close to AO direction.

$\therefore$  The change in velocity towards the center of the circle in

't' second =  $v \sin \theta - 0$

$$= v \sin \theta$$

As  $\theta$  is very small,  $\sin \theta = \theta$  (in radian)

The change in velocity in one second =  $\frac{v \sin \theta}{t}$

ie., Normal acceleration,  $a = \frac{v \sin \theta}{t}$

$\therefore$  Normal acceleration,  $a = \frac{v \theta}{t}$

Since  $\omega = \frac{\theta}{t}$ ,  $a = v \omega$

since  $v = r \omega$ ,  $a = r \omega \cdot \omega = r \omega^2$

since  $\omega = \frac{v}{r}$ ,  $a = v \cdot \frac{v}{r} = \frac{v^2}{r}$

Normal acceleration,  $a = \frac{v^2}{r}$  or  $a = r \omega^2$

### Centripetal force

According to Newton's II Law of motion,  $F = ma$ . As 'm' is the mass of the particle moving along a circular path of radius 'r',  
Centripetal force = mass X normal acceleration.

Normal acceleration,  $a = \frac{v^2}{r}$  or  $a = r \omega^2$

Centripetal force,  $F = \frac{mv^2}{r}$  or  $F = mr \omega^2$

### SOLVED PROBLEMS

(1) A ball weighing 0.5 kg is tied to one end of a string and is whirled at a constant speed of  $10 \text{ ms}^{-1}$  in a horizontal plane. If the length of the string is 1 m, calculate the normal acceleration and the tension in the string.

Given  $m = 0.5 \text{ kg}$ ,  $v = 10 \text{ ms}^{-1}$   $\ell = r = 1 \text{ m}$   $a = ?$   $F = ?$

$$\text{Normal acceleration} = \frac{v^2}{r} = \frac{(10)^2}{1} = 100 \text{ ms}^{-2}$$

Tension in the string (Centripetal force)

$$F = \frac{mv^2}{r} = 0.5 \times (10)^2 = 50 \text{ N}$$

(2) A body of mass 1 kg is tied to a string of length 1m revolves in a horizontal circle. If the angular velocity is  $3 \text{ rad s}^{-1}$ , calculate the centripetal force acting on the body.

Given,  $m=1\text{kg}$ ,  $\ell=r=1\text{m}$ ,  $\omega=3\text{rads}^{-1}$

$$\text{Centripetal force } F = m r \omega^2 = 1 \times 1 \times (3)^2 = 9 \text{ N}$$

(3) Find the centripetal force on a body of mass 500g when it revolves in a circle of radius 1.5m. The body makes 120 R.P.M.

Given  $m=500\text{g}=0.5\text{kg}$ ,  $r=1.5\text{m}$

No. of revolutions = 120 per minute

$$\text{i.e., } n = \frac{120}{60} = 2 \text{ revolutions per second}$$

Centripetal force  $F = m r \omega^2$

$$= m r (2\pi n)^2 \quad (\text{since } \omega = 2\pi n) \quad = 4\pi^2 m r n^2$$

$$= 4 \times 3.14 \times 3.14 \times 0.5 \times 1.5 \times 2^2 \quad F = 118.4 \text{ N}$$

(4). A mass of 5kg is moving in a circle of radius 1m. If the centripetal force acting on the mass is 20N, find the linear velocity and angular velocity.

Given:  $F=20\text{N}$ ,  $r=1\text{m}$ ,  $m=5\text{kg}$

$$F = \frac{mv^2}{r}$$

$$v^2 = \frac{Fr}{m} = \frac{20 \times 1}{5} = 4 \text{ m}^2 \text{ s}^{-2}$$

$$v = 2 \text{ ms}^{-1}$$

$$\text{Angular velocity, } \omega = \frac{v}{r} = \frac{2}{1} = 2 \text{ rad s}^{-1}$$

(5) A rope of length 1m can withstand a maximum weight of 10kg. Now a stone of mass 200g is tied to it and is whirled round in a horizontal circle.

Calculate the maximum permissible speed of the stone.

Given:  $r=1\text{m}$ ,  $m=200\text{g}=0.2\text{kg}$  and  $F=10 \times 9.8 \text{ N}$

$$F = \frac{mv^2}{r}$$

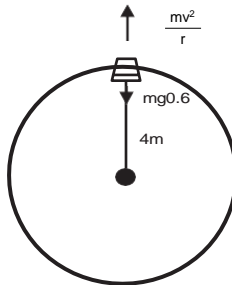
$$v^2 = \frac{Fr}{m} \therefore v = \sqrt{\frac{Fr}{m}} = \sqrt{\frac{10 \times 9.8 \times 1}{0.2}}$$

$$v = 7.07 \text{ ms}^{-1}$$

(6) One end of a string of length 0.64 m is attached to a bucket containing water and the bucket is rotated about the other end in a vertical circle. Find the minimum speed with which it can be rotated without spilling the water at the highest point. How many revolutions per minute is it making?

Given: Length of the string (or) radius of the circle = 0.64 m

For without spilling the water at the highest point



$$mg = \frac{mv^2}{r} \quad (\text{or}) \quad g = \frac{v^2}{r}$$

$$\therefore v^2 = rg = 0.64 \times 9.8 = 6.272$$

$$\therefore \text{The minimum speed } v = \sqrt{6.272} = 2.504 \text{ ms}^{-1}$$

Let  $n$  be the no. of revolutions per second.

$$n = \frac{\omega}{2\pi} = \frac{v}{r \cdot 2\pi} = \frac{2.504}{0.64 \times 2\pi} = 0.6231$$

No of revolutions per minute =  $0.6231 \times 60$

= 37.4 revolutions



### 3.3 APPLICATION OF CIRCULAR MOTION

#### **Banking of curved paths:**

Curved track is to be considered as a part of the circular path. Consider a motor car moving along a curved road. Now two forces are acting on the car, one is the weight of the car acting downward and second is the normal reaction acting upward. In addition to these two forces, a horizontal centripetal force should act on the car for safe negotiation of curved motion. This force is supplied by the pressure exerted by the road on the tyres of the car towards the inner side.

According to Newton's third law of motion, the tyres exert an equal and opposite pressure on the road towards the outside. Hence there must be a frictional force between the road and the tyres. This frictional force tries to balance the centrifugal force. If the road is in horizontal level, the friction may not be sufficient to balance the centrifugal force. Then the car may skid. To avoid this the outer edge of the road is raised a little above the level of the inner edge. This formation is called as banking of curved paths.

In the case of train moving along a curved track, the necessary centrifugal force is supplied by the pressure exerted by the rails on the flanges of the wheels. Also the flanges of the wheels exert an equal and opposite pressure on the rails. Hence there must be a large amount of friction and it leads to damage the rails and also to derailment of the train.

To avoid this damage and derailment, the outer rail is suitably raised over the inner rail to eliminate the flange pressure on the rails. This formation is called as banking of curved railway tracks.

The raising of outer edge of a curved path than the inner edge for providing the necessary centripetal force is known as banking of curved path.

#### **Angle of Banking**

It is the angle between the normal horizontal path and the curved banked path.

The angle of banking of a curved path is defined as the angle through which the outer edge of the road is raised over the inner edge.

The angle of banking of curved railway tracks is defined as the angle through which the outer rail is raised over the inner rail.

### Expression for Angle of Banking of Railway tracks

Consider a train moving along a curved path of radius  $r$ . In figure, ABCD shows the vertical cross-section of a railway carriage. If the outer rail is raised over the inner rail by an angle  $\theta$ , the carriage is inclined to the horizontal by the same angle. This angle is called angle of banking of the railway tracks.

Let  $R_1$  and  $R_2$  be the reactions of the inner and outer rails.

The total vertical component of  $R_1$  and  $R_2$  is

$$= R_1 \cos \theta + R_2 \cos \theta$$

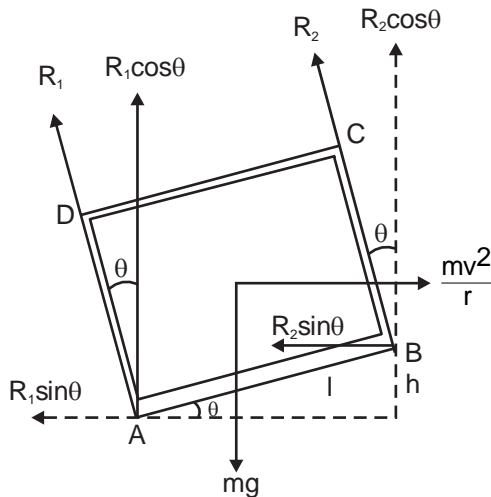
$$= (R_1 + R_2) \cos \theta$$

The total horizontal component of  $R_1$  and  $R_2$  is

$$= R_1 \sin \theta + R_2 \sin \theta$$

$$= (R_1 + R_2) \sin \theta$$

The vertical component balances the downward weight of the carriage. The horizontal component supplies the necessary centripetal force and it balances the centrifugal force. Let  $m$  be the mass of the carriage and  $v$  be the speed of the train.



The downward weight =  $mg$

The centrifugal force =  $\frac{mv^2}{r}$

Equating the forces  $(R_1 + R_2) \cos \theta = mg$  (1)

$(R_1 + R_2) \sin \theta = \frac{mv^2}{r}$  (2)

$(2) \frac{(R_1 + R_2) \sin \theta}{(R_1 + R_2) \cos \theta} = \frac{mv^2}{rmg}$

(1)  $\text{ie., } \tan \theta = \frac{v^2}{rg}$

∴ The angle of banking of railway track,

Let 'l' be distance (AB) between the rails.

$$\theta = \tan^{-1} \left( \frac{v^2}{rg} \right)$$

If 'θ' small,  $\sin \theta = \tan \theta$

$$\therefore \frac{h}{l} = \frac{v^2}{rg} \quad h = \frac{v^2 l}{rg}$$

While constructing curved railway tracks, the value of h is obtained for fixed values of v, l and r.

Therefore, on a curved rail track, every train has to move with a particular speed specified for the track.

**SOLVED PROBLEMS**

(1) A cyclist is riding at a speed of  $22 \text{ ms}^{-1}$  takes a turn on a curved circular road of radius 44m. What is his inclination to the vertical?

Given,  $V = 22 \text{ ms}^{-1}$ ,  $r = 44 \text{ m}$ ,  $g = 9.8 \text{ ms}^{-2}$

$$\begin{aligned} \tan \theta &= \frac{v^2}{rg} \\ &= \frac{(22)^2}{44 \times 9.8} = \frac{484}{431.2} \end{aligned}$$

$\tan \theta = 1.122$

$\theta = \tan^{-1}(1.122) = 48^\circ 18'$

(2) An electric train has to travel on a railway track with a curve of radius 120m with a speed of 36 kmph. Calculate the angle of banking of the rails.

$$\begin{aligned} \text{Given } v &= 36 \text{ kmph} = 36 \times \frac{1000}{3600} = 10 \text{ ms}^{-1} \\ r &= 120 \text{ m} \quad g = 9.8 \text{ ms}^{-2} \end{aligned}$$

$$\tan \theta = \frac{v^2}{g r} = \frac{(10)^2}{120 \times 9.8}$$

$$\tan \theta = 0.0850$$

$$\theta = \tan^{-1}(0.085) = 4^\circ 51'$$

(3) A bend in level road has a radius of 18m. Find the maximum speed with which a car negotiates the bend without skidding. The elevation of the road is  $7^\circ$ .

$$\text{Given, } R = 18 \text{ m, } \theta = 7^\circ \text{ and } g = 9.8 \text{ ms}^{-2}$$

$$\tan \theta = \frac{v^2}{rg}$$

$$\begin{aligned} v^2 &= rg \times \tan \theta = 18 \times 9.8 \times \tan 7^\circ \\ &= 18 \times 9.8 \times 0.1222 = 21.65 \text{ v} \\ &= 4.65 \text{ ms}^{-1} \end{aligned}$$

### 3.4 SIMPLE HARMONIC MOTION

Any motion in nature is classified into two types, viz., periodic and non-periodic motion. Periodic motion is that motion which repeats itself after equal intervals of time. The examples of periodic motion are the motion of planets around the sun, motion of hands of a clock, motion of the balance wheel of a watch.

If a body moves back and forth repeatedly about a mean position, it is said to possess oscillatory motion. Vibrations of guitar strings, motion of a pendulum bob, vibrations of a tuning fork, oscillations of mass suspended in a spring, vibrations of diaphragm in telephones are few

examples of oscillatory motion. In all the cases of vibrations of bodies, the path of vibration is always directed towards the mean or equilibrium position.

The oscillations can be expressed in terms of simple harmonic functions like sine or cosine function. A harmonic oscillation of constant amplitude and single frequency is called simple harmonic motion (SHM).

A particle is said to execute simple harmonic motion if its acceleration is directly proportional to the displacement from a fixed point and is always directed towards that point.

### **Important definitions of SHM**

#### **Displacement**

The distance travelled by the oscillating particle from its mean position at any instant is called displacement.

#### **Amplitude**

The maximum displacement of the oscillating particle from its mean position is called amplitude.

#### **Oscillation**

One complete to and from motion is called one oscillation.

#### **Period**

Time taken by the oscillating particle to complete one oscillation is called period.

#### **Frequency**

The number of oscillations or vibrations made by the vibrating particle in one second is called frequency.

#### **Phase**

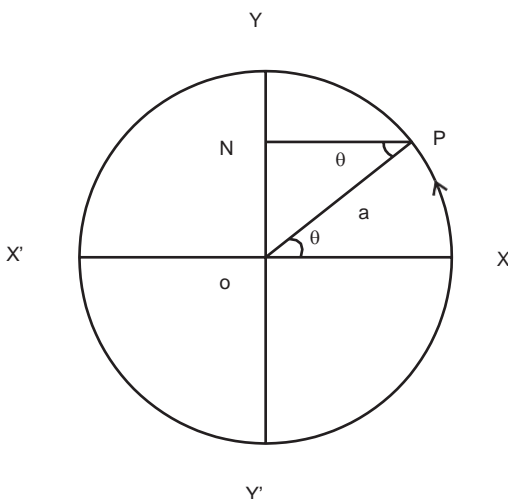
Phase of a particle vibrating in SHM at any instant is the state of vibrating particle regarding its position and direction of vibration at that particular instant.

### **The Projection of uniform circular motion on a diameter is SHM**

Consider a particle moving along the circumference of a circle of radius ' $a$ ' and centre  $O$ , with uniform speed ' $v$ ', in anticlockwise direction as shown in the figure. Let  $XX'$  and  $YY'$  be the two perpendicular diameters.

Suppose the particle is at P after a time  $t$  second. If  $\omega$  is the angular velocity, then the angular displacement in time  $t$  is given by  $\theta = \omega t$ . From P draw PN perpendicular to YY'. As the particle moves from X to Y, foot of the perpendicular N moves from O to Y. As it moves further from Y to X', then from X' to Y' and back again to X, the point N moves from Y to O, from O to Y' and back again to O. When particle completes one revolution along the circumference, the foot of the perpendicular point N completes one to and fro about the mean position O. This motion of the point N along the diameter YY' is simple harmonic.

Hence, the projection of a uniform circular motion on a diameter of a circle is simple harmonic motion.



## QUESTIONS

### PART-A

1. What is a projectile?
2. Define velocity of projection.
3. Define angle of projection.
4. Define trajectory of a projectile.
5. Define maximum height of a projectile.
6. Define time of flight of a projectile.
7. Define range of a projectile.

8. Give the condition for maximum range of a projectile.
9. What is a circular motion?
10. Define a radian.
11. Define angular velocity.
12. Define normal acceleration.
13. Define centripetal force.
14. What is centrifugal force?
15. Define angle of banking.
16. Why is the outer rail raised over the inner rail along a curved railway track?
17. Define Simple Harmonic Motion
18. Define amplitude
19. Define frequency
20. Define period
21. Define phase

### **PART B**

1. Derive an expression for the maximum height reached by the projectile.
2. Derive an expression for the time of flight of the projectile.
3. Derive an expression for the horizontal range of the projectile.
4. Derive the condition for the maximum range of the projectile (OR) Show that the range is maximum when the angle of projection is  $45^\circ$  for a given velocity of projection.
5. Prove that the path of a projectile is a parabola.
6. Derive the relation between angular velocity and linear velocity.
7. Derive the relations between angular velocity, period and frequency.
8. Obtain expressions for the normal acceleration and centripetal force of a body executing uniform circular motion.
9. Explain the banking of curved tracks.
10. Derive an expression for the angle of banking.
11. Show that the projection of uniform circular motion on the diameter of a circle is SHM.

### Exercise problems

1. A particle is thrown at an angle of  $30^\circ$  with a velocity of 36 kmph. Calculate its (1) time of flight and (2) horizontal range.  
[Ans: (i) 1.02s, (ii) 8.84m]
2. The range of a projectile is equal to 3 times the maximum height attained. Find the angle of projection.  
[Ans:  $53^\circ 7'$ ]
3. The maximum height of a projectile is 2 times its horizontal range reached. Find the angle of projection.  
[Ans:  $79^\circ 58'$ ]
4. A bomber flying horizontally with a constant velocity of 600 kmph at a height of 480m wants to hit a target on the ground. Find at what horizontal distance from the target should a bomb be dropped in order to hit the target.  
[Ans:  $R=1648\text{m}$ ]
5. A body of mass 2 kg tied to a string of length 2m revolves in a horizontal circle. If the angular velocity of the body is  $5 \text{ rad s}^{-1}$ , calculate the centripetal force acting on the body.  
[Ans:  $F=100\text{N}$ ]
6. In a circus, a man is riding a motor cycle in the form of a vertical circle inside the hollow spherical cave of diameter 6m. Find the minimum speed of the rider with which he can be riding without falling at the highest point.  
[Ans:  $v=7.668\text{ms}^{-1}$  or 27.6 kmph]
7. A string of length 1.5m can withstand a maximum weight of 5kg wt. A body of mass 2kg is tied to it and whirled round in a horizontal plane. Calculate the possible maximum number of revolutions per minute.  
[Ans:  $\text{RPM}=38.6$ ]
8. A scooter rider negotiates a curve of 100m radius on a level road with a speed of 72 kmph. Calculate the angle of inclination to the vertical he should make to avoid falling.  
[Ans:  $\theta=22^\circ 12'$ ]



9. An electric train has to travel a curved railway line of radius 50m with maximum speed of 36kmph. Calculate the angle through which the outer rail should be raised so that there is no lateral thrust on the rails.  
[Ans:=  $11^{\circ}31'$ ]
10. An aeroplane with a speed of  $120 \text{ ms}^{-1}$  negotiates a curve of radius 1200m. Calculate its angle of banking.  
[Ans:=  $50^{\circ}45'$ ]
11. The expected average speed of the traffic on a highway around the curve is 20kmph. Calculate the angle of banking of the highway if its radius of curvature be 12m.  
[Ans:=  $14^{\circ}42'$ ]

### **Brainteaser**

1. An electric train has to travel on a railway track with a bend of radius 120m with a speed of 45 kmph. Calculate the height through which the outer rail should be raised for safe running of the train, if the distance between the two rails is 1.6m.

## UNIT-IV

# DYNAMICS-II

### ROTATIONAL MOTION OF RIGID BODIES

#### Introduction

According to Newton's first law of motion, "Every body continues to be in a state of rest or in a state of uniform motion along a straight line, unless it is compelled to change that state by some external force".

This law consists of two parts. i.e., (1) A body at rest will not move on its own unless an external force acts on it and (2) A body which is in uniform linear motion will not change either its speed or direction of motion on its own, without the help of an external force.

The inability of a body to change its state on its own, without the help of an external force is termed as inertia. Inertia is a fundamental property of matter. The more is the mass of the body, the more will be inertia. Thus in translatory (linear) motion the mass of the body is a measure of coefficient of inertia.

#### Rigid body

If the distance between any two points in a body is not altered by applying a force, however large the force may be, the body is said to be a rigid body.

A rigid body may be defined as that body which does not undergo any change in its shape or size due to the application of force.

Actually, no body is a perfect rigid body. When the changes in the body are negligible, it can be considered as a rigid body.

#### Rotational motion

Each body is made of a large number of tiny particles. In the case of linear motion, all the particles present in the body will have same linear velocity.

When the body rotates about a fixed line (axis of rotation), its motion is known as rotatory motion.

The axis of rotation may lie within the body or outside the body. If all the particles of a body move in a circular path about the axis of rotation, the rigid body is said to have pure rotational motion.

When a body is in rotational motion about an axis, all the particles present in the body will have same angular velocity, but different linear velocities. The values of the linear velocities of these particles depend on the distance of the particles from the axis of rotation, since  $v = r\omega$

In rotational motion also, a rigid body which is free to rotate about an axis, opposes the change to  $\theta$  produced in its state. The measure of opposition not only depends on the mass of the body but also on the distances of the particles of the body from the axis of rotation.

Suppose a particle of mass 'm' is at a distance r from an axis, the product  $mr^2$  is called the moment of inertia of the particle about that axis.

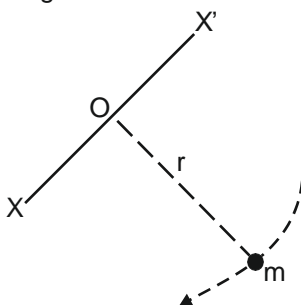
Thus in rotational motion the moment of inertia of the body about the given axis of rotation is a measure of coefficient of inertia.

Hence, the moment of inertia plays the same role in rotational motion as that of mass in translatory motion.

### Moment of inertia of a particle

The moment of inertia of a particle about an axis is equal to the product of the mass of the particle and square of its distance from the axis.

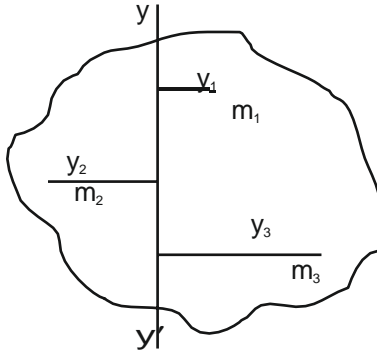
Consider a particle of mass 'm' is placed at a distance 'r' from the fixed axis. Then, the moment of inertia of the particle about the axis =  $mr^2$ . The S.I. unit for moment of inertia is  $\text{kgm}^2$ .



## Moment of Inertia of a rigid body

Inertia of a body is its inability to change by itself its state of rest or of uniform motion in a straight line.

Similarly Moment of inertia of a body is its inability to change by itself its state of rest or of uniform rotatory motion about an axis.



An external torque (rotating effect of force) is necessary to change its state.

Let us consider the rotation of a rigid body about an axis. It consists of a large number of particles of masses  $m_1, m_2, m_3$  etc., situated at distances  $r_1, r_2, r_3$  etc., from the axis  $yy'$ . Then  $m_1 r_1^2$  is known as the moment of inertia of the particle of mass  $m_1$  about that fixed axis.

Then, the moment of inertia of the rigid body = Sum of moments of inertia of all the particles present in the body.

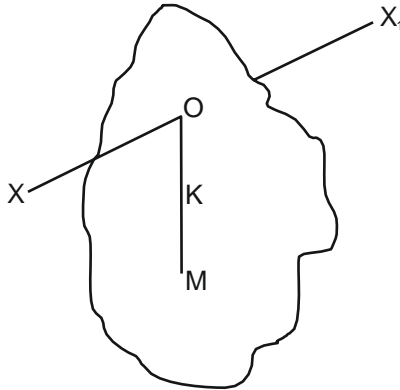
$$I = m_1 r_1^2 + m_2 r_2^2 + m_3 r_3^2 + \dots$$

$$I = \sum mr^2$$

Hence the moment of inertia of a rigid body about a fixed axis is the sum of the moment of inertia of all the particles of the rigid body.

## Radius of gyration

Radius of gyration is the distance between the given axis and the centre of mass of the body. The centre of mass of a body is a point where the entire mass of the body is supposed to be concentrated.



It is denoted by 'K'.

If M is mass of the body, then the moment of inertia =  $MK^2$ .

Hence,  $mr^2 = MK^2$

**Expression for Kinetic Energy of a Rigid body rotating about an axis:**

Consider a rigid body rotating about a fixed axis  $XOX'$ . The rigid body consists of a large number of particles. Let  $m_1, m_2, m_3, \dots$  etc., be the masses of the particles situated at distances  $r_1, r_2, r_3, \dots$  etc., from the fixed axis. All the particles rotate with the same angular velocity  $\omega$ . But the linear velocities of the particles are different.

$$\text{Kinetic Energy} = \frac{1}{2} mv^2$$

$$= \frac{1}{2} m(r\omega)^2 \quad (v = r\omega)$$

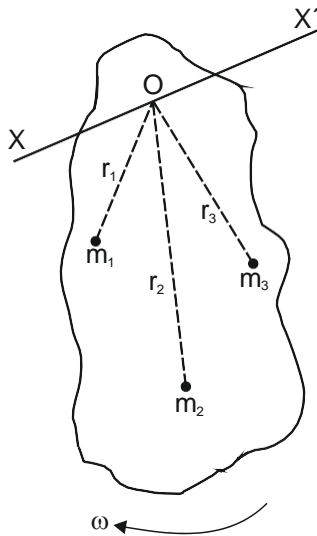
$$= \frac{1}{2} mr^2\omega^2$$

The kinetic energy of the first particle =  $\frac{1}{2} m_1 r_1^2 \omega^2$

The kinetic energy of the second particle =  $\frac{1}{2} m_2 r_2^2 \omega^2$

The kinetic energy of the third particle =  $\frac{1}{2} m_3 r_3^2 \omega^2$

and so on.



The kinetic energy of the whole body is equal to the sum of the kinetic energy of all the particles present in the body.

$$\begin{aligned} \text{The kinetic energy of the rigid body} &= \frac{1}{2} m_1^2 \omega^2 + \frac{1}{2} m_2^2 \omega^2 + \frac{1}{2} m_3^2 \omega^2 + \dots \\ &= \left( \frac{1}{2} m_1^2 + m_2^2 + m_3^2 + \dots \right) \omega^2 \\ &= \frac{1}{2} \sum m r^2 \omega^2 \end{aligned}$$

But,  $\sum m r^2 =$  Moment of inertia of the rigid body about the fixed axis  $= I$

$$\text{Therefore, The kinetic energy of the rigid body} = \frac{1}{2} I \omega^2$$

### Angular momentum

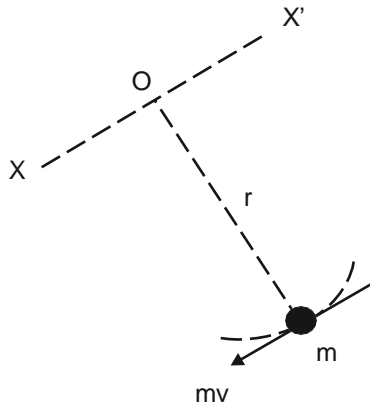
The moment of linear momentum is known as angular momentum.

Consider a particle of mass  $m$  is at a distance  $r$  from the axis of rotation. When a particle is in rotational motion about an axis, it has both linear velocity ' $v$ ' and angular velocity ' $\omega$ '.

Then, Angular momentum of the particle = linear momentum  $\times$  perpendicular distance between the particle and the axis of rotation

$$= m v \times r$$

$$= m r \omega \times r \quad (Q v = r \omega)$$

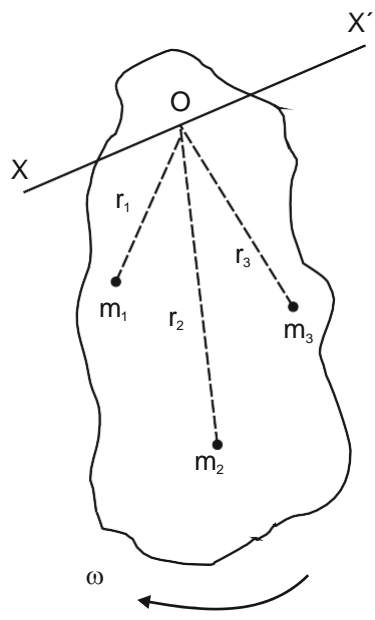


$\therefore$  Angular momentum =  $mr^2\omega$

Where  $\omega$  is the angular velocity of the particle.

The SI unit for angular momentum is  $\text{kgm}^2\text{s}^{-1}$

**Expression for Angular momentum of a Rigid body rotating about an axis:**



Consider a rigid body rotating about a fixed axis  $XOX'$ . The rigid body consists of a large number of particles. Let  $m_1, m_2, m_3, \dots$  etc., be the masses of the particles situated at distances  $r_1, r_2, r_3, \dots$  etc., from the fixed axis. All the particles rotate with the same angular velocity, but with different linear velocities depending on the values of  $r$ .

$$\begin{aligned} \text{Angular momentum} &= \text{moment of linear momentum} \\ &= \text{linear momentum} \times \text{distance} \\ &= mv \times r = mr\omega r \quad (\because v = r\omega) \\ &= mr^2\omega \end{aligned}$$

$\therefore$  The angular momentum of the first particle  $= mr_1^2\omega$   
 The angular momentum of the second particle  $= mr_2^2\omega$   
 The angular momentum of the third particle  $= mr_3^2\omega$

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...and so on.

The angular momentum of the whole body is equal to the sum of the angular momenta of all the particles present in the body.

$$\begin{aligned} \left. \begin{array}{l} \text{The angular momentum} \\ \text{of the rigid body} \end{array} \right\} L &= mr_1^2\omega^2 + mr_2^2\omega^2 + mr_3^2\omega^2 + \dots \\ &= (mr_1^2 + mr_2^2 + mr_3^2 + \dots)\omega \\ &= \sum mr^2\omega \end{aligned}$$

But,  $\sum mr^2 =$  moment of inertia of the rigid body  $= I$

$\therefore$  The angular momentum of the rigid body  $= I\omega$

### Law of conservation of angular momentum

When there is no external torque acting on a rotating body, the angular momentum of that body remains a constant. This is the statement of law of conservation of angular momentum.

i.e., if  $I_1$  and  $\omega_1$  are the initial moment of inertia and angular velocity of a rotating body and if  $I_2$  and  $\omega_2$  are new moment of inertia and angular velocity of the body, without the help of any external torque, then according to this law,

$$I_1\omega_1 = I_2\omega_2$$



**Examples**

- a) Consider a person standing on a turn-table with arms extended and a pair of weights, one in each hand. The table is made to rotate by a motor and then the motor is switched off. Now, if that person pulls his arms inwards, we can see a considerable increase in the speed of rotation. This is because, in the new position, his moment of inertia  $I$  about the axis of rotation decreases, due to the decrease in the value of  $r$ . Since the angular momentum is conserved, a decrease in the value of  $I$  results in an increase in the value of angular velocity  $\omega$ . Therefore the person is found to rotate faster.
- b) A circus acrobat, a diver or skater, all take advantage of this principle. Consider a diver has just left a diving board with his arms and legs extended, with a particular angular momentum. If he now pulls his arms and legs in, his moment of inertia decreases, leading to an increase in angular velocity  $\omega$ . By increasing  $\omega$ , the diver can perform more number of somersaults, before entering the swimming pool.

**GRAVITATION**

**Newton's laws of Gravitation**

Law 1: Any two particles of matter attract each other with a force

Law 2: The force of attraction between any two objects is

- i. directly proportional to the product of the masses
- ii. inversely proportional to the square of the distance between them.

If  $m_1, m_2$  are masses of two particles, separated by a distance 'd' then the force of attraction between the particles,

$$F \propto \frac{m_1 m_2}{d^2}$$

$$F = G \frac{m_1 m_2}{d^2}$$


where  $G$  is known as "universal gravitational constant" and the value of  $G = 6.6733 \times 10^{-11} \text{ Nm}^2 \text{ kg}^{-2}$

### Acceleration due to gravity

Galileo was the first to make a systematic study of the motion of a body under the gravity of Earth. He dropped various objects from the leaning tower of Pisa and made an analysis of their motion under gravity. He came to the conclusion that "in the absence of air, all bodies will fall at the same rate". It is the air resistance that slows down a piece of paper or a parachute falling under gravity. If a heavy stone and a parachute are dropped where there is no air, both will fall together at the same rate.

Experiments showed that the velocity of a freely falling body under gravity increases at a constant rate. i.e., with a constant acceleration. The acceleration produced in a body on account of the force of gravity is called acceleration due to gravity. It is denoted by 'g'. At a given place, the value of 'g' is the same for all bodies irrespective of their masses. It differs from place to place on the surface of the Earth. It also varies with altitude and depth.

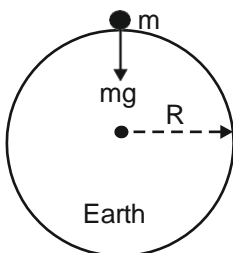
The value of g at sea-level and at a latitude of 45° is taken as the standard (i.e)  $g = 9.81 \text{ m s}^{-2}$ .

### Acceleration due to gravity at the surface of the Earth

Consider a body of mass m on the surface of the Earth as shown in the Figure. Its distance from the centre of the Earth is R (radius of the Earth).

The gravitational force experienced by the body,  $F = \frac{GMm}{R^2}$

where M is the mass of the Earth.



From Newton's second law of motion,  
Force  $F = mg$ .

Equating the above two forces,

$$\frac{GMm}{R^2} = mg$$

$$\therefore g = \frac{GM}{R^2}$$

This equation shows that  $g$  is independent of the mass of the body  $m$ . But, it varies with the distance from the centre of the Earth. If the Earth is assumed to be a sphere of radius  $R$ , the value of  $g$  on the surface of the Earth is given by

**Mass and Weight**  $g = \frac{GM}{R^2}$

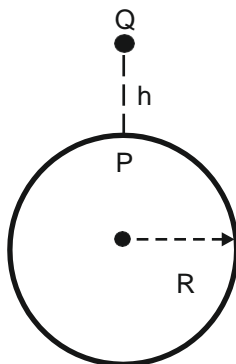
The mass of a body is defined as the quantity of matter contained in it. The mass of a body is always constant. It is measured in kilogram.

The weight of the body is usually defined as the force with which the earth attracts the body. Since the value of  $g$  varies slightly from place to place, the weight of a body is not a constant but varies slightly from place to place. However, at any particular place the weight of a body is proportional to its mass. Thus if  $w_1$  and  $w_2$  are the weights of two masses  $m_1$  and  $m_2$  at any place, then.

$$\frac{w_1}{w_2} = \frac{m_1}{m_2}$$

**Variation of acceleration due to gravity with altitude (height):**

Let  $P$  be a point on the surface of the Earth and  $Q$  be a point at an altitude  $h$ . Let the mass of the Earth be  $M$  and radius of the Earth be  $R$ . Consider the Earth as a spherical shaped body.



The acceleration due to gravity at P on the surface is

$$g = \frac{GM}{R^2} \quad \text{-----} > (1)$$

Let the body be placed at Q at a height h from the surface of the Earth. The acceleration due to gravity at Q is

$$g_h = \frac{GM}{(R+h)^2} \quad \text{-----} > (2)$$

dividing (2) by (1)

$$\frac{g_h}{g} = \frac{R^2}{(R+h)^2}$$

By simplifying and expanding using binomial theorem,  $g_h = g \left( 1 - \frac{2h}{R} \right)$

The value of acceleration due to gravity decreases with increase in height above the surface of the Earth.

### Weightlessness

Television pictures and photographs show astronauts and objects floating in satellites orbiting the Earth. This apparent weightlessness is sometimes explained wrongly as zero-gravity condition. Then, what should be the reason?

Consider the astronaut standing on the ground. He exerts a force (his weight) on the ground. At the same time, the ground exerts an equal and opposite force of reaction on the astronaut. Due to this force of action, he has a feeling of weight.

When the astronaut is in an orbiting satellite, both the satellite and astronaut have the same acceleration towards the centre of the earth. Hence, the astronaut does not exert any force on the floor of the satellite. So, the floor of the satellite also does not exert any force of action on the astronaut. As there is no reaction, the astronaut has a feeling of weightlessness.

### SATELLITES

A body moving in an orbit around another bigger body is called a satellite. A body that moves around a planet is called a satellite.

The objects that are moving in orbit by nature itself around a planet are called natural satellites. For example, moon is the natural satellite for the earth. The earth is a satellite for the sun. Man also has placed artificially some satellites to move in orbit around the desired planets. These satellites are called artificial satellite.

The artificial satellites are carried by rockets to the predetermined height, a few hundred kilometers above the surface of the earth. At this height the air is in a rarefied state and hence air friction is negligible.

After reaching the predetermined height, the satellite is given very high horizontal velocity so that it remains moving in an nearly circular orbit.

The artificial satellites are broadly classified as earth resources satellites, meteorological satellites and satellites carrying microwave sensors.

### **Escape velocity and Orbital velocity**

**Escape velocity ( $V_e$ ):** When a body is thrown vertically upwards, it will return to the earth's surface after attaining certain height. If the velocity of projection is increased, the height attained by the body becomes greater and then the body returns to the earth. This is due to the gravitational force of attraction of the earth. If the body is to be projected with a particular greater velocity, the body escapes from the gravitational pull so that it never returns to the earth. This velocity of projection is called escape velocity, it is different for different planets.

The minimum velocity required to project a body in upward direction from the surface of the earth so that it escapes from the earth's gravitational pull is called escape velocity.

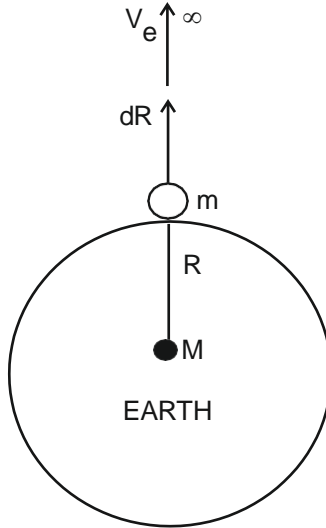
### **Orbital velocity ( $V_o$ )**

In order to put a satellite into the orbit around the earth, the satellite must be projected to the particular height and then it must be turned in a direction perpendicular to the line from the centre of the earth so that it moves in an orbit around the earth.

The velocity of the satellite along its orbit around the earth is called orbital velocity.

### Expression for Escape Velocity ( $V_e$ ):

Consider a body of mass  $m$  which is to be projected with escape velocity  $V_e$  from the surface of the earth. If the body is placed on the surface of the earth, then



The weight of the body =  $mg$  ... (1)

Let  $M$  be the mass of the earth and  $R$  its radius. According to Newton's law of gravitation,

The gravitational force,  $F = \frac{GMm}{R^2}$  ... (2)

where  $G$  is a constant known as gravitational constant. The gravitational force on a body in the earth's atmosphere is also equal to the weight of the body. Equating the eqns. (1) and (2)

$$\frac{GMm}{R^2} = mg$$

$$GM = gR^2$$

Since  $V_e$  is the escape velocity given to the body.

The given kinetic energy =  $\frac{1}{2} m V_e^2$  .... (3)

This kinetic energy is converted into work done to escape the body from gravitational pull.

If the body moves a distance  $R$  in vertical direction against the force  $F$ , then,  $\text{Work} = \text{Force} \times \text{distance}$

$$\begin{aligned} \therefore \text{Work done by the body} &= FdR \\ &= \frac{GMm}{R^2} \times dR \end{aligned}$$

In order to escape the body from the gravitational pull, the work done by the body to move from the distance  $R$  to infinity (the boundary of the gravitational field) is given by

$$\begin{aligned} W &= \int_R^\infty \frac{GMm}{R^2} dR \quad (4) \\ W &= \frac{GMm}{R} \\ \frac{1}{2} mV_e^2 &= \frac{GMm}{R} \end{aligned}$$

Equating the eqns. (3) and (4).

$$V_e = \sqrt{\frac{2GM}{R}}$$

Here,  $GM = gR^2$

$$\therefore V_e = \sqrt{\frac{2gR^2}{R}}$$

$\therefore$  The Escape velocity  $V_e = \sqrt{2gR}$

### Expression for orbital velocity of a satellite:

Consider a satellite of mass  $m$  is placed on the surface of the earth.

The weight of the satellite  $= mg$

..(1)

Let  $M$  be the mass of the earth, and  $R$  its radius. According to Newton's law of gravitation,

$$\text{The gravitational force } F = \frac{GMm}{R^2} \quad (2)$$

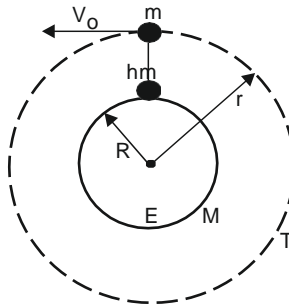
where  $G$  is a constant known as gravitational constant.

The gravitational force on the satellite is equal to its weight. Equating the eqns. (1) and (2).

$$\frac{GMm}{R^2} = mg \quad \therefore GM = gR^2$$

The satellite is to be projected vertically upto a particular height  $h$ , after it must be turned in a perpendicular direction, then the satellite revolves around the earth in a circular orbit of radius  $r$ , with an orbital velocity of  $V_0$ .

The centripetal force,  $F = \frac{mV_0^2}{r}$  ... (3)



The necessary centripetal force for circular motion of the satellite is given by the gravitational force. Hence, equating the equations (3) and (2)

$$\frac{mV_0^2}{r} = \frac{GMm}{r^2}$$

$$\therefore V_0 = \sqrt{\frac{GM}{r}} = \sqrt{\frac{gR^2}{r}} \quad (GM = gR^2)$$

The orbital velocity of the satellite,  $V_0 = \sqrt{\frac{gR^2}{R+h}} \quad (R = R+h)$

**Note:** When the satellite is launched in near earth orbit which is 200 km to 300 km,  $h$  is small when compared to  $R$  so that,  $R+h = R$ . Then

$$V_0 = \sqrt{gR}$$

### Expression for period of Revolution

Time taken by the satellite to complete one revolution round the earth is called period of revolution.

$$\text{Period } T = \frac{\text{Circumference of the orbit}}{\text{Orbital Velocity}}$$



$$\begin{aligned}
 T &= 2\pi(R+h) \sqrt{\frac{(R+h)g}{R^2}} \\
 T &= \frac{2\pi r}{v_0} \quad \text{herer=radiusoftheorbit=R+h} \\
 &= \frac{2\pi(R+h)}{\sqrt{\frac{gR^2}{R+h}}} \quad \left( QV_0 = \sqrt{\frac{gR^2}{R+h}} \right)
 \end{aligned}$$

If his very small when compared to R,

$$R+h=R, \text{ then period } T=2\pi \sqrt{\frac{R^3}{gR^2}}, \quad \therefore T=2\pi \sqrt{\frac{R}{g}}$$

### Geo- Stationary Satellite

A geo-stationary satellite is a particular type used in television and telephone communications. A number of communications satellites, which appear to remain in fixed position at a height of 36,000 km above the equator, are called synchronous satellites or geostationary satellites. Some television programmes or events occurring in other countries are often transmitted live with the help of these satellites.

The orbit at a height of 36,000 km above the equator, on which satellites are revolving, is called geosynchronous orbit. In this orbit, the satellite takes 24 hours for revolving round the earth once.

For a satellite to appear fixed at a position above a certain place on the Earth, its orbital period around the earth must be exactly equal to the rotational period of the earth about its axis. As the relative angular velocity of the earth and that of satellite, which is revolving from west to east over the equator, are identical, the satellite appears stationary to an observer on the earth.

The ground-based transmitting antenna, which is directed towards the satellite, can operate effectively without need for expensive equipment to track the satellite's motion. This implies enormous savings. TV waves originating from the ground-based station are re-transmitted from the satellite and are received at the destination. As the geostationary satellite can 'see' only one-third of the earth's surface, at least three such satellites are required to cover the entire globe.

## Polar Satellites

The polar satellites revolve around the earth in a north-south orbit passing over the poles as the earth spins about its north-south axis.

The polar satellites positioned nearly 500 to 800 km above the earth travel from pole to pole in 102 minutes. The polar orbit remains fixed in space as the earth rotates inside the orbit. As a result, most of the earth's surface crosses the satellite in a polar orbit. Excellent coverage of the earth is possible with this polar orbit. The polar satellites are used for mapping and surveying.

## Uses of Artificial Satellites

The artificial satellites are launched for many purposes by different countries. The important uses of artificial satellite are

- i. Collection of scientific data
- ii. Weather monitoring
- iii. Military Spying
- iv. Remote sensing
- v. Communication purpose – the satellite receives microwaves and TV signals from the earth and amplifies them and transmits them back to various stations on the earth.

## Worked Problems

1. If the radius of the earth is 6400 km and the acceleration due to gravity is  $9.8 \text{ ms}^{-2}$ . Calculate the escape velocity

Given  $R=6400 \text{ km}$  and  $g=9.8 \text{ ms}^{-2}$

$$\begin{aligned} \text{Escape velocity } V_e &= \sqrt{2gR} = \sqrt{2 \times 9.8 \times 6400 \times 10^3} \\ V_e &= 11.2 \text{ km per second} \end{aligned}$$

2. A satellite is revolving around the earth at a distance of 182 km from the surface of the earth. The radius of the earth is 6371 km and  $g$  is  $9.81 \text{ ms}^{-2}$ . Calculate the orbital velocity of the satellite.

Given:  $g=9.81 \text{ ms}^{-2}$ ,  $R=6371 \times 10^3 \text{ m}$ ,  $h=182 \times 10^3 \text{ m}$

$$(R+h) = (6371+182) \times 10^3 = 6553 \times 10^3 \text{ m}$$

$$V_o = \frac{gR^2}{(R+h)} = \sqrt{\frac{9.81 \times (6371 \times 10^3)^2}{6553 \times 10^3}}$$

$$V_o = 7795 \text{ ms}^{-1} \text{ (or) } 7.795 \text{ kms}^{-1}$$

3. An artificial satellite revolves around the earth at a distance of 3400 km. Calculate the period of revolution. Radius of the earth  $R = 6400$  km,  $g = 9.8 \text{ ms}^{-2}$

Given

$$R = 6400 \times 10^3 \text{ m}$$

$$h = 3400 \times 10^3 \text{ m}$$

$$g = 9.8 \text{ ms}^{-2}$$

$$= 2\pi \sqrt{\frac{(R+h)^3 g}{R^2}}$$

$$T = 2\pi \sqrt{\frac{(6400 \times 10^3 + 3400 \times 10^3)^3}{9.8 \times (6400 \times 10^3)^2}}$$

$$= 2\pi \sqrt{\frac{(9800 \times 10^3)^3}{9.8 \times (6400 \times 10^3)^2}}$$

$$= 2\pi \sqrt{\frac{9.41192 \times 10^{20}}{4.01408 \times 10^{14}}}$$

$$= 2\pi \sqrt{2.3447 \times 10^6}$$

$$= 2 \times 3.14 \times 1.53125 \times 10^3$$

$$= 9.616 \times 10^3 \text{ second}$$

## QUESTIONS

### Part– A

1. Define inertia
2. Define rigid body
3. Define moment of inertia of a particle
4. Define moment of inertia of a rigid body
5. Define radius of gyration
6. Define angular momentum
7. State the law of conservation of angular momentum
8. State Newton's law of gravitation
9. State Newton's law of gravitation
10. Define mass of a body
11. Define weight of a body

12. What is a satellite?
13. Define escape velocity
14. Define orbital velocity
15. Define period of revolution of a satellite
16. Give any two uses of artificial satellites.

### **Part–B**

1. Derive an expression for the moment of inertia of a rigid body about an axis.
2. Derive an expression for the kinetic energy of a rigid body rotating about an axis.
3. Derive an expression for the angular momentum of a rigid body rotating about an axis.
4. Obtain an expression for the acceleration due to gravity on the surface of the earth.
5. Obtain an expression for the variation of acceleration due to gravity with altitude.
6. Explain the concept of weightlessness.
7. Derive an expression for the escape velocity from the surface of the earth
8. Derive an expression for the orbital velocity of a satellite.
9. Derive an expression for the period of revolution of a satellite
10. Write notes on Geo-stationary and Polar satellites.

### **Exercise Problems**

1. Find the escape velocity at the surface of the moon, given that the radius of moon is  $2 \times 10^6 \text{ m}$  and acceleration due to gravity on the surface of moon is  $1.69 \text{ ms}^{-2}$ .
2. A satellite is revolving in circular orbit at a height of 800 km from the surface of the earth. Calculate the orbital velocity. The radius of the earth is  $6400 \text{ km}$  and  $g = 9.8 \text{ ms}^{-2}$
3. A satellite is revolving in circular orbit at a height of 1200 km from the surface of the earth calculate the period of revolution. Radius of the Earth is  $6400 \text{ km}$  and  $g = 9.8 \text{ ms}^{-2}$

### **Brainteaser**

1. The escape velocity from the surface of the earth is  $11.2 \text{ km s}^{-1}$ . Find the escape velocity from the surface of another planet where mass and diameter are twice that of earth.
2. What is the acceleration due to gravity at a distance from the centre of the Earth, which is equal to diameter of the Earth?.

## UNIT - V

# SOUND AND MAGNETISM

### SOUND

#### Introduction

Sound is produced due to the vibrations of the body. These vibrations are transferred to the air medium and propagated in all directions in the form of waves. When the vibrational sound waves reach our ear, the diaphragm of the ear vibrates with the equal vibrations produced by the body. Hence we are able to hear the sound by the sensation perceived by the nerves of the ear.

The number of vibrations made in one second is known as frequency of the sound. It is expressed in hertz (Hz). The range of the frequency between 20 Hz and 20,000 Hz is the audible range, Human ears cannot respond to the sound below and above this range. In this sense, all vibrating bodies cannot produce audible sound. The vibrations of frequency below 20 Hz are called infrasonic and above 20,000 Hz are called ultrasonic.

Sound is a mechanical wave and hence it requires a medium to propagate. So we can't hear the sound in vacuum. It travels with the velocity of  $330 \text{ ms}^{-1}$  in air. Sound waves are reflected and refracted like light waves.

#### Wave motion:

When a stone is thrown on a pond of water, ripples spread out in all directions on the surface of water. The stone disturbs the water medium at one place but the disturbance is transferred in all directions continuously. This continuous movement of the disturbance is called a wave.

If a wave passes in a medium, the particles of the medium vibrate about their mean position. The particles do not move along with the wave, only the vibrations are transferred from one particle to an adjacent particle of the medium in the form of energy.

There are two types of wave motion. They are

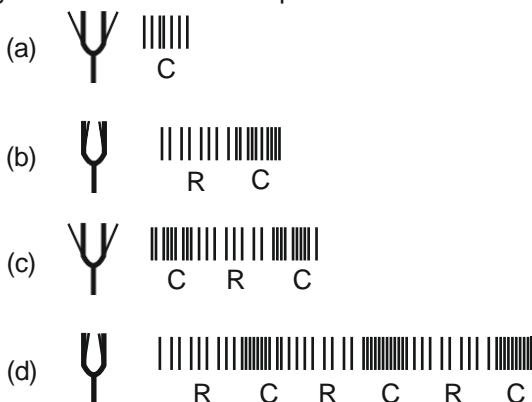
- 1) Longitudinal wave motion and
- 2) Transverse wave motion

### 1) Longitudinal wave motion:

If the particles of the medium vibrate parallel to the direction of propagation of the wave, the wave is known as longitudinal wave.

Examples:

1. The propagation of sound in air
2. The propagation of sound in gas
3. The propagation of sound inside the liquid



The longitudinal waves travel in a medium in the form of compressions and rarefactions. The place where the particles of the medium crowded together are called compressions and the places where the particles spread out are called rarefactions.

The compressions and rarefactions produced by a vibrating tuning fork are as shown in the figure. When the prong (arm) of the fork moves to the right, it compresses the medium in front of it to form compression. Meanwhile the prong returns to the left, a temporarily vacuum is created there. To fill it, the particles of the medium spread out in that place to form a rarefaction.

Thus as the prong of the fork vibrates to and fro, alternate compressions and rarefactions are transmitted in the medium. As a

result, the particles of the medium simply move back and forth about their mean position parallel to the direction of the propagation of the wave.

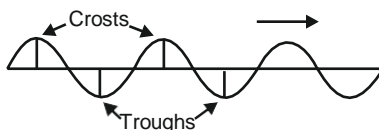
## 2) Transverse wave motion:

If the particles of the medium vibrate perpendicular to the direction of propagation of the wave, the wave is known as a transverse wave.

Examples:

1. Ripples travelling on the water surfaces.
2. Waves travelling along a rope.
3. Other waves like light waves, heat radiations, radio waves etc.

The transverse waves travel in a medium in the form of crests and troughs. The points where the particles of the medium displaced maximum in the upward direction are called crests. The points where the particles displaced maximum in the downward direction are called troughs.



The crests and troughs produced by the transverse wave motion are as shown in the figure. In a transverse wave, alternate crests and troughs are transmitted in the medium. As a result, the particles of the medium move up and down about their mean position perpendicular to the direction of propagation of the wave.

## Distinction between Transverse and Longitudinal Waves

Transverse Wave	Longitudinal Wave
1. It is a wave motion in which the particles of the medium are vibrating perpendicular to the direction of propagation.	It is a wave motion in which the particles of the medium are vibrating parallel to the direction of propagation.
2. Alternate crests and troughs are formed.	Alternate compressions and rarefactions are formed.



Transverse Wave	LongitudinalWave
3. Transverse waves can travel through solids and on liquidsurfaces.	Longitudinal waves can travel through any medium whether solid,liquidorgas.
4. Transversewavescanbe reflected, diffracted and polarized.	Longitudinalwavescannotbe polarized.

### **ProgressiveWaves:**

If a wave travels continuously in a medium without any disturbance, then the wave is said to be progressive wave. Longitudinal waves and Transverse waves are two types of progressive waves and they can travel continuously in any medium if there is no obstruction.

### **Amplitude:**

When sound wave propagates in a medium, the maximum displacement of the vibrating particles of the medium from their mean position is called amplitude.

### **Wavelength( $\lambda$ ):**

The wavelength is the distance between two consecutive particles of the medium which are in the same state of vibration.

It is also defined as the distance travelled by the wave during the time the vibrating particle completes one vibration.

In longitudinal waves, the wavelength is the distance between two successive compressions or rarefactions. In transverse waves, the wavelength is the distance between two successive crests or troughs.

### **Period(T):**

The time taken by the vibrating particle to make one vibration is called period.

### **Frequency(n):**

The frequency is the number of vibrations made by the vibrating particle in one second.

### **Velocity(v):**

The distance travelled by the sound wave in one second is known as velocity of sound.

### Relation between Wavelength, Frequency and Velocity of a Wave:

Let  $n$  be the number of vibrations made by the vibrating particle in one second. It is also known as its frequency.

Time taken for one vibration = period ( $T$ ) =  $1/n$ .

Let  $\lambda$  be the wavelength of the wave produced

Velocity of the wave is the distance through which the wave advances in the medium in one second

$$\therefore \text{Velocity of wave motion, } V = \frac{\text{distance travelled}}{\text{time taken}}$$

$$V = \lambda T = \lambda (1/n) = \lambda n$$

$$\therefore V = n\lambda$$

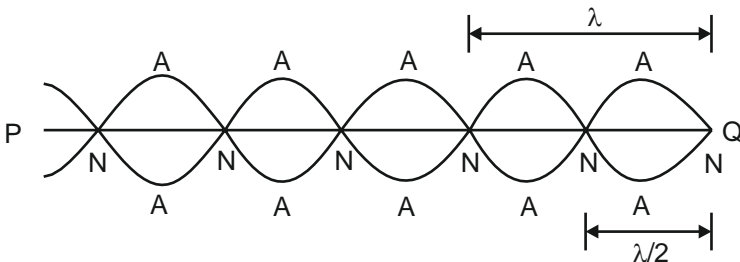
### Stationary Waves:

If a progressive wave travelling in a medium meets the surface of an obstacle, it is reflected. The reflected wave is superimposed on the incident wave to form a new type of wave called stationary wave.

Also, when two identical waves having equal wavelength and amplitude travel in opposite directions they superimpose on each other forming stationary wave.

A stationary wave formed by a vibrating string is as shown in the figure. Consider a string  $PQ$  fixed at the end  $Q$  and it vibrates up and down at the free end  $P$ . Then a transverse wave is set up and it proceeds towards the fixed point  $Q$  and gets reflected back to the end  $P$ . Now the stationary wave is formed in the string.

At certain points of the medium, the displacement due to the two waves cancel each other and those points remain at rest. Such points are called nodes ( $N$ ). At certain other points there is maximum displacement. Such points are called antinodes ( $A$ ). The distance between two successive nodes or antinodes is  $\lambda/2$ .



The distance between a node and the next antinode is  $\lambda/4$   
The longitudinal waves also produce the stationary waves.

## VIBRATIONS

### Free Vibrations:

The vibrations of any body with its natural frequency are called free vibrations.

When a body is set in vibration and left free, it executes vibrations. The frequency depends upon the dimensions and elastic constants of the body. Such vibrations are called free vibrations and the frequency of vibration is known as the natural frequency.

If a tuning fork is set in vibration, it vibrates with its own frequency. Such vibrations are called free vibrations or natural frequency.

### Forced Vibrations:

The vibrations of a body with a frequency induces vibrations on another vibrating agent are called forced vibrations.

Suppose a vibrating tuning fork is placed with its stem on a table, the vibrations of the fork are impressed on the table and the table is forced to vibrate. The vibrations set up on the table are called forced vibrations.

### Resonance:

When the forced vibrations given on the body is equal to its natural frequency of vibrations, the body vibrates with maximum amplitude. This phenomenon is called resonance.

When a vibrating tuning fork is kept on a table, the table is forced to vibrate with the frequency of the tuning fork. If the natural frequency of the table is equal to the frequency of the tuning fork, the table vibrates with its natural frequency and hence resonance occurs.

### Law of vibration in stretched strings:

**I Law:** The frequency of vibrating string is inversely proportional to its length, when the tension and linear density of the string are kept constant.

$$n \propto \frac{1}{l}$$

**II Law:** The frequency of vibration is directly proportional to the square root of the tension, when the length and linear density of the string are kept constant.

$$n \propto \sqrt{T}$$

**III Law:** The frequency of vibration is inversely proportional to the square root of the linear density of the string, when the tension and length of the string are kept constant.

$$n \propto \frac{1}{\sqrt{m}}$$

According to above laws,

$$n \propto \frac{1}{l} \cdot \sqrt{\frac{T}{m}} \quad (\text{or}) \quad n = k \frac{1}{l} \cdot \sqrt{\frac{T}{m}}$$

Where  $k$  is a constant and its value is equal to  $\frac{1}{2}$ .

$$\therefore n \propto \frac{1}{2l} \sqrt{\frac{T}{m}}$$

$$\text{If } T = Mg, n = \frac{1}{2l} \sqrt{\frac{Mg}{m}}$$

$$n = \frac{1}{2} \sqrt{\frac{Mg}{l^2 m}}$$

Note: The linear density,

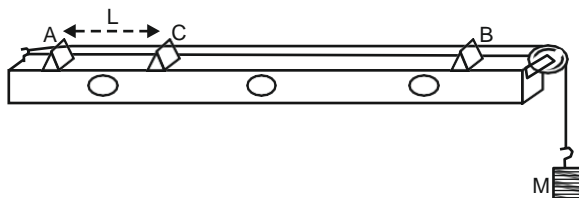
$$m = \frac{\text{mass}}{\text{length}} = \frac{\text{Volume} \times \text{density}}{\text{length}} = \frac{\pi r^2 l \rho}{l} = \pi r^2 \rho$$

### Sonometer:

The phenomenon of resonance is used in sonometer. In sonometer, the frequency of a tuning fork is equal to the frequency of the vibrating string. Here resonance takes place and the string vibrates with maximum amplitude.

### Experimental determination of frequency of a tuning fork using sonometer:

The sonometer consists of a hollow wooden box. A nail is fixed at one end and a smooth pulley is fixed at the other end of the box. One end of a sonometer string is tied to the nail and the other end of the string passes over the smooth pulley. The free end is attached to the weight hanger. A, B and C are three knife edges placed on the box under the string. A, B are fixed and C is movable.



As a suitable tension ( $T=Mg$ ) is applied to the string. A small paper rider is placed on the string in between A and C. The tuning fork of frequency 'n' is excited with a rubber hammer and its stem is kept on the sonometer box. Now the string is made to vibrate. The movable knife-edge C is adjusted such that the string vibrates with the same frequency of the fork. At that time, the paper rider placed on the string between A and C, is violently thrown off from the string. Now the vibrating length of the string  $AC=l$  is measured.

Let  $r$  be the radius of the string measured by using screw gauge and ' $\rho$ ' be the density of the material of the wire, then

The linear density,  $m = \pi r^2 \rho$

The experiment is repeated by changing the value of tension (load).

The readings are tabulated.

Sl.No.	Load (M)	Vibrating length $AC = l$	$f$	$\frac{M}{l^2}$

Mean:

The frequency of the tuning fork,  $n = \frac{1}{2l} \sqrt{\frac{Mg}{m}}$

The mean value of  $\frac{M}{l^2}$  is substituted in the above formula and

the frequency of the tuning fork 'n' is calculated

**ULTRASONIC:**

The sound waves above the frequency of 20000 Hz (20 KHz) are known as ultrasonic waves. Ultrasonic waves are inaudible to humans, because their high frequencies are unable to vibrate the human ear

drum. The ultrasonic waves are high frequency sound waves smaller wavelength. Its wavelength at room temperature is 0.0175m. Since they are sound waves, they require a material medium to travel and have most of the properties of sound waves. In addition to this, the ultrasonic waves are having high energy and it helps in large number of applications in engineering and medical fields.

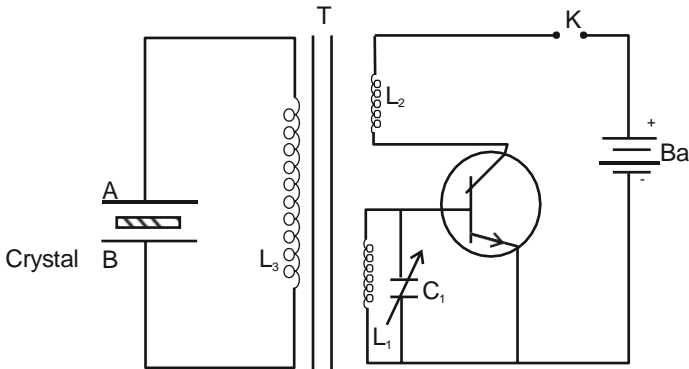
The ultrasonic waves are usually generated by the application of magnetostriction effect or piezo electric effect. The piezo electric generator is more efficient than the magnetostriction generator.

### **Production of ultrasonic waves – Piezoelectric method:**

In this method, the ultrasonic waves are produced by using the principle of piezoelectric effect. According to this effect, if a mechanical pressure is applied to one pair of opposite faces of certain crystals like quartz, cut with their faces perpendicular to its optic axis, the other pair of opposite faces develop equal and opposite electrical charges. The inverse of piezo electric effect is also true. According to this, if an AC-voltage is applied to one pair of opposite faces of the crystal, alternative mechanical elongations and contractions are produced in the other pair of opposite faces of the crystal. Inverse piezoelectric effect is used in this method to vibrate the crystals and to produce ultrasonic waves.

The experimental arrangement of piezoelectric oscillator is shown in the figure. It consists of a common emitter transistor circuit, a transformer (T) and two metal plates A and B. In transformer, three inductive coils are used,  $L_1$ ,  $L_2$  are the primary coils and  $L_3$  is the secondary coil. The coil  $L_1$  is connected to the variable capacitor C, to form a tank circuit in which frequency is generated. The coil  $L_2$  is connected to the collector of the transistor. The base of the transistor is coupled to the tank circuit. The coil  $L_3$  is connected to the parallel plates A and B. A quartz crystal is placed in between them.

When the circuit is switched on, current starts to flow in the tank circuit and setting up oscillations. These oscillations are mutually induced in the secondary coil  $L_3$ . The crystal is now under high frequency alternating voltage and it starts vibrating under inverse piezoelectric



effect. This causes a back emf to be formed across  $L_2$ . This emf is amplified by the transformer and fed back into the tank circuit, where the oscillations get sustained. The capacitor  $C_1$  is varied till the frequency of oscillations reaches with the natural frequency of the crystal. Due to resonance the crystal vibrates with large amplitude. The vibrations of the crystal produce high power ultrasonic waves in the surrounding air.

**Properties of Ultrasonic waves:**

1. The ultrasonic waves are high frequency sound waves.
2. They are having smaller wavelength.
3. They produce heating effect when pass through the medium.
4. They get reflected, refracted and absorbed by the medium similar to the ordinary sound waves.
5. They act as catalytic agent to accelerate chemical reactions.
6. They produce stationary wave pattern in the liquid while passing through it.

**Applications of ultrasonic waves:**

Ultrasonic waves have a large number of practical applications. Some important applications are given below.

1. The ultrasonic waves are used to find the depth of the sea, distance and direction of submarine and depth of the rocks in the sea.
2. They are used in non-destructive testing to detect cracks in the materials, welding, casting etc.

3. They are used in the ultrasonic microscope to detect concealed objects.
4. They are used to heat the substance.
5. They are used to bore holes in steel and other hard metals.
6. They are used to accelerate chemical reactions.
7. They are used to solder the aluminium. In ordinary soldering method aluminium cannot be soldered.
8. They are used in the mixing of alloys of different compositions.
9. They are broadly used in medical field such as to detect tumours, abnormal growth, cancer, broken teeth, relieving neuralgic pains, headache, body massage action and also used in bloodless surgery.
10. They are used in the sterilization of water and milk.

### **ACOUSTICS OF BUILDINGS:**

**Echo:** The first reflected sound known as echo. The sound produced by a source is propagated continuously in a medium if there is no disturbance. But if it meets the hard surface of an obstacle, it is reflected. The clear echo depends upon the following factors.

1. Good reflector of sound
2. Maximum surface area of the reflector and
3. The distance of the reflector from the source of sound.

**Reverberation:** The sound produced in a hall suffers multiple reflections before it becomes inaudible. As a result of these reflections, the listener continues to receive sound, even if the source of sound is cut off. This prolonged reflection of sound in a room even after the sound source has been stopped is called reverberation. It is the persistence of sound due to multiple reflections from the walls, floor and ceiling of a hall. The reverberation is also called multiple echoes. In a room, the walls, floor and other flat surfaces reflect sound with a small loss of energy.

**Reverberation time:** If a building is to be acoustically correct, its reverberation time must be in optimum level. It should not be too long or too short. If it is too short, then the room becomes dead in sound aspect. If it is too long, then the reverberation will be there inside the building for long duration. The reverberation produces continuous sound with decreasing intensity up to a particular time after that it disappears. This time is known as reverberation time.



The reverberation time is defined as the time taken by the sound to fall from its original intensity to one millionth of its original intensity.

**Sabine formula:** Sabine derived an equation for the reverberation time.

Where  $V$  is the volume of the hall,  $\alpha$  is the coefficient of absorption of each reflecting surface present in the hall and  $A$  is the area of the each

$$T = \frac{0.16V}{\alpha A} \quad \text{second}$$

sound absorbing surface present in the hall.

**Coefficient of absorption of sound energy:**

The co-efficient of absorption of sound energy of any surface is defined as the ratio of the sound energy absorbed by the surface to the total sound energy incident on the surface.

Let  $\alpha$  be the coefficient of absorption of sound energy of a surface, then

$$\alpha = \frac{\text{The sound energy absorbed by the surface}}{\text{The total sound energy incident on the surface}}$$

**Acoustics of Buildings (Architectural acoustics):**

The term 'Acoustics of building' describe the production, transmission and reception of sound inside the building. The following points should be taken into consideration for the acoustic design of the buildings, auditorium, cinema theatres etc.

1. The sound heard by the audience should be sufficiently loud in any part of the hall.
2. The quality of the speech and music should not be changed anywhere inside the hall.
3. There should not be focusing of sound due to walls and ceiling, in any part of the hall.
4. There should not be any vibrations due to resonance.
5. There should not be any other noise from other sources, both from outside the hall and from within the hall.

Two more important factors should also be considered for good acoustic design of the buildings.

## 1. Echo and 2. Reverberation

**1. Echo:** The direct sound from the source and the reflected sound (echo) from the walls produce confusion in certain buildings. A hall with large number of open windows is free from the defect. Echoes can be eliminated by making the walls rough. But in the case of musical hall, however echoes are desirable, to a certain extent.

**2. Reverberation:** The Reverberation is the prolonged reflections of sound from the walls, floor and ceiling of a hall. The sound produced in the hall undergoes multiple reflections before it becomes inaudible. As a result the clarity of the successive speech is affected. Due to this reason, the reverberation time should be optimum to hear the clear sound in the hall.

To minimize this defect, Sabine derived an equation for the reverberation time,

Where  $V$  is the volume of the hall,  $\alpha$  is the coefficient of absorption of

$$T = \frac{0.16V}{\alpha A} \text{ second}$$

each reflecting surface present in the hall and  $A$  is the area of the each sound absorbing surface present in the hall.

In the above Sabine's equation,  $T$  will be minimized by increasing the value of  $\alpha$ . So the reverberation is minimized by covering the inner surfaces of the hall using absorption materials like carpets, wall screen, porous tiles etc.

The reverberation time value depends on the use for which the building is designed. The acceptable limit for reverberation time is

For speech, 0.5s

For music, between 1.0 and 2.0s

For small theatres, between 1.0 and 1.5s

For larger theatres, up to 2.3 s

### **Noise Pollution:**

The unwanted sounds which are dumped into the atmosphere, thus producing an nuisance and adverse effect on the health of human beings, are known as noise pollution. Noise pollution leads to severe health

problems by creating either physical or mental problems. As a result, it affects the working efficiency, personal comfort, and in some cases it may lead to industrial accidents. There are varieties of noises such as domestic noise, traffic noise, aircraft noise, etc. Even though the sources of noise are numerous, the types of sources of noise are generally classified into two types.

(i) Industrial noises and (ii) Non-industrial noises

The noises which are reproduced due to the movements of car, motor truck, motor cycle, aircraft, machines, explosions, etc. are known as industrial noises. The noises which are reproduced due to the loudspeaker, construction work, radio, etc. are known as non-industrial noises. The industrial and non-industrial noises are generally measured by using the sound pressure level in decibel (dB).

Some important sound pressure level for domestic and industrial noise is listed below.

Source of noise.

<u>Source of noise.</u>	<u>Sound pressure level (dB)</u>
Garden	20
Library	30
Ring in alarm clock	80
Heavy traffic	90
Electric generator	76
Concrete mixer	85
Power saw	90
Noisy factory	100
Threshold of pain	140

**Noise control in Machines:**

In industry, the noise pollution produced by the machines is to be acceptable to the working environment for the employees. When the noise level (dB) is above the acceptable level, it will affect the workers efficiency, comfort, and sometimes may lead to industrial accidents. Therefore, one has to take necessary care for the design of machine and work place for quieter operations. The following are the some of the general principles to be followed in controlling the noise pollutions in machines:

1. The impacting parts of the machines must be enclosed within enclosures.
2. All the enclosures should have high sound absorbing materials.
3. Care must be taken to select the sound absorbing material to suit the environmental conditions.

4. The vibrating machines must be dynamically balanced.
5. The rotating or impacting machines must be fixed using a fixed bed.
6. The connections such as electricity, gas, air etc. should be provided within the vibration decouplers around the machine.
7. The water connections should be provided through pressure release chambers.
8. Engines like internal combustions should be properly silenced.
9. All the machines should be kept in a large work area which includes good sound absorbing materials as an enclosure.
10. The machine should be isolated from the offices, show rooms etc. Therefore, the noise pollution due to the machines will not have any influence on office/showrooms.

The other source of noise pollution in machines is the vibration of the machines. The impact noise produced by the machine under operating conditions will be transferred into the building in the form of vibration. These vibrations pass through the building and hence, the problem will be extended to the remote parts of the building. This can be reduced to a greater extent by providing discontinuities in the construction of the building.

### Worked Problems:

1. A wire 50 cm long and of mass  $6.5 \times 10^{-3}$  kg is stretched so that it makes 80 vibrations per second. Find the stretching tension.

$$m = \frac{\text{mass}}{\text{length}} = \frac{6.5 \times 10^{-3}}{0.5} = 13 \times 10^{-3} \text{ kg m}^{-1}$$

$$n = \frac{1}{2l} \sqrt{\frac{T}{m}} \quad \text{or } n^2 = \frac{T}{4l^2 m}$$

$$\therefore \text{Tension } T = 4l^2 m n^2 = 2 \times 0.5 \times 0.5 \times 13 \times 10^{-3} \times 80 \times 80 \\ = 83.2 \text{ N}$$

2. The density of a sonometer wire of radius 0.3 mm is  $7800 \text{ kg m}^{-3}$ . Find its linear density.

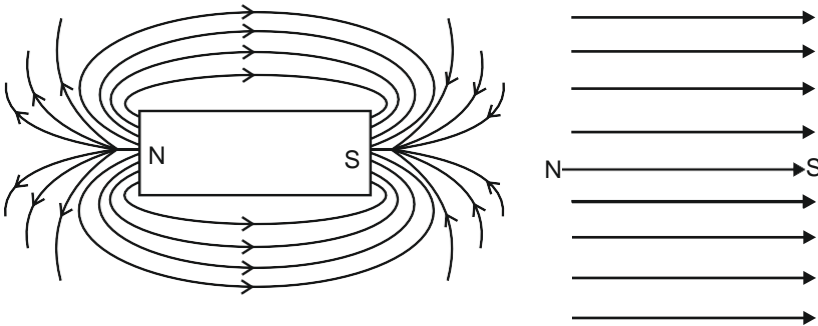
$$\text{Linear density } m = \pi r^2 \rho \\ m = 3.14 \times (0.3 \times 10^{-3})^2 \times 7800 \\ = 2204.28 \times 10^{-6} \\ = 2.204 \times 10^{-3} \text{ kg m}^{-1}$$

## MAGNETISM

The word 'magnetism' is derived from the name of a dark brown coloured ore called magnetite found in Magnesia. The ore possesses the property of attracting small pieces of iron. The ore magnetite available in nature is called as natural magnet. Artificial magnets also can be prepared from iron or steel materials. Iron materials are mostly used to prepare electromagnets and temporary magnets whereas the steel materials are used to prepare permanent magnets.

### Magnetic field:

The space around the magnet in which the magnetic lines of force act is called the magnetic field. The direction of magnetic field at a point is given by the direction of the force acting on the north pole of a magnetic needle placed at the point. If the magnetic lines of force formed are parallel, the field is called uniform magnetic field and have equal strength at all the points.



### Coulomb's Inverse Square Law:

- The force of attraction or repulsion between two poles is
- i) directly proportional to the product of the two poles,
  - ii) Inversely proportional to the square of the distance between the poles.

Let the two poles of strength  $P_1$  and  $P_2$  are placed in air (or) vacuum at a distance  $d$ . The force of attraction or repulsion between the poles

$$F \propto \frac{P_1 P_2}{d^2} \text{ (or) } F = k \frac{P_1 P_2}{d^2}$$

where  $k$  is a constant and  $k = \frac{1}{4\pi\mu_0}$

$\mu_0$  is permeability of the free space

$$\therefore F = \frac{1}{4\pi\mu_0} \frac{P_1 P_2}{d^2}$$

$$\mu_0 = 4\pi \times 10^{-7} \text{ henry/metre } F =$$

$$\frac{1}{4\pi \times 4\pi \times 10^{-7}} \times \frac{P_1 P_2}{d^2}$$

$$F = \frac{10^7}{16\pi^2} \frac{P_1 P_2}{d^2}$$

If  $P_1 = P_2 = 1$  and  $d = 1$  m

$$\text{Then, } F = \frac{10^7}{16\pi^2} \text{ newton}$$

### Unit Pole:

Unit pole is defined as that pole which is placed at a distance of 1 m in air (or) vacuum from an equal and similar pole repels it with a force of

$$\frac{10^7}{16\pi^2} \text{ newton}$$

### Pole Strength (P):

The pole strength of a magnet is defined as the force acting on the pole when it is placed in a uniform magnetic field of unit intensity.

Let  $P$  be the pole strength placed in a uniform magnetic field of intensity  $H$ , Force acting on the pole,  $F = HP$ . If  $H = 1$ , then  $F = P$

The unit of pole strength is ampere metre (Am.).

### Magnetic Induction (B) :

The magnetic induction or magnetic flux density is defined as the total number of magnetic lines of force passing normally, through unit area of cross-section.

The unit of magnetic induction is weber/metre<sup>2</sup>.

### **Intensity of magnetic field (H):**

The intensity of the magnetic field at a point in a magnetic field is defined as the force experienced by a unit north pole placed at that point.

The unit of Intensity of magnetic field is ampere/metre ( $\text{Am}^{-1}$ )

### **Magnetic permeability ( $\mu$ ):**

The magnetic permeability of the medium (material) is the ratio of the magnetic induction in the medium to the intensity of magnetic field

$$\text{Magnetic permeability} = \frac{\text{Magnetic induction}}{\text{Intensity of magnetic field}} \text{ i.e. } \mu = \frac{B}{H}$$

producing that induction.

The unit of magnetic permeability is weber/ampere-metre ( $\text{WbA}^{-1}\text{m}^{-1}$ ) or henry/metre.

### **Magnetic moment (m):**

Magnetic moment of a magnet is defined as the moment of the couple acting on a magnet when it is placed perpendicular to a uniform magnetic field of unit intensity.

Consider a magnet of length  $2l$  placed perpendicular to the uniform magnetic field of intensity  $H$ . If  $P$  is the pole strength of the magnet, Force experienced by the poles

$$F = HP$$

$$\text{If } H = 1, \text{ unit intensity of magnetic field, Force } F = P$$

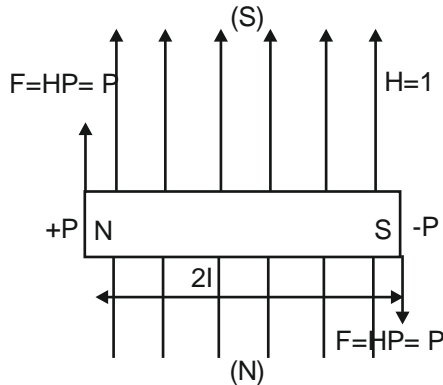
Two equal forces  $P$  and  $P$  acting on the magnet in opposite directions to form a couple.

$$\therefore \text{Moment of the couple} = P \times 2l$$

$$\text{i.e. Magnetic moment } M = 2lP$$

Hence, the magnetic moment of a magnet can also be defined as the product of pole strength and the distance between the poles. But, the distance between the poles is approximately equal to the length of the magnet. So, the moment can be taken as the product of pole strength and length of the magnet.

The unit of magnetic moment is ampere metre<sup>2</sup> ( $\text{Am}^2$ )



### Intensity of Magnetisation (M):

Intensity of magnetisation of a magnet is defined as the magnetic moment per unit volume of the magnet.

Let  $m$  be the magnetic moment of the magnet having volume  $v$ , then

$$\text{Intensity of magnetisation} = \frac{\text{Magnetic moment}}{\text{Volume of the magnet}}$$

$$\text{i.e. } M = \frac{m}{v}$$

The unit of intensity of magnetisation is ampere/metre ( $\text{Am}^{-1}$ )

### Hysteresis:

In electrical method, a magnetic material can be magnetised by keeping it inside an insulated coil (solenoid) through which a current is passed. Before passing the current, the magnetising field around the coil is zero. Then there is no magnetism induced in the material. If the magnetising field  $H$  is gradually increased by increasing the current ' $I$ ' the intensity of magnetisation induced in the material also increases up to a particular value. After that the intensity of magnetisation is not increased even though the magnetising field is increased. Now the material is said to be saturated.

If the magnetising field is gradually decreased to zero value, the intensity of magnetisation in the material also decreases but not to zero value. There is a tendency in the material to retain some amount of magnetism. Then, if the field is increased in the reverse direction from zero value, for a particular value the retained magnetism in the material is removed completely. Further increasing the field, the material attained the saturated value of magnetism in the reverse direction and soon.



**Definition:**The lagging in intensity of magnetisation behind the magnetising field is known as hysteresis.

A graphical curve which represents the variation in the intensity of magnetisation (M) with magnetic field intensity (H) is known as hysteresis loop (M-H curve).

**Retentivity, Coercivity and magnetic saturation:**

1. Retentivity or residual magnetism:

Retentivity or residual magnetism is the amount of intensity of magnetisation retained in the material after removing the magnetising field.

2. Coercivity or coercive force:

Coercivity or coercive force is the amount of magnetising field applied in the reversed direction to remove the residual magnetism completely from the material.

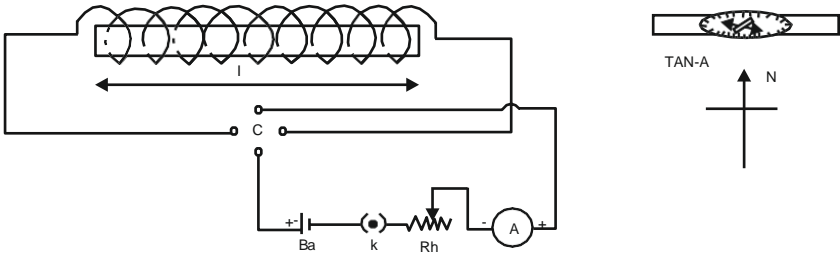
3. Magnetic saturation:

When the magnetising field is increased, the intensity of magnetisation induced in the material also increases up to a particular value. After that the intensity of magnetisation is not increased even though the magnetising field is increased. Now the material attains the saturation value of intensity of magnetisation.

**Experiment to draw the Hysteresis loop:**

The experimental arrangement is shown in the figure. A long solenoid is connected in a circuit. The direction of the current can be reversed by using a commutator C. The given specimen in the form of a rod is placed inside the solenoid. A deflection magnetometer is placed in a position at a short distance along the axis of the rod. The deflection magnetometer is adjusted to zero position.

The intensity of magnetic field H can be calculated by using the formula  $H = (NI)/L$ , where N is the number of turns in the solenoid, I the strength of the current passing through the solenoid and L, the length of the solenoid. The value of I can be measured by using an ammeter. The intensity of magnetisation M induced in the specimen can be calculated by using the formula  $M = k \tan \theta / V$  where k is constant,  $\theta$  is the deflection in the magnetometer and V is the volume of the magnet.



The strength of the current is gradually increased in steps of 0.2A upto a required value. Every time the deflection in the magnetometer is noted. Then the values of H and the corresponding values of M are calculated. The current is decreased in steps upto zero value and again the corresponding values of H and M are calculated. The experiment is also repeated by passing the current in the reverse direction.

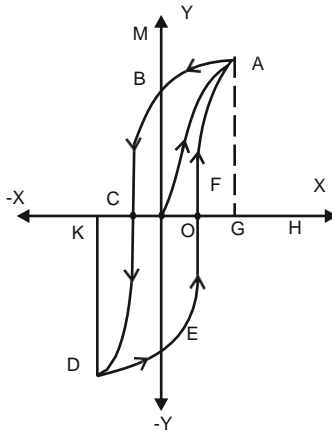
A graph is drawn by taking the values H along X-axis and M along Y-axis. A loop OABCDEF is obtained as shown in the figure.

In the figure, when H is increased from zero to G, the value of M is increased from zero to A represented by OA. At the point A, the specimen reached the saturated value.

If H is decreased from G to zero, the value of M decreases from A to B, but not to zero. Now the specimen retains some amount of magnetism known as retentivity (OB).

When H is increased in the reverse direction from zero to C, the value M reaches zero value at point C. The amount of H applied in the reverse direction to remove the residual magnetism is known as coercive force or coercivity (OC). Further repeating the process the remaining portion CDEF is obtained. The closed loop ABCDEF is called one cycle of hysteresis loop or M-H curve.

When the material is taken through a cycle of magnetisation, there is some expenditure of energy. This loss of energy dissipates in the form of heat in the specimen. The loss of energy is directly proportional to the area of the hysteresis loop.

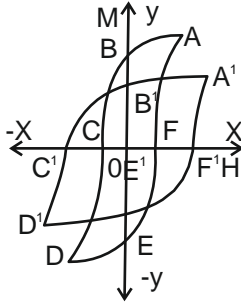


### **Selection of magnetic materials for permanent and temporary magnets (Uses of Hysteresis loops):**

The hysteresis loops are very useful in the selection of magnetic materials for permanent and temporary magnets and other industrial purposes. The hysteresis loops for different materials are compared to one another and then the materials are selected for suitable purposes.

For example, let ABCDEF A be the hysteresis loop for soft iron and A' B' C' D' E' F' A' the hysteresis loop for steel. Here the area of the hysteresis loop for steel is more than that of soft iron. Therefore the loss of energy is more in steel than that in soft iron. Hence the quantity of heat emitted by steel is more than that by soft iron. For this reason soft iron is selected as a core material for transformers, chokes, AC motors, dynamos, etc.

The retentivity of soft iron (OB) is more than that for steel (OB'). So, soft iron retains more magnetism than steel. But the coercivity for soft iron (OC) is less than the coercivity for steel (OC'). So the strength of the magnetic field needed to remove the residual magnetism in soft iron is less than in steel. That means it is easy to remove the residual magnetism in soft iron than steel. Hence steel retains the magnetism during long period. For the above reasons steel is used for permanent magnets and soft iron is used for temporary magnets and electromagnets.



## QUESTIONS

### Part-A

1. Define wavemotion.
2. Define transverse wavemotion.
3. Define longitudinal wavemotion.
4. Define progressive wave.
5. Define amplitude of a wave.
6. Define wavelength of a wave.
7. Define period of wavemotion.
8. Define frequency of a wave.
9. Define velocity of sound wave.
10. Define stationary wave.
11. Define free vibrations.
12. Define forced vibrations.
13. Define Resonance.
14. State any one of the laws of transverse vibrations in stretched strings.
15. What is the use of a sonometer?
16. What is ultrasonic?
17. Write any two properties of ultrasonic.
18. Write any two applications of ultrasonic.
19. What is an echo?

20. What is reverberation?
21. What is reverberation time?
22. Write Sabine's formula.
23. Define coefficient of absorption of sound energy.
24. State Coulomb's Inverse Square Law.
25. Define Pole Strength.
26. Define Unit Pole.
27. Define magnetic induction.
28. Define intensity of magnetic field.
29. Define permeability.
30. Define magnetic moment of a magnet.
31. Define intensity of magnetisation.
32. Define Hysteresis.
33. Define Retentivity
34. Define coercivity
35. Define magnetic saturation.

### **Part-B**

1. Explain transverse wave motion and longitudinal wave motion.
2. Distinguish between transverse and longitudinal wave motion.
3. Write short notes on stationary waves.
4. Explain the laws of transverse vibrations in a stretched string and obtain the expression for the frequency of vibration.
5. Describe how the frequency of a tuning fork is determined using a sonometer.
6. Explain the production of ultrasonic by piezo-electric method.
7. Explain the properties and applications of ultrasonic.
8. Write a note on acoustics of buildings.
9. Explain noise pollution and the methods of controlling industrial noise.

10. Describe the method of drawing hysteresis loop of a specimen using a solenoid.
11. Explain the uses of hysteresis loop.

### Exercise Problems

1. A sonometer wire of 0.5 m long gives vibrations of 256 Hz when stretched with a load of 5 kg. Find the linear density of the material of the wire.  
(Ans: 450N)
2. Find the frequency of sound produced by a string 25 cm long stretched by a load of 5 kg. The linear density of the wire is  $4.9 \times 10^{-3} \text{ kg m}^{-1}$   
(Ans: 43.20N)
3. A string 75 cm long and weighing 15 g produces a note of frequency 100 cycles per second on plucking. What is the tension in the string?  
(Ans: 209Hz)
4. A wire of 50 cm long and of mass  $6 \times 10^{-3} \text{ kg}$  is stretched so that it makes 60 vibrations per second. Find the tension in the wire.  
(Ans: 361.Hz)
5. The vibrating length of 0.75 m of a sonometer wire is in unison with a tuning fork when stretched by a weight of 5 kg. The linear density of the wire is  $0.5 \times 10^{-3} \text{ kg m}^{-1}$ . Calculate the frequency of the tuning fork.  
(Ans: 250Hz)
6. A wire 50 cm long and of mass  $0.6 \times 10^{-3} \text{ kg}$  is stretched by a tension of 4 kg wt. When sounded, it is found to vibrate in 2 loops. Calculate the frequency of the note emitted by the wire.  
(Ans: 250Hz)
7. A wire 0.5 m long vibrates 100 times a second. If the length of the wire is shortened to 0.4 m and the stretching force is increased to 4 times its original value, what will be the new frequency?  
(Ans: 250Hz)

### BrainTeaser

1. Thunder was heard 6 second after a flash of lighting was seen. If the velocity of sound is  $345\text{ms}^{-1}$ , calculate the distance at which flash occurred.
2. Twomagneticpoles,oneofwhichistwicestrongerthantheotherrepel oneanotherwithaforceof  $2 \times 10^{-5}\text{N}$  when kept separated at a distance of 20cm in air. Calculatethe strength of each pole.

**FIRST SEMESTER**  
**1004ENGINEERINGPHYSICS-I**  
**MODELQUESTIONPAPER-1**

Time:3Hrs

MaxMarks:75

**PART-A**

Marks15x1=15

**Note:** Answer any 15 Questions.

1. Define unit.
2. Define derived quantities.
3. Write Parallelogram Law of forces.
4. Define moment of a force.
5. Write Hooke's Law.
6. Define Co-efficient of viscosity.
7. Define Terminal velocity.
8. Write any two practical applications of capillarity.
9. Define trajectory.
10. Define angular velocity.
11. Define angle of banking.
12. Define frequency.
13. Define rigid body.
14. Write law of conservation of angular momentum.
15. Define escape velocity.
16. Write any two uses of artificial satellites.
17. Define resonance.
18. What is ultrasonic?
19. Define intensity of magnetisation.
20. Define retentivity.

**PART-B**

Marks 5x12=60

Note: i) Answer all Questions choosing any two subdivisions from each question.

ii) All subdivisions carry equal marks.

1a) List the conventions followed in SI.

b) Derive expressions for the magnitude and direction of the resultant of two forces acting at a point with an acute angle between them.



- c) Describe an experiment to determine the mass of the given body using principle of moments.
- II a) Describe an experiment to determine the Young's modulus of the material of a beam by uniform bending method.
- b) Derive Poiseuille's formula for the coefficient of viscosity of a liquid by dimensional method.
- c) A capillary tube of bore 0.5 mm is dipped vertically in water of surface tension  $0.072 \text{ Nm}^{-1}$ . Find the height of capillary rise.
- III a) A missile is fired at an angle of  $40^\circ$  to hit a target situated at a distance of 100 km. Find the velocity of projection of the missile.
- b) Derive an expression for the angle of banking of a curved path.
- c) Show that simple harmonic motion can be obtained by the projection of a uniform circular motion on any diameter.
- IV a) Derive an expression for the angular momentum of a rigid body rotating about an axis.
- b) Derive an expression for the orbital velocity of a satellite.
- c) Derive the expression for the variation of acceleration due to gravity with altitude.
- V. a) Explain the production of ultrasonic using Piezoelectric generator.
- b) A sonometer wire is loaded with a mass of 2 kg. The linear density of the wire is  $2 \times 10^{-3} \text{ kgm}^{-1}$ . When an excited tuning fork is placed on the sonometer box, the resonating length is found to be 15.4 cm. Find the frequency of the tuning fork.
- c) Explain the method of drawing hysteresis loop of a specimen taken in the form of a rod, using a solenoid.

**FIRST SEMESTER**  
**1004ENGINEERINGPHYSICS-I**  
**MODELQUESTIONPAPER-II**

Time:3Hrs

MaxMarks:75

**PART-A**

Marks15x1=15

**Note:** Answer any 15 Questions.

1. Write the two supplementary quantities.
2. Derive the dimensional formula for force.
3. Define coplanar forces.
4. Define moment of a couple.
5. Define Poisson's ratio.
6. What is the significance of Reynolds number?
7. Write any two practical applications of viscosity.
8. Define angle of contact.
9. Define time of flight of a projectile.
10. What is the relation between linear velocity and angular velocity?
11. What is banking of a curved path?
12. Define simple harmonic motion.
13. Define radius of gyration.
14. Write the Newton's second law of gravitation.
15. What is the difference between mass and weight of a body?
16. Define orbital velocity.
17. Write the relation between wavelength, frequency and velocity of a wave.
18. Write any two applications of ultrasonic.
19. Define reverberation time.
20. What is coercivity?

**PART-B**

Marks 5x12=60

Note: i) Answer all Questions choosing any two subdivisions from each question.

ii) All subdivisions carry equal marks.

- I a) Explain multiple and sub-multiple and the corresponding prefixes of units.

- b) Describe an experiment to verify Lami's theorem.
  - c) Find the magnitude and direction of the resultant of two forces 3 N and 4 N acting at a point, if the angle between the forces is  $60^\circ$ .
- II a) Explain the elastic behaviour of a material with the help of stress–strain curve.
- b) Describe an experiment to determine the coefficient of viscosity of a high viscous liquid by Stokes' method.
  - c) Derive an expression for the surface tension of the liquid by capillary rise method.
- III a) Show that the path of a projectile is a parabola.
- b) Derive expressions for the normal acceleration and centripetal force acting on a body executing uniform circular motion.
  - c) An aeroplane travelling in a curved path with a speed of 500 kmph tilt through an angle of  $30^\circ$ , for safe negotiation. What is the radius of the curve?
- IV a) Derive an expression for kinetic energy of a rigid body rotating about an axis.
- b) Explain geo–stationary and polar satellites.
  - c) Assuming the average radius of the earth as 6380 km, find the escape velocity of a body from the surface of the earth.
- V a) Describe an experiment to determine the frequency of a tuning fork using sonometer.
- b) Explain noise pollution and the measures to control the noise in machines.
  - c) Explain the uses of hysteresis loop in the selection of magnetic materials.

# FIRST SEMESTER

## PRACTICAL-I

### MICROMETER (SCREW GAUGE)

**ExperimentNo.**

**Date:**

**Aim:**

To measure the thickness of the given irregular glass plate, using a micrometer. To measure area of the glass plate using a graph sheet and to calculate the volume of the glass plate.

**Apparatus required:**

Screw gauge, graph sheet and given glass plate.

**Formula:**

Volume of the glass plate = Thickness  $\times$  area of the glass plate.

Thickness of glass plate = Observed Reading + Zero Correction  
 $= \{PSR + (HSC \times LC)\} + ZC$

Where PSR = Pitch scale reading  
 HSC = Head scale coincidence  
 LC = Least Count  
 ZC = Zero Correction

Head Scale Reading (HSR) = Head Scale coincidence  $\times$  Least Count

Observed Reading (OR) = Pitch Scale Reading + Head Scale Reading

Correct Reading (CR) = Observed Reading + zero Correction

**Description:**

The screw gauge consists of a screw working in a hollow cylindrical nut attached to one end of a frame G. On the hollow cylinder a scale is marked in mm. This scale is called the pitch scale (PS). There is a fixed stud A at one end of the frame and exactly opposite to A, there is a movable stud B of the screw. The other end of this screw has a head H. To the head, a hollow sleeve is attached. On the tapering of the sleeve 50 or 100 divisions are marked. This scale is called the Head Scale (HS). The Screw gauge head is having a ratchet arrangement, to avoid over-working of the screw.

## **Procedure:**

### **i) To find the least count (LC):**

Least count is the minimum measurement that can be made with the given instrument.

First, the pitch of the screw is determined; by giving a known number of rotations to the head scale and noting the distance advanced in the pitch scale in mm. Then the number of divisions in the head scale is also noted.

The pitch and least count are determined using the formulae.

### **ii) To find Zero Error (ZE) and Zero Correction (ZC):**

The screw gauge is checked to find whether there is any initial (zero) error in the instrument. If there is any initial error, suitable correction is to be made.

When the studs A and B touch each other, if the zero of the head scale lies on the same line as that of the pitch scale index line (I.L), the instrument has no error.

If the zero of the head scale is above the index line, it has negative error. So the zero correction is positive.

If the zero of the head scale is below the index line, it has positive error. So the zero correction is negative.

The type of error (ZE) and the suitable zero correction (ZC) for the given micrometer is determined with the help of figures and formulae given.

### **iii) To find the thickness of the glass plate:**

The given glass plate is placed gently in between the two studs A and B and the ratchet is rotated till the glass plate is firmly but gently gripped. Note the number of completed divisions in mm on the pitch scale as Pitch scale reading (PSR) and the divisions on the head scale, which coincides with the index line as head scale coincidence (HSC). The PSR and HSC are entered in the tabular column. Then the head scale reading (HSR), observed reading (OR) and correct reading (CR) are calculated. The procedure is repeated for different positions of the glass plate and the average thickness of the glass plate is calculated.

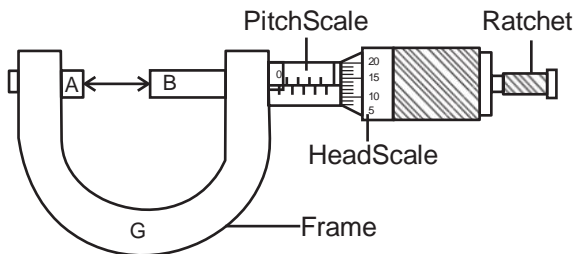
**iv) To measure the area of the glass plate:**

The given irregular glass plate is placed on a graph sheet. The trace of the glass plate is drawn on the graph sheet. The number of squares inside the trace in  $\text{mm}^2$  is counted. This gives the area of the glass plate in  $\text{mm}^2$ .

By multiplying the thickness of the glass plate and the area of the glass plate, the volume of the glass plate is determined.

**Result:**

The thickness of the given glass plate =  $\times 10^{-3}$  m  
 The area of the glass plate =  $\times 10^{-6}$   $\text{mm}^2$   
 The volume of the glass plate =  $\times 10^{-9}$   $\text{mm}^3$



**Observations:**

(i) Least Count:

$$\text{Least count} = \frac{\text{Pitch of the screw}}{\text{No. of head scale division}}$$

$$\text{Pitch of the screw} = \frac{\text{Distance moved in the pitch scale}}{\text{No. of rotations given to HS}}$$

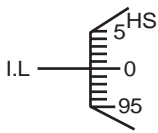
$$= \text{mm}$$

$$\text{Number of head scale divisions (HSD)} = \frac{\text{div}}{\text{pitch}}$$

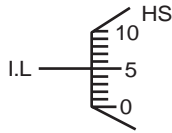
$$\text{Therefore, Least Count (LC)} = \frac{\text{pitch}}{\text{No. of HSD}} = \text{mm}$$

$$\text{LC} = \text{mm}$$

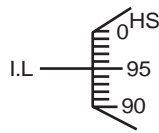
Three possible initial error and initial correction:



No error  
No correction



Positive error  
Negative correction  
HSC=5 div  
ZE=+(HSCxLC)  
=+(5x0.01mm)  
=+0.05mm  
ZE=+0.05mm  
ZC=-0.05mm



Negative error  
Positive correction  
HSC=95 div  
ZE=-(100-HSC)xLC  
=-(100-95)x0.01mm  
=-5x0.01mm  
ZE=-0.05mm  
ZC=+0.05mm

(ii) Zero error and zero correction for the given screw gauge:

**Tabular Column:**

To find the thickness of the given glass plate:

LC= mm ZE= mm ZC= mm

Sl. No.	PSR	HSC	HSR=HSCxLC	OR = PSR + HSR	Thickness CR=OR+ZC
Unit	mm	div	mm	mm	mm
1.					
2.					
3.					
4.					
5.					

The average thickness of the given glass plate = mm

=  $\times 10^{-3}$  m

Area of the glass plate using graph sheet = mm<sup>2</sup>

=  $\times 10^{-6}$  m<sup>2</sup>

Volume of the glass plate = Thickness x Area

=  
= m<sup>3</sup>

# VERNIERCALIPERS

**ExperimentNo:**

**Date:**

**Aim:**

To measure the length and diameter of the given solid cylinder using vernier calipers and to calculate the volume of the solid cylinder.

**Apparatusrequired:**

Vernier calipers and solid cylinder.

**Formula:**

$$\text{Volume of the solid cylinder} = \pi r^2 l$$

Where

l = length of the cylinder

r = radius of the cylinder

$$\begin{aligned} \text{Length/diameter of the solid cylinder} &= \text{Observed reading} + \text{Zero correction} \\ &= \{ \text{MSR} + (\text{VSC} \times \text{LC}) \} + \text{ZC} \end{aligned}$$

Where MSR = Main Scale Reading

VSC = Vernier Scale Coincidence

LC = Least Count

ZC = Zero Correction

$$\text{Vernier Scale Reading (VSR)} = \text{Vernier Scale Coincidence (VSC)} \times \text{Least Count (LC)}$$

$$\text{Observed Reading (OR)} = \text{Main Scale Reading (MSR)} + \text{Vernier Scale Reading (VSR)}$$

$$\text{Correct Reading (CR)} = \text{Observed Reading (OR)} + \text{Zero Correction (ZC)}$$

**Description:**

The vernier calipers consists of main scale (MS) graduated in cm along the lower edge of the frame and in inches along the upper edge of the frame. To one end of the frame, two fixed jaws are provided. There are two movable jaws, which slide over the main scale, and it can be fixed at any point using a screw S. The movable jaw attachment carries a vernier scale (VS).



## **Procedure:**

### **i) To find the Least Count:**

Least Count is the minimum measurement that can be made with the given instrument.

The value of 1 main scale division (MSD) and number of divisions (n) on the vernier scale are determined. The least count of the instrument is calculated using the formula

$$\text{Least Count (LC)} = \frac{1}{n} \times \text{value of 1 MSD}$$

### **ii) To find the Zero Error (Z.E) and Zero Correction (Z.C):**

The instrument is checked to find whether there is any initial error. If there is any initial error, suitable correction is to be made.

When the two jaws are in contact, the graduations are made such that the zero of the MS coincides with the zero of the VS. Now the instrument has no error.

If the zero of the VS is on the right side of the zero of the MS, it has positive error. Then the zero correction is negative.

If the zero of the VS is on the left side of the zero of the MS, it has negative error. Then zero correction is positive.

In each case, note the division of the vernier scale which is coinciding with any one of the main scale divisions. This is vernier scale coincidence (VSC). Applying this in the given formula, the value of zero error is determined. Hence zero correction (ZC) is also determined.

### **iii) To find the length and diameter of the solid cylinder:**

The given cylinder is gently placed in between the two lower jaws of the vernier calipers such that the length of the cylinder is parallel to the scale.

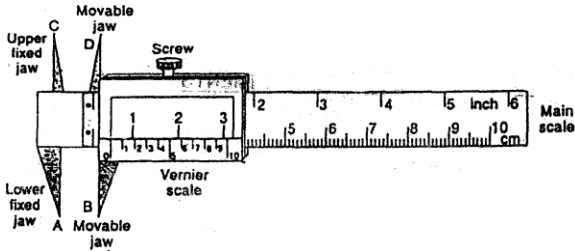
The completed main scale reading (MSR) is taken by noting the position of zero of the vernier scale. The vernier scale coincidence (VSC) is then noted. VSC is that particular division which coincides with any one of the main scale divisions, making a straight line.

The MSR and VSC are noted in the tabular column I for different settings of the cylinder. Then the Vernier Scale Reading (VSR), Observed Reading (OR) and Correct Reading (CR) are calculated using the formula given. The average value of CR is determined, which gives the length of the given cylinder.

Similarly the readings for diameter of the solid cylinder are taken by placing the cylinder suitably in between the lower fixed jaws and the readings are noted in the tabular column II.

**Result:**

The length of the solid cylinder  $l = x \times 10^{-2} \text{ m}$   
 The diameter of the solid cylinder  $d = x \times 10^{-2} \text{ m}$   
 The volume of the solid cylinder  $v = x \times 10^{-6} \text{ m}^3$



**Observations:**

(i) Least count:

Value of 10 MSD = cm

Value of 1 MSD = cm

No. of division on the vernier scale (n) =

(ii) The three possible initial error and initial correction:

$$\text{Least Count (LC)} = \frac{1}{n} \times \text{value of 1 MSD} = \text{cm}$$

No error

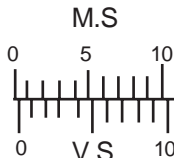
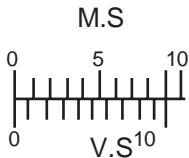
Positive error

Negative error

No correction

Negative correction

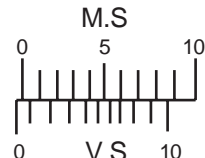
Positive correction



VSC = 3 div

$$\begin{aligned} & \times \text{LC} \\ & = +3 \times 0.01 \text{ cm} \\ & = +0.03 \text{ cm} \end{aligned}$$

$$\text{ZC} = -0.03 \text{ cm}$$



$$\text{VSC} = 7 \text{ div } \text{ZE} = +(\text{VSC})$$

$$\begin{aligned} \text{ZE} &= -(10 - \text{VSC}) \times \text{LC} \\ &= -(10 - 7) \times 0.01 \text{ cm} \\ &= -0.03 \text{ cm} \end{aligned}$$

$$\text{ZC} = +0.03 \text{ cm}$$

Zero error and zero correction of the given vernier calipers:

**Tabular column I:**

To find the length of the solid cylinder

LC =      cm      ZE =      cm      ZC =      cm

Sl. No	MSR	VSC	VSR = VSC x LC	OR = MSR + VSR	Length CR = OR + ZC
Unit	cm	div	cm	cm	cm
1					
2					
3					
4					
5					

The average length of the solid cylinder (L) =      cm  
 =       $\times 10^{-2}$  m

**Tabular column II:**

To find the diameter of the solid cylinder

LC =      cm      ZE =      cm      ZC =      cm

Sl. No	MSR	VSC	VSR = VSC x LC	OR = MSR + VSR	Diameter CR = OR + ZC
Unit	cm	div	cm	cm	cm
1					
2					
3					
4					
5					

The average diameter of the solid cylinder (D) =      cm  
 =       $\times 10^{-2}$  m

The average radius of the solid cylinder 'r' =       $\times 10^{-2}$  m

Volume of the solid cylinder =       $\pi r^2 l$   
 =       $\text{m}^3$

## CONCURRENT FORCES

**Experiment No:**

**Date:**

**Aim:**

To verify i) Parallelogram law of forces and ii) Lami's theorem.

**Apparatus required:**

Vertical drawing board, two pulleys with clamps, three sets of slotted weights, thread, white paper, scale, protractor, compass and pencil.

**Formula:**

1) To verify the parallelogram law of forces, it is to be shown that

Resultant = Equilibrant and they are opposite in directions.

i.e.  $OD = OC$  and  $\angle COD = 180^\circ$

where OD and OC are lines drawn representing the resultant and equilibrant of P and Q.

2) To verify Lami's theorem, it is to be shown that

$$\frac{P}{\sin \alpha} = \frac{Q \sin \beta}{\sin \alpha} = \frac{R}{\sin \gamma}$$

where P, Q and R are the forces,  $\alpha$ ,  $\beta$  and  $\gamma$  are the angles opposite to P, Q and R respectively.

**Description:**

A drawing board is supported vertically on a wall. The two smooth pulleys are fixed at the two top corners of the board, so that they can rotate freely. The two weight hangers P and Q are attached to the two ends of a long thread, after passing through the pulleys. The third weight hanger R is attached to the middle of the thread as shown in the figure.

**Procedure:**

The weights in the hangers P, Q and R are adjusted such that the point O is at rest. A small displacement is given to the system so as to overcome the friction in the pulleys. A drawing paper is fixed on the board, behind the strings. The trace of the three strings is taken on the sheet of paper, using an adjustable lamp arrangement in front of the threads. The trace sheet is removed and the lines are joined to get the image of the threads. These lines meet at a point O.

**i) To verify parallelogram law of forces:**

Taking a suitable scale (0.050 kg = 1 cm) OA, OB and OC are marked to represent the forces P, Q and R respectively both in magnitude and in direction. With OA and OB as adjacent sides, a parallelogram OADB is completed. The length of the diagonal OD is measured and entered in the tabular column. Also the  $\angle COD$  is measured and entered in the tabular column.

According to parallelogram law, the diagonal OD gives the resultant of P and Q, whereas OC is the equilibrant of P and Q.

The procedure is repeated thrice with different combinations of P, Q and R.

**ii) To verify Lami's theorem:**

In all the three figures drawn in the traces sheet, the angles between the forces are marked. The angle between the forces Q and R is marked as  $\alpha$ . The angle between the forces R and P is marked as  $\beta$ . The angle between the forces P and Q is marked as  $\gamma$  as shown in the figure. The angles are measured and entered in the corresponding tabular column. To check the accuracy of angle measurement, for each case, the sum of the three angles measured should be equal to  $360^\circ$ . i.e.  $(\alpha + \beta + \gamma) = 360^\circ$ .

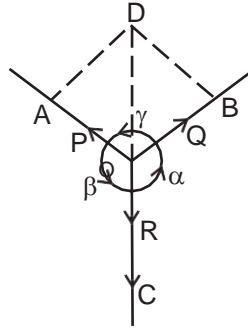
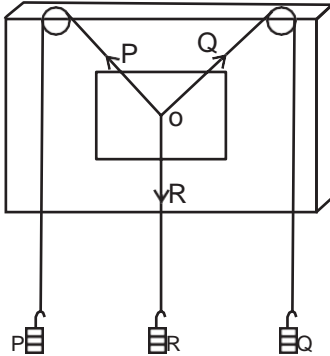
Then the calculations are done using the given formula and the values are entered in the tabular column.

**Result:**

1) In the tabular column I, in all cases, it is seen that Resultant = Equilibrant i.e.  $OD = OC$  and also  $\angle COD = 180^\circ$ , which verifies parallelogram law of forces.

2) In the tabular column II, in all cases, it is seen that the values in the last three columns are found to be equal. which verifies Lami's theorem.

$$\text{i.e. } \frac{P}{\sin \alpha} = \frac{Q}{\sin \beta} = \frac{R}{\sin \gamma}$$



**TabularColumnI:**

To verify parallelogram law of forces

Scale: 0.05kg.wt = 1cm

SI No.	Force P	Line OA	Force Q	Line OB	Force R	Equilibrant OC	Resultant OD	∠COD
Unit	kg. wt	cm	kg. wt	cm	kg. wt	cm	cm	degree
1.								
2.								
3.								

**TabularColumnII:**

To verify Lami's Theorem

SI No.	Forces			Angles			$\frac{P}{\sin \alpha}$	$\frac{Q}{\sin \beta}$	$\frac{R}{\sin \gamma}$
	P	Q	R	$\alpha$	$\beta$	$\gamma$			
Unit	kg. wt	kg. wt	kg. wt	degree	degree	degree	kg. wt	kg. wt	kg. wt
1.									
2.									

3.									
----	--	--	--	--	--	--	--	--	--

**Calculation:**

1.  $P = \text{kg.wt}, \alpha = \circ$        $Q = \text{kg.wt}, \beta = \circ$        $R = \text{kg.wt}, \gamma = \circ$

$$\frac{P}{\sin \alpha}$$

$$\frac{Q}{\sin \beta}$$

$$\frac{R}{\sin \gamma}$$

2.  $P = \text{kg.wt}, \alpha = \circ$        $Q = \text{kg.wt}, \beta = \circ$        $R = \text{kg.wt}, \gamma = \circ$

$$\frac{P}{\sin \alpha}$$

$$\frac{Q}{\sin \beta}$$

$$\frac{R}{\sin \gamma}$$

3.  $P = \text{kg.wt}, \alpha = \circ$        $Q = \text{kg.wt}, \beta = \circ$        $R = \text{kg.wt}, \gamma = \circ$

$$\frac{P}{\sin \alpha}$$

$$\frac{Q}{\sin \beta}$$

$$\frac{R}{\sin \gamma}$$



## POISEUILLE'S METHOD

**ExperimentNo.:**

**Date:**

**Aim:**

To determine the coefficient of viscosity of water by capillary flow method, using graduated burette without stopper.

**Apparatusrequired:**

Burette stand, graduated burette without stopper, Capillary tube, rubber tube, beaker, stopwatch, water, funnel, wooden bench, etc.

**Formula:**

The coefficient of viscosity of the water

$$\eta = \frac{\pi \rho g r^4 H t}{8 l v}$$

Where

$\rho$  = density of water

$g$  = acceleration due to gravity

$r$  = radius of the capillary tube

$l$  = length of the capillary tube

$v$  = volume of water comes out of the capillary flow (5 cc)

$t$  = time taken for 5 cc flow of water

$H$  = the average pressure head height of the liquid levels from the capillary tube.

$$H = \frac{h_1 + h_2}{2} - h_3$$

$h_1$  = height of initial level from the experimental table

$h_2$  = height of final level from the experimental table

$h_3$  = height of the axis of capillary tube from the table

**Description:**

The graduated burette without stopper is mounted vertically in the stand. A rubber tube is connected to the bottom of the burette. To the other end of the tube, a capillary tube is inserted and placed in a perfectly horizontal position.

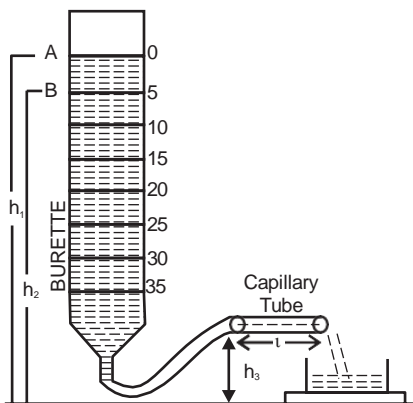
**Procedure:**

Pure water is poured into the burette using a funnel. The height of the liquid level in the burette is adjusted such that there is a streamline flow of water through the capillary tube i.e. 5 to 8 drops per minute. This can be achieved by raising or lowering the capillary tube. When the water level in the burette comes to zero mark a stop watch is started and the time for the water level to reach 5cc, 10cc, 15cc, 20cc and 25cc mark is noted and entered in the tabular column I.

From the above readings, the time of flow of 5cc water (t) at various ranges like 0cc to 5cc, 5cc to 10cc, 10cc to 15cc, 15cc to 20cc and 20cc to 25cc are tabulated in tabular column II. In each case the heights of the respective initial levels and final levels from the experimental table are noted. The average pressure head height is calculated and entered in the tabular column. Substituting the average value of Ht values in the formula given, the coefficient of viscosity of water is calculated.

**Result:**

The coefficient of viscosity of water =  $\text{Nsm}^{-2}$



**Observations:**

Density of water	=	$\text{kgm}^{-3}$
Acceleration due to gravity	=	$\text{ms}^{-2}$
Radius of the capillary tube	=	m
Length of the capillary tube	=	m
Volume of water comes out of the cap. flow	=	$5\text{cc} = 5 \times 10^{-6} \text{m}^3$

**TabularcolumnI:**

To notethetimetakenforthe flowofwater

Sl.No	WaterlevelIntheburette	Timeofflow
	cc	s
1	0	
2	5	
3	10	
4	15	
5	20	
6	25	

**TabularcolumnII:**

Heightoftheaxisofcap.tubefromthetable $h_3 =$  m

Sl. No	Initial water level	Final Water level	Heightof Initial level $h_1$	Height of final Level $h_2$	Ave. Pr. Headheight $\frac{h_1+h_2}{2} - h_3$	Time of Flow t	H.t
	cc	cc	m	m	m	s	ms

Mean

**Calculation:**

The coefficientofviscosityofwater

$$\eta = \frac{\pi \rho g r^4 H t}{8 l v}$$

# COEFFICIENT OF VISCOSITY OF HIGH VISCIOUS LIQUID – STOKES' METHOD

**Experiment No.:**

**Date:**

**Aim:**

To determine the coefficient of viscosity of high viscous liquid (Castor oil) by Stokes' method.

**Apparatus Required:**

A tall glass jar, castor oil, glass beads of different radii, stopwatch, screw gauge, etc.

**Formula:**

The coefficient of viscosity of the highly viscous liquid  $\eta = \frac{2(\rho - \sigma)g(r^2t)}{9h}$

Where

$\rho$  = the density of the material of the solid sphere (glass beads)

$\sigma$  = the density of the liquid (castor oil)

$g$  = the acceleration due to gravity

$r$  = the radius of the glass beads

$t$  = the time taken by the solid sphere to travel the marked distance  $h$ , inside the high viscous liquid.

$h$  = the distance between the marks B and C

**Description:**

The pure, transparent, high viscous liquid (castor oil) is taken in the tall glass jar. On the outer surface of the jar, two markings B and C are made at a distance of say 50 cm. The marking B should be well below the free surface of the liquid. Then only the sphere can attain terminal velocity when travelling from B to C.

**Procedure:**

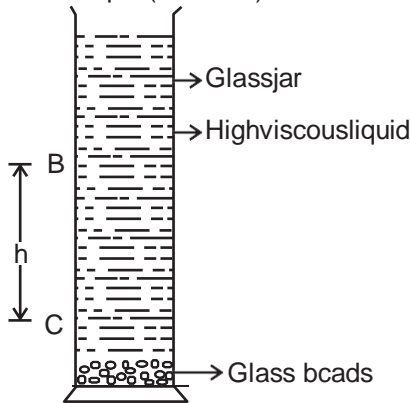
First the least count, zero error and zero correction of screw gauge are determined. Then using the screw gauge, the diameters of the given glass beads are measured. The radii of the beads are calculated. Values of  $\rho$  and  $\sigma$  are taken from standard tables.

The sphere is placed gently on the surface of the high viscous liquid and dropped. When it crosses the marking B, a stop clock is started and when it crosses the marking C, the clock is stopped. The time taken ( $t$ ) to travel by the sphere from B to C is recorded in the tabular column. The

distance between B and C is noted as 'h'. The same procedure is repeated for all the given spheres. Then the coefficient of viscosity of highly viscous liquid (castor oil) is calculated, using the formula.

**Result:**

The coefficient of viscosity of the high viscous liquid (castor oil) =  $\text{Nsm}^{-2}$



**Observations:**

The density of the material solid sphere (glass beads)  $\rho = \text{kgm}^{-3}$

The density of the viscous liquid (castor oil)  $\sigma = \text{kgm}^{-3}$

The acceleration due to gravity  $g = \text{ms}^{-2}$

The distance between B and C  $h = \times 10^{-2} \text{m}$

$$\text{Pitch of the screw} = \frac{\text{Distance moved in pitch scale}}{\text{No. of rotations given to HS}}$$

(I) Least Count:  $\text{Least Count} = \frac{\text{Pitch/No. of Head Scale Divisions}}{\text{mm} = \text{mm}}$

$$\text{Therefore, Least Count (LC)} = \frac{\text{pitch}}{\text{No. of HSD}} = \text{----- mm}$$

$$\text{Number of head scale divisions (HSD)} = \frac{\text{LC}}{\text{mm}} = \text{div}$$

(ii) Zero Error and Zero Correction:

**TabularcolumnI:**

To find the radii of the solid spheres (glass beads)

LC= \_\_\_\_\_ mm; ZE= \_\_\_\_\_ mm; ZC= \_\_\_\_\_ mm

Solid sphere	Sl. No.	PSR	HSC	HSR= (HSCxLC)	OR=PS R+HSR	CR= OR+ZC	Mean diameter d	Radius r= d/2
	Unit	mm	div	mm	mm	mm	mm	mm
Bead1	1.							
	2.							
	3.							
	4.							
	5.							
Bead2	1.							
	2.							
	3.							
	4.							
	5.							
Bead3	1.							
	2.							
	3.							
	4.							
	5.							

**TabularcolumnII:**

To find  $r^2t$

Glassbeads	Radius r	Radius <sup>2</sup> r <sup>2</sup>	Time taken t	r <sup>2</sup> t
Unit	mm	mm <sup>2</sup>		mm <sup>2</sup>
Bead1				
Bead2				
Bead3				

$$\begin{aligned} \text{The average of } r^2 &= \text{mm}^2 \\ &= \text{x } 10^{-6} \text{m}^2 \end{aligned}$$

Calculation:

The coefficient of viscosity of the highly viscous liquid

$$\begin{aligned} (\text{castor oil}) \eta &= \frac{2(\rho - \sigma)g(r^2 t)}{h} \\ &= \\ &= \text{Nsm}^{-2} \end{aligned}$$

## SURFACETENSION

**Experiment No.:**

**Date:**

**Aim:**

To determine the surface tension of water by capillary rise method.

**Apparatus required:**

Beaker with water, capillary tube, pointer, travelling microscope, retort stand with clamp, two hole rubber cork, wooden bench, etc.

**Formula:**

$$\text{The surface tension of water } T = \frac{hr\rho g}{2}$$

where

h = height of the liquid rise in the capillary tube

r = radius of the capillary tube

$\rho$  = density of water

g = acceleration due to gravity

**Description:**

A capillary tube of uniform cross-sectional area is cleaned well and clamped vertically with one end inside the water taken in a beaker. Due to surface tension, water rises to a definite height into the tube. This is called capillary rise. If there is any air bubble inside the tube, it should be removed. A pointer is mounted vertically so that its lower end just touches the surface of water in the beaker.

**Procedure:**

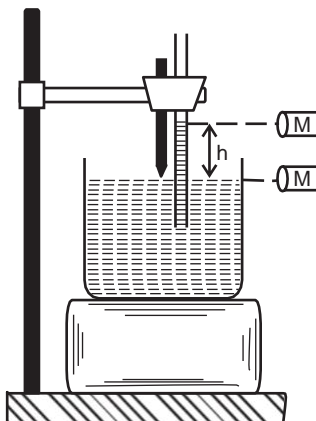
The least count of the microscope is determined as in the case of vernier calipers.

The microscope is adjusted such that the horizontal cross wire is tangential to the lower meniscus of the liquid inside the capillary tube. The MSR and VSC from the vertical scale of the microscope are noted. Then the beaker of water is removed, without disturbing the pointer. Now the microscope is focussed so that the horizontal cross wire is tangential to the tip of the pointer. The MSR and VSC readings are noted in the tabular column. The difference between these two observed readings (O.R) gives the height of capillary rise 'h'.

The capillary tube is mounted horizontally and the microscope is focussed to one end of the tube, so that we can see the bore of the capillary tube in the field of view.

Now, the horizontal cross wire is made tangential to one side of the bore (AA'). The readings are noted in the tabular column from the vertical scale. Using the minute adjustment screw of the microscope, the horizontal cross wire is made tangential to the other side of the bore (BB') and the readings are noted. The difference between these two observed readings (O.R) gives the diameter of the capillary tube. Hence the radius of the capillary tube is determined.

The surface tension of water is calculated using the formula.

**Result:**



The surface tension of water is found to be =

$\text{Nm}^{-1}$

**Observations:**

Density of water =  $\text{kgm}^{-3}$   
 Acceleration due to gravity  $g$  =  $\text{m}^{-2}$

**To find Least count of microscope:**

Least count:

Value of 20 MSD = cm

value of 1 MSD = cm

$$\text{Least Count (LC)} = \frac{1}{n} \times \text{value of 1 MSD}$$

No. of divisions on the vernier scale (n) =

=

= cm

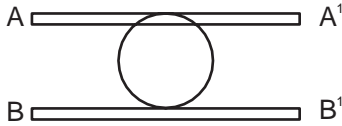
**Tabular column:**

To find the height of capillary rise  $h$

Position	Microscope Readings				Height of the capillary rise $h = (x - y)$
	MSR	VSC	VSR = VSC x LC	OR = MSR + VSR	
Unit	cm	div	cm	cm	cm
Horizontal cross wire at the lower meniscus				$x =$	
Horizontal cross wire at the tip of the pointer				$y =$	

Least count = cm



Height of the liquid rise in the capillary tube (h) = cm  
 =  $\times 10^{-2} \text{m}$



**TabularcolumnII**

To find the radius of capillary tube:

Least count = \_\_\_\_\_ cm

Position	Microscope Readings				Diameter 2r=(A~B)
	MSR	VSC	VSR=VSCxLC	OR=MSR+VSR	
Unit	cm	div	cm	cm	cm
 AA'				A =	
 BB'				B =	

The diameter of the capillary tube  $d =$  \_\_\_\_\_ cm  
 The radius of the capillary tube  $=d/2 =$  \_\_\_\_\_ cm  
 $r =$  \_\_\_\_\_  $\times 10^{-2}$  m

**Calculation:**

The surface tension of water  $T = \frac{hr\rho g}{2}$

Height of the liquid rise in the capillary tube  $h =$  \_\_\_\_\_  $\times 10^{-2}$  m

Radius of the capillary tube  $r =$  \_\_\_\_\_  $\times 10^{-2}$  m

Density of the water  $\rho =$  \_\_\_\_\_  $\text{kgm}^{-3}$

$T = \frac{hr\rho g}{2} =$  \_\_\_\_\_  $\text{Nm}^{-1}$

## SONOMETER

**ExperimentNo.:**

**Date:**

**Aim:**

To find the frequency of the given tuning fork using a sonometer.

**Apparatus Required:**

$$\text{The frequency of the tuning fork } n = \frac{1}{2} \sqrt{\frac{(M)g}{(l^2)m}}$$

Sonometer, screw gauge, tuning fork, hammer, weight hanger set and paper rider.

**Formula:**

Where

M = total mass at the end of the sonometer wire.

i.e. Mass of the weight hanger + Mass placed on it

g = acceleration due to gravity.

m = linear density of the wire

i.e., mass per unit length of the wire.  $l =$  first

resonating length.

The linear density of the wire,  $m = \pi r^2 \rho$

Where

$\rho$  = density of the material of the wire.

r = radius of the sonometer wire.

**Description:**

A sonometer consists of a hollow wooden box about one metre with bridges A and B fixed at two ends on the top surface. A uniform wire is stretched over the two bridges, one end of the wire being fixed to the nail at one end of the box. The other end passes over a smooth pulley and carries a weight hanger on which weights can be placed. M and N are two movable bridges in between the fixed bridges A and B. By changing their position, the length of the vibrating segment of the wire can be altered.

**Procedure:**

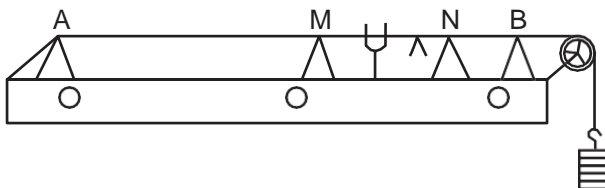
Asuitable load say 0.5 kg is placed on the hanger. The two movable

bridges M and N are placed close to each other. A light paper rider is placed at the centre of the segment MN of the sonometer wire. The given tuning fork is excited and its stem is pressed on the sonometer box. The length of the vibrating segment MN is gradually increased by moving the bridge M or N, till the paper rider flutters violently and falls down. The first resonating length  $l$  between M and N is measured using the metre scale fixed on the sonometer box. The load is changed to 1 kg, 1.5 kg, 2 kg, 2.5 kg, 3 kg etc and in each case the resonating length is determined and  $M/l^2$  is calculated.

The diameter of the wire is found using a screw gauge. The radius of the wire is calculated. The mass per unit length of the wire is  $m = \pi r^2 \rho$ . Here  $\rho$  is the density of the material of the wire.

The frequency of the given tuning fork is calculated by substituting the values of  $M/l^2$  and  $m$  in the formula.

**Result:**



The frequency of the given tuning fork (n) = \_\_\_\_\_ Hz.

**Observations:**

Acceleration due to gravity  $g =$  \_\_\_\_\_  $\text{ms}^{-2}$   
 Density of the material of the wire (steel)  $\rho =$  \_\_\_\_\_  $\text{kgm}^{-3}$

Pitch of the screw =  $\frac{\text{Distance moved in Pitch scale}}{\text{No. of rotations given to HS}}$

(i) Least Count:  $\text{Least Count} = \text{Pitch} / \text{No. of Head Scale Divisions}$

Therefore,  $\text{Least Count (LC)} = \frac{\text{pitch}}{\text{No. of HSD}} =$  \_\_\_\_\_ mm

Number of head scale divisions (HSD) = \_\_\_\_\_ mm = \_\_\_\_\_ mm  
 LC = \_\_\_\_\_ div = \_\_\_\_\_ mm

(ii) Zero Error and Zero Correction:

**Tabular column I:**

To find  $\frac{M}{\ell^2}$

Sl.No.	Mass placed at the end of the sonometer wire [mass of the hanger + mass placed on the hanger] M		First resonating length $\ell$		$\ell^2$	$\frac{M}{\ell^2}$
	Unit	Kg	cm	m		
1.	Hanger+0.5	1.0				
2.	Hanger+1.0	1.5				
3.	Hanger+1.5	2.0				
4.	Hanger+2.0	2.5				
5.	Hanger+2.5	3.0				

**Tabular Column II:**

To find the diameter of the wire

LC =                  mm      ZE =                  mm      ZC =                  mm

Sl.No.	PSR	HSC	HSR = HSC X LC	OR = PSR + HSR	CR = OR + ZC
Unit	mm	div	mm	mm	mm
1.					
2.					
3.					
4.					
5.					

Mean

The averaged diameter of the wire =                  mm

Radius of the wire  $r = \frac{d}{2}$  =                  mm

R =                   $\times 10^{-3}$  m

**Calculation:**

The Mass per unit length of the wire  $m = \pi r^2 \rho$

$$=$$

$$= \text{kgm}^{-1}$$

The frequency of the tuning fork  $n = \frac{1}{2} \sqrt{\frac{Mg}{l^2 m}}$

$$=$$

$$= \text{Hz}$$

## DEFLECTION MAGNETOMETER (TAN A – POSITION) COMPARISON OF MAGNETIC MOMENTS

**Experiment No.:****Date:****Aim:**

To compare the magnetic moments of the given two bar magnets using a deflection magnetometer in Tan A position, by equal distance method.

**Apparatus Required:**

$$\frac{m_I}{m_{II}} = \frac{\tan \theta_I t}{a n \theta_{II}}$$

Deflection magnetometer, two bar magnets, metre scale, etc.

**Formula:**

The ratio of the magnetic moments of two given short bar magnets

where

$m_I$  = the magnetic moment of first magnet

$m_{II}$  = the magnetic moment of second magnet

$\theta_I$  = mean deflection for first magnet

$\theta_{II}$  = mean deflection for second magnet

**Description:**

The magnetometer is arranged in a Tan A position. For arranging in a Tan A position the following adjustments are to be made.

1) The magnetometer is placed on a horizontal table and the arms of the

magnetometer are kept along the east-west direction and it is adjusted so that the arms are parallel to the length of the aluminium pointer.

- ii) Without disturbing the adjustment, the compass box is slowly rotated until the pointer reads  $0^\circ - 0^\circ$ .

**Procedure:**

Equal distance method:

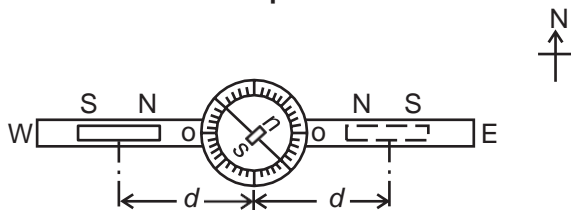
The magnetometer is arranged in tan A position. One of the two given bar magnets of magnetic moment  $m_1$  is placed on one side (Eastern arm) of the two arms with its axis along the east-west direction in the magnetometer. The distance of the magnet is suitably adjusted so that the deflection lies between  $30^\circ$  and  $60^\circ$ . The distance between the centre of the magnet and the centre of the magnetometer is measured as 'd'. The readings of both the ends of the pointer are taken as  $\theta_1$  and  $\theta_2$ . The magnet is reversed from pole to pole, in the same place and two more readings  $\theta_3$  and  $\theta_4$  are taken.

The magnet is then kept at the same distance on the other side of the arm (Western arm) and four deflections  $\theta_5, \theta_6, \theta_7$  and  $\theta_8$  are noted, as mentioned above. The average of the eight deflections is found and taken as  $\theta_1$ . The same procedure is repeated with the second magnet of magnetic moment  $m_2$ , keeping at the same distance and the average deflection  $\theta_2$  is found. The experiment is repeated for different values of d and the readings are tabulated. The ratio of magnetic moments ( $m_1/m_2$ ) is calculated using the formula.

**Result:**

- i) The magnetic moments of the given two bar magnets are compared.
- ii) The ratio of the magnetic moments of two given bar magnets by equal distance method =

**Tan A position:**



**Equal distancemethod.**

To find  $\frac{m_I}{m_{II}}$  by equal distancemethod

Distance d	Magnetic Moment	Deflections in degrees										$\frac{m_I \tan \theta_I}{m_{II} \tan \theta_{II}}$
		Easternarm				Westernarm				Mean $\theta$		
		Direct		Reversed		Direct		Reversed				
		$\theta_1$	$\theta_2$	$\theta_3$	$\theta_4$	$\theta_5$	$\theta_6$	$\theta_7$	$\theta_8$			
	$m_I$										$\theta =$	
	$m_{II}$										$\theta =$	
	$m_I$										$\theta =$	
	$m_{II}$										$\theta =$	
	$m_I$										$\theta =$	
	$m_{II}$										$\theta =$	

Average of  $\frac{m_I}{m_{II}}$

**Calculation:**

a.)  $d =$  cm

$$\frac{m_I \tan \theta_I}{m_{II} \tan \theta_{II}}$$

b.)  $d =$  cm

$$\frac{m_I \tan \theta_I}{m_{II} \tan \theta_{II}}$$

c.)  $d =$  cm

$$\frac{m_I \tan \theta_I}{m_{II} \tan \theta_{II}}$$

d.)  $d =$  cm

$$\frac{m_I \tan \theta_I}{m_{II} \tan \theta_{II}}$$



## MODEL QUESTION PAPER

1. Measure the thickness of the given irregular glass plate using micrometer. Determine the area of the glass plate using a graph sheet and calculate the volume of the glass plate.
2. Measure the length and diameter of the given solid cylinder using vernier calipers and then calculate the volume of the solid cylinder.
3. Verify the parallelogram law of forces and Lami's theorem using concurrent forces.
4. Determine the coefficient of viscosity of water by capillary flow method, using a graduated burette.
5. Determine the coefficient of viscosity of a high viscous liquid by Stokes' method.
6. Determine the surface tension of water by capillary rise method.
7. Determine the frequency of the given tuning fork using a sonometer.
8. Compare the magnetic moments of the two bar magnets using deflection magnetometer in Tan-A position, by equal distance method.

# SEMESTER II

## UNIT I

### HEAT

#### TRANSFER OF HEAT

##### **Introduction:**

Heat is a form of energy, which is responsible for the hot condition or the cold condition of a body. Since it is a form of energy, the quantity of heat is measured in 'joule' with symbol 'J'.

We can realize the hotness or coldness of a body by our sense of touch. But there should be a method to determine the degree of hotness or coldness.

Temperature is a measure of the relative degree of hotness or coldness of the body. For quantitative measurement of temperature some standard temperatures are required, which can be easily reproduced in any laboratory. Generally, we take the melting point of ice as lower fixed point and the boiling point of water (steam point) as upper fixed point.

To measure the temperature, different scales are in use.

##### **1. Centigrade scale :**

In this scale, the temperature is measured in 'degree Celsius' with symbol  $^{\circ}\text{C}$ . In this scale, the lower fixed point is taken as  $0^{\circ}\text{C}$  for ice point and the upper fixed point is taken as  $100^{\circ}\text{C}$  for steam point.

##### **2. Fahrenheit scale :**

In this scale, the temperature is measured in 'degree Fahrenheit' with symbol  $^{\circ}\text{F}$ . Here, the lower fixed point (ice point) is taken as  $32^{\circ}\text{F}$  and the upper fixed point (steam point) is taken as  $212^{\circ}\text{F}$ .

##### **3. Absolute scale or Kelvin scale:**

In this scale, the temperature is measured by the unit 'kelvin' with symbol 'K'. Usually, the temperatures are measured in centigrade scale in laboratories. To convert the temperature from centigrade scale to Kelvin scale, we have to add 273.15, because  $0\text{K}$  corresponds to  $-273.15^{\circ}\text{C}$ .

Ex: Theroomtemperature=32°C

$$=(32+273.15)K=305.15K$$

The relation between Celsius 'C', Kelvin 'K' and Fahrenheit 'F' is

$$\frac{C}{100} = \frac{F-32}{180} = \frac{K-273}{100}$$

### Heat Transfer:

There are three methods of transfer of heat from one place to another. They are Conduction, Convection and Radiation. Conduction takes place both in solids and fluids. Convection cannot occur in solids. Radiation does not require the presence of any medium.

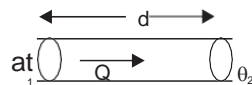
### Conduction:

In Conduction, the molecules of the body are responsible for the heat transfer. Here there is no actual movement of molecules from one place to another place. When a rod is heated at one end, the molecules at the hot end vibrate about their mean position and transfer the heat energy to the neighbouring molecules and thus the heat energy reaches the other end of the rod.

Conduction is the process by which heat is transmitted from one point to another without the actual movement of the particles from their equilibrium positions. Conduction takes place in solids, liquids and gases.

### Coefficient of thermal conductivity:

Let us consider a metallic rod of cross sectional area  $A$ . Let the two ends be separated by a distance  $d$ , maintained at temperatures  $\theta_1$  and  $\theta_2$ . Let  $\theta_1$  be greater than  $\theta_2$ . Heat flows from the  $\theta_1$  end the higher temperature to the end at the lower temperature.



When the steady state is reached, the quantity of heat  $Q$  conducted is

- i) directly proportional to the area of cross section  $A$
- ii) directly proportional to the difference in temperature between the ends ( $\theta_1 - \theta_2$ )
- iii) directly proportional to the time for which the heat is conducted ( $t$ )
- iv) and inversely proportional to the distance between the two ends.

$$\text{Hence } Q \propto A \frac{(\theta_1 - \theta_2)t}{d} \text{ or}$$

$$Q = \lambda A \frac{(\theta_1 - \theta_2)t}{d}$$

where  $\lambda$  is a constant known as the coefficient of thermal conductivity of the material and  $(\theta_1 - \theta_2)/d$  is known as temperature gradient.

$$\text{when } A=1, \quad \frac{(\theta_1 - \theta_2)}{d} = 1 \quad \text{and } t = 1, \quad \text{then } Q = \lambda$$

Hence, Coefficient of thermal conductivity ( $\lambda$ ) of the material of a conductor is defined as the quantity of heat conducted per second per unit area per unit temperature gradient at the steady state.

unit of  $\lambda$ :

$$Q = \lambda A \frac{(\theta_1 - \theta_2)t}{d} \quad \text{or}$$

$$\lambda = \frac{Qd}{A(\theta_1 - \theta_2)t}$$

Substituting the units,

$$\text{Unit of } \lambda = \frac{\text{Jm}}{\text{m}^2\text{Ks}} = \text{Js}^{-1}\text{m}^{-1}\text{K}^{-1} = \text{Wm}^{-1}\text{K}^{-1} \quad (\text{as Js}^{-1} \text{ is W})$$

**Selection of good and poor thermal conductors:-**

Substance	$\lambda(\text{Wm}^{-1}\text{K}^{-1})$	Substance	$\lambda(\text{Wm}^{-1}\text{K}^{-1})$
Silver	420	Water	0.6
Copper	390	Redbrick	0.6
Aluminium	240	Rubber	0.5
Steel	14	Wood	0.04 to 0.1
Lead	35	Glass	0.8
Concrete	1.1	Ebonite	0.2

We require both good and poor thermal conductors according to the need. Materials having high values of coefficient of thermal conductivity are termed as good thermal conductors and the materials having lower values of coefficient of thermal conductivity are called as poor thermal conductors or insulators. The presence of free electrons in metals helps the easy transfer of heat from one part of the metal to the other part. So, all metals are good thermal conductors of heat. Some important materials and their values of coefficient of thermal conductivity are listed above.

#### **Uses of thermal conductors:-**

1. Handles made of wood or ebonite are provided for cookers and hot water vessels.
2. Hot water bottles made of rubber are able to keep hot water at high temperature for a considerable period of time.
3. Use of double windows with a thin layer of air enclosed in between them keeps the room warm in cold countries.
4. Wool, cork and ebonite are used for the purpose of heat insulation in refrigeration.
5. Woolen clothes are used in winter to keep the body warm.
6. Sawdust and jute sheet is used to cover icetop to prevent it from melting.
7. Vessels made of copper, aluminum, etc., are used for cooking purpose as they easily conduct heat.
8. Copper is used in boilers and radiators, because of its good conductivity.

#### **Convection:-**

In the case of heat transfer by convection, there is actual movement or motion of molecules of the fluid (liquid or gas). The portions of the fluid that get warmed up by contact with the heat source, expand and so move up through the body of the fluid due to the decrease in density. There is an inflow of cooler molecules to take the place of heated mass of the fluid which has moved up. This circulatory motion of the fluid mass by which heat is transferred from place to place is called Convection.

Convection is the process in which heat is transmitted from one place to the other by the actual movement of heated particles.

Convection takes place only in liquids and gases. It cannot take place in solids.

### **Applications of Convection.**

1. The wind flow is due to the convection currents in the atmosphere. During day time, parts of earth get heated by the Sun. As the air expands, it rises up and its place is taken by the flow of air from colder areas.
2. The land breeze and sea breeze are due to the convection in the atmosphere. During day time, land mass is heated to a higher level than the sea. So, the warm air over the land rises giving place to the cool air flow from the ocean. This gives the sea breeze. During night time, the land mass cools quickly than the water in the sea. So, cool air flows from land mass towards sea which gives the land breeze.

### **Radiation:-**

For heat transfer by radiation, no medium is required. The heat from the Sun is received by earth only by radiation. Both heat and light travel through space with same velocity and they are similar in nature. Both are generated and transmitted in the same way. They are transverse waves and electro-magnetic in character, but they differ in wavelength and frequency. The thermal radiations are detected in the "infra-red" region of the electro-magnetic spectrum.

Radiation is the process by which heat energy is transmitted from one place to another without the aid of an intervening material medium.

### **Properties of thermal radiation:**

The nature of thermal radiation is similar to that of light. Following are some of the properties of thermal radiation.

1. Thermal radiation travels with the velocity of light, which is  $3 \times 10^8 \text{ ms}^{-1}$ .
2. Thermal radiation obeys the same laws of reflection, refraction etc., as light.
3. Thermal radiation travels through vacuum.
4. It obeys the law of inverse square, as light.
5. It travels in straight lines.
6. When thermal radiation falls on any body, which can absorb it, then converted into ordinary heat, which raises its temperature.
7. It is absorbed by dark rough surfaces and reflected by light smooth surface.

## **Applications of Radiation.**

1. White coloured dresses are used in hot countries to keep the inside cool.
2. In some countries, shining aluminium sheets are used to cover the roof of the house to reflect back the radiant heat and to keep the inside cool.
3. A cooking vessel is painted black at the bottom for greater absorption of heat, but polished at the top to minimize radiation losses.
4. In cold countries, hot air or water runs through the pipes along the walls inside a building and the radiant heat energy keeps the occupants warm by 'Central Heating'.

## **1.2 KINETIC THEORY OF GASES**

A gas consists of a large number of particles called molecules. The molecules are in the state of continuous motion. The pressure of a gas is due to the collision of individual molecules on the walls of the container. The behaviour of the gas can be explained on the basis of kinetic theory of gases. The kinetic theory of gases has found better applications to explain various properties of gases. This theory was developed by Clausius, Boltzmann and Maxwell.

### **1. Postulates of Kinetic Theory of gases:-**

1. A gas consists of a large number of tiny, rigid particles called molecules.
2. The gas molecules are identical in all respects, like mass, size, etc.
3. The molecules are perfectly elastic spheres.
4. The molecules are in random motion and travel with all possible velocities in all possible directions.
5. The molecules collide with each other and also with the walls of the container.
6. Between two successive collisions, the molecules travel in the straight line.
7. The average distance between two successive collisions is called the mean free path.
8. The time of collision is negligible when compared to the time taken by the molecule to travel its mean free path.

9. The volume of the molecules is negligible when compared to the volume of the gas.
10. The force of attraction between the molecules is negligible.
11. As the molecules of a gas are always in random motion, the energy of the gas is in the form of kinetic energy and it is directly proportional to the absolute temperature of the gas. When the temperature of the gas remains constant, the molecular velocities will not change.

**Mean square velocity and Root Mean Square velocity (R.M.S. velocity):-**

Let  $c_1, c_2, c_3, \dots, c_n$  be the velocities of 1<sup>st</sup>, 2<sup>nd</sup>, 3<sup>rd</sup>, ..., n<sup>th</sup> molecule present in the gas, then the mean square velocity is,

$$\overline{C^2} = \frac{c_1^2 + c_2^2 + c_3^2 + \dots + c_n^2}{n}$$

The mean square velocity is defined as the average of the squares of velocities of all the molecules.

Now R.M.S. velocity is obtained by taking square root of the mean square velocity of the molecules. R.M.S. velocity is

$$\sqrt{\overline{C^2}} = \sqrt{\frac{c_1^2 + c_2^2 + c_3^2 + \dots + c_n^2}{n}}$$

The R.M.S. velocity is defined as the square root of the average of squares of velocities of all molecules

**Expression for the pressure of a gas:-**

Consider a cubical vessel of side one metre containing a perfect gas. The gas contains  $n$ -number of molecules and  $m$  be the mass of each molecule. The molecules are in continuous motion and they collide with the wall of the container. The force experienced by the wall per unit area, due to the molecular collision gives the pressure of the gas.

$$\text{Pressure} = \frac{\text{Force}}{\text{Area}}$$

Consider a single molecule moving in the X-direction with velocity  $v$ . It collides with the wall BCGF of the container and then it rebounds with velocity  $-v$ .

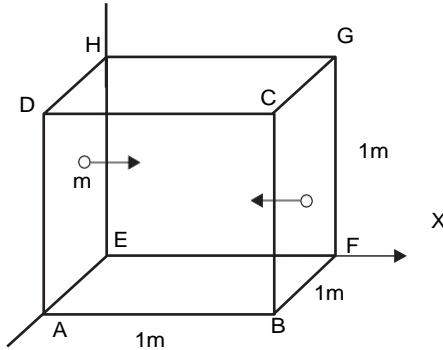


The momentum of the molecule before collision =  $mv$

The momentum of the molecule after collision =  $-mv$

$$\therefore \text{The change in momentum for single molecule} = mv - (-mv) = 2mv.$$

After recoil from the wall BCGF, the molecule travels towards the wall AEHD and collide then rebound back to the wall BCGF. So, the molecule travels  $2m_e \cdot t$  for every next collision on the same wall.



$$\left. \begin{array}{l} \text{Time required to travel} \\ 2m \text{ for next collision} \end{array} \right\} t = \text{distance} / \text{velocity} = 2/v \text{ second}$$

The rate of change of momentum = change in momentum / time taken

$$= \frac{2mv}{2/v} = mv^2$$

According to Newton's second law of motion,

Force = Rate of change in momentum.

$$\therefore \text{Force acting on the wall due to 1 molecule} = mv^2$$

An average  $n/3$  molecules can be considered to move in each of the three directions X, Y and Z.

$$\text{The force acting on the wall due to } n/3 \text{ molecules} = n/3(mv^2)$$

Pressure on the wall = Force / Area

$$= \frac{(n/3)mv^2}{1}$$

{Since the area of any side  $A = 1 \times 1 = 1 \text{ m}^2$ }

Since the pressure will be the same on all sides,

Pressure of the gas  $P = \left(\frac{1}{3}\right) nmv^2$

But all the molecules are not moving with the same velocity  $v$ .

Let  $c_1, c_2, c_3, \dots, c_n$  be the velocities of the molecules,

The mean square velocity  $C^2 = \frac{(C^2 + C_1^2 + C_2^2 + \dots + C_n^2)}{n}$

Substituting this average square of the velocity  $C^2$  for  $v^2$

Then, Pressure of the gas  $P = \frac{1}{3} n m C^2$

As we have taken the mass of each molecule and there are  $n$  molecules in the gas,

The total mass of the gas present in the vessel of  $1 \text{ m}^3$  volume =  $nm$

But, mass per unit volume of the gas =  $nm = \rho$  the density of the gas.

$\therefore$  Pressure of the gas  $P = \frac{1}{3} \rho C^2$

### Relation between Pressure and Kinetic Energy of a Gas:

Consider a perfect gas having unit volume with density  $\rho$  and  $C^2$  be the mean square velocity of gas molecules, then

$\therefore$  pressure of the gas  $\rho = \frac{1}{3} P C^2$   
 $\therefore$  The Pressure,  $P = \frac{1}{3} \rho C^2$

The density  $\rho = \frac{M}{V} = M$  {Since volume  $V = 1 \text{ m}^3$ }

Dividing and multiplying by 2,

$P = \frac{2}{2} \times \frac{1}{3} \rho C^2 = \frac{2}{3} \times \frac{1}{2} \rho C^2$

The mean Kinetic energy per unit volume of the gas

K.E. =  $\frac{1}{2} \rho C^2$

The Pressure  $P = \frac{2}{3} \text{K.E.}$

The pressure of a gas is equal to two-thirds of the mean kinetic energy per unit volume of the gas.

But, the mean square velocity of the molecules  $\overline{C^2}$  is directly proportional to the absolute temperature of the gas.

Hence, the pressure of a gas is directly proportional to the absolute temperature of the gas.

### Mean kinetic energy per molecule of the gas:

One mole of any gas contains  $N$  (Avogadro) number of molecules, so the gas equation from kinetic theory of gas is

$$P = \frac{1}{3} \frac{M \overline{C^2}}{V}$$

[But  $M = Nm$ ]

$$PV = \frac{Nm \overline{C^2}}{3} \quad \text{---(1)}$$

For ideal gas, the equation is  $PV = RT$  ---(2)

$$\text{From (1) and (2) } PV = \frac{Nm \overline{C^2}}{3} = RT$$

$$\overline{mC^2} = \frac{3RT}{N}$$

Kinetic energy of one molecule

$$\frac{1}{2} m \overline{C^2} = \frac{3RT}{2N}$$

(Dividing by 2 on both sides)

$$KE = \frac{3RT}{2N}$$

$$= \frac{3}{2} KT$$

where 'K' is the Boltzmann constant which is equal to

$$K = \frac{R}{N} = 1.38 \times 10^{-23} \text{ JK}^{-1}$$

**Worked Problems:**

1. Find the r.m.s. velocity of hydrogen molecule at S.T.P. if density of hydrogen is  $0.0899 \text{ kg m}^{-3}$  at S.T.P.

Given:  $\rho = 0.0899 \text{ kg m}^{-3}$

Known: At S.T.P,  $P = 0.76 \text{ m of Hg} = 0.76 \times 13.6 \times 10^3 \times 9.8$

$$P = \frac{1}{3} \rho C^2, \quad \text{R.M.S. Velocity, } C = \sqrt{3P/\rho}$$

$$\text{R.M.S. Velocity, } C = \frac{3 \times 0.76 \times 13.6 \times 10^3 \times 9.8}{0.0899} = 1.838 \times 10^3 \text{ ms}^{-1}$$

2. The root mean square velocity of argon gas molecule at S.T.P. is  $434 \text{ ms}^{-1}$ . Find the density of argon gas at S.T.P.

Given; R.M.S. Velocity,  $c = 434 \text{ ms}^{-1}$

Known At S.T.P  $P = 0.76 \text{ m of Hg}$

$$P = \frac{1}{3} \rho C^2$$

$$\rho = \frac{3P}{C^2}$$

$$\rho = \frac{3 \times 0.76 \times 13.6 \times 10^3}{434 \times 434} \times 9.8$$

$$= 1.6133 \text{ kg m}^{-3}$$

3. Find the temperature at which the root mean square velocity of hydrogen will double its value at S.T.P. pressure remaining constant.

Let  $t$  be the temperature at which the r.m.s. velocity of hydrogen will double its value at S.T.P. Let  $C_0$  and  $C_t$  be the R.M.S. Velocity of molecules at S.T.P. and at temperature ' $t$ '. i.e.  $C_t = 2C_0$

Known,  $C \propto \sqrt{T}$        $C_t \propto \sqrt{(273+t)}$  and  $C_0 \propto \sqrt{273}$

$$C_t = 2C_0 = \sqrt{(273+t)}$$

$$2 \times \sqrt{273} = \sqrt{273 + t}$$

Squaring both sides

$$4 \times 273 = 273 + t$$

$$\therefore t = 1092 - 273 = 819 \text{ K or } 546^\circ\text{C}$$

### 1.3 SPECIFIC HEAT CAPACITY

When a substance is heated, it gains heat and its temperature increases. The quantity of heat gained by different substances is specified in terms of the value known as specific heat capacity.

#### Specific heat of solid and liquid:

To increase the temperature of a substance, heat energy should be supplied to it. It is found that the quantity of heat energy supplied to the substance (or) taken by the substance ( $Q$ ) is directly proportional to

i) the mass of the substance ( $m$ ) and

ii) increase in temperature of the substance ( $T_2 - T_1$ )

where  $T_1$  initial temperature of the substance and  $T_2$  final temperature of the substance.

$$Q \propto m(T_2 - T_1)$$

$$(or) Q = Sm (T_2 - T_1) \quad (1)$$

Where  $S$  is the constant of proportionality and is called as the specific heat capacity of the substance.

In equation (1) when  $m = 1 \text{ kg}$  and  $(T_2 - T_1) = 1 \text{ K}$ , then,  $Q = S$

$\therefore$  Specific heat capacity of solid or liquid (or substance) is defined as the quantity of heat energy required to increase the temperature of  $1 \text{ kg}$  of the substance through  $1 \text{ K}$ .

#### Unit for specific heat capacity:

We know  $Q = Sm(T_2 - T_1)$

$$\therefore S = \frac{Q}{m(T_2 - T_1)}$$

Substituting the respective units in R.H.S, the unit for specific

heat capacity is  $\frac{\text{joule}}{\text{kilogram. kelvin}}$  with symbol  $\text{J kg}^{-1} \text{K}^{-1}$ .

For example, the specific heat capacity of water is  $4190 \text{ J kg}^{-1} \text{K}^{-1}$ . It means, to increase the temperature of  $1 \text{ kg}$  of water through  $1 \text{ K}$ ,  $4190 \text{ J}$  of heat energy is required.

Similarly, the specific heat capacity of copper is  $385 \text{ J kg}^{-1} \text{K}^{-1}$ . It means, to increase the temperature of  $1 \text{ kg}$  of copper through  $1 \text{ K}$ ,  $385 \text{ J}$  of heat energy is required.

### Specific heat capacities of gases:

When a solid or liquid is heated, its temperature alone is increased, neglecting the little increase in its volume. But, when a gas is heated, its pressure and volume change in addition to its temperature. So, to study the effect of heating on a gas, we have to keep either the pressure or volume of the gas constant. Accordingly, for each and every gas, there are two specific heat capacities.

#### 1. Specific heat capacity at constant pressure ( $C_p$ ):

The specific heat capacity of a gas at constant pressure is the quantity of heat energy required to increase the temperature of one mole of the gas through 1K, when the pressure is kept constant.

#### 2. Specific heat at constant volume ( $C_v$ ):

The specific heat capacity of a gas at constant volume is the quantity of heat energy required to increase the temperature of one mole of the gas through 1K, when the volume is kept constant.

The unit for  $C_p$  and  $C_v$  is  $J K^{-1} mol^{-1}$

#### Ratio of specific heat capacities ( $\gamma$ ):

For a gas, the ratio of the specific heat capacity at constant pressure to the specific heat capacity at constant volume is a constant. i.e.

$$\frac{\text{Specific heat capacity at constant pressure } (C_p)}{\text{Specific heat capacity at constant volume } (C_v)} = \gamma (\text{constant})$$

Since  $C_p$  is always greater than  $C_v$ , the value of  $\gamma$  is always greater than one for any gas.

#### 3. Ratio between $C_p$ and $C_v$

The ratio of the specific heat capacity of a gas at constant pressure to the specific heat capacity of a gas at constant volume is a constant and is denoted by the letter ( $\gamma$ )

$$\text{Then, } \gamma = \frac{C_p}{C_v}$$

When a gas is heated at constant volume the heat supplied to the gas is wholly used to raise the temperature of the gas whereas when the gas is heated at constant pressure, a part is used to do external work to keep the pressure constant.

For monoatomic gases  $\gamma = 5/3$  for diatomic gases,  $\gamma = 7/5$  and for triatomic gases  $\gamma = 4/3$

**Explanation for  $C_p$  greater than  $C_v$ :**

The Specific heat capacity of a gas at constant pressure ( $C_p$ ) is the heat energy required to increase the temperature of 1 mole of a gas through 1K, at constant pressure.

The Specific heat capacity of a gas at constant volume ( $C_v$ ) is the heat energy required to increase the temperature of 1 mole of a gas through 1K, at constant volume.

When a gas is heated, keeping the volume of the gas as constant, the entire heat energy supplied is used only to increase its internal energy i.e. temperature.

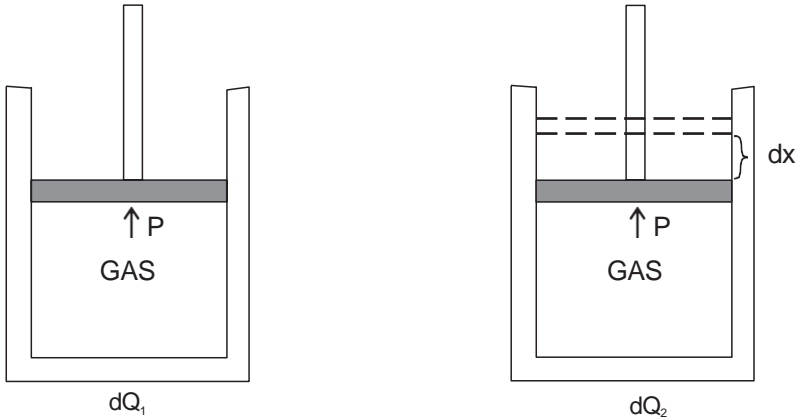
When a gas is heated, keeping the pressure of the gas as constant, the supplied heat energy is simultaneously used for two purposes. Apart of heat energy is used to increase the temperature of the gas and rest of the energy is used to do the external work of pushing the piston up, to maintain the constant pressure. Hence, in this case more heat energy is to be supplied to the gas. Thus  $C_p$  is greater than  $C_v$ .

**Meyer's relation  $C_p - C_v = R$  (Relation between  $C_p$  and  $C_v$ )**

Consider 1 mole of a perfect gas inside a cylinder in which there is a tight piston moving without any friction. The internal energy of the gas is directly proportional to its absolute temperature. Let  $dQ_1$  be the heat energy supplied to the gas at constant volume. Consequently, the temperature increases by  $dT$ .

We know, quantity of heat supplied =  
mass  $\times$  specific heat capacity  $\times$  increase in temperature

So,  $dQ_1 = 1 C_v dT$  (1)



Here, the entire heat is used to increase the temperature of the gas i.e. to increase the internal energy of the gas and no external work is done.

Let us now consider one mole of a perfect gas enclosed in a cylinder as before. Let the pressure and volume of the gas be  $P$  and  $V_1$ , respectively. Let  $A$  be the area of the piston. A quantity of heat  $dQ_2$  be given to the gas at constant pressure, such that the increase in temperature is  $dT$ .

$$\text{Now } dQ_2 = 1C_p dT \quad (2)$$

As the pressure of the gas is constant, the volume increases to  $V_2$ . So, some external work is done by the gas on the piston. Here the heat supplied is used for two purposes.

A portion of the heat  $dQ_2$  is used to increase the internal energy and the rest is used to do external work. i.e.  $dQ_2$  is used to

- (i) to increase the temperature of the gas by  $dT$ . and
- (ii) to push the piston up.

$$\therefore dQ_2 = dQ_1 + \text{work done in pushing the piston up.}$$

$$dQ_2 = dQ_1 + \{ \text{Force} \times \text{displacement of the piston} \} \quad (3)$$

We know that, Pressure = Force / Area

$$\therefore \text{Force acting on the piston} = \text{Pressure} \times \text{Area} = PA$$

Let  $dx$  be the distance moved by the piston, as shown in fig(ii).

Therefore (3) becomes,



$dQ_2 = dQ_1 + PAdx$ . But  $Adx = dV$   
 where  $dV = V_2 - V_1$  is the increase in volume of the gas.

$$dQ_2 = dQ_1 + PdV \quad (4)$$

Using (2) & (1), (4) becomes, 1

$$C_p dT = 1 C_v dT + PdV$$

$$C_p dT = C_v dT + PdV$$

$$(C_p - C_v) dT = PdV$$

The equation for a perfect gas is  $PV = RT$  On

differentiating,  $PdV = R dT$

So, the above expression becomes

$$(C_p - C_v) \cdot dT = R \cdot dT; \quad (C_p - C_v) = R \cdot \frac{dT}{dT}; \quad \text{But } \frac{dT}{dT} = 1$$

$$\therefore (C_p - C_v) = R$$

Note: Since value of the Universal gas constant  $R$  is positive, the value of  $C_p$  is always greater than  $C_v$  for any gas.

**Value of Universal Gas Constant: (R)**

Calculate the value of  $R$  from the gas equation  $PV = RT$  (OR)

Calculate the universal gas constant of given gas, given one gram molecule of a gas at N.T.P. occupies a volume of 22.4 litres.

At N.T.P. Pressure  $P = 0.76 \text{ m of Hg}$

$$= 0.76 \times 13.6 \times 10^3 \times 9.8$$

Volume  $V = 22.4 \text{ litres}$

$$= 22.4 \times 10^{-3} \text{ m}^3$$

Temperature  $T = 273 \text{ K}$

= Gas equation  $PV = RT$

$$R = \frac{PV}{T}$$

$$= \frac{0.76 \times 13.6 \times 10^3 \times 9.8 \times 22.4 \times 10^{-3}}{273}$$

Universal gas constant  $R = 8.31 \text{ JK}^{-1}$  per gram mole

(or)  $R = 8310 \text{ JK}^{-1}$  per k.mole

### Worked Examples:

1. A gas at 1.5 atmospheres occupies a volume of  $2\text{m}^3$  at  $25^\circ\text{C}$ . What will be the volume of gas at  $40^\circ\text{C}$  and at 3 atmospheres?

$$\text{Gasequation, } PV = RT \text{ (or) } \frac{PV}{T} = R$$

$$\text{Also } \frac{P_1 V_1}{T_1} = \frac{P_2 V_2}{T_2}$$

$$P_1 = 1.5 \text{ Atm. pre.}, T_1 = 25^\circ\text{C} + 273 = 298\text{K}$$

$$P_2 = 3 \text{ Atm. pre.}, T_2 = 40^\circ\text{C} + 273 = 313\text{K}$$

$$\text{And } V_1 = 2 \text{ m}^3$$

$$V_2 = \frac{(P_1 V_1 T_2)}{P_2 T_1} = \frac{(1.5 \times 2 \times 313)}{(3 \times 298)} \\ = 1.05 \text{ m}^3$$

2. The ratio between two specific heats of a gas is 1.36. The specific heat capacity of the gas at constant volume is  $23094 \text{ JK}^{-1}$  per k.mole. Find the value of Universal gas constant.

$$C_p / C_v = 1.36 \text{ and } C_v = 23094 \text{ JK}^{-1} \text{ per k.mole}$$

$$C_p = 1.36 \times C_v = 1.36 \times 23094$$

$$= 31407.84 \text{ JK}^{-1} \text{ per kmole.}$$

$$\text{By Mayer's relation } C_p - C_v = R$$

$$\text{i.e., } R = 31407.84 - 23094$$

$$R = 8313.84 \text{ JK}^{-1} \text{ per k.mole}$$

3. Find the specific heat capacity at constant pressure of helium if  $C_v = 12525 \text{ JK}^{-1}$  per k.mole. One gram molecule of the gas occupies  $22.42$  litres at NTP. Given  $v = 9.8 \text{ ms}^{-2}$  and density of mercury =  $13600 \text{ kgm}^{-3}$

$$\text{Given; } C_v = 12525 \text{ JK}^{-1} \text{ per k.mole}$$

$$V = 22.42 \text{ litres } = 22.42 \times 10^{-3} \text{ m}^3$$

$$\text{At N.T.P. } P = 0.76 \times 136000 \times 9.8 \text{ and } T = 273\text{K}$$

$$\text{Gasequation } PV = RT$$

$$R = \frac{PV}{T} = \frac{0.76 \times 136000 \times 9.8 \times 22.42 \times 10^{-3}}{273}$$

$$= 8.313 \text{ JK}^{-1} \text{ per mole}$$

$$= 8313 \text{ JK}^{-1} \text{ kmole}^{-1}$$

By Meyer's relation,  $C_p - C_v = R$

$$C_p = R + C_v \\ = 8313 + 12525$$

i.e.,  $C_p = 20838 \text{ JK}^{-1}$  per k. mole

4. The ratio of two specific heats of Ammonia is 1.31 and universal gas constant is  $8314 \text{ JK}^{-1}$  per k. mole. Calculate the specific heat of the gas at constant pressure and at constant volume.

The ratio,  $C_p/C_v = \gamma$ ,  $C_p = \gamma C_v$

By Meyer's Relation,  $C_p - C_v = R$ ,  $C_p = R + C_v$

Then  $R + C_v = \gamma C_v$

$$R = \gamma C_v - C_v = C_v(\gamma - 1)$$

$$C_v = R / (\gamma - 1) = 8314 / (1.31 - 1)$$

i.e.,  $C_v = 26819 \text{ JK}^{-1}$  per k. mole

$$C_p = R + C_v = 8314 + 26819$$

i.e.,  $C_p = 35133 \text{ JK}^{-1}$  per k. mole

## QUESTIONS

### Part-A

- 1) What is heat?
- 2) What is temperature?
- 3) What are the three important scales of temperature?
- 4) Define Conduction.
- 5) Define Convection.
- 6) Define Radiation.
- 7) Give examples for two good and two poor thermal conductors.
- 8) Define the coefficient of thermal conductivity.
- 9) Give the relation between pressure and kinetic energy of a gas.
- 10) Define mean square velocity of molecules.
- 11) Define Root Mean Square velocity of molecules.
- 12) Give the relation between two specific heats of a gas with universal gas constant.

- 13) State any two postulates of kinetic theory of gases.
- 14) Define mean free path.
- 15) Define specific heat capacity of a solid.
- 16) Define specific heat capacity of a liquid.
- 17) Define specific heat capacity of a gas at constant pressure.
- 18) Define specific heat capacity of a gas at constant volume.
- 19) Define ratio of specific heat capacities of gases.

### Part-B

- 1) Explain the process of conduction, convection and radiation.
- 2) Explain the selection and uses of good and poor thermal conductors.
- 3) List the properties and uses of thermal radiation.
- 4) List the postulates of kinetic theory of gases.
- 5) Derive an expression for the pressure of a gas on the basis of Kinetic theory of gases.
- 6) Derive the relation between pressure and kinetic energy and absolute temperature of a gas.
- 7) Derive the expression for the mean kinetic energy per molecule of a gas.
- 8) Derive Meyer's relation between two specific heat capacities  $C_p$  and  $C_v$  of a perfect gas. (or) Derive  $C_p - C_v = R$ , for a perfect gas.
- 9) Calculate the value of universal gas constant from the equation  $PV = RT$ .

### Exercise problems:

1. Find the RMS velocity of hydrogen molecules at S.T.P., if the density of hydrogen is  $0.0899 \text{ kg m}^{-3}$  at S.T.P.

(Ans:  $1838.52 \text{ ms}^{-1}$ )

(Hint: -S.T.P. means standard temperature and pressure  
Standard pressure = 1 atmospheric pressure = 76 cm of mercury  
 $P = h\rho g = 0.76 \times 13,600 \times 9.81 = 101396.16 \text{ Nm}^{-2}$ )

2. At atmospheric pressure, the density of Nitrogen gas is  $1.25 \text{ kg m}^{-3}$ . Find the R.M.S. velocity of Nitrogen molecules.

(Ans:  $493.05 \text{ ms}^{-1}$ )

3. The density of carbon dioxide gas at S.T.P. is  $1.977 \text{ kg m}^{-3}$ . Find the R.M.S. velocity of carbon dioxide molecules.

(Ans:  $392.05 \text{ ms}^{-1}$ )

4. Calculate R.M.S. velocity of air molecules at N.T.P., if the density of air is  $1.29 \text{ kg m}^{-3}$ .  
(Ans:  $485.35 \text{ ms}^{-1}$ )
5. The density of Helium gas at S.T.P. is  $0.1786 \text{ kg m}^{-3}$ . Calculate the RMS velocity of Helium molecules.  
(Ans:  $1304.4 \text{ ms}^{-1}$ )
6. The molecular weight of carbon di-oxide is 44. Calculate the RMS velocity of  $\text{CO}_2$  molecules at N.T.P.  
(Ans:  $393.32 \text{ ms}^{-1}$ )
7. The specific heat capacity of a gas at constant volume is  $23,094 \text{ J K}^{-1} \text{ k mol}^{-1}$ . The ratio of specific heat capacities of the gas is 1.36. Calculate the specific heat capacity of the gas at constant pressure and also the value of the universal gas constant R for that gas.  
(Ans:  $8313.84 \text{ J K}^{-1} \text{ k mol}^{-1}$ )
8. The specific heat capacity of Helium gas at constant volume is  $12,525 \text{ J K}^{-1} \text{ k mol}^{-1}$ . Find the specific heat capacity of the gas at constant pressure. 1 gram molecule of the gas occupies 22.4 litres at N.T.P.  
(Ans:  $20836.2 \text{ J K}^{-1} \text{ k mol}^{-1}$ )
9. The ratio of specific heat capacities of Helium is 1.667. Calculate its specific heat capacity at constant volume and at constant pressure. Given  $R = 8.3151 \text{ J K}^{-1} \text{ mol}^{-1}$ .  
(Ans:  $C_v = 12.430 \text{ J K}^{-1} \text{ mol}^{-1}$  &  $C_p = 20.72 \text{ J K}^{-1} \text{ mol}^{-1}$ )
10. Calculate the value of gas constant for nitrogen if its density is  $1.251 \text{ kg m}^{-3}$  at N.T.P. (Molecular mass of nitrogen is 28g.)  
(Ans:  $R = 8.3048 \text{ J K}^{-1} \text{ mol}^{-1}$ )

### Brain Teasers:

- 1) Calculate the value of gas constant for hydrogen, given density of hydrogen at NTP is  $0.09 \text{ kg m}^{-3}$  and molecular mass is 2g  
(Ans:  $R = 8.140 \text{ J K}^{-1} \text{ mol}^{-1}$ )
- 2) The ratio of the specific heat of methane is 1.313. The specific heat capacity at constant pressure is  $35200 \text{ J K}^{-1}$  per kmole. Calculate the value of specific heat at constant volume and hence calculate the value of gas constant.  
(Ans:  $C_v = 26808.8 \text{ J K}^{-1} \text{ k mol}^{-1}$  &  $R = 8911 \text{ J K}^{-1} \text{ k mol}^{-1}$ )

# UNIT - II

## THERMODYNAMICS, LIQUEFACTION OF GASES & NON-CONVENTIONAL ENERGY

### THERMODYNAMICS

#### Introduction:

Energy exists in various forms in nature. They are mechanical energy (kinetic and potential), heat energy, electrical energy, radiant energy, chemical energy, etc.

In a hydroelectric power station, mechanical energy is converted into electrical energy. In a steam engine, heat energy is converted into mechanical energy. When a metal piece is hammered mechanical energy is converted into heat energy.

Thermodynamics deals with the conversion of heat energy into mechanical energy and vice versa. But in a mere general sense, it includes the relation of heat to other forms of energy like electrical energy, chemical energy, etc. In our discussions we will be concerned mainly about the conversion of heat energy into mechanical energy.

The science of thermodynamics is guided by the three laws of thermodynamics which will be discussed here.

#### Zeroth Law:-

Consider a gas enclosed in a cylinder fitted with a piston. If the pressure and volume of the enclosed gas is  $P$  and  $V$  at the temperature of the surroundings, these values of  $P$  and  $V$  will remain constant as long as the external conditions of temperature and pressure remain unaltered. The gas is said to be in thermal equilibrium with the surroundings.

If two systems  $A$  and  $B$  are separately in thermal equilibrium with a third system  $C$ , then the three systems ( $A$ ,  $B$  and  $C$ ) are in thermal equilibrium with each other.

**Zeroth Law of thermodynamics states that two systems which are individually in thermal equilibrium with a third one, are also in thermal equilibrium with each other.**

**First Law of Thermodynamics:-**

When a certain quantity of heat is supplied to a system, the whole of the heat energy may not be converted into work. Part of the heat may be used in doing external work and the rest of the heat might be used in increasing the internal energy of the molecules. Let the quantity of heat supplied to the system be  $\Delta Q$ ,  $\Delta W$  is the work done by the system and  $\Delta U$  is the change in internal energy of the system.

$$\Delta Q = \Delta W + \Delta U$$

**Hence the first law of thermodynamics states that the amount of heat energy supplied to a system is equal to the sum of the change in internal energy of the system and the work done by the system.**

According to First Law of Thermodynamics, heat energy can be converted into mechanical work and vice versa.

**ISOTHERMAL AND ADIABATIC CHANGES:**

In general, when a quantity of heat is given to the gas a part of the heat is used to increase the internal energy of the gas and the rest is used to do external work by the gas. Let  $dQ$  be the quantity of heat given to the gas,  $dE$  the increase in internal energy and  $W$ , the external work done by the gas. Then

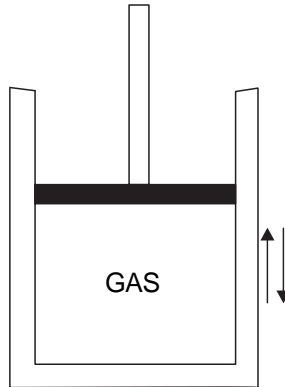
$$dQ = dE + W$$

This equation is called Gas-energy equation.

**Isothermal Change**

Any change in the pressure and volume of a gas at constant temperature is called as isothermal change.

Consider a gas inside the cylinder with a piston moving in it without friction. When the gas is compressed slowly, the temperature of the gas increases due to the work done on the gas. If the cylinder and the piston are made up of conducting materials then the gas gives out the excess heat to the surroundings. Hence there is no rise in temperature.



When the piston moves upwards slowly, the gas expands and the temperature of gas decreases due to the work done by the gas. Then the gas gained shortage of heat from the surroundings. Hence the temperature and therefore there is no change in internal energy. The gas-energy equation for Isothermal change becomes,

$dQ = \text{change in internal energy} + \text{External work for isothermal change, } dQ = 0 + W$

For a perfect gas, isothermal change can be represented by the equation,

$$PV = \text{a constant Also } P_1 V_1 = P_2 V_2$$

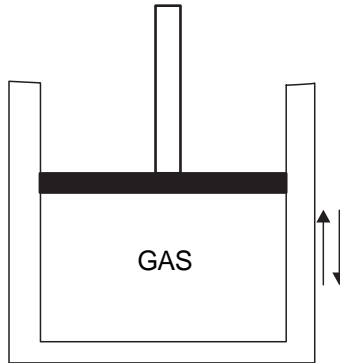
## 2. Adiabatic Change:

When the pressure and volume of a gas change, but no heat is allowed to enter or leave the gas, the change is called as adiabatic change.

Consider a gas inside the cylinder with a piston moving in it without friction. When the gas is compressed suddenly, the temperature of the gas is increased and the work is done on the gas. Similarly when the piston moves upward suddenly, the temperature of the gas is decreased and work is done by the gas. Due to the sudden compression or expansion, the gas neither gives out heat to the surroundings nor receives heat from the surroundings, when the cylinder and the piston are made of nonconducting materials.

During the adiabatic change, no heat leaves or enters the gas. That is  $dQ = 0$ . The work is done at the cost of internal energy. The gas energy equation, for the adiabatic change becomes,





$dQ = \text{Change in internal energy} + \text{External work}$

$$0 = dE + W \text{ (or) } W = dE$$

For a perfect gas, adiabatic change can be represented by the equation,

$$PV^\gamma = \text{a constant} \quad \text{Also} \quad PV_1^\gamma = PV_2^\gamma$$

The equation relating the temperature and volume of the gas is

$$TV_1^{\gamma-1} = T_2V_2^{\gamma-1}$$

The equation relating the temperature and pressure of the gas is

$$T_1 P_1^{\gamma-1} = T_2 P_2^{\gamma-1}$$

Where  $\gamma = C_p / C_v$  is the ratio of the two specific heat capacities of the gas.

### 3. Worked Problems:

1. A gas at 2 atm is compressed to half of its original volume. Calculate the final pressure, if the compression is (i) isothermal and (ii) adiabatic

( $\gamma = 1.4$ )

Let  $V$  be the original volume

Given;  $P_1 = 2$  atmosphere,  $V_1 = V$  and  $V_2 = \frac{1}{2}V$

- 1) For isothermal change, the equation is  $P_1 V_1 = P_2 V_2$

$$\therefore P_2 = \frac{P_1 V_1}{V_2} = \frac{2 \times V}{(1/2)V} = 4 \text{ atmospheric pressure}$$

ii) Adiabatic equation is  $P_1 V_1^\gamma = P_2 V_2^\gamma$

$$\therefore P_2 = \frac{P_1 V_1^\gamma}{V_2^\gamma} = \frac{2 \times V_1^\gamma}{\frac{V_1}{2}^\gamma} = \frac{2 \times V_1^\gamma \times 2^\gamma}{V_1^\gamma} = 2 \times 2^\gamma = 2 \times 2^{1.4}$$

$= 2 \times 2^{1.4} = 5.278$  Atmospheric Pressure.

2. Air at a pressure of 0.75 m of mercury and of volume 1 litre is compressed to a pressure of 1.5 m of mercury under isothermal process. Calculate the resulting volume.

The equation for isothermal change is  $P_1 V_1 = P_2 V_2$  Given

en;  $P_1 = 0.75$  m of Hg

$P_2 = 1.5$  m of Hg

$V_1 = 1$  litre

$$\therefore V_2 = (P_1 V_1) / P_2 = (0.75 \times 1) / 1.5 = 0.5 \text{ litre}$$

3. A certain mass of gas at 3 atmosphere is compressed adiabatically to half of its volume. Calculate the resulting pressure if  $\gamma = 1.4$

Given;  $V_1 = V, V_2 = \frac{1}{2} V$  and  $P_1 = 3$  At. pr.  $P = P V^\gamma / V^\gamma$

$$P_2 = P V_1^\gamma / (V_2)^\gamma$$

$$P_2 = (3 \times V^{1.4}) / (\frac{1}{2} V)^{1.4} = 3 \times 2^{1.4} = 7.92 \text{ At. pr.}$$

### Second Law of Thermodynamics:-

A heat engine is mainly used for the conversion of heat energy into mechanical work. A refrigerator is a device used to cool a certain space below the temperature of its surroundings. The first law does not contradict the existence of 100% efficient heat engine or a refrigerator, which are not attainable in practice. These phenomena led to the formulation of second law of thermodynamics. There are two versions of this second law of thermodynamics.

#### a. Clausius Statement:

It is impossible for a self acting mechanism working in a cyclic process unaided by an external agency to transfer heat from a body at a lower temperature to the body at a higher temperature.

This part of the law is applicable in the case of ice plants and refrigerators. i.e., Heat itself cannot flow from a body at a lower temperature to a body at a higher temperature, on its own.

### b. Kelvin's Statement:

It is impossible to derive a continuous supply of work from a body by cooling it to a temperature lower than that of its surroundings. This part of the law is applicable in the case of Heat engines. In a heat engine the working substance does some work and rejects the remaining heat to the sink. The temperature of the source must be higher than the surroundings. The engine will not work when the temperatures of the source and the sink are same.

### 4. Concept of Carnot's reversible Heat Engine and Efficiency:-

A machine that converts heat continuously into work is called a heat engine. Carnot conceived a theoretical engine which is free from all the defects of the practical engines. It is an ideal heat engine which has maximum efficiency.

For any engine, there are three essential requisites. They are

1. Source: The source should have a fixed higher temperature from which the heat engine can draw heat. The source has unlimited thermal capacity so that it can deliver any amount of heat required by the engine continuously, at constant temperature.
2. Sink: The sink should be at a fixed lower temperature to which any amount of heat can be rejected. The sink has unlimited capacity to receive the heat and its temperature remains constant.
3. Working substance: The working substance is usually a perfect gas inside a cylinder with non-conducting sides and conducting bottom. A perfect non-conducting and frictionless piston is fitted into the cylinder. A perfectly non-conducting stand is also provided so that the working substance can undergo adiabatic operation.

The ratio of part of heat converted into work to the part of heat absorbed from the source is the efficiency of the engine.

$$\text{Efficiency} = \frac{\text{output}}{\text{input}}$$

$$\text{Efficiency } \eta = \frac{(Q_1 - Q_2)}{Q_1}$$

where ' $Q_1$ ' is the Quantity of heat absorbed by the substance  
' $Q_2$ ' is the Quantity of heat rejected to the sink  
and ' $Q_1 - Q_2$ ' is the Quantity of heat converted into work.

## 2.2. LIQUEFACTION OF GASES

The work on the liquefaction of gases started in 1823 by Faraday. Gases like chlorine, hydrogen chloride, hydrogen sulphide, sulphur dioxide, ammonia and carbon dioxide were liquefied under high pressure with the aid of simple freezing mixtures. But this method cannot be used to liquefy oxygen, nitrogen, hydrogen and helium.

In 1863, Andrew Shadd discovered the phenomenon of critical state of the gas. According to Andrew's discovery, a gas cannot be liquefied when it is at a above a particular temperature. That particular temperature is called critical temperature. The pressure and volume of the gas corresponding to that temperature are called the critical pressure and critical volume. After Andrew's discovery, the liquefaction was done easily by cooling the gas below its critical temperature, even though the process is difficult to liquefy oxygen, nitrogen, hydrogen and helium since they have very low critical temperatures. For example, the critical temperature for oxygen is  $-119^{\circ}\text{C}$  and for hydrogen is  $-240^{\circ}\text{C}$ .

Then in 1877, Pictet liquefied the oxygen by cascade (step by step) process in which low temperatures are obtained in step by step by the evaporation of suitable liquids. In 1896, Linde performed a process to liquefy air by using Joule-Thomson effect.

### Critical constants:

The critical temperature, critical pressure and critical volume are called critical constants  $T_c$ .

- 1. Critical Temperature ( $T_c$ ):** The critical temperature of gas is defined as that temperature above which the gas cannot be liquefied however great the pressure applied may be.
- 2. Critical pressure ( $P_c$ ):** The critical pressure is the minimum pressure required to liquefy a gas when it is at its critical temperature.
- 3. Critical volume ( $V_c$ ):** The critical volume is the volume of unit mass of a gas when it is at its critical temperature and critical pressure.

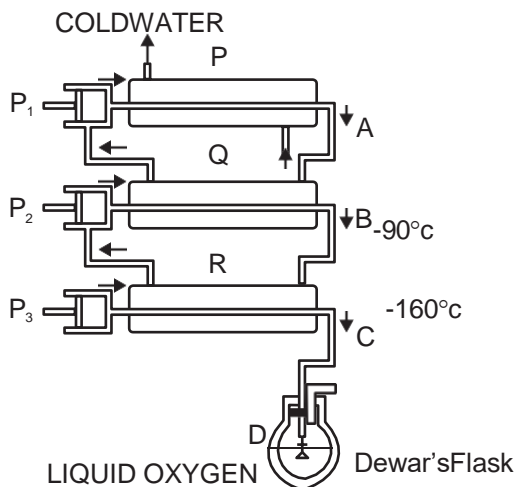
**Note:** Below the critical temperature the gas is termed as vapour and above the critical temperature it is termed as gas. The gas can be liquefied easily when it is in the state of vapour. Gases like oxygen, nitrogen, hydrogen and helium are termed as permanent gases since they have very low critical temperatures.

## 2. Cascade Process-Liquefaction of oxygen:

The critical temperature for oxygen is  $-119^{\circ}\text{C}$  and critical pressure is  $49.7\text{atm}$ .

Principle: When a liquid is allowed to evaporate under reduced pressure, it produces high cooling.

The apparatus arrangement used in this process is shown in the figure. It consists of three narrow tubes, A, B and C enclosed by three outer jackets P, Q and R respectively. The narrow tubes and the outer jackets are linked with the compression pumps  $P_1, P_2$  and  $P_3$  as shown in the figure. The methyl chloride gas of critical temperature  $145^{\circ}\text{C}$  is compressed by the pump  $P_1$  through the tube A. It is cooled by the cold water circulating in the jacket P. Here the methyl chloride reaches the temperature lower than its critical temperature. Then it is liquefied under high pressure. The liquid methyl chloride is collected in the jacket Q and evaporates under reduced pressure lowering the temperature to  $-90^{\circ}\text{C}$ .



The ethylene gas of critical temperature  $10^{\circ}\text{C}$  is compressed by the pump  $P_2$  through the tube B. It is cooled to  $-90^{\circ}\text{C}$  by liquid methyl chloride. Then it is liquefied under high pressure. The liquid ethylene is collected in the jacket R and evaporates under reduced pressure lowering the temperature to  $-160^{\circ}\text{C}$ .

The oxygen gas of critical temperature  $-119^{\circ}\text{C}$  is compressed to 50 atmospheric pressure by the pump  $P_3$  and passed through the tube C. It is cooled to  $-160^{\circ}\text{C}$  by liquid ethylene in R. Then it is liquefied and the liquid oxygen is collected in the Dewar flask D.

**Disadvantage:** The lowest temperature obtained in this process is  $-160^{\circ}\text{C}$ . Therefore the cascade process cannot be used to liquefy hydrogen and helium whose critical temperatures are lower than  $-160^{\circ}\text{C}$ .

### 3. Joule-Thompson Effect:

Using the principle of Joule - Thompson effect, air, hydrogen and helium were also liquefied. Joule performed an experiment to study the effect of change in the temperature of a gas when it is allowed to pass through a porous plug from the high pressure side to the low pressure side.

**Joule-Thompson effect:** When a gas is allowed to expand from a high pressure side to a low pressure side through a porous or nozzle or jet then it produces intense cooling, if the temperature of the gas is less than the temperature of inversion. This effect is called Joule - Thompson effect.

**Temperature of inversion:** If a gas is at a particular temperature no cooling or heating is observed when it is passed through the porous plug. If the temperature of the gas on the high pressure side is more than that temperature, heating will take place instead of cooling. That particular temperature is called the temperature of inversion.

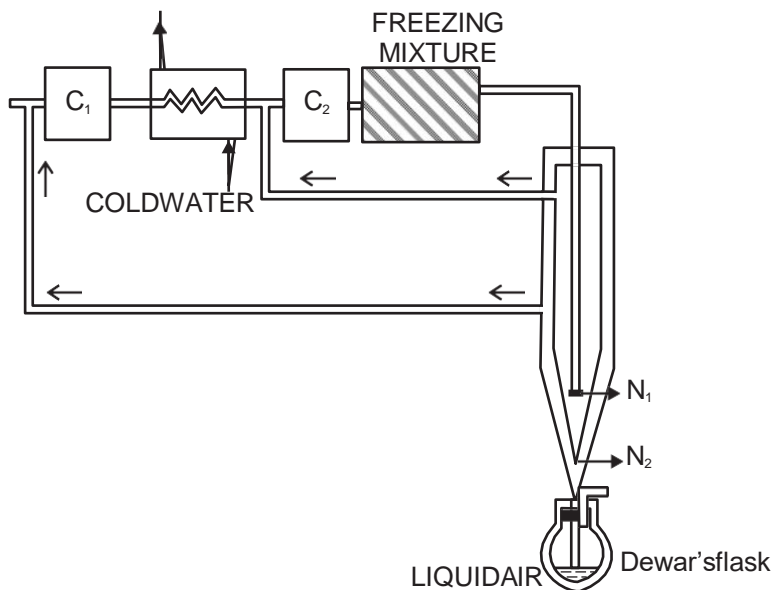
Temperature inversion may be defined as the temperature at which there is neither cooling nor heating due to Joule-Thompson effect.

### 4. Linde's Process-Liquefaction of Air:

The principle of Joule-Thompson effect is used in this process. The apparatus arrangement used in this process is shown in the figure. It consists of two compressors  $C_1$  and  $C_2$ , a cold water jacket, freezing mixture jacket, narrow tube having nozzle  $N_1$  and a wide tube having nozzle  $N_2$ .

Air is compressed to 25 atmospheres by the compressor  $C_1$ . The compressed air is passed through a tube surrounded by a cold water

jacket. Again this air is compressed to 200 atmospheres by the compressor  $C_2$ . This compressed air is passed through a spiral tube surrounded by a freezing mixture jacket. Now the air reaches to a temperature of  $-20^\circ\text{C}$ . This cooled air at high pressure is allowed to expand through an nozzle  $N_1$ , and it is further cooled by Joule-Thompson effect. This cooled air is circulated back to the compressor  $C_2$  and is compressed again.



As the process continues, after a few cycles, the air gets cooled to a sufficiently low temperature. Finally the air expands from high pressure side to the low pressure side through the nozzle  $N_2$ . Again the temperature is reduced by Joule - Thompson effect and the air gets liquefied. The liquid air is collected in the Dewar flask. The unliquefied air is again circulated back to the compressor  $C_1$ , and the process is repeated. The whole apparatus arrangement is packed in cotton or thermal wool to avoid any conduction and radiation.

## 2.3. NON CONVENTIONAL ENERGY

### Introduction:

Energy is an important input for all sectors of a country's economical development. For India, the energy needs is increasing rapidly day by day. It is a well-known fact that the currently available conventional energy resources like coal, oil, natural gas, etc., are depleting very fast. 54% of our energy need is provided by coal, which may be available for another 80 years at the current rate of usage. 32% of energy need is fulfilled by oil, which may last up to 22 years and the rest 14% is met by hydro, nuclear, natural gas, etc. So, we will have to increasingly depend on other non conventional energy resources like solar, wind, tidal, geothermal, hot dry rock, bio mass, etc.

These sources of energy utilised by human beings fall into two categories:

- 1. Non-renewable** energy resources are sources of power that cannot be replaced once they are used, because the energy source has taken millions of years to form (ex. coal, oil). These energy resources are also called as **finite** energy resources.
- 2. Renewable** energy resources are sources of power that quickly replenish (fill again) themselves and can be used again and again repeatedly (ex. Wind, tidal). These energy resources are also called as **infinite** energy resources.

### Conventional or Non-renewable sources of energy:

The conventional or non-renewable energy sources are reclassified as primary and secondary energy sources. The primary energy sources are which provide a net supply of energy, like coal, oil, natural gas, uranium, etc. These fuels are called primary fuels, from which the energy yield ratio is very high. i.e., the energy required to obtain these fuels is much less than what they can produce by combustion or nuclear reaction. These secondary fuels like wood, animal wastes, agricultural wastes, etc., produce net energy.

### Non-conventional or renewable sources of energy:

To meet our future energy requirements, we have to tap other available non conventional or renewable energy sources. The renewable sources of energy are also called as non conventional (or) alternate



sources of energy. They are Solar, Wind, Tidal, Geothermal, Bio mass, Hot dry rock, Chemical, etc. Various technologies are designed and developed to tap these earth's natural energy resources which are renewable. They are called renewable because, they are available in plenty and freely.

### **Solar energy:**

The sun is a large sphere of very hot gases, the energy being generated by various kinds of nuclear fusion reactions.

Sun gives us 1000 times more power than we need. Therefore, solar energy can be a major source of power. Even if we could tap one thousandth of solar power, that will meet our requirements. But so far it could not be developed on a large scale.

Energy produced and radiated by the sun, which reaches the earth, is termed as "Solar energy". Energy is radiated by the sun, as electromagnetic waves. Solar energy, received in the form of electromagnetic radiation, can be converted directly or indirectly into other forms of energy, such as heat and electricity, which can be utilised by human being. Since the sun can radiate energy at a constant rate for a few billion years, it may be regarded as an inexhaustible source of useful energy.

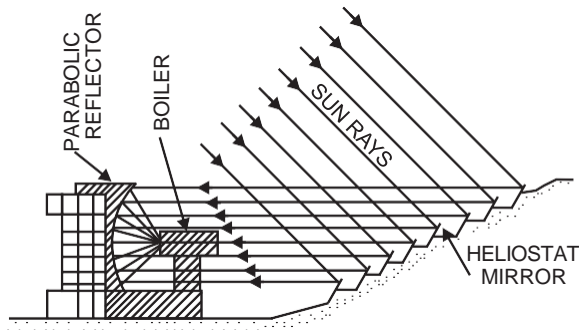
As India is situated near equator, where sun light is abundant for a major part of the year, utilising the solar energy to a greater extent is the best option available to us.

Some important applications of solar energy are

(i) Solar water heating, (ii) Heating and cooling of residential buildings, (iii) salt production by evaporation of seawater, (iv) Solar cookers, (v) Solar furnaces, (vi) Solar pumps for water pumping and (vii) Solar electric power generation.

From solar energy electricity can be produced with the help of photovoltaic or solar cells. The solar cells directly generate electricity by photovoltaic process. When the photons from the sun's light is absorbed by special type of semiconducting material they create free electrons.

These free electrons induce higher energy electrons to flow out of the semiconductor. The flow of higher energy electrons in the circuit constitutes the electric current.



### Solar water heating system

This system is used to drive a turbo-generator to produce electricity. A plant of this type is operating in France and produces 1000 kW.

## 2. Wind energy:

Wind is basically caused by the solar energy irradiating the earth. Energy of wind can be economically used for the generation of electrical energy. Wind is caused due to two main factors:

- (a) Due to air in motion. Air motion arises from a pressure gradient. The surface winds flow from the pole to the equator. Solar radiation heats the air near the equator, and this low density heated air is buoyed up. At the surface, it is displaced by cooler, more dense higher pressure air flowing from the poles. In the upper atmosphere near the equator, the air tends to flow towards the poles and away from the equator. The net result is a flow of surface winds from north to south in the northern hemisphere. Similarly, the flow of surface winds is from south to north in the southern hemisphere of the earth.
- (b) Due to the rotation of the earth with respect to the atmosphere, and its motion around the sun.

Wind energy, which is an indirect source of solar energy conversion, can be utilized to run a windmill, which in turn drives a generator to produce electricity. Wind can also be used to provide mechanical power, such as for water pumping.

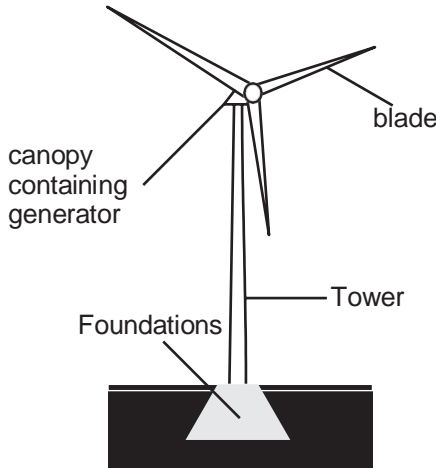
Some characteristics of wind energy are stated below:

- (i) It is a renewable source of energy.
- (ii) Wind-power systems are non-polluting, so it does not affect the environment.

- (iii) Wind energy systems avoid fuel provision and transport.
- (iv) Large areas are needed to install wind farms for electrical power generation and wind energy needs storage means because of its irregularity in the speed of flow.
- (v) Manufacture and implementation of wind farms can be costly.

As our country has a lot of wind throughout the year, the energy in the wind can be used to generate power that can be used for domestic and industrial uses. The windmill is the form in which we can convert the wind energy to electricity. The potential of wind energy is of the order of  $1.6 \times 10^7$  MW.

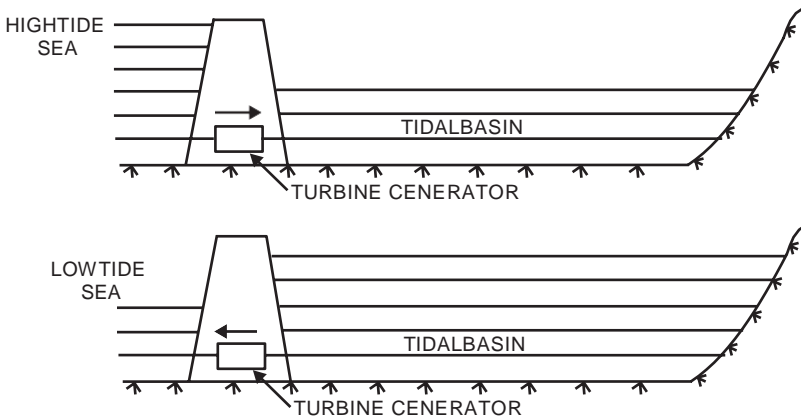
The windmill generally consists of a turbine (a dynamo which can produce electricity when rotated) connected to larger rotating blades. The blades rotate by using the wind.



The setting up cost of these windmills is high but the running cost is very low. Use of this would help in producing cheap power. Tamil Nadu has set up a large number of wind mills in the state especially in places like Tiruppur, Muppandal in Kanyakumari, Coimbatore and their surroundings.

### 3. Tidal energy:

Tide is a periodic rise and fall of the water level of sea due to gravitational effect of the sun and moon on the water of the earth. Due to fluidity of water mass, there is a periodic rise and fall in levels of seawater which is in rhythms with the daily cycle of rising and setting of sun and moon. As the earth rotates, the position of a given area relative to the moon changes, and thus a periodic succession of high and low tides occur. This periodic rising and falling of the water level of sea is known as tidal power.



The large scale up and down movement of sea water due to tide represents an unlimited source of energy. If some part of the vast energy can be converted into electrical energy it would be an important source of hydro – power. Using a hydraulic turbine, the tidal energy could be converted into electrical energy by means of an attached generator. Water at the time of a high tide is at a higher level and can be let into a basin to be stored at a high level there. The same water can be let back into the sea during the low tide through the turbines, thus producing power.

The advantages of tidal power are

- 1) It is free from pollution, as it does not use any fuel and does not produce any unhealthy waste like gases, ash, atomic refuse, etc.
- 2) These power plants do not demand large areas of valuable land because they are on the bays (Seashore).

Its own limitations are

- 1) Sea water is corrosive and it was feared that the machinery may get corroded. But using vinyl paint exhibited good results.
- 2) Cost is not favourable compared to the other sources of energy and it is feared that the tidal power plant would hamper the other natural uses such as fishing or navigation.

### **3. Geothermal energy:**

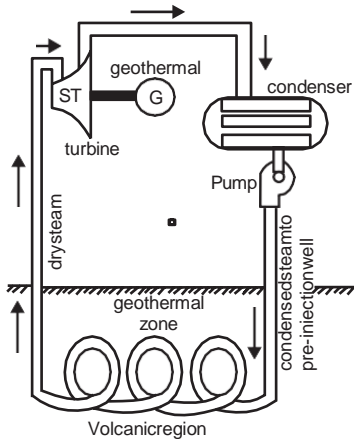
Energy present as heat (i.e. thermal energy) in the earth's crust, in the uppermost 10 km or so of the crust constitutes a potentially useful and almost inexhaustible source of energy.

Geothermal source is defined as "all of the heat stored in the earth's crust above 15°C to a depth of 10 km". It occurs when the immense heat energy in the core of earth rises close to the surface of the earth due to cracks or faults in the crust and heats the surrounding rock. These are tapped to use the stored heat energy, for power generation and several other uses, depending upon the temperature of occurrence and other parameters.

Hot molten rock, called "Magma" is commonly present at depths greater than 24 to 40 km. The hot magma near the surface solidifies into igneous rock which is formed by volcanic action. The heat of the magma is conducted upward to this igneous rock. Groundwater that finds its way down to this rock through fissures (cracks) in it, will be heated by the heat of the rock or by mixing with hot gases and steam let out from the magma. The heated water will then rise convectively upward into a porous and permeable reservoir above the igneous rock. The reservoir is capped by a layer of impermeable solid rock that traps the hot water in the reservoir. The solid rock, however, has fissures that act as vents of the giant underground boiler. The vents show up at the surface as geysers or hot spring. A well taps steam from the fissures for use in a geothermal power plant.

Geothermal energy has the highest energy density and it is classified as renewable because the earth's interior will continue to be hot for an indefinite future.

But it is very expensive to set up, and works in areas of volcanic activity. Dangerous underground gases have to be disposed of very carefully in this source of energy.

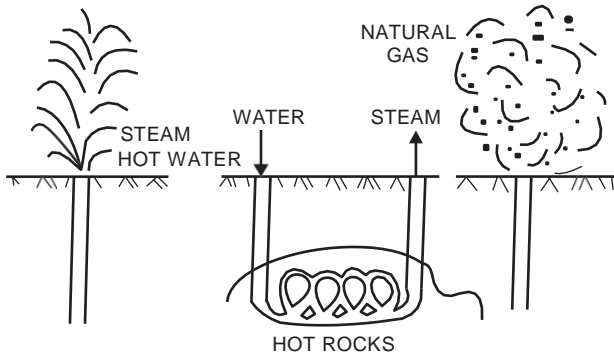


Geothermal power stations use the naturally produced steam to drive turbines and generators. Once used, the steam turns back into water and is pumped back into the ground again to be reheated by the earth.

#### 4. Hot dry rock resources (HDR):

This system is composed of hot dry rock (HDR) without underground water. They represent the largest geothermal resource available. The hot rock, occurring at moderate depths, has very low permeability and needs to be fractured to increase its heat transfer to surface. Hot dry rocks exist because they are impermeable to water or because water does not have access to them. Heat is recovered from such hot dry rocks by breaking up or cracking the rock to make it permeable and then introducing water from the surface. The water is heated up by the rock and is returned to the surface where the heat is utilized.

There are two methods to tap this geothermal energy. One possible method is to detonate a high explosive at the bottom of a well drilled into the rock. This may be a nuclear explosive. Another method is to use hydraulic fracturing to produce the heat transfer surface and permeability required to extract energy at a high rate from hot dry rock. Heat can then be extracted from the hot rock by circulating water through the crack. The thermal energy of HDR is used in a power plant to produce electricity.



The limitation of HDR is, two wells are to be drilled deeper and in much harder rock. This is expected to make the exploitation very costly, unless the underground rock targeted is very hot.

### 5. Energy from Biomass:

Biomass is organic matter produced by plants, both terrestrial and aquatic and their derivatives. It includes forest crops and residues, crops grown especially for their energy content on energy farms and animal manure. Biomass can be considered a renewable energy source because plant life renews and adds to itself every year. Biomass means organic matter in which photochemical approach is used to harness solar energy by photosynthesis. Solar energy is stored in the form of chemical energy.

Solar energy → Photosynthesis → Biomass → Energy conversion

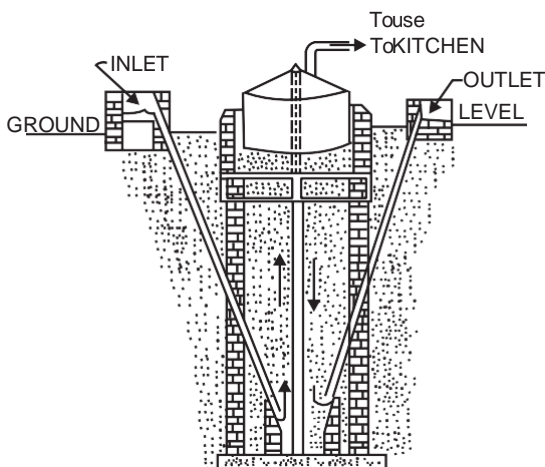
### Biomass resources fall into three categories:

- (i) Biomass in its traditional solid mass (wood and agriculture residue) - In this category biomass is burnt directly to get the energy.
- (ii) Biomass in non-traditional form (converted into liquid fuels) - In this category, the biomass is converted into ethanol and methane to be used as liquid fuels in engines.
- (iii) The third category is to ferment the biomass to obtain a gaseous fuel called bio-gas.

Biomass conversion can take many forms: (1) direct combustion, such as wood waste and bagasse, (2) thermochemical conversion, and (3) biochemical conversion.

Biomass includes wood waste which has potential of generating substantial electric power. All these biomass are highly dispersed and bulky and contain large amounts of water ( 50 to 90%).Thus it is not economical to transport them over long distances, and conversion into usable energy must take place close to the source, which is limited to particular regions. When biomass is burnt, it gives off atmospheric pollutants,includinggreenhousegasesandbiomassisonlyarenewable resourceifcropsarereplanted.

Rottingplantscanproducemethanegaswhichisthesameas“natural gas”. Thisgascanbecollectedfromsuchplacesandpipedtopeople's homesforheatingorcookingorburntinasmallpowerstationtoproduce electricity.



**Gobargas plant**

Wastes that are degradable can be used for generation of power. Gobargaswhichisproducedfromthewastesofanimalshasbeenused asasubstituteforLPGinvillagesforalongtime.

The waste is fed through the inlet pipe. The waste is fed till the partition wall is submerged in the waste.The methane gas bubbles out andiscollectedandtakenoutforusethroughapipelineprovidedatthetopof the drum. The outlet pipe is lower than that of inlet pipe to facilitate the used waste to flow out easily.



### **Advantages of renewable energy:**

1. Use of renewable energy resources is less pollutant.
2. Energy from renewable resources can be produced quickly.
3. Several renewable options are financially and economically competitive for certain applications in remote locations, where the costs of transmitting electrical power or transporting conventional fuels are high.
4. The plants which use renewable sources of energy can be built nearer to the site, where energy is needed, and reducing transportation cost of conventional fuels.
5. They can be produced in large numbers and introduced quickly, unlike large power stations which take often 10 years or more to become operational.
6. They can be built on, or close to the site where the energy is required. This minimises the transmission costs.
7. The important advantage is the availability in plenty, almost inexhaustible.

### **Disadvantages of renewable energy:**

1. Even a bad weather can interrupt the supply of electricity produced by wind, tidal, solar energy, etc.
2. The electrical energy produced by these renewable sources has to be stored suitably for continuous reliability.
3. Nonexistence of a proper planning, policy of finance, no a agency for implementing renewable energy systems, etc. at the government level.
4. Lack of qualified trained persons to design, manufacture, market, operate and maintain renewable energy sources plants.
5. Inadequate documentation and evaluation of past experience and a lack of clear priorities for future work.

## **QUESTION**

### **Part– A**

1. Write the Zeroth law of thermodynamics.
2. Write First law of thermodynamics.
3. What is an isothermal change?
4. What is an adiabatic change?

5. Write the Clausius statement of second law of thermodynamics.
6. Write the Kelvin's statement of second law of thermodynamics.
7. Define efficiency of Carnot's heat engine.
8. Define Critical temperature.
9. Define Critical pressure.
10. Define Critical volume.
11. What is the difference between a gas and a vapour?
12. What is the principle used in a cascade process?
13. What is the disadvantage of a cascade process of liquefaction of gases?
14. What is the temperature of inversion?
15. What is Joule–Thomson effect?
16. What are the renewable sources of energy?
17. What are the non-renewable sources of energy?
18. Define conventional energy with examples.
19. Define non-conventional energy with examples.
20. Write any two advantages of Non-conventional sources of energy.
21. Write any two disadvantages of Non-conventional sources of energy.

### **Part–B**

1. Explain isothermal and adiabatic change.
2. Describe the concept of Carnot's reversible engine and efficiency of a heat engine.
3. Describe the cascade process for the liquefaction of oxygen.
4. Describe Linde's process for the liquefaction of air.
5. What are the advantages and disadvantages of non-conventional energy over conventional energy?
6. Write a note on solar energy, wind energy and tidal energy.
7. Write a note on geothermal energy, hot dry rocks and biomass energy.

### **Exercise Problems:**

1. Air at 2 atmospheric pressure is compressed isothermally to one-fourth of its original volume. Find the final pressure. [Ans:  $P_2 = 8 \text{ At. pr}$ ]
2. A gas at a pressure of 0.75 m of mercury and volume one litre is compressed slowly to a pressure of 1.5 m of mercury. Find the resulting volume (Ans: 0.5 litre)

3. Air at a pressure of 5 atmosphere is compressed slowly to half of its original volume. Find the resulting pressure. (Ans:10atm)
4. A certain mass of gas at 1 atm pressure is suddenly compressed to one-fourth of its original volume. Find the resulting pressure.  
Given  $\gamma=1.4$  (Ans:6.964atm)
5. A certain mass of gas at 50atm is suddenly expanded to 10 times of its initial volume. Find the resulting pressure. Given  $\gamma=1.4$   
(Ans:1.9905atm)
6. A gas at 2 atm pressure is compressed to half of its original volume. Calculate the final pressure, if the compression takes place, (i) slowly and (ii) suddenly. Given  $\gamma=1.4$ . (Ans:(i)4atm(ii)5.2780atm)
7. Ten litres of gas at 50atmospheric pressure is suddenly expanded to 5 atmospheres. Find the resulting volume. Given  $\gamma=1.4$   
(Ans:51.7129litre)
8. Five litres of a gas at 5atmospheric pressure is suddenly compressed to a pressure of 5 times the original pressure. Find the resulting volume. Given  $\gamma=1.4$ . (Ans:1.5837litre)

**Brain Teasers:**

1. The ratio of the specific heat capacities of an ideal gas is 1.4. One atmosphere of the gas is compressed adiabatically to one-fifth of its original volume. Find its final pressure.  
[Ans:9.52At.pr]
2. Air at 5atmospheres is suddenly compressed to half of its volume. Find the resulting pressure of the air ( $\gamma=1.4$ )  
[Ans: $P_2=13.19\text{At.pr}$ ]

## UNIT - III

# LIGHT AND REMOTE SENSING

### LIGHT

Light is a form of energy radiated by luminous bodies. When it enters our eyes, it stimulates the retina to a sense of vision. An object is visible only by the scattering of light from the object to reach the eyes.

The objects which emit light rays (energy) are called light sources. For example, Sun, Stars, lamp etc are the light sources. Light can pass through the transparent medium like glass, air, water etc. Light rays undergo reflection, refraction interference, diffraction, polarization, etc. Light is propagated with a finite velocity  $3 \times 10^8 \text{ ms}^{-1}$ . Light plays a dual role. In some places it behaves like particles and in some places it behaves like waves, traveling in a straight line.

### OPTICAL FIBRE CABLE

#### Refraction

When a ray of light travels from one transparent medium into another medium, it bends while crossing the interface, separating the media. This phenomenon is called refraction.

#### Laws of Refraction

**I Law:** The incident ray, the refracted ray and the normal to the refracting surface at the point of incidence are in the same plane.

**II law:** The ratio of the sine of the angle of incidence to the sine of the angle of refraction is a constant for any two given medium for one particular colour of light. This second law is known as Snell's law.

#### Refractive index of a medium

When a ray of light passes from air or vacuum to the given medium, the ratio between the sine of the angle of incidence and the sine of the angle of refraction is called the refractive index of the given medium. If 'i' is the angle of incidence and 'r' is the angle of refraction,

$$\text{Refractive index of the medium } \mu = \frac{\sin i}{\sin r}$$

According to wave theory of light, if 'c' is the velocity of light in air or vacuum and 'v' is the velocity of light in the medium, then

Refractive Index of the medium

$$\mu = \frac{\text{Velocity of light in air (or) vacuum ('c')}{\text{Velocity of light in the medium ('v')}$$

## **FIBRE OPTICS**

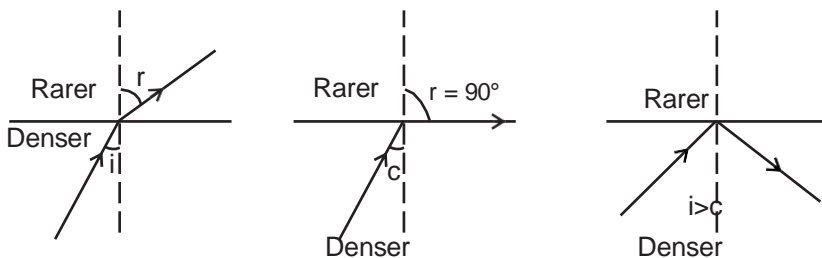
Fibre optics deals with the light propagation through thin glass fibres. Fibre optics plays an important role in the modern communication systems. In this system, television voice and digital data signals are transmitted from one place to another place in the form of light waves. The light waves are retransmitted along a thin cylindrical glass fibre by using the phenomenon of total internal reflection. It was first demonstrated by John Tyndall in 1870. But in the field of communication it was tried only from 1927. After the discovery of laser beam in 1960, fibre optical communication further developed and extended to medical field.

A bundle of fine fibres is used in the fibre optical communication. The bundle has the advantage of flexibility. A beam entering at one end can be transmitted at the other end through the fibre even though the fibre is curved. The study of such properties of a bundle of fibres is known as fibre optics.

### **Total Internal Reflection**

When a ray of light passes from a denser to a rarer medium, it is bent away from the normal. When the angle of incidence is gradually increased, the corresponding angle of refraction also increases. At a particular angle of incidence, the angle of refraction reaches to  $90^\circ$ . That particular value of angle of incidence is called the critical angle of the medium.

Critical angle is defined as a particular angle of incidence in the denser medium for which the corresponding angle of refraction in the rarer medium is  $90^\circ$ .



### Refraction

### Critical angle

### Total Internal Reflection

When the angle of incidence is further increased above the critical angle, the ray is totally reflected back in the same denser medium itself. This phenomenon is called total internal reflection.

Two conditions are to be satisfied for total internal reflection.

- 1) Light should travel from denser medium to rarer medium.
- 2) The angle of incidence should be greater than the critical angle of the medium.

According to Snell's law,

$$\mu_r^d = \frac{\sin i}{\sin r}$$

The refractive index of the rarer medium with respect to denser medium,

$$\mu_d^r = \frac{\sin r}{\sin i}$$

Also, the refractive index of the denser medium with respect to rarer medium

In critical angle position,  $i = c$  and  $r = 90^\circ$

$$\therefore \mu_d^r = \frac{\sin r}{\sin i} = \frac{\sin 90^\circ \sin i}{\sin c}$$

$$\text{i.e. } \mu_d^r = \frac{1}{\sin c} \quad (\text{Q } \sin 90^\circ = 1)$$

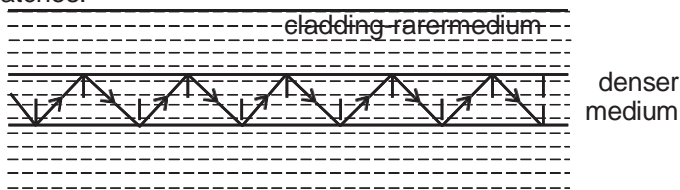
The sparkling of the diamond, the phenomenon of mirage and the propagation of light waves along optical fibre are due to the total internal reflection.

Sl.No.	Medium	Refractiveindex $\mu_{w.r.tair}$	Criticalangle:C
1	Water	1.33	$48^{\circ}45'$
2	Glass	1.5	$41^{\circ}48'$
3	Diamond	2.42	$24^{\circ}24'$

### Opticalfibrecableasawaveguide:

The optical fibre cables are used as wave guides for transmitting electromagnetic waves at optical frequencies. These frequencies are in the order of  $10^{15}$  Hz higher than that of radio wave frequency of  $10^{10}$  Hz. The number of information transmitted is directly proportional to the frequency of the transmitted signals. Hence 15,000 and more simultaneous telephone conversations are possible to transmit along a single optical fibre cable. But in a single pair of metallic cable, only 48 simultaneous telephone conversations can be transmitted in the form of electric current.

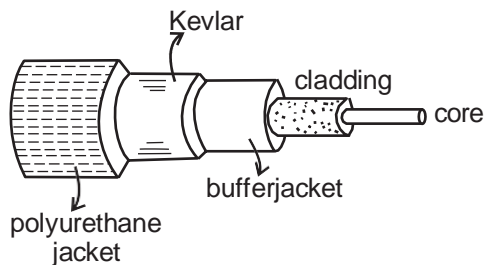
An optical fibre cable is a cylindrical core made of silica glass of radius 0.0002 cm to 0.001 cm. Around this core cladding is formed by the addition of small amount of boron, germanium or phosphorus to form the rarer medium. The cladding is necessary to retain the light waves within the core and also to provide mechanical strength. Further the cladding is coated with buffer jacket made of plastic to protect the fibre from moisture and scratches.



The buffer jacket is also surrounded by a thick tougher like Kevlar that provides toughness and tensile strength to the fibre. Finally the cable is covered by black polyurethane outer jacket that provides flexibility to the fibre.

When the optical signal enters at one end of the fibre, it is incident on the layer separating the core and cladding. The angle of incidence is greater than the critical angle. Since the core is a denser medium and the

cladding is a rarer medium, total internal reflections take place continuously throughout the core region as shown in the figure. Finally it comes out from the other end of the fibre. Hence the optical fibre functions as a waveguide.



### Advantages of optical fibre cables (O.F.C):

- 1) The fibre material is made of silica glass which is available in surplus in the world compared to copper or aluminium.
- 2) Silica is very much cheaper than copper.
- 3) There is no electrical danger, since no electric current flows through the fibre. It carries only optical signal.
- 4) Optical fibre can be handled more easily because of their lightweight, smaller in size and flexibility.
- 5) The optical signal transmitted through the fibre cannot be disturbed by the radio frequency interference.
- 6) Tapping (overhears secrets) of information is impossible in fibre optic systems.
- 7) A large number of signals can be simultaneously transmitted along the optical fibre cables compared to the metallic cables.
- 8) The loss of intensity of energy is less in optical fibre cables compared to the metallic cables. In the optical fibre cable the loss is 12.5 dB per km, whereas in the metallic cable the loss is 49 dB per km.
- 9) The optical fibre cables are having additional advantages of longer life, easy maintenance, low cost transmission, high quality signal transmission and efficient transmission.
- 10) The optical fibre cables are electrical insulators. They are not affected by any other electromagnetic wave and lightning.



## LASER

An atom consists of an electron cloud revolving around the central nucleus having protons and neutrons. The atoms are constantly in motion, they vibrate, move and rotate. Also they are in different states of excitation and hence have different energies. If some external energy in the form of light, heat, or electricity is supplied to an atom, the electrons from low energy level called ground state may jump to higher energy level called excited state. The electrons in the higher energy orbit cannot stay there for a long time. Therefore they return to the ground state spontaneously releasing the absorbed energy in the form of radiations.

### LASER

The term LASER stands for Light Amplification by Stimulated Emission of Radiation. The LASER is a device which amplifies light waves similar to the transistor which amplifies electric current. The laser is a powerful source of light.

### Characteristic of LASER

Laser radiation is different from normal light radiation. The characteristics of laser are given below:

1. Laser radiation is monochromatic. It contains only one particular wavelength (colour) of light. The wavelength is determined by the amount of energy released.
2. Laser radiation is coherent. In emitted radiation all the photons are having the same phase and amplitude.
3. Laser radiation is highly directional. Laser beam of light can travel very long distance, without much divergence.
4. Laser radiation is very intense. The intensity and hence the brightness of the laser radiation is high.

The radiation having single wavelength, same amplitude and same phase is called 'coherent radiation'.

### Principle of LASER

Laser works on the principle of quantum theory of radiation. When an atom in the lower energy  $E_1$  absorbs the photon energy of incident radiation, it is excited to the higher energy level of energy  $E_2$ . This process is called absorption. The number of absorption transition take place per

unit time per unit volume is directly proportional to the number of atoms in the lower energy level and the number of photons in the incident radiation.

The energy difference  $E_2 - E_1$  is called excitation energy and it is equal to the energy of the photon absorbed by the atom. In higher energy level ( $E_2$ ), the atom cannot remain long time, since it has a tendency to drop energy. So the atom jumps to the lower energy level ( $E_1$ ). At that time, the energy difference  $E_2 - E_1$  is emitted in the form of radiation as photon.

The energy of the photon released,  $E_2 - E_1 = h\gamma$ , where  $\gamma$  is the frequency of the radiation and  $h$ , the Planck's constant.

### **Spontaneous emission**

Spontaneous emission is the process of photon emission takes place immediately without any inducement during the transition of atoms from higher energy levels to lower energy levels. This is a random process in which the photons are emitted at different time. So, in the emitted radiation each photon has different phase and are incoherent.

The number of spontaneous transition ( $N_{sp}$ ) taking place per unit time per unit volume is directly proportional to the number of atoms in the higher energy level ( $N_2$ ).

$$N_{sp} \propto N_2 \text{ (or) } N_{sp} = AN_2$$

where  $A$  is a constant known as the probability of spontaneous transition per unit time. Also it is called Einstein's co-efficient of spontaneous emission.

### **Stimulated emission**

Einstein found a new process called stimulated emission to increase the number of transition of atoms from higher energy levels to lower energy levels.

Stimulated emission is the process of photon emission takes place by an inducement given by another photon incident on the atoms in higher energy levels. The energy of the photon emitted is equal to the energy of the photon incident.

Consider an atom in the higher energy level ( $E_2$ ). When an external radiation of photon energy  $E_2 - E_1$  is incident on the excited atom, the

photon stimulates the atom to make transition from high to lower energy level. As a result the same photon energy  $E_2 - E_1$  is emitted in the form of radiation. During this process, the stimulating photon and the photon emitted by the excited atom are emitted simultaneously in the same direction. Hence they are identical in phase, direction and frequency and are coherent. This process of stimulated emission is used to produce laser beam.

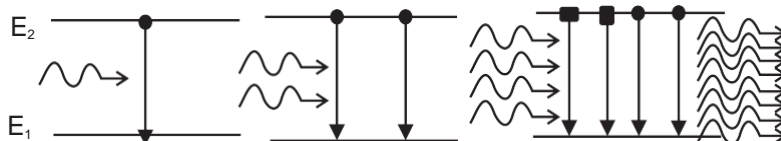
The number of stimulated transition ( $N_{st}$ ) taking place per unit time per volume is directly to

- 1) the number of atoms in the higher energy level ( $N_2$ )
- 2) The number of photons in the incident radiation ( $Q$ )

ie  $N_{st} \propto N_2 Q$  (or)  $N_{st} = B N_2 Q$

Where  $B$  is a constant known as the probability of stimulated transition per unit time. Also it is called Einstein's coefficient of stimulated emission.

The stimulated emission can be multiplied through a chain reaction as shown in the figure. When a single photon hits an atom in the higher energy level, two photons are emitted by stimulated emission.



Then these two photons hit on two atoms of higher level, four photons are emitted. This process is continued as a chain reaction and the photons are getting multiplied. Finally it leads to the emission of the powerful, coherent, monochromatic and highly directional beam of laser light. This is called Light amplification by stimulated emission of radiation. This amplification takes place only if there are more number of atoms in the excited state (higher energy level) than in the ground state (lower energy level).

## Population inversion

Population inversion is the establishment of the situation to make more number of atoms in the excited state required for the chain reaction of stimulated emission.

For achieving the population inversion, the number of atoms must be raised continuously to the excited state. When a sufficient energy is given to the atoms in the ground state, they are excited to the higher energy levels. This process is called pumping. If the atoms in the ground state are brought to the excited state by means of light energy, the process is called optical pumping.

## Types of Laser

The important types of laser are solid state lasers, gas lasers, liquid lasers, dye lasers and semiconductor lasers

Solid state lasers	: Ruby laser, Nd-YAG laser
Gas lasers	: He-Ne laser, CO <sub>2</sub> laser, Argon-iron laser
Liquid lasers	: Organic liquid lasers
Dye lasers	: Coumarin dye laser, Rhodamine 6G laser
Semiconductor lasers	: GaAs diode laser

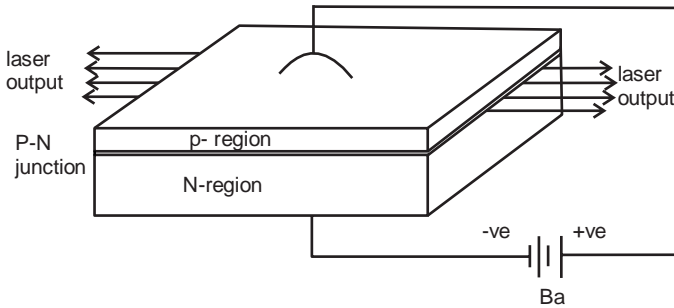
## Production of laser-GaAs-Semiconductor diode laser

In semiconductors, the lower energy level called valance band and the higher energy level called conduction band are separated by a certain amount of energy called energy gap or forbidden gap. If this energy is supplied to the electron in the valance band, it crosses the energy gap and enters into the conduction band by leaving holes in the valance band.

When the external photon energy equal to the energy gap is given to the semiconductor, it stimulates the more number of electrons to move from the conduction band to the valance band and recombine with the holes. Due to this recombination, energy is released at the junction in the form of coherent laser beam. Certain semiconductors like germanium and silicon are indirect band gap semiconductors in which there is no possibility of direct recombination of holes and electrons. Certain semiconductors like gallium and arsenide are direct band gap semiconductors and there is a large possibility of direct recombination of holes and electrons. Hence, GaAs is used to make light emitting diodes

and lasers. In the case of PN junction diode, the electron and hole densities can be increased at the junction by applying forward bias. The electrons and holes are repelled by the applied potential and they cross the PN junction. Then the holes and electrons recombine with each other producing coherent photons. These photons will stimulate more and more charge carriers for recombination and this results a light amplification.

A GaAs semiconductor diode laser is shown in the figure. It consists of a heavily doped PN-junction diode of crystalline gallium arsenide in the form of a plate of thickness 0.5mm. Electric current can be applied through the electrodes provided in the upper and lower surfaces of the diode. The sides of the junction are well polished and kept parallel to each other. It is convenient for the emitted photon to come out.



When the PN-junction diode is biased in forward direction, the population inversion is obtained. The electrons in the n-region are repelled by the applied -ve voltage and injected across the junction and enter into the p-region. If the current is small, a limited number of carriers undergo recombination. When a large current of order  $10^4$  ampere per  $\text{cm}^2$  is passed through the PN-junction, the rate of recombination increases and more number of photons are emitted. The emitted photons further stimulate more recombination. As a result the stimulated laser radiation is obtained along the PN junction. Thus GaAs - laser convert electricity into light. The efficiency of laser emission increases when GaAs diode is cooled.

### Uses of Laser:

1. Laser is used to find the long distances by direct pulse reflection method. Using this method, the distance between the moon and earth was measured accurately.
2. It is used to produce high temperature to melt and vapourise the metals in a very short period.
3. It is used to bore holes in hard substances like tungsten and diamond.
4. It can be modulated to transmit hundred messages at a time on radio, television and telephone.
5. The big size objects like aeroplanes, missile etc can be destroyed in a few seconds by passing a powerful laser beam on to them. For this reason, laser is called a "Death ray" instrument.
6. In surgery, a laser beam is used in spot-welding of a detached retina in the eye ball. It is also used for painless drilling and welding the surface of a tooth to prevent its decay.
7. The laser beam is not easily absorbed by water. So it can be used in underwater communication between submarines.

### PHOTOELECTRIC EFFECT

The emission of electrons from a metal surface when light or any other radiations of suitable wavelength incident on them is called photoelectric effect.

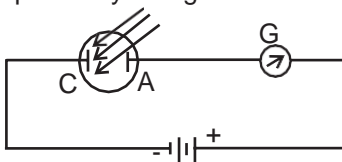
The photoelectric effect was discovered by Hertz in 1887. After that in 1888, Hallwachs found that only negatively charged particles have been emitted by the surface of the metal when ultra-violet radiations incident on that surface. Later in 1916, Einstein explained the photoelectric effect on the basis of Planck's quantum theory of light. For this, Einstein was awarded Nobel Prize in 1921.

According to quantum theory, light consists of photons and each photon has energy  $h\nu$ , where  $h$  is Planck's constant and  $\nu$  is the frequency of light. When a photon collides with an electron in the metal surface, the photon energy is transferred to the electron. Due to the addition of external energy, electrons are ejected out of the metal. The metal surface which emits photo electrons is called photo sensitive metal surface.

### Hallwachs experiment:

In the study of photoelectric effect, Hallwachs used ultraviolet radiations for the incidence on different zinc plates.

- 1) When ultraviolet radiations incident on a neutral zinc plate, the plate becomes positively charged.
- 2) When ultraviolet radiations incident on a negatively charged zinc plate, it lost its charge rapidly.
- 3) When ultraviolet radiations incident on a positively charged zinc plate, the plate becomes more positively charged.



Hallwachs also explained the photoelectric effect by using an experiment. In this arrangement, two zinc plates A and C are kept in an evacuated quartz bulb. The plate A acted as an anode which is connected to the +ve terminal of the battery. The plate C acted as a cathode which is connected to the -ve terminal of the battery.

A sensitive galvanometer is connected in series with the circuit. A window is provided on the quartz bulb to allow the ultraviolet radiations incident on the zinc plates. In the absence of ultraviolet radiations the galvanometer shows null deflections. That means there is no current flow through the circuit. When the ultraviolet radiations are incident on the cathode plate C, the galvanometer shows the deflection, due to the flow of electric current through the circuit. This electric current is called photoelectric current. When the radiations are incident on the anode plate A, there is no current flow through the circuit.

By this experiment Hallwachs came to the conclusion that only negatively charged particles can be emitted by the zinc plate under the action of ultraviolet radiations.

### Einstein's photoelectric equation

Einstein used the law of conservation of energy and Planck's quantum theory of light to explain the photoelectric effect. Einstein assumed that one photon ( $h\nu$ ) incident on the surface is used for two purposes.

Consider a photon of light of frequency  $\gamma$  is incident on a metal surface. The energy of the photon ( $h\gamma$ ) incident on the surface is used for two purposes. A part of the energy of the photon is used for the liberation of electron from the metal surface. This amount of energy is called photo electric work function ( $W_0$ ). The remaining energy of the photon is used in giving kinetic energy to the emitted electron.

Let  $v$  be the velocity of the emitted photoelectron and  $m$  its mass.

If the incident energy is very less, no electron is emitted from the metal surface. The metal surface required particular minimum energy ( $W_0$ ) just to emit the photo electron. Its value is different for different metals.

The kinetic energy of the photoelectron =  $\frac{1}{2}mv^2$

According to law of conservation of energy,  $h\gamma = W_0 + \frac{1}{2}mv^2$

$$(or) \quad \frac{1}{2}mv^2 = h\gamma - W_0 \quad (1)$$

Let  $\gamma_0$  be the minimum frequency of the light called threshold frequency having minimum energy ( $h\gamma_0$ ) to start photoelectric effect.

i.e.

The photoelectric work function  $W_0 = h\gamma_0$ . So

substituting the value of  $W_0$  in equation (1),

The above equation is called Einstein's photoelectric equation.

$$\frac{1}{2}mv^2 = h\gamma - h\gamma_0$$

i.e.  $\frac{1}{2}mv^2 = h(\gamma - \gamma_0)$

### Law of Photoelectric emission

The experimental study of photoelectric effect expresses the following results as fundamental laws of photoelectric emission.

- 1) The number of photo electrons emitted per second (photo electric current) is directly proportional to the intensity of the incident radiation.
- 2) For each metal surface there must be a certain minimum frequency of incident radiation to start photoelectric effect. This frequency is called threshold frequency.



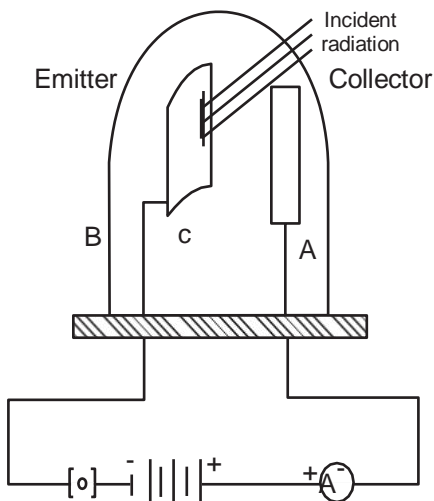
- 3) The maximum velocity of photoelectron is independent of the intensity of the incident radiation but depends upon the frequency of the incident radiation.
- 4) The photo electric emission is instantaneous process. The time lag between the incident of radiation and emission of photo electrons is very small and it is less than  $10^{-9}$  second.

### Photoelectric cells

A device which converts light energy into electrical energy is called photoelectric cell. The photoelectric cells are mainly classified into three types. They are

1. Photoemissive cells,
- 2) Photovoltaic cells and,
- 3) Photoconductive cells.

**1) Photo emissive Cell:** Photo emissive cell is based on the emission of electrons from a metal cathode, when the radiations incident on it. Hence the current in the circuit increases. The photoemissive cell is shown in the figure. It consists of a glass bulb B containing a semi cylindrical photocathode C.



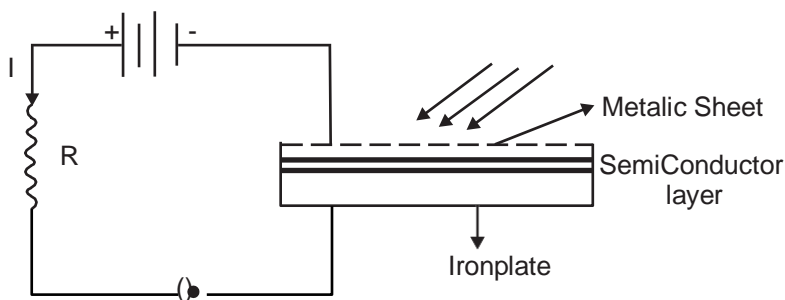
It acts as an emitter and coated with some photosensitive material like sodium oxide, cesium oxide etc. The anode A is in the form of a rod made of platinum or nickel fixed along the axis of the semi-cylindrical

cathode. It acts as a collector. The cathode is connected to the -ve potential and the anode is connected to the +ve potential.

When the radiations of suitable frequency are incident on the cathode, it emits photo electrons. These electrons are attracted by the anode and the circuit is completed. Hence the current is set up in the circuit. The strength of the current increases with the intensity of the incident radiation.

**2 Photo conductive cell:** Photo conductive cell is based on the principle that resistance of some semi conductor materials decreases when the radiations incident on it. Hence the current in the external circuit increases.

The photoconductive cell is shown in the figure. It consists of a thin layer of semiconductors such as selenium, lead sulphide rigidly fixed

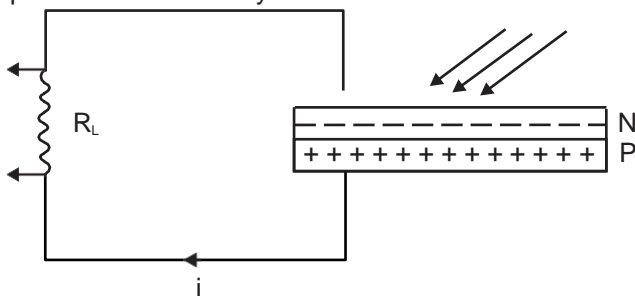


between an iron plate and transparent metallic sheet. A battery is connected between the iron plate and the top metallic sheet through a resistance  $R$ . If the circuit is closed, a weak current may be set up in the circuit. When the radiations are incident on the top sheet, the resistance of the semiconductor decreases and the current increases. The strength of the current is proportional to the intensity of radiations.

**3. Photo Voltaic Cell:** Photo voltaic cell is based on the development of voltage across the PN-junction of a semiconductor diode when the radiations incident on it. The photo voltaic cell is shown in the figure. It is made of semiconductors of silicon or selenium crystals in the form of PN-junction diode. The diode is connected externally by means of a load resistance

$R_L$ . When the radiations are incident on the N-type

semiconductor, the photo electrons are emitted and circulated through the load resistance. This constitutes photoelectric current 'i' and voltage is developed across the load resistance. The strength of the current is directly proportional to the intensity of the incident radiations.



### Applications of photocells:

Some of the important applications of photocells are given below.

- 1) The photoelectric cells are used in the exposure device of camera to determine the correct time of exposure.
- 2) Today photo voltaic cells commonly called solar cells are used to generate electric power in a large scale to run the electric motor and to glow electric lamps.
- 3) They are used for reproducing sound in cinematography.
- 4) They are used in automatic opening and closing of door.
- 5) They are used in fire alarm and burglar alarm.
- 6) They are used in space ships for harnessing solar energy.

## REMOTE SENSING

### Introduction

Remote sensing is the process of sensing and measuring the objects in a long distance without physical contact with them. A device used for the above purpose is called remote sensor.

Remote sensors are mounted on the aircraft or satellite platforms. These sensors receive the reflected or emitted radiations from a particular place on the earth's surface and measure the energy. These measurements are being made at a large number of points from the image of the earth's surfaces. These sensed images are interpreted in

terms of geographical variations. The active power of the remote sensors depends upon the detection of energy emitted or reflected by the object.

**Types of remote sensing (Active and Passive):**

The system of remote sensing is classified into two types. They are

- 1) Active remote sensing
- 2) Passive remote sensing

(1) Active remote sensing is the process of sensing the distant target in which the sensors direct their own generated radiation towards the target and then it detects the amount of energy reflected by the target.

Examples: Radar, Radar Altimeter, Wind scatterometer, Synthetic Aperture Radar (SAR) and Camera with flash.

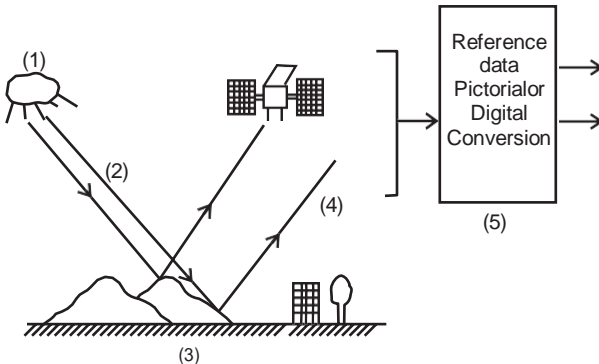
(2) Passive remote sensing is the process of sensing the distant object in which the sensors detect only the amount of energy radiated by the object.

Examples: Aerial photography, Camera without flash and Microwave radiometer.

**Components of remote sensing:**

The electromagnetic remote sensing process is divided into two components namely

- (1) Data acquisition and (2) Data analysis

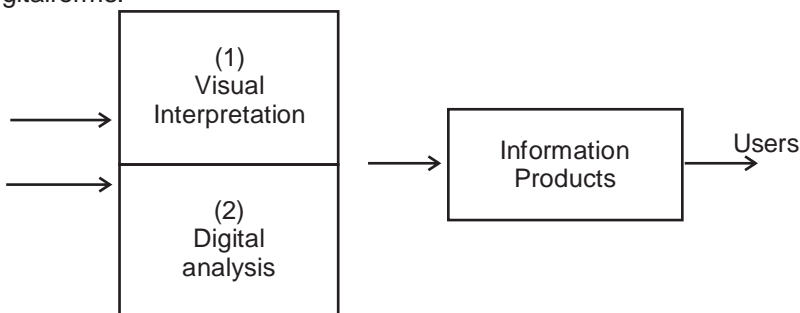


**Date acquisition:** The schematic diagram of data acquisition is shown in the figure. In this process, five distinct elements are involved.

They are

- 1) Energy Sources.
- 2) Propagation of energy through the atmosphere.
- 3) Energy interaction on earth's surface.
- 4) Spaceborne sensors record the reflected energy.
- 5) Conversion of sensor data into the form of pictures or digital information.

The electromagnetic radiations emitted by the energy source are propagated through the atmosphere and incident on the surface of the earth. Then the radiations are reflected back towards the space borne remote sensors. These sensors generate the receiving data into picture or digital forms.



**Data analysis:** This schematic diagram of data analysis is shown in the figure. The data analysis is divided into two separate processes. They are

- 1) visual interpretation technique
- 2) Digital analysis

In visual interpretation techniques, various viewing instruments are used for examining and analysing the pictorial data. In digital analysis the computers are used to analyse the digital data. All analyzing data are viewed in the information products by users. The data analysis can be carried only with the help of referenced data.

The process that involves examination and analysis of pictorial data is called visual image interpretation techniques. This process involves the use of fundamental picture elements like tone, texture, pattern, shape and size in order to identify the various objects. Stereoscopic instruments are used for this purpose.

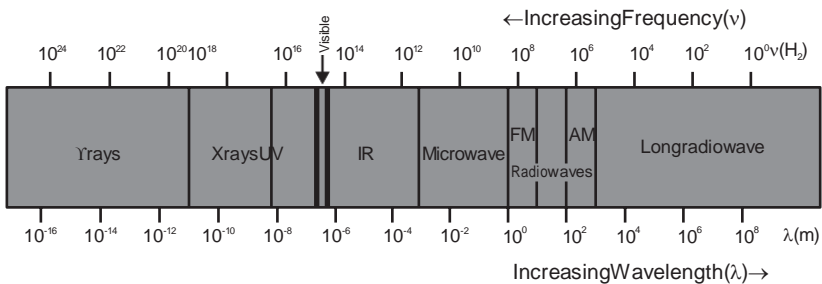
Use of computers to analyse digital data is known as digital image processing techniques. If the data is available in digital form, it can be analysed on interactive computer systems for getting statistical data. This statistical data is analysed using computers by comparing with the actual “signatures” of the objects collected through field visits. This system of classification of objects is quite accurate.

**Referencedata:**

Whether it is pictorial data analysis or digital data analysis, referencedata should be generated as supporting data. Referencedata is called ground truth or field check. The reference data is used to analyse and interpret remotely sensed data and also to verify information obtained from remote sensing data. The acquisition of referencedata is through actual field visits.

The collection of referencedata may depend upon time-critical or time-stable measurements. The time-critical measurements are those vegetation conditions, water pollutants and climatic conditions. The time-stable are other geological measurements.

**The Electromagnetic Spectrum**



Radio waves, microwaves, infrared radiation, light rays, ultraviolet rays, X-rays, and gamma rays are components of electromagnetic waves. All these electromagnetic rays are collectively called as electromagnetic spectrum.

The electromagnetic spectrum is a continuum of all electromagnetic waves arranged according to frequency and wavelength.

## **Radiowaves**

Radiowaves are used to transmit radio and television signals. Radio waves have wavelengths that range from less than a centimetre to tens or even hundreds of metres. Radio waves with wavelengths of a few centimetres can be transmitted from a satellite or airplane antenna. The reflected waves can be used to form an image of the ground in complete darkness or through clouds.

## **Microwave**

Microwave wavelengths range approximately from one millimetre (the thickness of a pencil lead) to one metre. Microwaves are emitted from the Earth, atmosphere, and from objects such as cars and planes. These microwaves can be detected to give information, about the object that emitted the microwaves.

## **Infrared Radiation**

Infrared is the region of the electromagnetic spectrum that extends from the visible region to about one millimetre in wavelength. Infrared waves are thermal radiations. Infrared images obtained by sensors in satellites and airplanes can yield important information on the health of crops and can help to see forest fires even when they are covered by thick smoke.

## **Visible light**

The visible light is the portion of the electromagnetic spectrum with wavelengths between 400 to 700 nanometre. It is only the part of the electromagnetic spectrum that we can see. Visible waves have great utility for the remote sensing of vegetation and for the identification of different objects by their visible colours, during daytime.

## **Ultraviolet Radiation**

Ultraviolet radiation has a range of wavelengths from 400 billionths to about 10 billionths of a metre. Sunlight contains ultraviolet waves which can burn our skin. Most of these are blocked by ozone in the Earth's upper atmosphere. A small dose of ultraviolet radiation is beneficial to humans, but larger doses cause skin cancer and cataracts.



## **X-rays**

X-

rays are high energy waves which have great penetrating power and are used extensively in medical applications and in inspecting welds. The wavelength range is from about ten billionth to about 10 trillionth of a metre.

## **Gamma rays**

Gamma rays have wavelengths of less than about ten trillionth of a metre. They are more penetrating than X-rays. Gamma rays are generated by radioactive materials and in nuclear explosions. They are used in many medical applications, like cancer treatment.

## **Microwave remote sensing:**

The microwave remote sensing is one of the systems of remote sensing. In this system electromagnetic radiations are used in the wavelength region of microwave. This system of remote sensing provides useful information about the earth's atmosphere, land and ocean.

The region of the spectrum composed of electromagnetic radiation with wavelength between 1 mm and 1 m is called the microwave band. The microwave band is a valuable region for remote sensing in view of two distinctive features.

1. Microwaves are capable of penetrating the atmosphere under almost all conditions. Depending on the wavelengths involved, microwaves can 'see through' haze, light rain, snow, clouds and smoke.
2. Microwave reflections or emissions from earth materials do not interfere with the microwave radiation emitted by sun, as it is very negligible.

The examples for microwave remote sensing are

- 1) Microwave radiometer
- 2) Radar altimeter
- 3) Wind scatterometer
- 4) Synthetic aperture radar (SAR)

A microwave radiometer is a passive device which records the natural microwave emission from the earth. It is used to measure the total water content of atmosphere within the field of view. A radar altimeter sends out pulses of microwave signals and then records the reflected signal from the earth's surface. The height can be measured by using the time delay of the return signals.

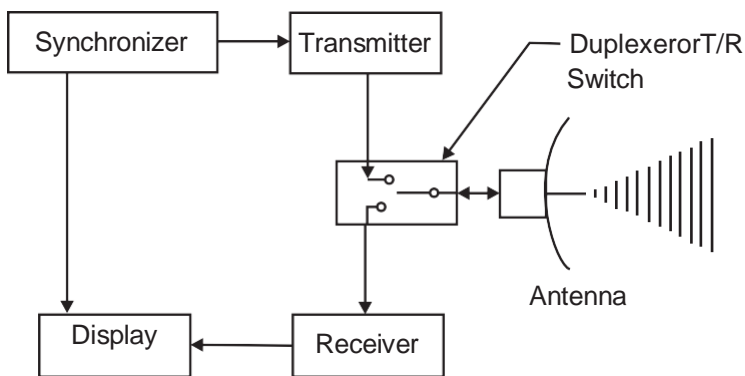
A wind scatterometer sends out pulses of microwave along several directions and records the magnitude of the signals scattered back from the ocean surface. It can be used to measure wind speed and direction over the ocean surface. Synthetic aperture radar is a device used in airborne or space borne platforms to send out pulses of microwave signal towards the earth and then record the reflected signal from the earth's surface. It can be used to generate high resolution images of the earth's surface.

The microwave remote sensing system is also classified into two types. There are

- 1) Active sensor type and
- 2) Passive sensor type

Imaging radar, wind scatterometer and radar altimeter are the examples for active type and microwave radiometer is an example for passive type.

**RADAR**



The RADAR is an acronym derived from Radio Detection and Ranging. Radar is an electromagnetic device for detecting the presence and location of objects. The presence of the object, their range and direction are determined by the transmission and return of electromagnetic energy. The radar antenna transmits short burst pulses of energy to the target and echoes from this target carry information about the position and quality of the illuminated object. The block diagram of an elementary pulsed radar set is given above.

Radar consists of a transmitter and a receiver each connected to a directional antenna through a Duplexer or Transmitter/Receiver switch. The transmitter sends out microwave power through the antenna. The receiver collects as much energy as possible from the echoes reflected in its direction by the target. In most of the radar systems the receiving antenna is same as the transmitting antenna.

The transmitter generates a short pulse and sends it to the antenna via a duplexer. As soon as a small fraction of the pulse is fed to the duplexer, it disconnects the transmitter from the antenna and connects the receiver to it. The transmitted pulse is reflected by the target and a portion of it reaches the antenna. Now the duplexer connects the receiver to the antenna so that the returning echoes can be processed. The processed pulses are fed to the necessary display devices which give the nature and position (range) of the target. In most radars, the antenna moves in a predetermined pattern and this process is known as scanning.

## QUESTION

### Part-A

1. What is Retraction?
2. Write the laws of retraction.
3. Define refractive index.
4. What is total internal reflection?
5. Define critical angle.
6. Expand the acronym LASER.
7. Write any one of the characteristics of LASER.
8. What is the principle of LASER?
9. What is spontaneous emission?
10. What is the stimulated emission?
11. What is population inversion?
12. Write any two uses of LASER.
13. What is photoelectric effect?
14. What are the three types of photoelectric cells?
15. Write one of the applications of photocells.
16. What is remote sensing?
17. What is active remote sensing?
18. What is passive remote sensing?
19. What are the important components of remote sensing?

20. What are the uses of referenced data?
21. What is microwave remote sensing?
22. Why microwave remote sensing is preferred?
23. Expand the acronym RADAR.

### **Part B**

1. Explain the phenomenon of total internal reflection with the help of ray diagram and give the relation between critical angle and refractive index.
2. Explain with diagram how optical fibre cable functions as a wave guide.
3. List the advantages of optical fibre cables (O.F.C)
4. Write a note on spontaneous emission, stimulated emission and population inversion.
5. Describe the production of LASER using GaAs semiconductor diode source.
6. List the uses of Laser
7. Explain Hallwach's experimental arrangement used to study the photoelectric effect.
8. Derive Einstein's photoelectric equation.
9. Explain the laws of photoelectric emission.
10. Explain photoemissive cell with neat circuit diagram.
11. Explain photoconductive cell with neat circuit diagram
12. Explain photovoltaic cell with neat circuit diagram.
13. Explain the applications of photocells.
14. Describe the data acquisition and data analysis of remote sensing.
15. Describe the components of electromagnetic spectrum used in remote sensing.
16. Explain microwave remote sensing.
17. Explain the working of RADAR with block diagram.

## UNIT - IV

# ELECTRICITY

### Introduction

The flow of free electrons in a conductor constitutes electric current. The current is defined as the rate of flow of charges across any cross sectional area of a conductor. The current is expressed in ampere and it is a scalar quantity. The direction of current is the direction of flow of positive charges or opposite to the direction of flow of electrons. The electric current is the flow of positive charge from a point at a higher potential to a point at a lower potential. Current flows only if there are charge carriers present in a conductor which can move freely when an electric field source is connected between the ends of the conductor. When charges are in motion, they produce magnetic effects. The current flowing through a conductor is one ampere if a charge of one coulomb flows through any section of it per second.

### ELECTRICAL CIRCUITS

In an electrical circuit there must be a source of energy which provides energy to electrons so that they can move and constitute an electric current. For easy flow of current, the connecting wires must be made of conductor material. The energy source creates a potential difference across the circuit and therefore it forces electric charges to move along the conductor. When a voltage ( $V$ ) is applied across a conductor, a current ( $I$ ) flows in the conductor. The magnitude of current in a conductor depends on its resistance and the resistance plays an important part in the conduction.

### Ohm's law

George Simon Ohm in 1826 established the relationship between potential difference and current in a conductor which is known as Ohm's law. It states that the current ( $I$ ) flowing through a conductor is directly proportional to the potential difference ( $V$ ) across the ends of the conductor, if its temperature remains the same.

$$i \propto V$$

$$\text{or } V \propto i$$

$$\text{or } V = Ri$$

Where  $R$  is called the resistance of the conductor.

The unit of resistance is Ohm with symbol ' $\Omega$ '

There are several materials which do not obey Ohm's law. Conductor which do not obey ohm's law are called non-ohmic conductors. This is due to increase in the resistance of the conductor when the current increases. Transistors, crystal rectifier etc., do not obey Ohm's law.

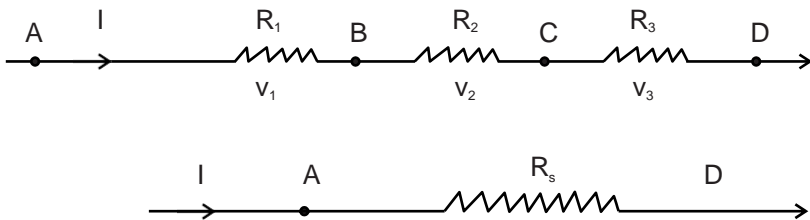
### Laws of Resistances

When more than one resistor is to be connected in an electric circuit they can be connected either in series or in parallel. The equivalent resistance offered in the circuit by a group of resistor depends on the way of their combination.

### Resistances in Series

The resistances are connected in series combination will increase the effective resistance in the circuit. When a number of conductors are connected end to end so that the same current flows through all of them, then they are said to be in the series combination.

Consider  $AB$ ,  $BC$  and  $CD$  are three conductors of resistances  $R_1$ ,  $R_2$  and  $R_3$  respectively connected in series as shown in the figure. Let  $i$  be the current flowing through them and  $V_1$ ,  $V_2$  and  $V_3$  be the potential differences across  $R_1$ ,  $R_2$  and  $R_3$  respectively.



Let  $R_s$  be the equivalent resistance ( $R_s$ ) carrying the same current  $I$  and having potential difference  $V$  across the ends  $A$  and  $D$  such that

$$V = V_1 + V_2 + V_3$$

Applying Ohm's law we get,

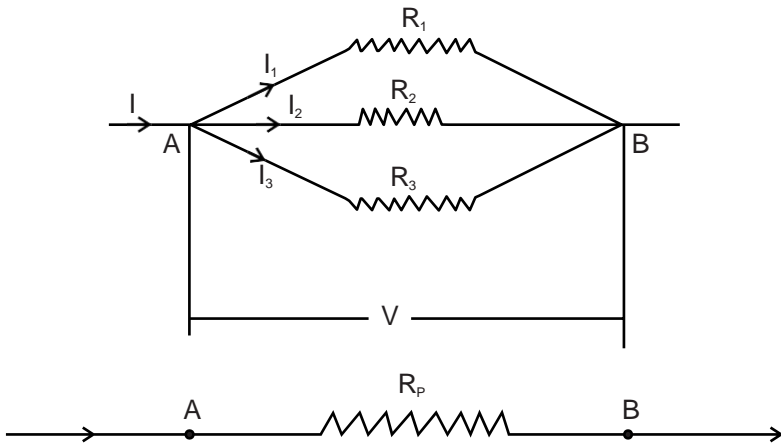
$$IR_s = IR_1 + IR_2 + IR_3$$

i.e.  $R_s = R_1 + R_2 + R_3$

Thus, when a number of conductors are connected in series, the equivalent resistance is equal to the sum of the individual resistances.

### Resistances in Parallel

The resistances are connected in parallel combination to decrease the effective resistance in the circuit. When a number of resistors are connected to common terminals so that the potential difference across



each conductor is the same, then they are said to be in parallel combination.

Consider  $R_1$ ,  $R_2$  and  $R_3$  are the resistances of the conductors connected between the point  $A$  and  $B$  and  $V$  is the common potential

difference across AB. Let  $I_1$ ,  $I_2$  and  $I_3$  be the current through the individual branches.

Let AB be the equivalent resistance ( $R_p$ ) having a potential difference  $V$  and carrying a current  $I$  such that,

$$I = I_1 + I_2 + I_3$$

Applying Ohm's law we get,

Thus, when a number of conductors are connected in parallel, the reciprocal of the equivalent resistance is equal to the sum of the reciprocal of the individual resistances.

$$\frac{V}{R_p} = \frac{V}{R_1} + \frac{V}{R_2} + \frac{V}{R_3}$$

ie,  $\frac{1}{R_p} = \frac{1}{R_1} + \frac{1}{R_2} + \frac{1}{R_3}$

### Resistivity

The resistance of a conductor is directly proportional to the length of the conductor and inversely proportional to the area of cross section of the conductor.

is called the resistivity or specific resistance of the material of the conductor.

$$R \propto \frac{\ell}{A}$$

ie  $R \propto \frac{\ell}{A}$

$$R = (\rho) \frac{\ell}{A}$$

$$\therefore \rho = \frac{RA}{\ell}$$

conductor.

The resistivity or specific resistance of a material is defined as the resistance offered by a conductor of unit length having unit area of cross section. The unit of resistivity is Ohm metre ( $\Omega m$ ).

### Conductivity

The reciprocal of resistivity of the material of a conductor is called as its conductivity. It is denoted by  $\sigma$



Based on the above expression of conductivity, the unit of conductivity is mho/m.

$$\sigma = \frac{l}{\rho}$$

$$\sigma = \frac{l}{RA}$$

### Kirchoff's current and voltage laws

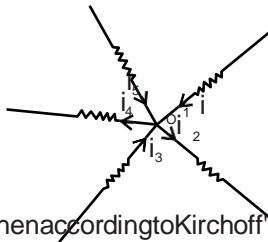
In electrical circuits which consist of simple arrangement of resistors in series or in parallel with a given source of emf, all calculations can be carried out using Ohm's law. But in more complicated circuits where there are junctions between more conductors and where more than one source of emf are active, then Kirchoff's laws can be useful for calculation purposes. There are two laws called Kirchoff's current law and Kirchoff's voltage law.

#### First law – Kirchoff's Current law or junction law

“In any network of conductors, the algebraic sum of the currents meeting at any junction is zero”.

The convention applied in the junction rule is that the current flowing towards a junction is taken as positive and the current flowing away from the junction is taken as negative.

Let  $i_1, i_2, i_3, i_4$  and  $i_5$  be the current passing through the five conductors



and meeting at the junction 'O'. Then according to Kirchoff's first law,  $i_1 -$

$$i_2 + i_3 - i_4 + i_5 = 0$$

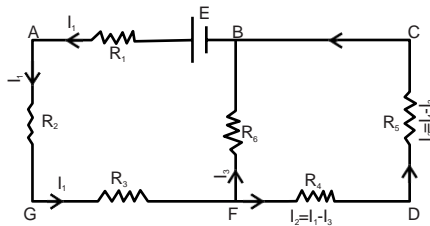
$$\text{or } i_1 + i_3 + i_5 = i_2 + i_4$$

That is, the sum of the currents flowing toward the junction is equal to the sum of the currents flowing away from the junction.

## Secondlaw–Kirchoff'sVoltagelawormeshlaw

“In any closed electrical network, the algebraic sum of the products of current and resistance in each branch of a network is equal to the algebraic sum of the emfs in that network.”

The current direction may be assumed either clockwise or anti-clockwise. The convention applied is, if the current flows in the same direction in which we are considering the network or mesh, it is taken as positive and the current in opposite direction is taken as negative. The emf which sends the current in the same direction in which the mesh is



considered, it is taken to be positive and emf which sends the current in the opposite direction is taken to be negative.

For the network BAGFB we have

$$I_1 R_1 + I R_2 + I_1 R_3 + I_3 R_6 = E$$

For the network FB CDF we have

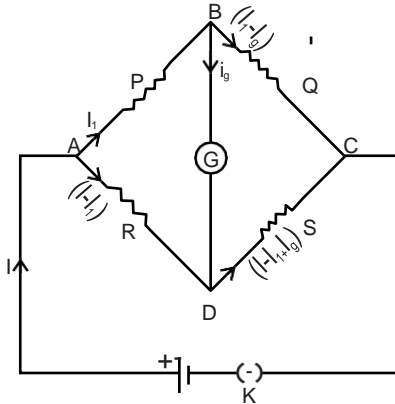
$$I_3 R_6 - (I_1 - I_3) R_5 - (I_1 - I_3) R_4 = 0$$

## Condition for balancing the Wheatstone's network

Wheatstone bridge is a very accurate and precise method of determining an unknown resistance and it is an important application of Kirchoff's laws. Since it is a null method, the measurement of unknown resistance is not affected by the internal resistance of the battery used.

Wheatstone bridge or network consists of four resistances P, Q, R and S connected in the form of a quadrilateral ABCD. A galvanometer of resistance G is connected between the opposite end points B and D. A cell of emf  $E$  and a key are connected between the opposite end points A and C.

When the circuit is closed different currents flow through various branches of the network as shown in the diagram. According to Kirchoff's



current  $I_g$  is the current passing through the galvanometer. Considering the network ABDA, applying Kirchoff's second law,

$$I_1 P + I_g G - (I - I_1) R = 0$$

or  $I_1 P + I_g G = (I - I_1) R$  (1)

Considering the network BCDB, applying Kirchoff's second law

$$(I - I_g) Q - (I - I_1 + I_g) S - I_g G = 0$$

or  $(I - I_g) Q = (I - I_1 + I_g) S + I_g G$  (2)

The four resistances P, Q, R and S are adjusted so that there is no deflection in the galvanometer.

i.e.  $I_g = 0$ . Now the bridge is said to be balanced.

Now, equation (1) becomes

$$I_1 P = (I - I_1) R$$
 (3)

and equation (2) becomes

$$\frac{I_1 P}{(I - I_1) S} = \frac{(I - I_1) R}{(I - I_1) S}$$

$$\frac{P}{Q} = \frac{R}{S}$$

$$I_1 Q = (I - I_1) S$$
 (4)

dividing (3) by (4) we get

This is the balancing condition of Wheatstone's bridge network.

### Solved Examples

- 1) Calculate the resistance of a wire of length 1.5 m and diameter 0.8 mm.

The specific resistance of the material of the wire is  $45 \times 10^{-8}$  ohm m

$$l = 1.5 \text{ m}$$

$$d = 0.8 \text{ mm} \quad r = 0.4 \text{ mm}$$

$$= 0.4 \times 10^{-3} \text{ m}$$

$$\rho = 45 \times 10^{-8} \text{ ohm m}$$

We know specific resistance  $\rho = \frac{RA}{l}$

$$\text{Resistance } R = \frac{\rho l}{A}$$

$$\text{ie } R = \frac{\rho l}{\pi r^2}$$

$$\frac{45 \times 10^{-8} \times 1.5}{3.14 \times (0.4 \times 10^{-3})^2} = 1.343 \text{ ohm.}$$

Given

$$R = 1.5 \text{ ohm}$$

$$\text{Diameter} = 0.32 \text{ cm} \quad r = 0.16 \text{ cm}$$

$$r = 0.16 \times 10^{-2} \text{ m}$$

$$l = 80 \text{ cm}$$

$$= 80 \times 10^{-2} \text{ m.}$$

We know specific resistance  $\rho = \frac{RA}{l}$

$$= \frac{R \pi r^2}{l}$$

$$\frac{1.5 \times 3.14 \times (0.16 \times 10^{-2})^2}{80 \times 10^{-2}}$$

$$= 15.07 \times 10^{-6} \text{ ohm m}$$

2) Calculate the resistivity of a material of a wire whose resistance is 1.5

ohm with a diameter of 0.32 cm and length 80 cm. Given

## HEATING AND CHEMICAL EFFECTS OF CURRENT

If a positive charge is moved from a point at a higher potential to a point at a lower potential, energy is liberated. When current flows in a conductor, the electrons collide with the ions because of the acceleration of electrons and energy is transferred to the ions. Therefore, there is an increase in the energy of the ions and the temperature of the conductor rises and thereby it is heated. This is called heating effect of electric current. This heating effect was studied by Joule.

### Joule's law of heating

The quantity of heat developed ( $H$ ) in a conductor is

- (i) directly proportional to the square of the current passing through the conductor i.e.  $H \propto I^2$
- (ii) directly proportional to the resistance of the conductor i.e.  $H \propto R$
- (iii) directly proportional to the time of flow of current i.e.  $H \propto t$

$$\begin{aligned} \therefore H &= I^2 \frac{E}{I} t \\ &= I^2 R t \\ H &= I^2 R t \end{aligned}$$

where  $I$  is in ampere,  $R$  is in ohm and  $t$  is in second

$$\text{But } R = \frac{E}{I}$$

$$H = E I t \text{ joule}$$

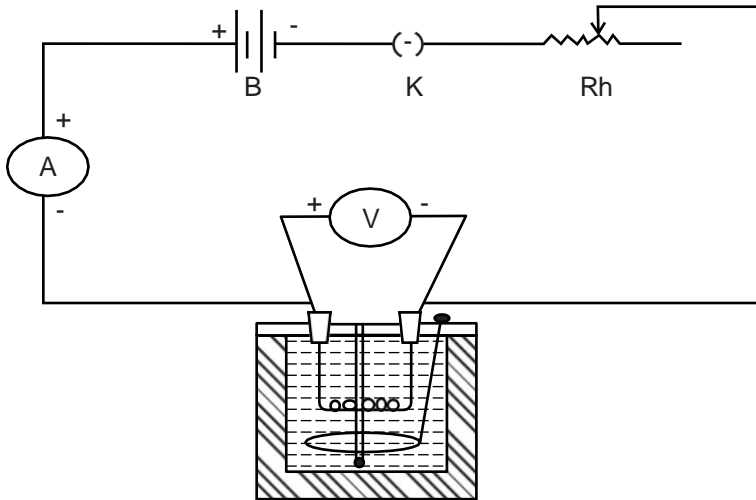
The heating effect of electric current is applied in the working principle of electric fuse wire, electric heater, electric welding and electric furnace.

### Experimental determination of specific heat capacity of a liquid using Joule's calorimeter

The specific heat capacity of a liquid can be determined experimentally using Joule's calorimeter.

The joule's calorimeter is a polished copper or aluminum vessel containing the given liquid.

It is covered by a non-conducting lid and a suitable resistance coil is immersed in the liquid and its ends are connected to connecting terminals provided in the lid. There are holes in the lid for a stirrer and inserting a thermometer. The calorimeter is placed inside a wooden box and it is filled with cotton or wool to minimize the radiation losses.



The mass of the clean, empty calorimeter with stirrer is determined and it is taken as  $m_1$  kg. Now the calorimeter is filled with the given liquid and its mass is determined with stirrer and it is taken as  $m_2$  kg. The heating coil should be immersed well in the liquid. The calorimeter is now placed inside the wooden box.

A battery, an ammeter, a key and a rheostat are connected in series with the coil and a voltmeter is connected in parallel across the coil. The current is adjusted for about one ampere. The circuit is kept open. The ammeter reading and the voltmeter reading are noted as  $I$  and  $E$ . Now, a thermometer is inserted through the hole in the lid. The initial temperature of the liquid is noted ( $\theta_1$  °C). When the circuit is closed, simultaneously a stop clock is started. The liquid is stirred well and when the temperature rises about 5 °C, the circuit is kept open and the stop

clock is stopped and the final temperature  $\theta_2^\circ$  is noted. The time is noted as  $t$  second.

The heat produced by the heating coil =  $EIt$  joule.

Heat is absorbed by both the calorimeter and the liquid.

Heat absorbed by the calorimeter =  $m_1s(\theta_2 - \theta_1)$  joule

where 's' is the specific heat capacity of the material of the calorimeter.

[Heat absorbed = mass  $\times$  specific heat capacity  $\times$  rise in temperature]

$$\frac{EIt}{(m_2 - m_1)(\theta_2 - \theta_1)} = \frac{m_1s}{(m_2 - m_1)} \text{ J kg}^{-1} \text{ K}^{-1}$$

Heat absorbed by the liquid =  $(m_2 - m_1)S(\theta_2 - \theta_1)$  joule.

Heat produced = Heat absorbed

i.e.  $EIt = m_1s(\theta_2 - \theta_1) + (m_2 - m_1)S(\theta_2 - \theta_1)$

The specific heat capacity of the liquid  $S$  is given by

## CHEMICAL EFFECT OF CURRENT

The phenomenon of conduction of electricity through liquid was first studied by Faraday in 1833. Pure liquid does not conduct electricity, but the conductivity of a liquid increases if the solution contains an acid. The molecules of a salt in solution are supposed to break up into two parts, positive and negative ions and this process is called dissociation.

When an electric field is applied between two electrodes dipping in the solution, the positively charged ions move towards the negative electrode (cathode) and they are called as Cations. The negatively charged ions move towards the positive electrode (anode) and they are called as Anions.

The process of chemical decomposition of an electrolyte by the passage of electric current through it is called electrolysis.

The liquid containing the compound is called an electrolyte and the vessel containing the electrolyte is called a voltameter.

### Faraday's laws of electrolysis

The two laws of electrolysis are stated as follows.

### First law

The mass of an element liberated at cathode from an electrolyte by passing of electric current is directly proportional to

(i) The strength of the current (I)

(ii) The time for which the current is passed through the electrolyte (t) i.e.

$$\propto I$$

$$\text{and } m \propto t$$

$$\text{i.e. } m \propto It \text{ or}$$

$$m = ZIt$$

Here, the constant Z is known as the electrochemical equivalent (e.c.e) of the element.

### Second law

If the same current passes through different electrolytes for the

$$m_1 \propto E_1;$$

$$m_2 \propto E_2 \text{ and } m_3 \propto E_3$$

$$\text{or } \frac{m_1}{E_1} = \frac{m_2}{E_2} = \frac{m_3}{E_3}$$

same time then the masses of the elements liberated from electrolytes are directly proportional to their respective chemical equivalent weights.

Let  $m_1$ ,  $m_2$  and  $m_3$  be the masses of the elements liberated by the passage of the same current through the electrolytes for the same time whose chemical equivalent weights are  $E_1$ ,  $E_2$  and  $E_3$ , then according to second law,

### Electro-chemical equivalent (e.c.e) of an element

According to Faraday's First law of electrolysis

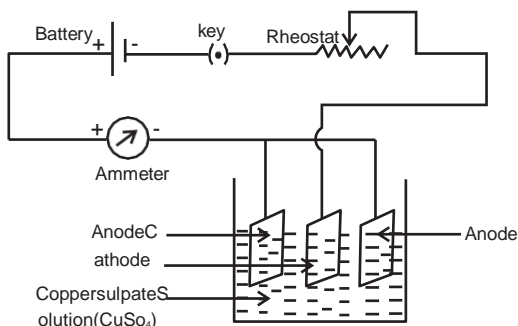
$$m \propto It$$

$$\text{or } m = ZIt$$

Where Z is a constant called electrochemical equivalent of the given element



When  $i = 1$  ampere and  $t = 1$  second, then  $Z = m$



Therefore the electro-chemical equivalent of an element is defined as the mass of the element liberated from an electrolyte when one ampere current is passed through it for one second. Since  $Z = m/It$  the unit of e.c.e is kilogram per ampere second or  $\text{kg C}^{-1}$  (ampere second = coulomb).

### Experimental determination of e.c.e of copper

Copper voltameter is used for the determination of electro-chemical equivalent of copper. It consists of a glass vessel containing copper sulphate solution and this is the electrolyte. There are three electrodes of copper plates dipped in the solution. The two outer plates are connected to the positive terminal of the battery through the ammeter and they act as anodes. The central copper plate is connected to the negative terminal of the battery and this plate is the cathode plate.

A battery, a plug key, a rheostat and an ammeter are connected in series with the copper voltameter. The circuit is closed and the rheostat is adjusted to send suitable current for a while and then the circuit is kept open.

The cathode plate is taken out, polished well with sand paper, washed and dried. The mass of the cathode is determined carefully correct to a milligram, using a digital balance. The cathode is then kept in its position and when the circuit is closed, simultaneously a stop clock is started. The current value is maintained the same by adjusting the rheostat, if necessary, and it is allowed to pass for thirty minutes. After the

time the current is stopped and the cathode plate is taken out, rinsed with water, dried and then weighed correct to a milligram.

The difference between the final and the initial mass gives the mass of copper deposited on the cathode plate and this is taken as  $m$ . Let  $I$  be the current that is passed through the electrolyte and  $t$  the time for which current is passed. Then electro-chemical equivalent

$$m = 1.2 \text{ g}$$

$$= 1.2 \times 10^{-3} \text{ kg}$$

$$t = 1 \text{ hour}$$

$$= 3600 \text{ s}$$

$$I = 1 \text{ ampere}$$

$$\text{ECE of copper } Z = \frac{m}{It}$$

$$= \frac{1.2 \times 10^{-3} \times 1 \times 3600}{1 \times 3600}$$

$$= 3.33 \times 10^{-7} \text{ kg C}^{-1}$$

$$Z = m/It$$

Thus, the e.c.e. of copper is determined using the above equation.

### Solved Examples:

1) Calculate the electro-chemical equivalent of copper when a current of

$$I = 1$$

$$\text{ampere } t = 30$$

$$\text{minutes}$$

$$= 30 \times 60$$

$$= 1800 \text{ s}$$

$$Z = 3.33 \times 10^{-7} \text{ kg C}^{-1}$$

$$\text{We know } Z = \frac{m}{It}$$

$$m = ZI t$$

$$m = 3.33 \times 10^{-7} \times 1 \times 1800$$

$$m = 5.994 \times 10^{-4} \text{ kg.}$$

one ampere passing through a copper volt meter for one hour liberates 1.2 g of copper.

Given,

2) Calculate the mass of copper deposited when a current of one ampere flows for 30 minutes in a copper voltmeter. e.c.e of copper =  $3.33 \times 10^{-7}$

$$m = 3.3g$$

$$= 3.3 \times 10^{-3} \text{kg}$$

$$t = 45 \text{ minutes}$$

$$= 45 \times 60$$

$$= 2700 \text{ s}$$

$$Z = 1.118 \times 10^{-6} \text{kgC}^{-1}$$

We know,

$$Z = \frac{m}{It}$$

It

$$I = \frac{m}{Zt}$$

Zt

$$= \frac{3.3 \times 10^{-3}}{1.118 \times 10^{-6} \times 2700}$$

$$I = 1.093 \text{ ampere}$$

kgC<sup>-1</sup>

<sup>1</sup>Given,

$$m = 2 \times 10^{-3} \text{kg}$$

$$I = 1.2 \text{ ampere}$$

$$Z = 0.68 \times 10^{-6} \text{kgC}^{-1}$$

We know,

$$Z = \frac{m}{It}$$

It

$$\therefore t = \frac{m}{ZI}$$

ZI

$$t = \frac{2 \times 10^{-3}}{0.68 \times 10^{-6} \times 1.2}$$

$$t = 2451 \text{ s.}$$

3) Calculate the current required to deposit 3.3g of silver in 45 minutes.  
 e.c.of silver =  $1.118 \times 10^{-6} \text{ kgC}^{-1}$

Given,

4)  $2 \times 10^{-3} \text{ kg}$  of gold is to be deposited electrolytically by passing a current of 1.2 ampere. What will be the time taken? e.c.of gold is

$0.68 \times 10^{-6} \text{ kgC}^{-1}$

Give R,

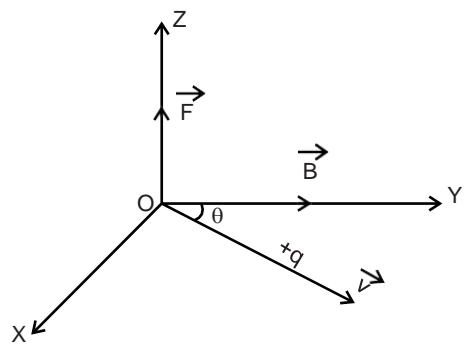
**MEASURING INSTRUMENTS**

When charges are in motion they produce magnetic effect. An electric current produces magnetic field.

Therefore moving charged particles in magnetic field experience force  $\vec{F} = q\vec{v} \times \vec{B}$

**Motion of a charged particle inside a uniform magnetic field**

If an electric charge +q moves with velocity  $\vec{v}$  through a magnetic field then the magnetic field exerts a force  $\vec{F}$  on the charge. It has been



found experimentally that the magnitude of this force is.

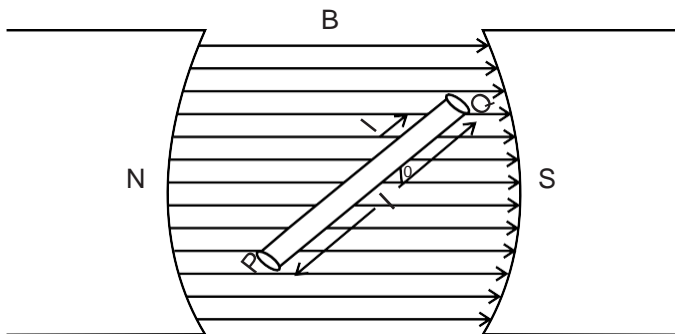
$F = Bqv \sin \theta$

Where  $\theta$  is the angle between the direction of  $\vec{v}$  and  $\vec{B}$

The force is exerted perpendicularly to both  $\vec{v}$  and  $\vec{B}$  and  $\vec{F}$  is perpendicular to the plane containing  $\vec{v}$  and  $\vec{B}$ . The magnitude of magnetic force on a moving charge in a magnetic field is directly proportional to (i) magnitude of charge q (ii) velocity of the charge (iii) strength (B) of the magnetic field and (iv) sine of angle between  $\vec{v}$  and  $\vec{B}$

**Expression for the force acting on a current carrying straight conductor placed in a uniform magnetic field**

Let us consider a small element PQ of a straight conductor of length



$l$ , which carries a current  $I$ . This is placed in a uniform magnetic field of magnetic induction  $B$ . The current passing through the conductor is due

$$\text{ie Current} = \frac{\text{Charge}}{\text{time}}$$

$$I = \frac{nq}{t}$$

to the flow of electric charges. Let  $\theta$  be the angle between the direction of current flow and the magnetic induction  $B$ . Let there be 'n' number of free charges each of charge 'q' moving with velocity  $v$  through the conductor.

We know force on each charge  $F = Bqv \sin \theta$

Since there are  $n$  numbers of charges moving in the conductor of length  $l$ ,

$$\text{ie } v = \frac{l}{t}$$

$$F = BIt \frac{l}{t} \sin \theta$$

$$F = BI l \sin \theta$$

The total force experienced by the conductor is given by

$$F = Bnqv \sin \theta$$

In a conductor current is the rate of flow of charges

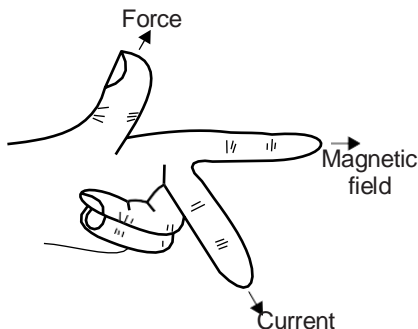
Here  $t$  is the time taken by a charge to travel a distance  $l$  of the

conductor.

From the above relation  $q = It$

The total force experienced by the conductor is given by  $F = B I l \sin \theta$

The velocity of charges is the distance travelled by them in time  $t$



This expression gives the force on a straight conductor placed in a uniform magnetic field.

Suppose if the straight conductor is placed at right angles to the magnetic field then  $\theta = 90^\circ$

The force acting on the conductor is given by

$$F = B I l \sin 90^\circ$$

ie  $F = B I l (\sin 90^\circ = 1)$

The direction of the force is given by Fleming's left hand rule.

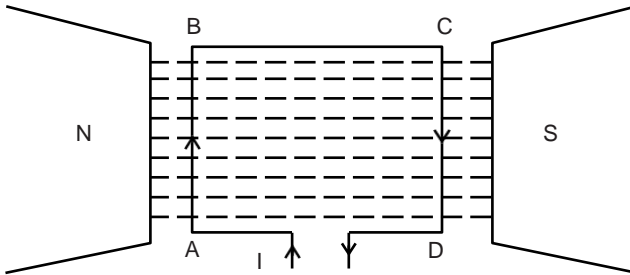
### Fleming's Left Hand Rule

The direction of motion of current carrying conductor placed in a uniform magnetic field is given by Fleming's Left Hand Rule. According to this rule, when the forefinger, middle finger and thumb of the left hand are stretched mutually perpendicular to each other and if the forefinger represents the direction of magnetic field, the middle finger represents the direction of current in the conductor, then the thumb gives the direction of motion of the conductor.

### Expression for the torque experienced by a rectangular current

## carrying coil placed in a uniform magnetic field

Consider a rectangular coil ABCD placed in a uniform magnetic field



of induction  $B$ . Let  $l$  be the length of the coil and  $b$  its breadth. The magnetic field acts parallel to the plane of the coil and perpendicular to the sides AB and CD. Let  $I$  be the current passing through the coil.

Now the side AB of length  $l$  experiences a force  $BIl$ , since the angle between this side and the magnetic field is  $90^\circ$ . The direction of this force is given by Fleming's left hand rule. The side CD also experiences an equal force of  $BIl$  but in the opposite direction, since the current on this side flows in opposite direction.

But the sides BC and AD are parallel to the direction of the magnetic field and hence there is no force acting on these two sides.

Therefore, the two equal unlike and parallel forces acting on the sides AB and CD constitute a couple. Torque is the moment of the couple in the coil.

Moment of the couple = One of the forces  $\times$  the perpendicular distance between the forces.

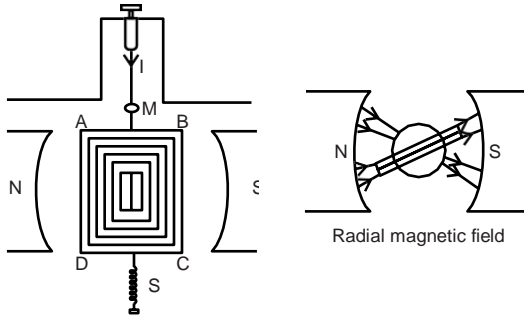
This perpendicular distance is the distance between the two sides AB and CD, measured along the direction of the field. It should be noted that this distance is not a constant and varies from position to position of the coil. The breadth BC of the coil ( $b$ ) gives the maximum distance.

The maximum torque on the coil =  $BI \times b = BIA$

(where  $\times b = A$ , The area of cross section of the rectangular coil)

If there is 'n' number of turns in the coil then the maximum torque is given by  $nBIA$ .

### Working of a Moving coil galvanometer



The principle of the moving coil galvanometer is that a coil of conductor carrying current placed in a magnetic field experiences a torque.

The moving coil galvanometer consists of a rectangular coil of wire of many turns wound on a non-magnetic metallic frame and is suspended between two strong permanent magnetic poles N and S. The coil is suspended by means of a fine phosphor bronze wire from a movable torsion head. The coil is held tight by a hairspring S. A small light mirror M is attached to the suspension fibre and the angle of deflection of the coil can be measured with a lamp and scale arrangement. A small soft iron cylinder is fixed at the centre of the coil for producing a radial magnetic field.

Now, when a current of lampere passes through the vertical limb AD of the rectangular coil ABCD, it experiences a force perpendicular to it and directed outwards according to Fleming's left hand rule. Similarly the limb BC experiences an equal and opposite force to that of AD and directed inwards. It should be noted that in all the position of the coil, the forces on the sides AD and BC will always be perpendicular to the magnetic field, since the plane of the coil remains parallel to the magnetic field. No force acts on the sides AB and CD, since they are parallel to the magnetic field. The coil will come to rest in a deflected position.

According to Fleming's left hand rule the forces experienced by the



sides AD and BC are  $Bl$  and  $B'l$  where,  $l$  is the length of the sides and  $B$  is the magnetic induction. These forces constitute a couple and the controlling couple is provided by the torsion in the phosphor bronze suspension fibre.

For  $n$  turns of the coil, the deflection couple produced =  $Bl \times b$

$$I = \frac{C}{\theta} \cdot nBA$$

The current required ( $I/\theta$ ) to produce unit deflection is

$$\text{called the figure of merit of the Galvanometer, } K = \frac{CnB}{A}$$

Since  $C, n, B$  and  $A$  are all constants  $I = K\theta$  i.e.,

$$I \propto \theta$$

where  $b$  is the breadth of the coil.

For  $n$  turns of the coil, the deflection couple produced =  $nBl \times b$

$$= nBIA$$

where  $l \times b = A$ , the area of coil.

If  $C$  is the couple per unit twist of the suspension wire then for a twist  $\theta$ , the restoring couple =  $C\theta$

(The deflection  $\theta$  can be measured from the lamp and scale arrangement)

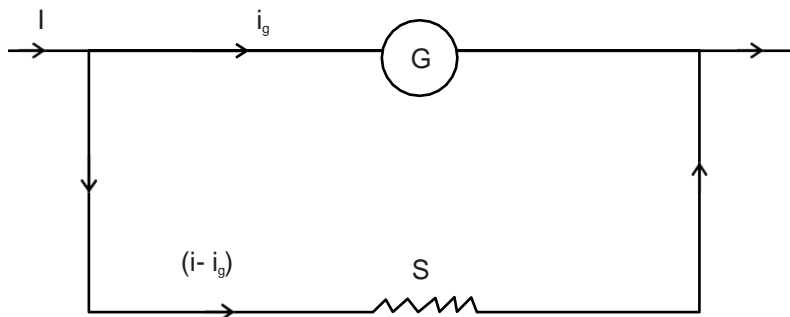
For equilibrium position,

Deflecting Couple = Restoring couple; i.e.;  $nBIA = C\theta$

The current is directly proportional to the deflection produced. For a small value of  $I$ , if the deflection  $\theta$  is large, then the galvanometer is said to be sensitive.

### Merit of moving coil galvanometer

- (i) The external magnetic fields do not alter the deflection of the coils since there is a strong magnetic field between the pole pieces, and the galvanometer can be placed in any position.



(ii) Since the deflection of the coil is directly proportional to the current, a uniformly divided scale can be used to measure the current.

### Conversion of a galvanometer into an ammeter

A galvanometer is a very sensitive device used to detect the strength of flow of current and large current cannot be passed through it, because large current flow can easily damage the coil. Therefore a low resistance called shunt is connected in parallel to the galvanometer to avoid the damage of the coil and only a fraction of the current will pass through the

$$i_g G = (i - i_g) S$$

$$\therefore S = \frac{i_g G}{i - i_g}$$

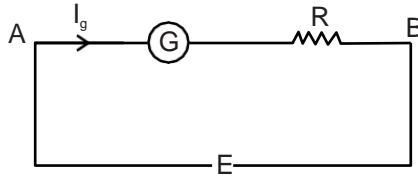
galvanometer and most of the current passes through the shunt.

A galvanometer can be converted into an ammeter by connecting a suitable low resistance called shunt in parallel to it.

Consider a galvanometer of resistance  $G$  ohm. If the galvanometer is to be converted into an ammeter, a low resistance called shunt  $S$  should be connected in parallel with the galvanometer.

Let  $I_g$  be the maximum current that can be passed through the galvanometer and this current will give the full scale deflection in the galvanometer. Therefore out of the total current of  $I$  ampere in the circuit, the current passing through the shunt is

$$i_s = (I - I_g)$$



Since the galvanometer and the shunt are parallel.

The current is a fraction of  $i$  and the scale is marked in ampere. In an electrical circuit, an ammeter is always to be connected in series.

$$E = I_g(G+R)$$

$$\text{i.e. } \frac{E}{G+R} = I_g$$

$$\text{and } R = \frac{E}{I_g} - G$$

### Conversion of a galvanometer into a voltmeter

The voltmeter should have high resistance, then only it will not draw any appreciable current from the circuit. It is used to measure the potential difference across two points of an electrical circuit.

A galvanometer can be converted into a voltmeter by connecting a high resistance in series with the galvanometer.

Consider a galvanometer of resistance  $G$  ohm. If it is to be converted

Full scale deflection  $I_g = 15 \text{ mA} = 15 \times 10^{-3} \text{ A}$  Galvanometer resistance  $G = 5 \text{ ohm}$

$$i = 1.5 \text{ A}$$

$$\text{We know, } S = \frac{i_g G}{i - i_g}$$

$$S = \frac{1.5 \times 10^{-3} \times 5}{1.5 - (15 \times 10^{-3})} = 0.05 \text{ ohm}$$

into a voltmeter, a high resistance  $R$  should be connected in series with it to read  $E$  volt.

Let  $I_g$  be the current required to produce full scale deflection in the

Potential difference  $E = 1.5$  volt

$$\begin{aligned} \text{We know, } R &= \frac{E - G}{I_g} \\ &= \frac{1.5}{15 \times 10^{-3}} - 5 \\ &= 95 \text{ ohm} \end{aligned}$$

galvanometer. A and B are two terminals of the voltmeter. Since the high resistance  $R$  is connected in series with the galvanometer, the current  $I_g$  passes through both the galvanometer and the high resistance.

The scale is marked in volt. In an electrical circuit, a voltmeter is always connected in parallel between the two terminals across which the p.d. is to be measured.

$$\begin{aligned} G &= 200 \text{ ohm} \\ E &= 10 \text{ volt} \end{aligned}$$

$$I_g = 1 \text{ mA} = 1 \times 10^{-3} \text{ A}$$

$$\text{We know, } R = \frac{E - G}{I_g}$$

$$R = \frac{10}{10^{-3}} - 200$$

$$R = 10000 - 200 = 9800 \text{ ohm}$$

### Solved Examples

- 1) A galvanometer has a resistance of 5 ohm and full scale deflection is produced by 15 mA. Calculate what resistance must be used to enable it to read (i) 1.5 ampere (ii) 1.5 volt.

Given,

A shunt of 0.05 ohm is to be connected in parallel with the

$$G = 100 \text{ ohm}$$

$$I_g = 100 \text{ mA} = 100 \times 10^{-3} \text{ A}$$

$$I = 10 \text{ A}$$

$$\text{We know, } S = \frac{I_g G}{I - I_g}$$

$$S = \frac{100 \times 10^{-3} \times 100}{10 - 100 \times 10^{-3}}$$

$$S = \frac{10}{10 - 0.1}$$

$$S = 1.01 \text{ ohm}$$

galvanometers so as to measure up to 1.5 ampere.

A resistance of 95 ohm is to be connected in series with the galvanometers so as to measure up to 1.5V.

2) A galvanometer has a resistance of 200 ohm and when a current of 1mA passes through it, produces a deflection of 1 division. How this can be converted into a voltmeter to read 10 volt per division?

Given,

A high resistance of 9800 ohm is to be connected in series with the galvanometers so as to measure up to 10 volt/division

3) The resistance of a galvanometer is 100 ohm and it can withstand a maximum current of 100 mA. How this can be converted into an ammeter to read 10A.

Given,

A shunt of 1.01 ohm should be connected in parallel with the galvanometer to measure up to 10A.

## **ELECTROMAGNETIC INDUCTION**

Oersted in the year 1820 discovered that a current flowing through a conductor is accompanied by a magnetic field around it. Many attempts were made after that to verify its reverse effect of the possibility of producing an induced emf by means of a magnetic field. The possibility of producing emf across the ends of a conductor with the changing magnetic flux linked with the conductor was demonstrated by Michael Faraday. Thus electromagnetic induction is the phenomenon according to which an induced emf is produced in a coil by the change of magnetic flux linked with it.

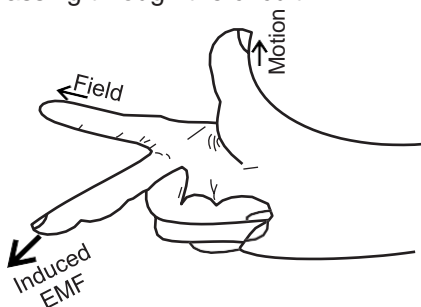
### **Faraday's laws of electromagnetic induction**

1. Whenever there is a change in the magnetic flux or the magnetic lines of force passing through a circuit, an induced emf is produced in the circuit.
2. The induced emf lasts only for the period during which the magnetic flux or the magnetic lines of force passing through the circuit actually

$$\therefore e = - \frac{Nd\phi}{dt}$$

change.

3. The magnitude of the induced emf is equal to the rate of change of magnetic flux passing through the circuit.

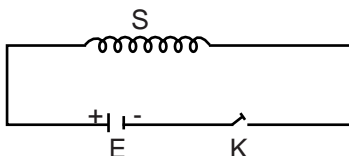


### Lenz's law

There is a closer relationship between the direction of the changing magnetic field and the direction of the current induced by it. The Russian Scientist, Lenz formulated an important law in 1835 giving the direction of the induced current produced in a circuit.

Lenz's law states that the induced current produced in a circuit always flows in such a direction that it opposes the change or causes that produces it.

If there are N number of turns in a coil and  $\phi$  is the magnetic flux linked with each turn of the coil then, the total magnetic flux linked with the



coil at any time is  $N\phi$ . Then the induced e.m.f. is

The negative sign is a consequence of Lenz's law. Lenz's law is in accordance with the law of conservation of energy.

### Fleming's Right Hand Rule

Fleming suggested a rule which readily gives the direction of the induced e.m.f. in a conductor moving at right angles to a magnetic field.

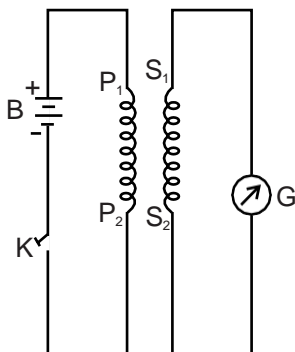
According to this rule, when the forefinger, the middle finger and the thumb of the right hand are stretched mutually perpendicular to each other and if the forefinger represents the direction of the magnetic field and the thumb represents the direction of motion of the conductor, then the middle finger points the direction of the induced current.

This is an important rule and has wide practical applications. It is sometimes called the dynamo or generator rule.

### Self Induction

Self induction is the property of a coil which enables it to produce an opposing induced e.m.f. when the current in the coil changes.

Let us consider a coil  $S$  connected to a cell  $E$  and a tap key  $K$ . When there is no current flowing through the coil, there is no magnetic flux linked with it. When the circuit is closed by pressing the tap key  $K$ , the current in the circuit changes from zero to maximum and the magnetic flux passing



through the coil also increases from zero to a maximum value. Due to this variation in magnetic flux linked with the coil, an induced e.m.f. is produced in the coil itself, which according to Lenz's law, opposes the growth of the

current. Thus the current does not attain its maximum value instantaneously on closing the circuit but takes some time. Thus the effect of self-induction is to delay the rate of the growth of current.

### **Mutual induction**

This is the phenomenon according to which an induced emf is produced in a circuit by the change of current in a neighbouring circuit. Whenever there is a change in the magnetic flux linked with a coil, there is also a change of flux linked with the neighbouring coil and thus producing an induced emf in the second coil.

Let us consider two coils  $P_1, P_2$  and  $S_1, S_2$  placed close to each other. The coil  $P_1, P_2$  connected in series with a battery B and a key K is called primary, whereas the coil  $S_1, S_2$  with no cell in it is called secondary and a galvanometer G is connected in series with these secondary coil.

When the key K is pressed, the current starts flowing in the primary coil and gradually reaches its maximum value and the magnetic flux linked with it also increases. As a result, the magnetic flux passing through the secondary changes and hence an induced current is produced in the secondary coil. According to Lenz's law, the direction of the induced current is such as to oppose the growth of the current in the primary.

When the key K is released, current starts decreasing from maximum to zero value and therefore the magnetic flux linked with the primary coil decreases. Therefore magnetic flux linked with secondary coil also decreases and hence an emf is induced in the secondary coil. According to Lenz's law, the induced current in the secondary flows in such a direction so as to oppose the decrease in current in the primary coil.

## **QUESTIONS**

### **Part– A**

1. Write Ohm's law
2. Write the law of resistances when the resistances are connected in series
3. Write the law of resistances when the resistances are connected in parallel
4. Define the term Resistivity



5. Define the term Conductivity
6. Write Kirchoff's current law
7. Write Kirchoff's voltage law
8. What is a Wheatstone's network?
9. What is meant by heating effect of electric current?
10. Write Joule's law of heating
11. Define electrolysis
12. Write Faraday's first law of electrolysis
13. Write Faraday's second law of electrolysis
14. What is a Voltmeter?
15. Define Electro-chemical equivalent (e.c.e) of an element
16. Write Fleming's left hand rule
17. Write the principle of a moving coil galvanometer
18. What is meant by the figure of merit of a galvanometer?
19. What are the merits of a moving coil galvanometer?
20. Write Lenz's law
21. Write Fleming's right hand rule

### **Part-B**

1. Derive the condition for balancing the Wheatstone's network.
2. Describe an experiment to determine the specific heat capacity of a liquid using Joule's calorimeter.
3. Describe an experiment to determine the electrochemical equivalent of copper using a copper voltameter
4. Derive an expression for the force acting on a current carrying straight conductor placed in a uniform magnetic field
5. Derive an expression for the torque experienced by a rectangular current carrying coil placed inside a uniform magnetic field.
6. Describe the construction and working of a moving coil galvanometer with a diagram
7. Describe how a galvanometer can be converted into i) an ammeter and ii) a voltmeter.
8. What is electromagnetic induction? Write Faraday's laws of electromagnetic induction.
9. Explain self induction and mutual induction.

## Exercise problems

1. Calculate the resistance of the wire of length 1.5m and diameter 0.5mm if the resistivity of the material of the wire is  $49 \times 10^{-8}$  ohm metre.  
[Ans 3.745 ohm]
2. Calculate the resistivity of the material of a wire of length 25.6m and diameter 0.84mm whose resistance is 0.72 ohm  
[Ans  $1.557 \times 10^{-8}$  ohm metre]
3. Calculate the resistance of a wire of 1.5m long and area of cross-section  $0.4 \text{ mm}^2$  whose resistivity is  $46 \times 10^{-8}$  ohm metre  
[Ans 4.312 ohm]
4. Calculate the electrochemical equivalent of a metal when a current of one ampere liberates in one hour 1.2g of metal from its electrolyte  
[Ans  $3.33 \times 10^{-7} \text{ kg C}^{-1}$ ]
5. What will be the time taken for  $2.2 \times 10^{-3}$  kg of gold to be deposited electrolytically by passing a current of one ampere? Given, e.c.e of gold =  $0.68 \times 10^{-6} \text{ kg C}^{-1}$   
[Ans 3235s]
6. Calculate the amount of current that should be passed through a copper voltameter to deposit 3.12g of copper in one hour (e.c.e of copper =  $3.33 \times 10^{-7} \text{ kg C}^{-1}$ )  
[Ans 2.6A]
7. A galvanometer of resistance 30 ohm shows maximum deflection for a current of 2mA. How this can be converted into a voltmeter to read 10V?  
[Ans 4970 ohm connected in series with the galvanometer.]
8. An ammeter of resistance 0.60 ohm can measure current upto 1A. What must be the shunt resistance to enable the ammeter to measure current upto 5A?  
[Ans 0.15 ohm]

## Brain Teaser

1. In the electrification of a house how the various lamps are reconnected? Explain Why?
2. Why does a silver voltameter give accurate results than a copper voltameter?

# UNIT - V

## BASICS OF ELECTRONICS

### 5.1. CAPACITORS

#### Introduction:

Apart from resistors and insulators, a capacitor is also the other basic passive component commonly used in electrical circuits. A capacitor is a device which has the ability to store charges in the form of electrical energy, which neither a resistor nor an insulator can do.

An ideal capacitor may retain the charges stored in it for a long time, after power supply is removed from the circuit. That is, it can store charges forever. This stored charge can discharge through a conductor (or) wire. This discharge can cause shock or damage to the connected equipment, during the misuse of capacitors in a circuit.

#### Construction of a capacitor:

Two similar metallic conductors separated by an insulator (or) dielectric medium forms a capacitor. Conductor – insulator – conductor sequence is the basic idea behind a capacitor.

The basic difference between an insulator and dielectric medium is:

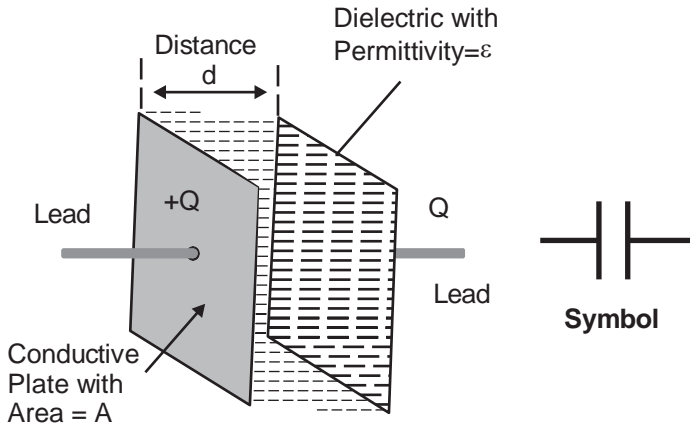
The materials which do not have free electrons in them are unable to conduct electricity and they are termed as insulators.

Some material like ceramic behaves as a dielectric medium. Charges are induced in the opposite faces of it, because of the external applied electric field. Such material is called dielectric.

All dielectric materials are insulators, but not all insulators make good dielectric materials.

The dielectric medium could be air, mica, ceramic, paper, polyester, Teflon, etc.

In the figure shown below two metallic conductors of equal area "A" separated by a dielectric medium, forms a two-terminal device, named as Capacitor.



The plates are kept perfectly parallel to each other. So it is called a parallel plate capacitor. The conducting plates of a capacitor can be either square or circular or rectangular or cylindrical or spherical in shape. Shape selection depends on its application and voltage rating.

**Capacitance:**

An ideal capacitor is characterized by a single constant value named as capacitance. The capacitance of a capacitor is defined as amount of charge required to create a unit potential difference between its plates or capacitance can also be defined as the ratio of charge to potential difference.

$$C = \frac{\text{Charge}}{p.d} = \frac{Q}{\text{farad V}}$$

If **Q = 1 coulomb** and **V = 1 volt**

Then **C = 1 farad (F)**

Capacitance of a capacitor depends upon the area of the plate, distance of separation between the plates "d" or thickness of the dielectric medium used.

The unit of capacitance is 'farad' with symbol 'F' named after British Physicist Michael Faraday.

**Definition of 'farad':-**

One farad is defined as the capacitance of a capacitor which requires a charge of one coulomb to establish a p.d of one volt between the plates.

How many electrons are present in one coulomb of charge?

Let there be "n" electrons in one coulomb. But charge  $q = ne$ , where 'e' is the charge of one electron.

Therefore 
$$n = \frac{q}{e} = \frac{1}{1.6 \times 10^{-19}} = 6.25 \times 10^{18} \text{ electrons}$$

So to establish one volt p.d. between the two plates, large number of electrons have to move from one plate to another plate. So one farad is too large for practical purposes.

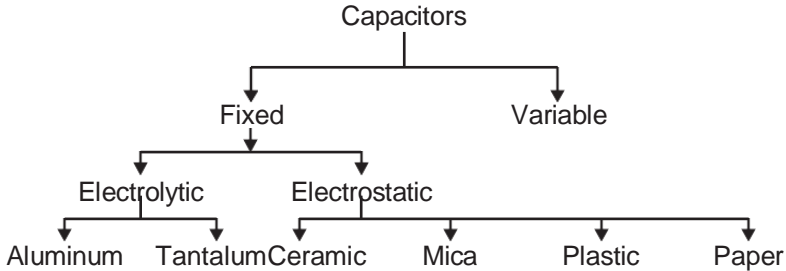
In most of the cases we have smaller unit like

one millifarad =  $1 \text{ mF} = 1 \times 10^{-3} \text{ F} = 0.001 \text{ F}$

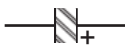
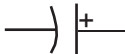
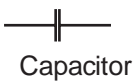
one microfarad =  $1 \mu\text{F} = 1 \times 10^{-6} \text{ F} = 0.000001 \text{ F}$

one nanofarad =  $1 \text{ nF} = 1 \times 10^{-9} \text{ F} = 0.000000001 \text{ F}$

**Types of Capacitors:-**



The symbol for capacitors is not unique. In the modern electronic world, the symbols for capacitors are given below.



Polarized Capacitor



Variable Capacitor



Trimmer capacitor

## Energy stored in a capacitor

Energy stored in a capacitor is equal to the work done to charge it. The work done is stored in the form of energy. That is work done to move the electron from one plate to the other plate.

For moving a small amount of charge "dq" from one plate to the other plate against the p.d, between the plates, work is to be done.

Work done = p.d x charge But,

$$\begin{aligned} \text{p.d.} V &= \frac{q}{c} \\ \text{Work done } W &= \int dq \frac{q}{c} \end{aligned}$$

where, W is the work measured in joule

q is the charge measured in coulomb and c is the capacitance measured in farad

Energy stored in a capacitor can be obtained by integrating this equation. Starting with an uncharged capacitance (q = 0) and moving charge Q from one plate to the other plate,

The required work "W" is

$$W = \int_0^Q \frac{q}{c} dq = \frac{1}{2c} Q^2 = \frac{1}{2} CV^2 \therefore Q = CV$$

This work is stored in the form of electrical energy.

## POWER GENERATED BY A CAPACITOR

Since power is energy dissipated in a time, the power generated by a capacitor can be expressed as

$$P = \frac{W}{\text{Time}} = \frac{\text{Energy}}{\text{dissipation time}} = \frac{\text{Energy}}{E}$$

### Example:-

The energy stored in a 10 μF capacitor charged to 230 V can be calculated as

$$\begin{aligned} W &= \frac{1}{2} CV^2 = \frac{1}{2} \times 10 \times 10^{-6} \times (230)^2 \\ \text{Energy} &= 0.26 \text{ J} \end{aligned}$$

When this energy dissipated within  $5\mu\text{s}$  means, the power generated is

$$P = \frac{0.26}{5 \times 10^{-6}} = 52000 \text{ W} = 52 \text{ KW}$$

Because of this stored energy, we feel electric shock when we touch the two terminals of a capacitor simultaneously.

**PRECAUTION:**

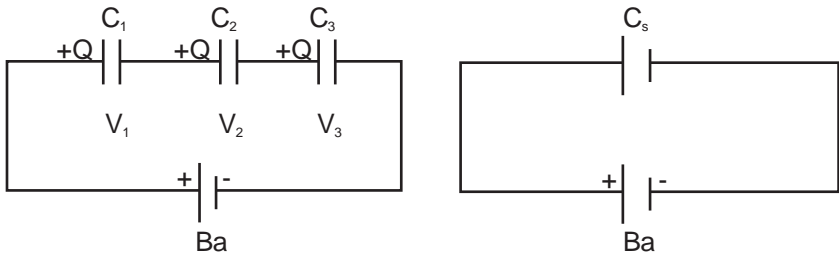
The following precautions are to be taken

(i) All power capacitors must be fully discharged before working on the equipment.

(ii) All cables, motor windings, and other components which can exhibit capacitance must be discharged before working on the components or associated circuitry.

**Capacitors in series:-**

Let us consider three capacitors of capacitances  $C_1, C_2$  and  $C_3$  respectively connected in series as shown in the figure.



Due to the induction of charges on the plates by the applied p.d.,

- 1) Charge on each capacitor is same irrespective of its capacitance.
- 2) p.d. across each capacitor is different
- 3) Sum of the voltage across the capacitors is equal to the applied battery Voltage "V".

So,  $V = V_1 + V_2 + V_3$  (1)

We know  $C = \frac{Q}{V}$  (or)  $V = \frac{Q}{C}$

So the p.d. between the first capacitor is

$$C_1 = \frac{Q}{V_1} \text{ (or) } V_1 = \frac{Q}{C_1}$$

Similarly for second and third capacitors

$$V_2 = \frac{Q}{C_2} \text{ and } V_3 = \frac{Q}{C_3}$$

substituting all in equation (1)

$$\frac{Q}{C_s} = \frac{Q}{C_1} + \frac{Q}{C_2} + \frac{Q}{C_3}$$

If  $C_s$  is the effective capacitance of this series connection, then

$$\frac{Q}{C_s} = Q \left( \frac{1}{C_1} + \frac{1}{C_2} + \frac{1}{C_3} \right)$$

Canceling  $Q$  on both sides

$$\frac{1}{C_s} = \left( \frac{1}{C_1} + \frac{1}{C_2} + \frac{1}{C_3} \right)$$

When capacitors are connected in series the reciprocal of the effective capacitance is equal to the sum of the reciprocals of the individual capacitances.

Hence connecting capacitors in series is equivalent to increasing the thickness of the electric medium. When we increase the thickness, distance increases. So capacitance decreases.

Capacitors connected in series block DC and allow AC.

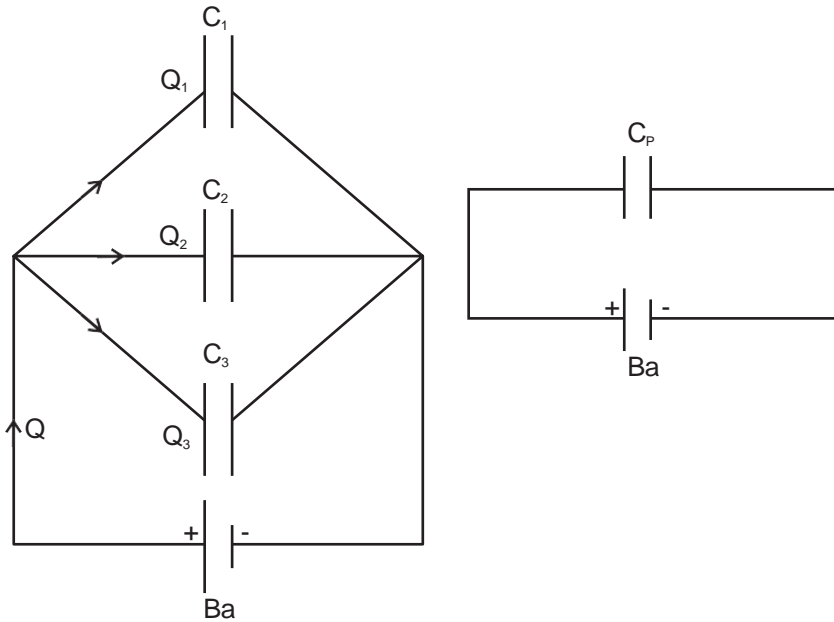
### Capacitors Connected in Parallel:-

Let us consider three capacitors of capacitances  $C_1$ ,  $C_2$  and  $C_3$  respectively connected in parallel as shown in the diagram.

The charge  $Q$  supplied by the battery is divided among the three capacitors as  $Q_1$ ,  $Q_2$  and  $Q_3$ . Therefore,

- 1) Charges across each capacitor are different.
- 2) p.d. across all capacitors is same
- 3) Sum of the charges stored in the capacitors is  $Q$





$$\therefore Q = Q_1 + Q_2 + Q_3 \quad \text{-----} \rightarrow (1)$$

$$\text{We know } C = \frac{Q}{V} \text{ (or) } Q = CV$$

The charge stored on the first capacitor is

$$Q_1 = C_1 V.$$

Similarly for the other two capacitors

$$Q_2 = C_2 V \text{ and } Q_3 = C_3 V$$

Substituting this in equation (1) above

$$CV = C_1 V + C_2 V + C_3 V$$

If \$C\_p\$ is the effective capacitance of the parallel connection,

$$\text{then } C_p V = V(C_1 + C_2 + C_3)$$

Cancelling the "V" on both sides

$$C_p = C_1 + C_2 + C_3$$

When capacitors are reconnected in parallel, the effective capacitance is equal to the sum of the individual capacitances.

Connecting capacitors in parallel is equivalent to adding their plate areas. Hence, combined capacitance equals to sum of the individual capacitances.

### **Application of capacitors:-**

Capacitors are used to accurately measure the fuel level in airplanes fuel tank. As the fuel level between the plates varies, the Capacitance also varies.

A motor capacitor such as a **starting capacitor** or **running capacitor**, alters the current through one or more windings of an electric motor to create a rotating magnetic field. Around dual run capacitor is used in some air conditioner compressor units, to boost both the fan and compressor motor function.

Starter in a fluorescent tube is nothing but a simple capacitor. It is connected in parallel to the tube. Fluorescent tube contains no noble gases. Initially tubes requires instant spark of discharge current, so that the electrodes get hot enough to ionize the low pressure gas in the tube. Once these get hot enough it will continue to ionize the gas. The starter is provided to give that instant discharge current. After lighting up of the tube by discharge process, capacitor starter has no roll until the tube is switched off.

In Radio, the variable capacitors are used with inductance to tune the receiver to a different resonant frequency for receiving the programmes of different stations transmitting different frequency waves.

Capacitors with a flexible plate can be used to measure strain (or) pressure.

A camera flash unit contains a capacitor. The light energy for each "flash" from the camera, is from electrical energy stored in the capacitor. Here electrical energy is converted as light energy.

Capacitors are manufactured in various sizes and shapes depending upon their use in a circuit. Capacitors are used in FM, TV Circuit, Radio, Cellphone, iPod, Computer and hi-fidelity devices. A capacitor is a part of all modern electronic gadgets.

**Solved Problems:**

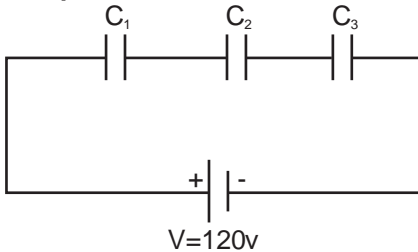
1. Three capacitors each of capacitance 9 pF are connected in series. Calculate the effective capacitance of the capacitors and potential difference across each capacitor, when they are powered by 120V DC supply.

**Given values:-**

**$C_1=9\text{pF}, C_2=9\text{pF}$  and  $C_3=9\text{pF}$**

In series connection with p.d. of 120V.

**$C_1=C_2=C_3=9\text{pF}$**



For series connection,

$$\frac{1}{C_s} = \frac{1}{C} + \frac{1}{C} + \frac{1}{C} = 3 \left( \frac{1}{C} \right)$$

$$\frac{1}{C_s} = 3 \times \frac{1}{9 \times 10^{-12}} = \frac{1}{3 \times 10^{-12}}$$

Therefore  $C_s = 3 \times 10^{-12} \text{F} = 3\text{pF}$

In series connection, sum of the p.d across the capacitor is equal to supply voltage 120V

$$V = V_1 + V_2 + V_3$$

$$120 = \frac{q}{C_1} + \frac{q}{C_2} + \frac{q}{C_3} \quad C \quad C \quad C$$

$$120 = 3 \frac{q}{9 \times 10^{-12}} = \frac{3q}{9 \times 10^{-12}}$$

$$120 = \frac{q}{3 \times 10^{-12}}$$

$$q = 3 \times 120 \times 10^{-12} = 360 \text{ pC}$$

Therefore, potential difference across each capacitor

$$V = \frac{q}{\text{Capacitance}} = \frac{360 \times 10^{-12}}{9 \times 10^{-12}} = 40 \text{ V}$$

2. Calculate the effective capacitance of three capacitors of capacitance 1 mF, 2 mF and 4 mF respectively, when they are connected in series and parallel configuration in a circuit.

**Given values:-**

$$C_1 = 1 \text{ mF}, C_2 = 2 \text{ mF} \text{ and } C_3 = 4 \text{ mF}$$

- (a) When they are connected in series

$$C_s = \frac{1}{\left( \frac{1}{C_1} + \frac{1}{C_2} + \frac{1}{C_3} \right)}$$

$$= \frac{1}{\left( \frac{1}{1} + \frac{1}{2} + \frac{1}{4} \right)} \times 10^{-3}$$

$$= \frac{1 + 0.5 + 0.25}{10^{-3}} = \frac{1.75}{10^{-3}}$$

$$C_s = \frac{1 \times 10^{-3}}{1.75} = \frac{1}{1.75} \times 10^{-3}$$

$$C_s = 0.571 \times 10^{-3} = 0.571 \text{ mF}$$

- (b) When they are connected in parallel

$$C_p = C_1 + C_2 + C_3 = (1 + 2 + 4) \times 10^{-3} \text{ F}$$

$$C_p = 7 \times 10^{-3} \text{ F} \quad C_p = 7 \text{ mF}$$

## 5. 2.SEMI CONDUCTORS

### Introduction:

The branch of science and engineering dealing with current conduction through vacuum tube or semiconductor is known as electronics.

Initially, vacuum tubes were used in the process of electrical signal generation, amplification and transmission. With the advent of semiconductor devices like diode, transistor and other solid state electronic components, the vacuum tubes were replaced completely from all fields of applications. All modern gadgets like television, computer, CD player, automatic washing machine, etc., use microprocessors having integrated chips which consist a large number of logic gates, diodes, transistors, resistors, etc.,

Materials are classified as conductors, insulators and semi conductors according to their electric conductivity. This classification is purely based on the available number of free electrons within the materials, apart from the bonded orbital electrons.

So the word conductivity is used to describe a material's ability to transport electricity. The metals like copper and gold are good conductors. Glass and plastics, on the other hand, are very bad conductors. Semi conducting materials like germanium and silicon has conductivity somewhere between good conductors and insulators.

In semiconductors, charges movement can be manipulated according to our need to make electronic devices. So silicon and germanium are used as a base material for making electronic components like diode, transistors, etc.

New emerging organic semi conductors with similar properties resembling conventional semiconductors are available now. In general, nowadays semiconductors are classified as inorganic and organic semi conductors. One can fabricate diode, transistors and other devices using organic semiconductors which are biodegradable electronic devices.

Devices fabricated using inorganic semi conductors like silicon, Germanium, Gallium Arsenide, Silicon carbide etc., are not bio

degradable which poses threat to our green environment. So the new organic semi conductors will give us biodegradable materials, to lead greenelectronicworld.

### **EnergyBandinSolids:**

During solid formation, the outer most energy levels of the atoms overlapwitheachothertoformabandofenergy. Anenergybandconsists ofcloselyspacedenergylevels, whichareconsideredtobecontinuous.

The outermost electrons of an atom are called valence electrons and the band of energy occupied by the valence electrons is known as valence band (V.B.) This valence band may be partially filled or completely filled but it cannot be empty.

The next higher permitted band of energy is called the conduction band (C.B.). It is located above the valence band. The electrical conductivityofsolidsdependsonthenumberofelectronsresentinthe conduction band of their atoms. The conduction band may be empty or partially filled and can never be completely filled. When the electrons reachtheconductionbandfromthevalenceband, theycanfreelymove, sotheyarecalledasfreeelectronsorconductionelectrons. These free electronsareresponsibleforflowofelectriccurrentthroughthesolid. The gap between the valence band and the conduction band is called the forbiddenenergygap(F.G.).

Materials are classified as good conductors, insulators and semiconductorsbasedontheirabilitytoconducelectricity.

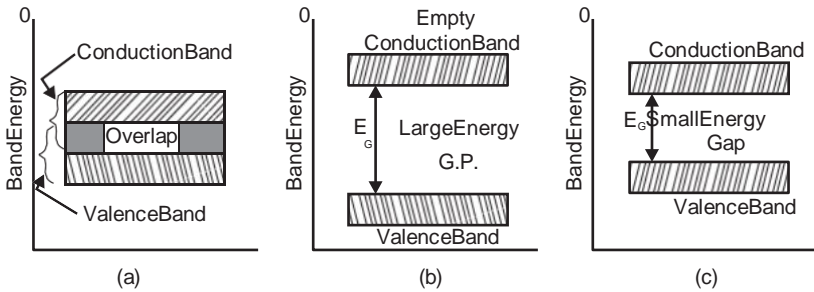
### **Energy band diagram for Conductors, insulators and semi conductors:-**

Our interest is focused on conducting property of a material. In general, physicalandchemicalpropertiesofanelementaredecidedby the valence orbital electrons. For example, if valence orbit is filled, that element is inert like He, Ne, etc. On the other hand the materials with unfilled valence orbit exhibit electrical and magnetic properties like metals. So, onlyvalencebandisconsideredforfurtherstudies. Theother low level bands are not contributing any significant changes in the conducting properties.

The energy band occupied by the valence electrons is called valence band (V.B.)

Next, the conducting band (C.B) lies above the valence band. Conduction band is due to the free electrons from the atoms of the crystal or solid. These free electrons possess kinetic energy and act as a carrier within the crystal.

The energy band diagrams for good conductors, insulators and semiconductor are shown in the figure below.



The above energy band diagram has a gap between V.B. and C.B., named as forbidden energy gap. So based on the energy gap between the valence band and conduction band, the materials are classified as good conductor, bad conductor and semiconductor.

**(a) For good conductors,  $E_g = 0$ . Here valence band and conduction band is overlapping each other.**

Zinc is a good example for conductor which has overlap of valence band and conduction band. In Zinc, valence band is completely filled and the conduction band is empty, but they overlap each other. Because of overlap, it is considered as single band. Hence conduction band is assumed as partially filled. Therefore, on applying a small electric field, the metals conduct electricity.

**(b) For insulator,  $E_g$  value is large.**

In insulators like diamond, the forbidden energy gap is quite large. The forbidden energy gap value is 6 eV, so a minimum of 6 eV energy is required for electron to move from valence band to conduction band. Normally conduction band is empty and valence band is full for insulators.

Since there is no electron for conduction, these type of materials behave as insulators.

**(c) For Semiconducting materials, the  $E_g$  value is very small.**

The energy band structure of the semi conductor is similar to the insulator, but the forbidden gap is very narrow. The forbidden energy gap is 1.1 eV for Silicon. These semiconductor acts as an insulator at 0K and at room temperature the thermal energy absorbed from the surroundings is enough for the electron to jump from the valence band to the conduction band. So as semiconductor starts conducting at room temperature.

**Fermi level:-**

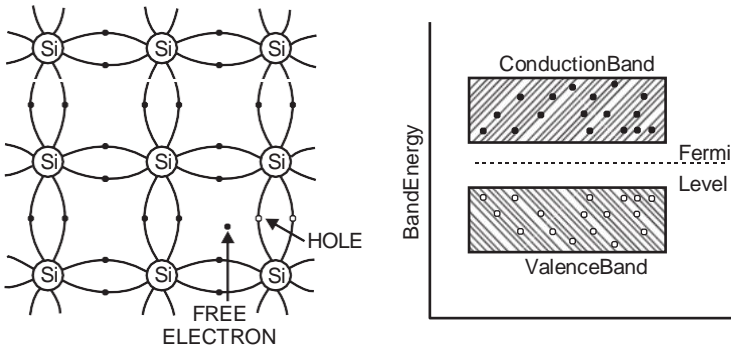
At 0K, the electrons start filling the energy levels in valence band, from bottom to top level. Few electrons fill the bottom level and other electrons above this level. So filling of energy starts from bottom of the band to top of the band.

The highest energy level, which an electron can occupy in the valence band at 0K, is named as Fermi level.

Fermi level is also defined as the energy level corresponding to the average energy of electrons and holes present in the crystal.

**Intrinsic Semiconductors:-**

In general, semiconductors are classified as intrinsic and extrinsic semiconductors. Semiconductors in pure form are called intrinsic semiconductor. Example:- A crystal formed by silicon atoms alone. The silicon atoms are arranging themselves by sharing an electron between the neighbouring atoms. Such bond is covalent bond.





In intrinsic semiconductor like silicon forbidden energy gap value is 1.1 eV. This energy is available for semiconductors placed at room temperature. Due to that thermal energy some covalent bond within the crystal breaks (or) some electrons are pumped from valence band to conduction band. In the bond from which electron is freed, a vacancy is created there. This absence of electron is named as hole. Electron and hole pair is created. Thus at room temperature, a pure semiconductor will have both electrons and holes wandering in random directions. This electron and holes are called intrinsic carriers and such a semiconductor is called intrinsic semiconductor.

As the crystal is electrically neutral, the number of electrons equal to number of holes, in intrinsic semiconductor.

∴ Fermi level lies at the middle of the forbidden gap.

The vacant place created in the valence band due to the jumping of electron from the valence band to conduction band is called 'hole', which is having positive charge.

### **Doping:-**

Pure semiconductor at room temperature possesses free electrons and holes but their number is so small that conductivity offered by the pure semiconductors cannot be used for any practical purpose like device making.

By the addition of certain selected impurities to the pure semiconductor in a very small ratio ( $1:10^6$ ), the conductivity of a silicon or germanium crystal can be remarkably improved.

The process of adding impurity to a pure semiconductor crystal to improve its conductivity is named as doping.

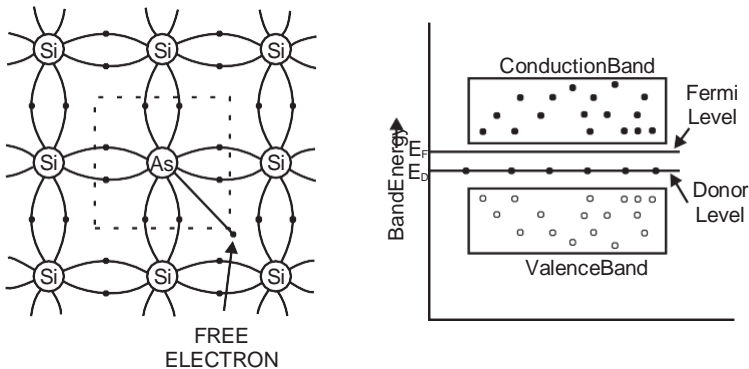
### **Extrinsic Semiconductor:-**

The impurity added semiconductors are named as extrinsic semiconductors. The extrinsic semiconductors are classified as P-type and N-type semiconductors, based on the type of impurity atoms added to these semiconductors.

**N-Type(Extrinsic)Semiconductor:-**

Pentavalent element like antimony (Sb) or arsenic (As) is added to pure silicon crystals. These impurity atoms replace some of the silicon atoms, here and there in the crystal. The added arsenic (As) atom shares its five electrons with the neighbor atoms and releases its fifth electron to the crystal for conduction. So these pentavalent elements are called donor impurities, as they donate electrons, without creating holes.

In silicon, electron needs 1.1 eV to move from valence band to conduction band. This energy becomes available to the semiconductor even at room temperature. So at room temperature few covalent bonds within the material are broken by the thermal energy from the surrounding and some electrons from the valence band are pumped to conduction band. This process leaves some absence of electrons in the valence band. Electron and hole pairs are recreated. At the same time, the number electrons in the conduction band are increasing further by the addition of pentavalent impurities without any addition of holes which already exist within the crystal. So the numbers of electrons are donor electrons plus thermal electrons at room temperature. This thermal excitation produces very less pair of electrons and holes, whereas the added impurity donates more electrons. The majority charge carriers electrons are of the order of  $10^{24}$ , whereas the minority charge carriers holes are of the order of  $10^8$  at 300K.

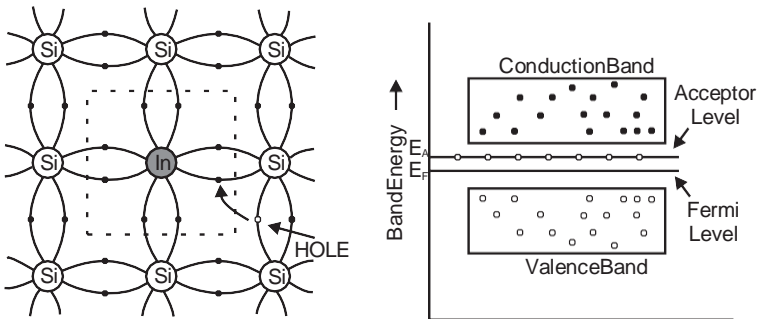


Hence, the majority charge carriers are electrons in this material. The electron carries negative charge, so it is named as N-type semiconductor and conduction is due to a large number of electrons.

As the number of electrons in the conduction band is more than the number of holes in the valence band, in an N-type semiconductor, the Fermi level lies nearer to the conduction band.

**P-type Semiconductor:-**

When trivalent elements like indium, aluminium, boron are doped with pure silicon, the added impurity atoms replace some of the silicon atoms, here and there in the crystal and establish covalent bonds with the neighboring atoms. Indium has three electrons but that indium is covered by four silicon atoms as shown in the figure below. So, one of the covalent bonds is not completed by sharing of electrons between them. There is an absence of electron which creates a hole.



Indium needs one more electron to complete its covalent band. So indium is an acceptor of electrons. Now, this extrinsic semiconductor gains thermal energy from the surrounding at room temperature. So some electron absorbs this thermal energy and jumps to the conduction band. This creates electrons and holes pair due to thermal excitation. So the total number of holes in the valence band is due to donor's atom plus thermally generated holes.

The holes are majority and electrons are minority. Hence conduction is due to majority charge carriers which are holes. Here the holes are behaving like positive charge carriers. So this material is named as P-type.

As the number of holes in the valence band is more than the number of electrons in the conduction band, the Fermi level lies nearer to the valence band, in P-type semiconductor.

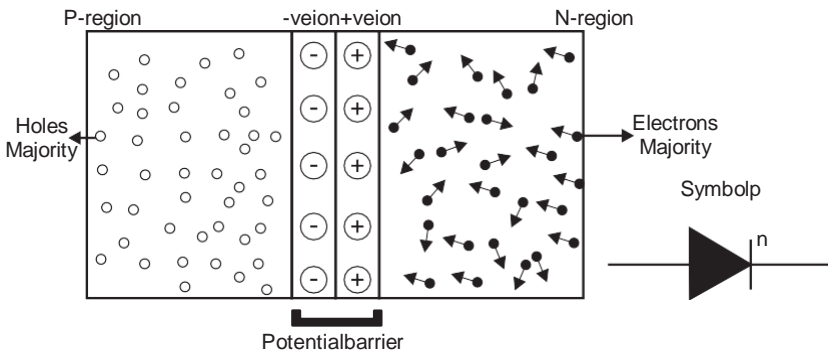
### 5.3. DIODES AND TRANSISTORS

#### Introduction:-

To understand electronic devices and circuits, a brief idea about semiconductor diodes is a must. The semiconductor diode is a fundamental two-terminal electronic device, similar to a resistor. The voltage-current (V-I) relationship of a resistor is linear. However, the V-I characteristic of a diode is not only non-linear but also depends on the operating condition. That is, a resistor allows the charge carriers at any condition and behaves like a passive element. A diode allows current to pass through it in one direction and acts as a switch in electronic circuits.

#### P-N junction diode:-

At room temperature, a piece of P-type material has a majority of holes and N-type material has a majority of electrons. When a part of an intrinsic semiconductor piece is doped with pentavalent impurities and the remaining part is doped with trivalent impurities, a P-N junction diode is formed.



### **Potential barrier and depletion region:**

In the P-N junction diode as shown in the above figure, the P region has circles with a negative sign indicating immobile ions and the mobile holes are represented by small circles.

In the N region, the circles with a positive sign inside represent immobile ions and the mobile free electrons are represented by small dots.

During the junction formation, the free electrons and holes on both sides of the junction migrate across the junction by the process of diffusion. The electrons passing through the junction from the N region into the P region recombine with holes in the P region very close to the junction.

Similarly, the holes crossing the junction from the P region into the N region recombine with the electrons in the N region very close to the junction. This recombination of free or mobile charges produces a narrow region on either side of the junction of width about  $10^{-4}$  cm to  $10^{-6}$  cm. This region is called the depletion region, where there are no mobile charges available.

In the depletion region, the atoms on the left side of the junction become negative ions (-) and the atoms on the right side of the junction become positive ions (+). Thus an internal potential difference (p.d.) is produced across the junction. This p.d. is called the internal potential barrier. This prevents the further flow of charge carrier from P-region to N-region. The potential barrier for germanium is 0.3V and that for silicon is 0.7V. The potential barrier of a P-N junction diode can be decreased or increased by applying external voltage.

This barrier acts like a battery. However, the value of potential barrier depends upon the number of diffused impurity atoms within the silicon crystal (or) depends upon the dopant concentration.

The arrow mark or arrow-head represents holes current flow direction (Conventional Current flow direction), when they form a circuit with an external source.

## Forward and Reverse biasing of P-N junction diode:

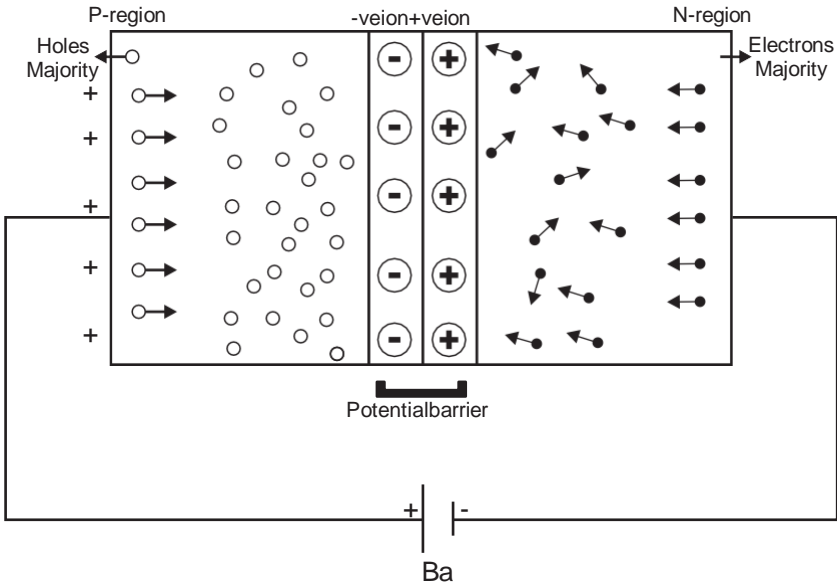
### Biasing:

Applying a suitable d.c. voltage to a diode is known as biasing. It can be done in two ways.

### Forward Biasing:

When the positive terminal of the battery is connected to the P-type semiconductor and the negative terminal to the N-type semiconductor of the P-N junction diode, the junction is said to be forward biased.

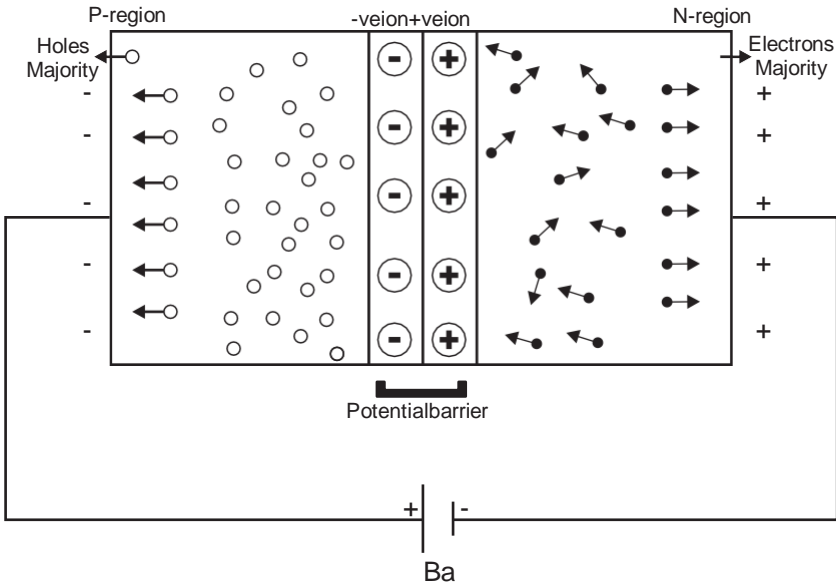
When the applied voltage is increased from zero, the holes and the electrons move towards the junction. Therefore the depletion layer is decreased and disappears i.e., the potential barrier is disappeared.



While the holes and the electrons move across the junction, they combine together and get neutralized. Electrons from the battery move into the N-type semiconductor. Further they move across the junction and come out of the P-type semiconductor. Hence there is a flow of current. The current is due to flow of electrons. Thus the P-N junction diode conducts electricity. As the battery voltage increases, the current also increases. The current is of the order of milliampere (mA).

**Reverse biasing:**

When the negative terminal of the battery is connected to the P-type semiconductor and the positive terminal to the N-type semiconductor of the P-N junction diode, the P-N junction diode is said to be reverse biased. The negative potential of the battery attracts the holes. Similarly, the positive potential of the battery attracts the electrons. Therefore the holes and electrons move away from the junction. Hence the width of the depletion layer increases and there is no current flow through the junction diode during reverse bias. However, as the reverse bias voltage increases, the minority charge carriers move across the junction. Therefore there is a very feeble current, which is of the order of microampere ( $\mu A$ ).



Diode acts as a switch. In forward biased state diode is in "ON" position and in the reverse biased condition the same diode is in "OFF" position. So diode is an electronic device which allows the conventional current flow in one direction alone.

Diodes can be used in a number of ways. For example, a device that uses batteries often contains a diode that protects the device if you insert

the batteries backward. This diode simply blocks any current from leaving the battery if it is reversed. It protects the sensitive electronic devices.

### **P-N junction diode – Rectification:**

A P-N junction diode conducts electricity when it is forward biased. It does not conduct electricity when it is reverse biased. Hence it is used to rectify alternating voltage (A.C.).

The process in which an AC voltage is converted into a unidirectional (D.C) voltage is known as rectification and the circuit used for the conversion is called a rectifier.

An electronic device which converts a.c. power into d.c. power is named as rectifier.

### **Principle:-**

It is based on the principle that a junction diode offers low resistance path for input signal, when forward biased and high resistance under reverse biased condition. When a.c. input is applied to a junction diode, it gets forward biased during the positive half cycle and reverse biased for negative half cycle.

### **The Full Wave Bridge Rectifier**

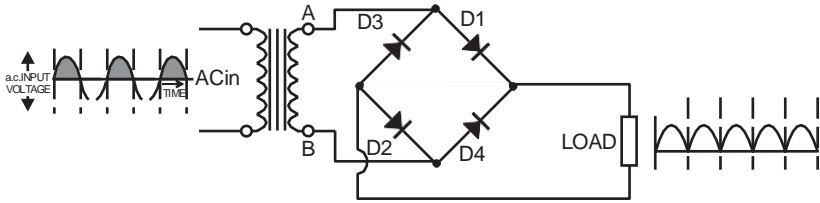
This type of rectifier uses four diodes connected in a closed loop "bridge" configuration to produce the desired output. The main advantage of this bridge circuit is that it does not require a special centre tapped transformer, normally used in the two diodes based full wave rectifier, thereby reducing its size and cost. The secondary winding is connected to one pair of opposite ends of the diode bridge network and the load resistance to the other two opposite corners of the bridge as shown below.

### **The Diode Bridge Rectifier**

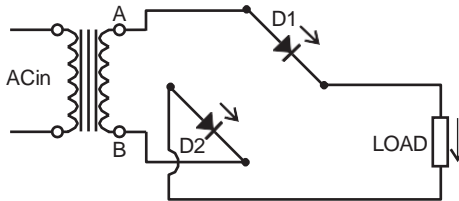
The two diodes labeled  $D_3$  and  $D_1$  are arranged in series pair with respect to each other. Similarly the other two diodes  $D_2$  and  $D_4$  are in series pair with respect to each other. Here the diodes act as a switch, during the forward bias. During the reverse biased state it acts as a non-conductor. So



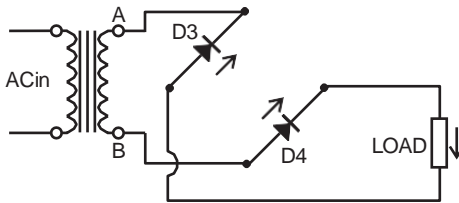
two diodes are in “ON” and other two diodes are in “OFF” condition for each half of the input signal. Only two diodes are conducting current during each half cycle. During the positive half cycle of the supply, diodes D1 and D2 conduct in series while diodes D3 and D4 are reverse biased and the current flows through the load resistance as shown below.



**The Positive Half-cycle**



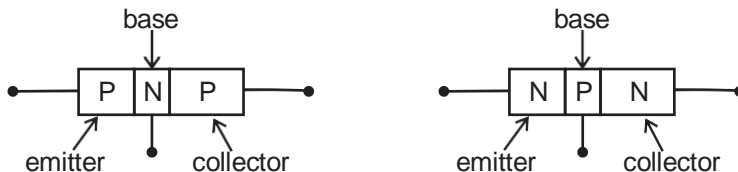
**The Negative Half-cycle**



During the negative half cycle of the supply, diodes D3 and D4 conduct in series, but diodes D1 and D2 switched off as they are now reverse biased. The current flowing through the load resistance is in the same direction as before. That is current flow direction in the load resistance is from top to bottom. As the current flowing through the load is unidirectional, the voltage developed across the load is also unidirectional and thus we get d.c. voltage as output across the load resistor.

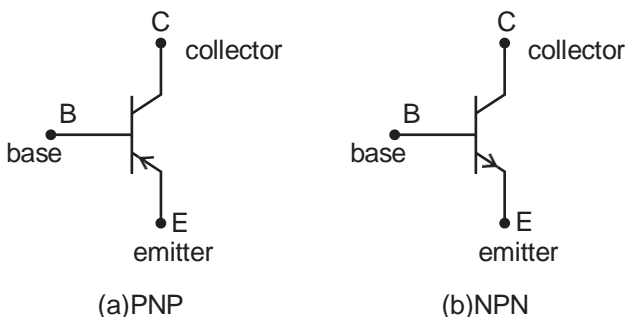
## Transistor:

A junction diode cannot be used for amplifying a signal. For amplification another type of semiconductor device called 'transistor' is used. Transistor is a three-sectioned semiconductor. Transistor is a solid state device. Two P-N junction diode placed back to back form a three layer transistors.



The three sections of the transistors are called emitter[E], base[B] and collector[C]. In a transistor the emitter is heavily doped, since emitter has to supply majority carriers. The base is lightly doped. Two type of transistors are available, namely N-P-N and P-N-P transistor.

## Symbol for transistors:-

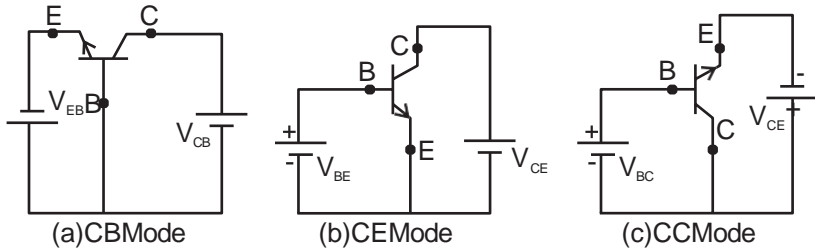


In the symbolic representation for a transistor, the arrow mark is placed on the emitter in the direction of conventional current flow, i.e., from P to N direction.

## Three different Configurations:-

In an electronic circuit a transistor can be connected in three different ways. They are,

- (i) Common base (CB)
- (ii) Common emitter (CE) and
- (iii) Common Collector (CC)



The term common is used to denote the lead that is common to the input and output circuits. The three different modes are shown above for NPN transistor.

For proper working of a transistor, the input junction should be forward biased and the output junction should be reverse biased.

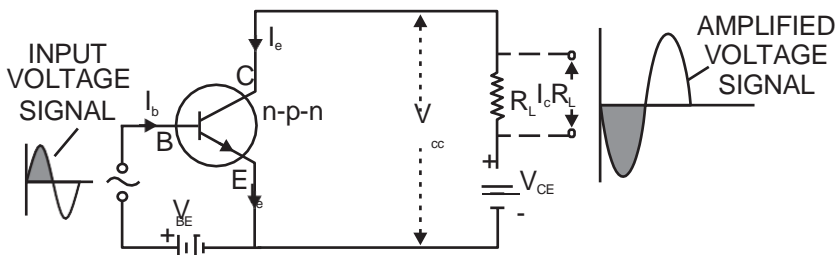
### Why do we need amplification?

When we cannot hear a stereo system, we have to increase the volume, when picture in our television is too dark, we should increase the brightness control. In both of these cases, we are taking a relatively weak signal and making it stronger (i.e., increasing its power). The process of increasing the power of an a.c. signal is called amplification.

The circuit used to perform this function is called amplifier. An amplifier may also be defined as a device, which amplifies the input weak signal. The input signal may be obtained from a phonograph, tape head or a transducer such as thermocouple, pressure gauge etc.

### Common Emitter Amplifier:-

In common emitter amplifier, the emitter is common to both the input and output circuits. The emitter is forward biased by using base-bias battery  $V_{BE}$  and due to the forward bias, the resistance of input circuit is low. The collector is reverse biased by using bias battery  $V_{CE}$ . The low input voltage signal is applied in base-emitter circuit (input circuit) and the amplified output is obtained across the collector and emitter.



The emitter current, base current and collector current are related to each other by the equation.

$$I_e = I_b + I_c$$

Also, corresponding to collector current  $I_c$ , the collector voltage i.e. voltage across the collector and emitter will be

$$V_{ce} = V_{cc} - I_c R_L$$

The variation in input signal voltage causes variation in emitter current, which in turn produces change in collector current and hence in collector voltage. These variations in collector voltage, which are proportional to the input emitter current, appear as amplified output voltage. The output and input voltage will be out of phase by  $180^\circ$  as shown in the above figure.

### A.C. voltage gain:-

It is the ratio of the change in output voltage to the change in input voltage. It is denoted by  $A_v$ .

Suppose that on applying an a.c. input signal, the input current changes by  $\Delta I_b$  and corresponding change in the output current by  $\Delta I_c$ . Then,

$$A_v = \frac{\text{Change in output voltage}}{\text{input voltage}} = \frac{\Delta I_c \times R_{out}}{\Delta I_b \times R_{in}} = \frac{\Delta I_c}{\Delta I_b} \times \text{resistance gain}$$

### A.C.powergain:

It is the ratio of the change in output power to the change in input power.

$$\begin{aligned} \text{A.C.powergain} &= \frac{\text{Change in output power}}{\text{Change in input power}} \\ &= \frac{(\Delta I_C)^2 R_{out}}{(\Delta I_b)^2 R_{in}} \end{aligned}$$

### Advantages of common emitter configuration:

On comparing the three different configurations of an amplifier, with help of their characteristics like input impedance, output impedance, current gain, voltage gain, power gain and phase reversal the following are the major advantages of common emitter configuration when compared with other configurations.

- (a) Medium input impedance
- (b) Medium output impedance
- (c) High current gain
- (d) High voltage gain
- (e) High Power gain
- (f) Phase reversal

In common emitter amplifier, the output signal is  $180^\circ$  out of phase with input signal.

## 5.4 DIGITAL ELECTRONICS

### Introduction

With the invention of the transistor in 1948 by W.H.Brattain and I.Bardeen, the electronic circuits became considerably reduced in size. It was due to the fact that a transistor is cheaper, easily available, smaller compare to vacuum tube, less power consuming. It is widely used in many electronic circuits.

With the development of printed circuit board (PCB) which further reduced the size of electronic equipments in the early 1960's a new field of "microelectronics" was born primarily to meet the requirements of the military, which wanted to reduce the size of its electronic equipment. The drive for extremely small size of microelectronic circuits ends up with an invention of integrated circuits (ICs). Integrated circuits can contain anything from one to millions of logic gates, flip-flops, multiplexers, and other circuits in a few square millimetre. The small size of these circuits allows high speed, low power dissipation and reduced manufacturing cost compared with board-level integration (PCB).

This increased capacity per unit area can be used to decrease cost and increase functionality.

In electronics, an **integrated circuit** (also known as **IC chip**, or **microchip**) is a miniaturized electronic circuit (consisting mainly of semiconductor devices, as well as passive components) that has been manufactured on the surface of a thin substrate of semiconductor material. Integrated circuits are used in almost all electronic equipment in use today and have revolutionized the world of electronics.

The integrated circuit was invented by Jack Kilby and Robert Noyce. This invention is a boon for digital technologies like computer, mobile phones, MP3 and DVD's. This list could be almost infinite.

### **Digitalelectronics:**

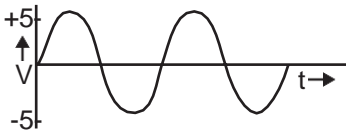
In the modern electronic world, signal processing plays vital role. Different types of signals with different shapes are generated by different devices. Two shapes of signals or waves are considered here.

### **Analog signal and Digital signal:**

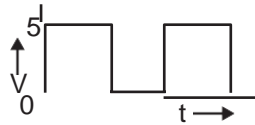
The voltage signals which vary continuously with time are called continuous or analog voltage signals. The Fig(a) below shows a typical voltage signal, varying as a sinusoidal wave of 0 to 5v.

In general, symmetrical square wave forms a digital signal. Fig(b) below shows a digital voltage signal, which does not vary continually with time. The values will be equal to 0V or 5V. By representing these two voltage levels as a binary numbers, 0 and 1 can be formed.

Digital signal processing is familiar and formed a digital world. The



(a)



(b)

counters, computers, etc are outcome of digital electronics.

### Positive and Negative Logic:

In computing system, the binary numbers symbols "0" and "1" represent two possible states of a circuit or an electronic device.

In positive logic, the "1" represents

- |                    |                        |                     |
|--------------------|------------------------|---------------------|
| 1. an "ON" circuit | 2. a "CLOSED" switch   | 3. a "HIGH" voltage |
| 4. a "PLUS" sign   | 5. a "TRUE" statement. |                     |

Consequently, the "0" represents

- |                     |                         |                    |
|---------------------|-------------------------|--------------------|
| 1. an "OFF" circuit | 2. an "OPEN" switch     | 3. a "LOW" voltage |
| 4. a "MINUS" sign   | 5. a "FALSE" statement. |                    |

In Negative logic, just opposite conditions prevail.

Suppose, a digital system has two voltage levels like 0V and 5V.

If we say that value 1 stands for 5V and value 0 for 0V, then we have positive logic system.

If on the other hand, we decide that "1" should represent 0V (low voltage) and 0 should represent 5V (high voltage), then we have negative logic system.

### Logic gates:

The logic gates are building blocks of digital electronics. They are used in digital electronics to change one voltage level (input voltage) into another (output voltage) according to some logical statement relating them.

Thus, logic gate is a digital circuit, which works according to some logical relationship between input and output voltage.

The logic gate may have one or more inputs, but only one output.

Truth table of a logic gate is a table that shows all possible input combinations and the corresponding output for the logic gate.

The logical statements that logic gates follow are recalled "Boolean

expressions”.

1 In Boolean algebra, the addition sign (+) is referred as OR. The Boolean expression is

$$y = A + B$$

This Boolean expression is read as y is equal to A 'OR' B.

2. The multiplication sign (.) is referred as AND in Boolean algebra. The Boolean expression is

$$y = A . B$$

This Boolean expression is read as y is equal to A 'AND' B.

3. The bar sign (-) is referred to as NOT in Boolean algebra. The Boolean expression is

$$y = \bar{A}$$

This expression is read as y is equal to 'NOT' A.

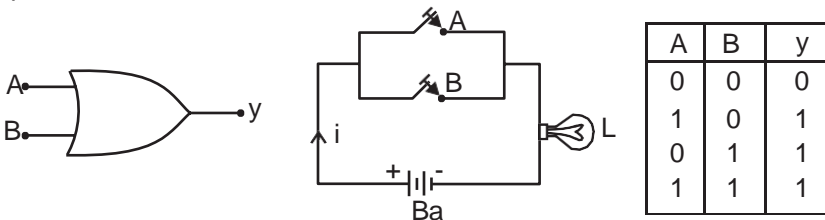
### BASIC LOGIC GATES:

The following three gates viz., OR gate, AND gate and NOT gate are called as basic logic gates.

#### OR-GATE

In general, the simple OR gate is a two input and one output logic gate. It combines the inputs A and B to give the output y, following the Boolean expression  $y = A + B$ . The symbol, truth table and equivalent electrical circuit are shown below.

The function of the OR gate is, the output is “TRUE” if any one of its inputs is in 'TRUE' condition.



Two parallel switches and lamp is connected as shown above.

Lamp “ON” is equal to 1

Lamp “OFF” is equal to 0.



Case1:-

If both the switches A and B are open. No current flows through the external wire. So Lamp is "OFF" i.e. equal to "0".

Case2:-

If the Switch A is closed and B is open. The conventional current passes through "A" Lamp is "ON" i.e. equal to 1

$$1+0=1$$

Case3:-

If Switch A is open and Switch B is closed. Current passes through "B", So Lamp is "ON" i.e. equal to 1.

$$0+1=1$$

Case4:-

If both the switches are closed, Lamp is "ON", i.e. equal to 1.

$$1+1=1$$

So, for all possible input and the obtained output values are tabulated. It is called truth table.

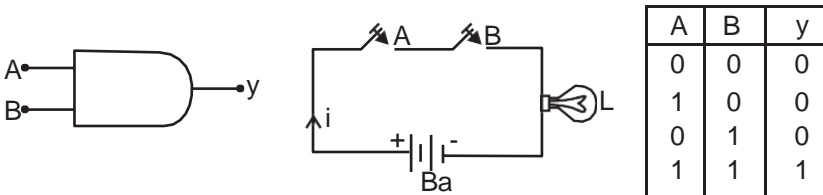
**AND gate:-**

In general, the simple AND gate is also a two-input and one output logic gate. It combines the inputs A and B to give the output Y, following the Boolean expression

$$y = A \cdot B$$

i.e.  $y$  is equal to A "AND" B. The symbol, equivalent circuit and truth table are shown below.

The function of the AND gate is, the output is 'TRUE' if and only if all the inputs are in 'TRUE' condition.



In the above equivalent diagram the switches and the lamp is in series connection mode.

**Case1:-**

If both the switches A and B are OPEN (i.e., A=0, B=0) then the lamp will not glow (i.e. y=0). The current will not pass through the lamp.

**Case2:-**

If the switch “A” is closed and “B” is open, current will not pass through the lamp. The lamp is OFF, so y=0.

**Case3:-**

If the switch “B” is closed and “A” is open, current will not pass through the lamp. The lamp is OFF, so y=0.

**Case4:-**

If both switches A and B are closed, current will pass through the circuit. Now the lamp is “ON” and glowing. So y=1.

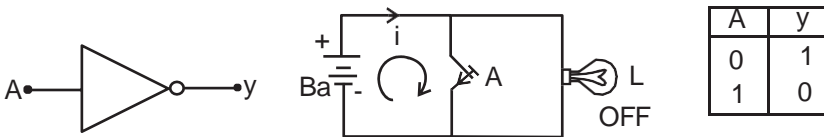
**NOT gate:-**

The NOT gate is a one input and one output logic gate. It inverts or complements the input A to give output y following the Boolean expression.

$$y = \overline{A}$$

This gate is also called as 'inverter'. The symbol, equivalent circuit and truth table are shown below. The equivalent diagram is shown below.

The function of the NOT gate is to invert or complement its input. i.e., the output is 'TRUE' if input is 'FALSE' and vice versa.



**Case1:-**

If the switch “A” is open, the current will pass through the lamp and lamp glows.

So, y=1 when A=0

**Case2:-**

If the switch “A” is closed, the current will take shortest path. That is current will pass through the switch above. Hence the lamp is “OFF”, so  $y = 0$  when  $A = 1$ .

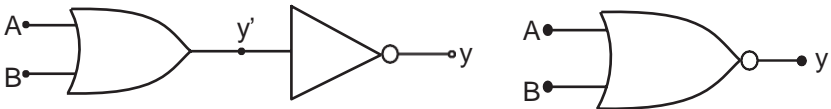
**UNIVERSAL LOGIC GATES:**

Any logic gate can be made from combinations of NAND gate or NOR gates. Hence, the NOR gate and NAND gate are called as Universal logic gates.

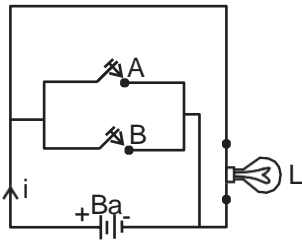
**NOR gate**

It is a logic gate in which OR gate is followed by a NOT gate. The symbol, equivalent circuit and truth table are shown below.

The function of this gate is 'inverting' the output of the OR gate.



It combines the inputs A and B to give the output y, by the following Boolean expression  $y = \overline{A+B}$



A	B	y'	y
0	0	0	1
1	0	1	0
0	1	1	0
1	1	1	0

A	B	y
0	0	1
1	0	0
0	1	0
1	1	0

**Case1:-**

If both the switches A and B are open, the current flows through the lamp, Lamp is “ON”. So  $y = 1$ .

**Case2:-**

If the switch A is closed and B is open, the current will pass through the shortest path or low resistance path. Hence conventional current flow through the switch “A” and reaches the cathode of the battery. The lamp is “OFF”, so  $y = 0$ .

**Case3:-**

If the switch A is open and B is closed, the conventional current will pass through the shortest path. The current will pass through the switch B and reach the cathode of the battery. Current will not pass through the lamp, so lamp is OFF, i.e.  $y=0$ .

**Case4:-**

If the switch A and B are closed, the current from the battery will pass through two parallel switches and reach the cathode of the battery. Current will flow through shortest path (or) low resistance path. The lamp is "OFF", so  $y = 0$ .

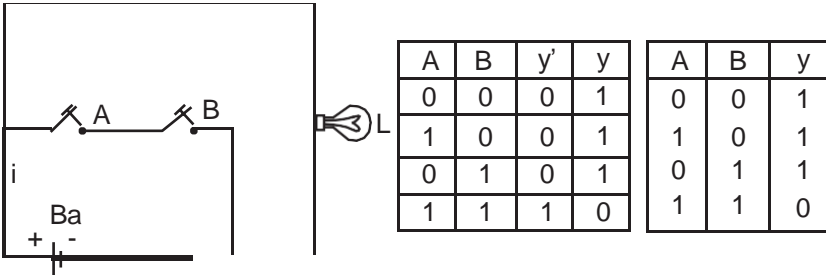
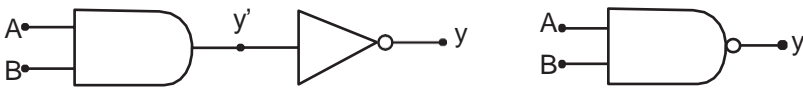
**NAND gate:-**

It is logic gate in which AND gate is followed by NOT gate. The symbol, equivalent circuit and truth table are shown below.

The function of this gate is "inverting" the output of the AND gate.

It combines the inputs A and B to give the output y, by the following Boolean expression

$$y = \overline{A \cdot B}$$



**Case1:-**

If both switches A and B are open, the current will flow through the lamp. The lamp is ON, so  $y=1$

**Case2:-**

If the switch A is closed and B is open, the current will again flow through the lamp. The lamp is ON, so  $y=1$ .

**Case3:-**

If the switch A is open and B is closed, the current flows through the lamp. The open switch A acts as a circuit breaker. The lamp is ON, so  $y=1$ .

**Case4:-**

If the switch A and B are also in closed position, the current will pass through the switches and reach the cathode of the battery. Current will flow through shortest path (or) low resistance path. Hence, the lamp is OFF, So,  $y=0$

**Simple application of NAND gate:-**

NAND gate is used to design an interior lighting system of a car. The door position logic is used here. That is, Door "open" is equal to "0" input. The door "closed" is equal to "1" input. Now a table is drawn for different combinations of these inputs

Door A	Door B	Lamp $y = \overline{A \cdot B}$
0 open	0 open	1 (ON)
1 closed	0 open	1 (ON)
0 open	1 closed	1 (ON)
1 closed	1 closed	0 (OFF)

From the above table it is clear that the lamp inside a car goes off automatically, when both doors are closed. The same NAND gate logic can be extended for all the four doors of a car with four inputs.

**SPECIAL LOGIC GATES:**

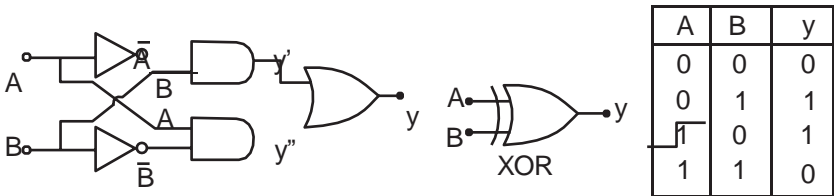
The following two logic gates, viz., the XOR and XNOR gates are called as special logic gates.

**1. The XOR gate:-**

EXCLUSIVE OR gate is abbreviated as XOR gate. The function of this gate is to produce the TRUE condition on its output, when any one of its input conditions is TRUE, but not both. i.e. the output of the gate is HIGH (ON condition) only when the inputs are different.

An OR gate recognises words with one or more 1s. The XOR gate recognises only words that have odd number of 1s. This is why the circuit is known as EXCLUSIVE OR gate.

An XOR gate is obtained by using OR, AND & NOT gates. It is a combination of all the three basic logic gates. The diagram, symbol and truth table is shown below



The Boolean expression to represent XOR operation is

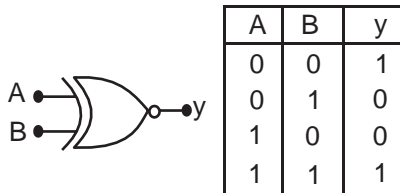
$$Y = A + B \oplus$$

which is read as A XOR B = AB + AB

**2. The "XNOR" gate:-**

The XNOR gate (sometimes spelled 'exnor' or 'enor') is a digital logic gate whose function is the inverse of the exclusive OR (XOR) gate. The symbol and truth table for XNOR gate are shown below.

The Boolean expression is  $Y = A + B \ominus$



## **Integrated Circuits:-**

An Integrated Circuit (IC) is just a compact electronic circuit. It has both active and passive components.

A discrete circuit is one that is built by connecting separate components. In this case, each component is produced separately and then all are assembled together to make the electronic circuit. For example, consider house wiring. In that, basically switches, lamp and wires are used. They are different components made by different manufactures. We assemble them to form a circuit to energise the lamp and other equipments.

In IC, all components are manufactured on the same silicon wafer. Diodes, transistors, resistors, capacitors and their interconnection is also established well to perform a specific task. Hence all active and passive components are formed on the same chip, using microelectronic techniques.

## **Scale of Integration:-**

In scale of integration, the circuit density and component density is playing important role. For example, 1 lakh components can form 100 circuits. Another circuit designer can form 1000 circuits using the same 1 lakh components. Hence comparing, circuit density and component density is not unique. It all depends upon the application and circuit designer's idea.

An approximate method of classifying the amount of circuit or components density is as follows:-

### **1. SSI – Small Scale Integration**

In this case, the number of circuits packed in one IC is less than 12 (or) number of components is less than 50.

### **2. MSI – Medium Scale Integration:**

Here, number of circuits per package is between 13 and 99 or number of components is between 50 and 500.

### **3. LSI – Large Scale Integration:-**

In this case, circuit density is between 100 and 9999 or components density is between 5000 and 100000 (hundred thousand).

**4. VLSI–VeryLargeScaleIntegration:-**

Herethenumberofcircuitsperpackageisbetween10,000to 99,999ornumberofcomponentsisbetween100000and1000000.

**5. ULSI–UltraLargeScaleIntegration:-**

In this case, the circuit density is between 100 000 to 999 999 or numberofcomponentsbetween1000000–10000000.

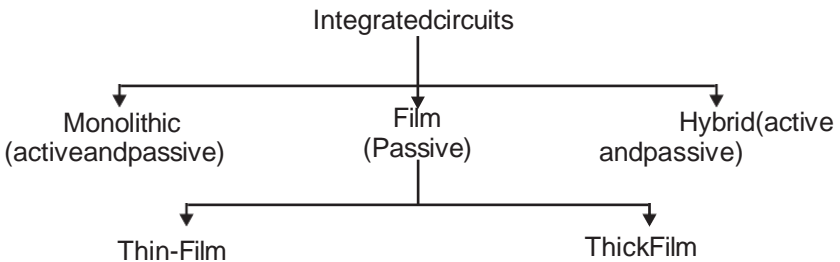
**6. GSI–GigaScaleIntegration:-**

Herethenumberofcircuitsperpackageis1000000ormore. Numberofcomponentsare100000000(Ninedigit)

InSummary;

<b>CircuitsperChip</b>	
SSI	<12
MSI	12-99
LSI	100–9999
VLSI	10000–99999
ULSI	10000–9999999
GSI	>1000000

**ClassificationofICs:-**

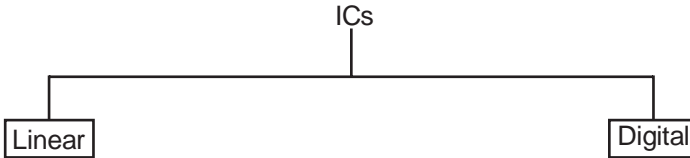




### **Classification of ICs function:-**

The earlier classification of ICs was based on their method of construction, based on circuit density and component density.

The hi-fi world is based on the signal processing. In general two types of signals are referred here, analog and digital signals. Now, these ICs are classified further based on the signal they process.



#### **(a) Linear ICs:-**

The Linear Integrated Circuits (LICs) are also referred as analog ICs. Sinewaves are used and processed in this ICs. LICs find wide use in military and industrial applications.

#### **(b) Digital Integrated Circuits:-**

In Digital Integrated Circuits (DICs), signal processed is digital signals. Digital signals are usually binary. DICs are used in calculator chips, micro processors, memory chips, etc.

### **Advantage of ICs:-**

#### **ICs have the following advantages:-**

1. Extremely small physical size. Often the size is thousands of times smaller than discrete circuit.
2. Weightless when compared with discrete circuit. Since many circuit functions can be packed into a small area, in many applications where weight and space are critical, such as in air craft, space shuttle, satellite, etc., the IC invention and application brought a very big change in the satellite communication and satellite launching.
3. Lifetime of ICs is long which is most important from both military and consumer application point of view. Most significant factor is the absence of soldered connections.

Extremely high reliability. IC damages are less under proper usage.

4. In signal processing, the signal transfer from one circuit to another circuit or from one device to another device is fast when we use ICs. Moreover, because of short distance between the internal circuits, the stray signal pick is less and in many ICs. Response time and speed are more, on the other hand.
5. Low power consumption.
6. Easy replacement
7. Cost effective due to mass production of IC and less failure rate in mass production.

#### **Disadvantages of ICs:-**

1. Fabrication of coil and inductor is not possible on a silicon wafer.
2. ICs function at fairly low voltages.
3. They handle limited amount of power.
4. ICs are temperature sensitive and usage of ICs in hot environment is a major disadvantage. So ICs are highly insulated from heat and radiation in space technology applications.
5. The various components and their inter connections are tiny. So powerful microscopes are needed to check the ICs.
6. IC failures may be due to defects in semiconducting materials. Apart from defects in the crystals used for IC fabrication, the dust particles will also cause a major failure in ICs fabrication. So high class clean room is needed.
7. Single crystal growth without defects needs sophisticated machines.

## **QUESTIONS**

### **Part– A**

1. Define capacitance of a capacitor.
2. Define farad.
3. Write the dielectric medium used in a capacitor.
4. Write any two applications of capacitors.
5. Write the energy stored in a capacitor, with details.
6. Write the expressions for effective capacitance when two capacitors are connected in Series and in parallel.

7. Find the effective capacitance when two equal capacitors are connected in series.
8. What is forbidden energy gap?
9. What is Fermi level?
10. What is Doping?
11. What is the purpose of doping?
12. Name the two types of extrinsic semiconductor and write the majority and minority carriers in each.
13. What is a hole in P-type semiconductors?
14. What is potential barrier in P-N Junction diodes?
15. What is depletion region in P-N junction diodes?
16. How will you connect the battery for forward bias? For this bias, does the diode act as a conductor or non-conductor?.
17. How will you connect the battery for reverse bias? For this bias, does the diode act as a conductor or non-conductor?
18. What is the advantage of a bridge type rectifier?
19. Draw the symbol for PNP and NPN transistor.
20. Write the three different configurations for a transistor used in electronic circuits.
21. Write any two advantages of common emitter configuration.
22. What is amplification?
23. What is the difference between analog and digital waves?
24. What is positive logic?
24. What is negative logic?
25. What are the basic logic gates?
26. Draw the symbol for OR gate and write the Boolean Expression for the same.
27. Draw the symbol for AND gate and write the Boolean Expression for the same.
28. Draw the symbol for NOT gate and write the Boolean Expression for the same.
29. What are the universal logic gates?
30. Draw the symbol for NAND gate and write the Boolean Expression for the same.

31. What are the special logic gates?
32. Draw the symbol for NOR gate and write the Boolean Expression for the same
33. Draw the symbol for XOR gate and write the Boolean Expression for the same
34. Draw the symbol for XNOR gate and write the Boolean Expression for the same.
35. What is an Integrated Circuit?
36. Scaling of ICs is based on what factors?
37. What is linear IC?
38. What is digital IC?

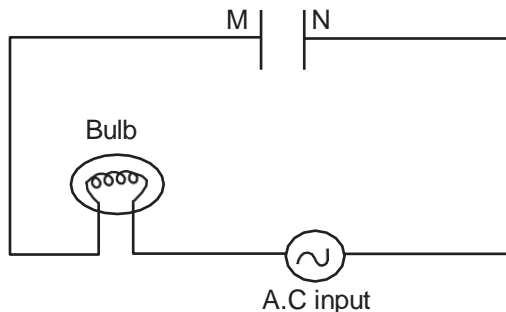
### **Part-B**

1. Derive the expression for effective capacitance when three capacitors are connected in series.
2. Derive the expression for effective capacitance when three capacitors are connected in parallel.
3. Show that when two equal capacitors are connected in parallel, the system has four times the capacitance of that obtained, when they are connected in series ( $C_p = 4C_s$ )
4. Find the resultant capacitance of three capacitors of capacitance 20mF, 30mF and 50mF, when they are connected in series and in parallel.
5. The capacitors of capacitance 3mF and 5mF are connected in series. Calculate the effective capacitance of the capacitors and potential difference across each Capacitor, when they are powered by 60VDC supply.
6. Explain how to identify good conductors, semiconductors and insulators using energy band diagram.
7. Write short notes about intrinsic semiconductor.
8. Explain the Formation of Extrinsic P-type semiconductors.
9. Explain the Formation of Extrinsic N-type semiconductors.
10. Explain the P-N junction diode formation, depletion region and barrier potential.

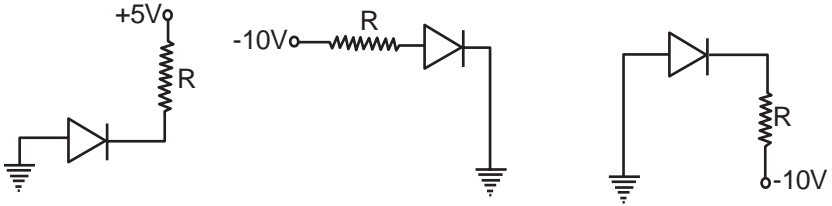
11. Explain the forward bias action in a P-N junction diode, with a neat diagram.
12. Explain the reverse bias action in a P-N junction diode, with a neat diagram.
13. Explain the full wave bridge rectifier with a neat circuit diagram and input and output waveforms.
14. Explain the working of a transistor connected in common emitter configuration, as an amplifier.
15. Explain all the three basic logic gates with symbols, Boolean expressions and Truth tables.
16. Explain the NAND and NOR gates with symbols, Boolean expressions and truth tables.
17. Explain the XOR and XNOR gates with symbols and truth tables.
18. Explain level of integration in ICs based on circuit densities and component densities.
19. Discuss the advantages and disadvantages of ICs.

**Brainteasers:**

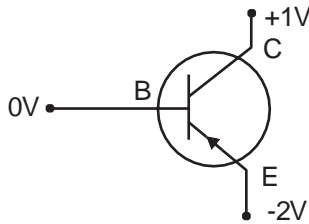
1. Explain how a reversed bias diode acts as a variable capacitor.
2. Two capacitors of capacitance 3 m F and 6 m F are connected in series and charged by connecting a battery of 10v. Next replace the battery from the circuit, connect a wire between the terminals (show circuiting process). What is the final charge stored on the capacitors?
3. Analyze the circuit given below and explain the current flow direction through the circuit. When will the lamp glow connected in series with the capacitor?



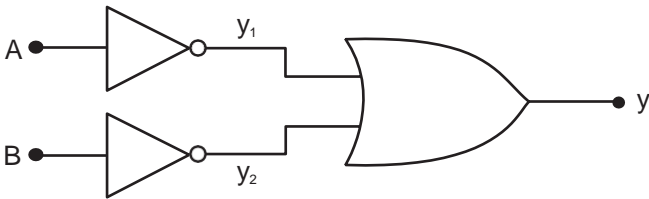
4. In the following circuits, which of the diodes is forward-biased and which is reverse-biased?



5. How the P.N.P transistor shown below is biased?



6. Prove that the following diagram's truth table is equal to NAND gate truth table.



**SECOND SEMESTER**  
**2004 ENGINEERING PHYSICS-II**  
**MODEL QUESTION PAPER-1**

Time: 3 Hrs

Max Marks: 75

**PART-A**

Marks 15x1=15

**Note:** Answer any 15 Questions.

1. Define convection.
2. Define coefficient of thermal conductivity.
3. Write the expression for root mean square velocity of molecules.
4. Define specific heat capacity of a substance.
5. Write First Law of thermodynamics.
6. Define isothermal change.
7. Define temperature of inversion.
8. Write any two advantages of renewable energy.
9. Write the Snell's law of refraction.
10. What is population inversion?
11. Write any two applications of photoelectric cells.
12. What is remote sensing?
13. Define resistivity.
14. Write Joule's law of heating.
15. Write Fleming's Left Hand Rule.
16. Define mutual induction.
17. Define 'farad'.
18. What is a 'hole'?
19. Write the three possible configurations in which a transistor can be connected.
20. Write the various levels of integration.

**PART-B**

Marks 5x12=60

**Note:** i) Answer all Questions choosing any two subdivisions from each question.

ii) All subdivisions carry equal marks.

1a) List the properties of thermal radiation.

b) Derive an expression for the pressure of a gas on the basis of postulates of kinetic theory of gases.

- c) The ratio of specific heat capacities of a gas is 1.66. Assuming the value of universal gas constant  $R$  as  $8.12 \text{ JK}^{-1} \text{ mol}^{-1}$ , find the values of specific heat capacity at constant pressure and specific heat capacity at constant volume of the gas.
- II a) 10 litres of gas at 5 atmospheric pressure is suddenly compressed to half of its initial volume. Find the resulting pressure, assuming  $\gamma = 1.667$ .
- b) Explain the cascade process of liquefaction of oxygen.
- c) Explain solar energy, wind energy and tidal energy.
- III a) Explain the application of total internal reflection in optical fibre cable as a waveguide.
- b) Explain the production of LASER using a Semiconductor diode source.
- c) Explain the working principle of RADAR with a block diagram. IV
- a) Derive the condition for balancing Wheatstone's network.
- b) Derive an expression for the torque acting on a rectangular current carrying coil placed in a uniform magnetic field.
- c) A Galvanometer of resistance  $75 \Omega$  shows full scale deflection for a current of  $100 \text{ mA}$ . How will you convert it into i) an ammeter to read a maximum of  $5 \text{ A}$  and ii) a voltmeter to read a maximum of  $10 \text{ V}$ ?
- V a) Three capacitors of values  $10 \text{ F}$ ,  $20 \text{ F}$  and  $30 \text{ F}$  are used in a circuit. Find the effective capacitance when they are connected i) in series and ii) in parallel.
- b) Explain the working of a Full Wave rectifier with a neat circuit diagram using P-N junction diodes in bridge type Configuration.
- c) Explain the working of a NPN transistor as an amplifier in common emitter configuration, with a neat circuit diagram.



**SECOND SEMESTER**  
**2004 ENGINEERING PHYSICS-II**  
**MODEL QUESTION PAPER-II**

Time: 3 Hrs

Max Marks: 75

**PART-A**

Marks 15x1=15

**Note:** Answer any 15 Questions.

1. What is conduction?
2. Write any two properties of thermal radiation.
3. What is the relation between the pressure and kinetic energy of a gas?
4. Define specific heat capacity of a gas at constant pressure.
5. Write Clausius statement of second law of thermodynamics.
6. What is an adiabatic change?
7. What is the principle of the cascade process?
8. What is solar energy?
9. Define refractive index of a medium.
10. Write any two advantages of Optical Fibre Cable.
11. Write Einstein's photoelectric equation and explain each term in that equation.
12. Expand the acronym RADAR.
13. Write the statement of Ohm's law.
14. Write Faraday's first law of electrolysis.
15. What is a shunt?
16. Write the statement of Lenz's law.
17. Write any two applications of capacitors.
18. What is doping?
19. Define rectification.
20. Draw the symbolic representation of a NAND gate and write its Boolean expression.

## PART-B

Marks 5x12=60

Note: i) Answer all Questions choosing any two subdivisions from each question.

ii) All subdivisions carry equal marks.

- I a) Explain the selection of good and poor thermal conductors based on their thermal conductivities.  
b) The density of a gas is  $1.25 \text{ kg m}^{-3}$  at N.T.P. Find the Root mean square velocity of its molecules.  
c) Derive Myer's relation  $C_p - C_v = R$ .
- II a) 9 liters of a gas at 1 atm pressure is compressed slowly to pressure of 3 atm. Find the resulting volume.  
b) Explain liquefaction of air by Linde's process.  
c) Explain geothermal energy and biomass.
- III a) Explain spontaneous and stimulated emissions.  
b) Explain the working of photoemissive cell and photovoltaic cell.  
a) Explain the components of remote sensing.
- IV a) Calculate the length of the wire required to make a 10 ohm resistance, if the radius of the wire is 0.25 mm and the resistivity of the material of the wire is  $1 \times 10^{-6} \Omega \text{m}$ .  
b) Describe an experiment to determine the electro chemical equivalent of copper using copper volta meter.  
c) Describe the construction and working of a moving coil galvanometer.
- V a) Derive an expression for effective capacitance when capacitors are connected in parallel.  
b) Explain the energy band diagram of good conductors, insulators and semiconductors.  
c) Explain various levels of integration.

## SECOND SEMESTER

### PRACTICAL -II

#### REFRACTIVE INDEX

**Experiment No.:**

**Date:**

**Aim:**

To determine the refractive index of a transparent liquid (water) using a travelling microscope.

**Apparatus required:**

Given liquid, glass beaker, travelling microscope, lycopodium powder, pin, etc.

**Formula:**

The refractive index of liquid (water)

$$\mu = \frac{\text{Real depth of the liquid}}{\text{Apparent depth of the liquid}}$$
$$= \frac{\text{Reading 3} - \text{Reading 1}}{\text{Reading 3} - \text{Reading 2}}$$

**Description:**

A travelling microscope is an ordinary microscope fixed on a stand in such a way that it may be made to travel in vertical as well as horizontal direction without disturbing its adjusted focus. The readings are recorded by means of main scale and vernier scale of high accuracy (0.001 cm) attached to the instrument.

**Procedure:**

The least count of the microscope is determined as in the case of vernier calipers.

A small pin is fixed horizontally with wax or cellotape at the bottom of the empty beaker. The tip of the pin is focussed clearly on the microscope and the corresponding main scale reading (MSR) and vernier scale coincidence (VSC) in the vertical scale are noted in the tabular column. (Reading 1)

The given liquid is taken in the beaker. Now the pin is apparently raised through a height. Therefore, it will be no longer in focus. The microscope is moved up without changing the adjusted focus, so that the image of the pin is clearly seen through the microscope. The corresponding main scale reading (MSR) and vernier scale coincidence (VSC) in the vertical scale are noted in the tabular column. (Reading<sub>2</sub>)

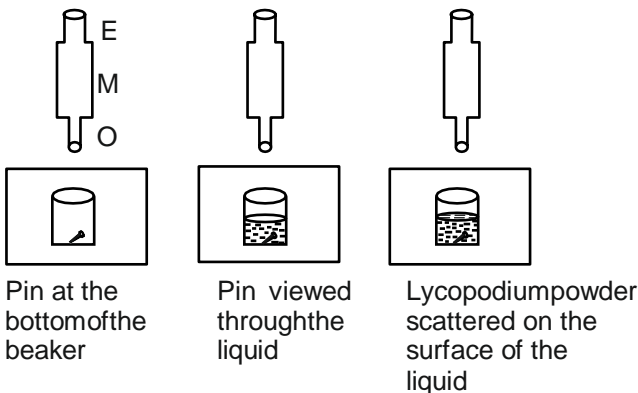
Finally, a little lycopodium powder or saw dust which can float on liquid is scattered on the surface of the liquid. The microscope is further moved up without changing the adjusted focus, so that the clear image of lycopodium powder is seen through the microscope. The corresponding main scale reading (MSR) and vernier scale coincidence (VSC) in the vertical scale are noted in the tabular column. (Reading<sub>3</sub>)

The difference between this Reading<sub>3</sub> and Reading<sub>1</sub> gives the real depth of the liquid, whereas the difference between Reading<sub>3</sub> and Reading<sub>2</sub> gives the apparent depth of the liquid.

By substituting the readings in the formula, the refractive index of the liquid (water) is determined.

**Result:**

The refractive index of the liquid (water) =



**To find Least count of microscope:**

Least count: Value of 20 MSD = cm  
 Value of 1 MSD = cm  
 No. of divisions on the vernier scale (n) =  
 =  
 Least Count (LC) =  $\frac{1}{n}$  x value of 1 MSD  
 = cm

**Tabular column:**

Least count = cm

Position	Microscope Readings			
	MSR	VSC	VSR = VSC x LC	OR = MSR + VSR
Unit	cm	div	cm	cm
Clear image of tip of the pin fixed at the bottom of the beaker.				(Reading 1)
Clear image of tip of the pin, after the liquid is taken in the beaker.				(Reading 2)
Clear image of lycopodium powder scattered on the surface of liquid.				(Reading 3)

**Calculation:**

The refractive index of liquid (water)

$$\mu = \frac{\text{Real depth of the liquid}}{\text{Apparent depth of the liquid}}$$

Real depth of the liquid = Reading 3 – Reading 1  
 = cm - cm  
 = cm

Apparent depth of the liquid = Reading 3 – Reading 2  
 = cm - cm  
 = cm

Refractive index of the liquid  $\mu =$

## SOLAR CELL

**Experiment No.:**

**Date:**

**Aim:**

To draw the V-I characteristics curve for the given solar cell.

**Apparatus required:**

Solar cell, milliammeter (0–100 mA), voltmeter (0–2 V), variable resistor (0–1 k $\Omega$ ), 60 W bulb, connecting wires, plug key.

**Theory:**

The solar cell is a semiconductor device that converts solar energy (electromagnetic radiation) into electrical energy by photovoltaic effect. Solar cells are widely used to produce electrical energy from sun's radiation.

A solar cell consists of a p-n junction diode, in which electrons and holes are generated by the incident photons (i.e. electromagnetic radiation). When an external circuit is connected through the p-n junction device, a current passes through the circuit. Therefore, the device generates power when the electromagnetic radiation is incident on it.

**Procedure:**

A voltmeter, milliammeter and a variable resistor of 0 to 1000 ohm and a plug key are reconnected as shown in the circuit diagram.

The solar cell is illuminated by a filament bulb of 60 W. Now the two pins of the solar cell are plugged into the two terminals (A and A') of the

milliammeter and the short circuit current  $I_{sc}$  is noted, keeping the plug key in open condition.

Then the two pins of the solar cell are plugged into the two terminals (V and V) of the voltmeter and the open circuit voltage  $V_{oc}$  is noted, keeping the plug key in open condition.

Now the plug key is closed to include the variable resistor in the circuit. By slowly changing the value of resistance included in the circuit, the milliammeter readings are noted for various values of voltmeter readings in steps of 0.1V or 0.2V.

The readings of both the milliammeter and voltmeter are entered in the tabular column. With these readings, the V-I characteristics graph of the solar cell is drawn.

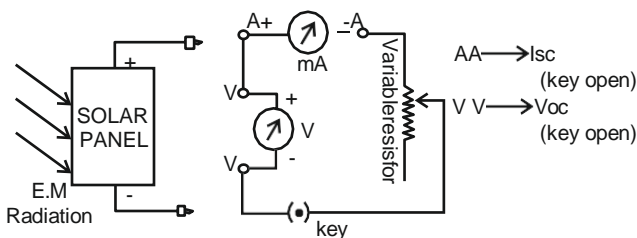
**Result:**

Thus the V-I characteristic of the solar cell is drawn.

**Note:**

- i) The  $V_{oc}$  and  $I_{sc}$  values depend on the type of solar cell and number of solar cells used in the solar panel.
- ii) These values also depend on the source of electromagnetic radiation, which we use to irradiate the solar cell.
- iii) Care should be taken so that the bulb and solar cell do not come into direct contact. (I.e. touch each other). In such case the readings of the meters may go beyond the range and the solar cell may get damaged.

**Circuit Diagram**



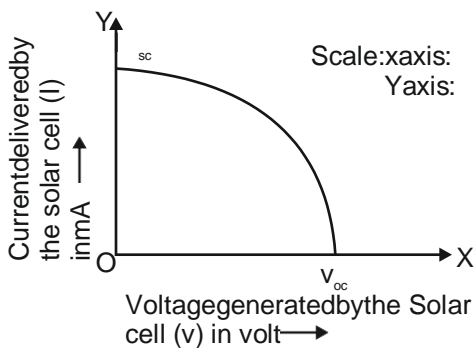
**TabularColumn:**

Shortcircuitcurrent  $I_{sc}$  = mA

Opencircuitvoltage  $V_{oc}$  = volt

S.No	Voltagegeneratedbythe solar cell (V)	Currentdeliveredbythe solar cell (I)
Unit	volt	milliampere

**ModelGraph:**



## SPECTROMETER

**ExperimentNo.:**

**Date:**

**Aim:**

To determine the angle of the prism and the angle of minimum deviation using spectrometer and to calculate the refractive index of glass.

**Apparatusrequired:**

Spectrometer, given glass prism, Sodium vapour lamp, Reading lens, blackcloth, etc.



**Formula:**

The refractive index of the material of the glass prism

Where  $A$  = angle of the prism

$D$  = angle of minimum deviation

**Description:**

$$\mu = \frac{\sin\left[\frac{A+D}{2}\right]}{\sin\left[\frac{A}{2}\right]}$$

Spectrometer is an optical instrument having three parts

(i) Collimator (ii) Telescope and (iii) Prism table. The following initial adjustments should be carried out before starting any experiment with a spectrometer.

- a) The telescope is turned towards a white wall of background and the eyepiece is gently pushed in or pulled out so that a clear image of the cross wires is seen without parallax.
- b) To adjust the telescope for parallel rays, the telescope is directed towards a distant object like a distant tree or building and the distance between the objective lens and the eyepiece is altered till the clear image of the distant object is seen in the field of view of the telescope.
- c) To adjust the collimator for parallel rays, the slit of the collimator is illuminated with a monochromatic source of light (sodium vapour lamp). The telescope is brought in line with the collimator in order to catch the image of the slit in the telescope. The distance between the convex lens in the collimator and the slit is altered to get the clear image of the slit through the telescope, so that the width of the slit is minimized.
- d) The prism table is made perfectly horizontal by adjusting the screws provided with the help of the spirit level.

## **Procedure:**

### **Angle of the prism A:**

After making initial adjustments, the least count of the spectrometer is determined (as in Vernier Calipers).

Having a very fine slit, the telescope is adjusted for direct ray. A prism is placed on the prism table as shown in the figure. The telescope is turned to catch the reflected image from one polished face AB of the prism. Fixing in that position, the tangential screw is adjusted till the vertical cross wire is made to coincide with the fixed edge of the image of the slit. The readings in the scales I and II are noted.

The telescope is then turned to the other polished face AC of the prism. The readings in the scales I and II are noted when the vertical cross wire coincides with the fixed edge of the image of the slit. The difference between both the scale I readings is  $2A$ . Similarly the difference between both the scale II readings is  $2A$ . Then the average of the angle of the prism A is calculated.

### **Angle of minimum deviation D:**

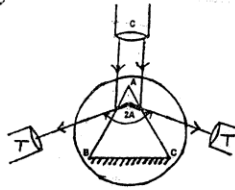
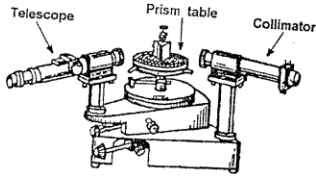
The prism is placed on the prism table as shown in the figure. The ray of light from the collimator incident on one polished face of the prism gets refracted and emerges out of the other polished face. Viewing the image of the slit through the telescope, the prism table and telescope are rotated simultaneously through small angles such that the image of the slit moves towards the direct ray. At one stage, it is found that the image of the slit is stationary for a moment and on rotating the prism table further in the same direction the image begins to retrace its path. The telescope is fixed at that position, the cross wire is made to coincide with the fixed edge of the image of the slit. At this minimum deviation position, the readings of scales I and II are noted.

The prism is removed from the prism table. The telescope is brought in line with the collimator and the image of the slit is seen through the telescope. The vertical cross wire is made to coincide with the fixed edge of the image of the slit. The readings in the scales I and II are noted.

The difference between the two readings of scale I or scale II gives the angle of minimum deviation D. Hence the refractive index of the material of the prism is calculated using the formula.

**Result:**

The refractive index of the material of the prism  $\mu =$



**Observations:**

Least count of Spectrometer:

Value of 20 MSD

=

Value of 1 MSD

=

No. divisions on the Verniers scale (n)

=

=

Least Count (LC)

$= \frac{1}{n} \times \text{value of 1 MSD}$

=

**Tabular column:**

To determine the angle of the prism A

LC =

Ray	Spectrometer readings in degree							
	Scale I				Scale II			
	MSR	VSC	VSR = VSC $\times$ LC	OR = MSR + VSR	MSR	VSC	VSR = VSC $\times$ LC	OR = MSR + VSR
Reading of the reflected ray (left)				$X_1 =$				$X_3 =$
Reading of the reflected ray (right)				$X_2 =$				$X_4 =$

$$2A = (x_1 - x_2) =$$

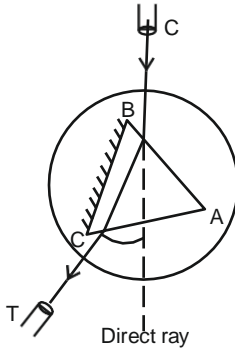
$$A = \frac{\quad}{2} =$$

The average angle of the prism A

$$2A = (x_3 - x_4) =$$

$$A = \frac{\quad}{2} =$$

$$= \frac{\quad + \quad}{2}$$



**Tabular column II:**

To determine the angle of the minimum deviation D

LC =

Ray	Spectrometer readings in degree							
	Scale I				Scale II			
	MSR	VSC	VSR = VSC × LC	OR = MSR + VSR	MSR	VSC	VSR = VSC × LC	OR = MSR + VSR
Reading of minimum deviated ray				$d_1 =$				$d_3 =$
Reading of Direct ray				$d_2 =$				$d_4 =$

The angle of the minimum deviation of the prism  $D = \frac{D_1 + D_2}{2}$

$$D_1 = (d_1 - d_2) =$$

$$D_2 = (d_3 - d_4) =$$

=

**Calculation:**

$$\mu = \frac{\sin\left[\frac{A+D}{2}\right]}{\sin\left[\frac{A}{2}\right]} =$$

The refractive index of the material of the glass prism

## OHM'S LAW

### Verification of Laws of Resistances

**Experiment No.:**

**Date:**

**Aim:**

To verify the laws of resistances when the two given standard resistances are connected i) in series and ii) in parallel, using Ohm's law.

**Apparatus required:**

Two known standard resistances, battery, key, rheostat, ammeter, voltmeter, connecting wires, etc.

**Formulae:**

i) When two resistances  $R_1$  &  $R_2$  are connected in series, their effective resistance is given by the formula  $R_s = R_1 + R_2$

$$R_s = \frac{E}{I}$$

The formula for determining the effective resistance by experiment is

ii) When two resistances  $R_1$  &  $R_2$  are connected in parallel, their

$$R_p = \frac{R_1 \times R_2}{R_1 + R_2}$$

effective resistance is given by the formula

$$R_p = \frac{E}{I}$$

The formula for determining the effective resistance by experiment is where

$R_1$  = the value of the one standard resistance

$R_2$  = the value of the other standard resistance

$E$  = Potential difference between the ends of the resistances  
(Voltmeter reading)

$I$  = Current flowing through the resistances (Ammeter reading)

### **Procedure:**

#### **Resistances in series:**

The given two resistances  $R_1$  and  $R_2$  are connected in series. This combination is again connected in series with a battery eliminator, key, rheostat and an ammeter. The voltmeter is connected in parallel to the combination of resistances to measure the potential difference across the resistances.

Now the circuit is closed. By adjusting the rheostat for different values of current, the corresponding values of potential difference are noted in tabular column. The effective resistance  $R_s$  is calculated in each case using the formula. Then the average value of effective resistance  $R_s$  is calculated.

#### **Resistances in parallel:**

The given two resistances  $R_1$  and  $R_2$  are connected in parallel. This combination is again connected in series with a battery eliminator, key, rheostat and an ammeter. The voltmeter is connected in parallel to the combination of resistances, to measure the potential difference across

the resistances.

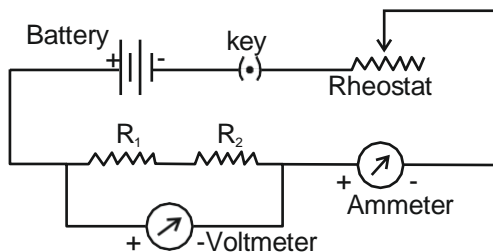
Now the circuit is closed. By adjusting the rheostat for different values of current, the corresponding values of potential difference are noted in tabular column II. The equivalent resistance  $R_p$  is calculated in each case using the formula. Then the average value of effective resistance  $R_p$  is calculated.

**Result:**

- 1) When the resistances are reconnected in series:
  - i. The effective resistance by calculation  $R_s =$  ohm
  - ii. The effective resistance by experiment  $R_s =$  ohm
- 2) When the resistances are reconnected in parallel:
  - i. The effective resistance by calculation  $R_p =$  ohm
  - ii. The effective resistance by experiment  $R_p =$  ohm

The laws of resistances are verified, using Ohm's law.

**Circuit diagram for series connection**



Resistance  $R_1 =$  ohm; Resistance  $R_2 =$  ohm

Effective resistance  $R_s = R_1 + R_2$   
 $=$  ohm  $=$  ohm

**Tabularcolumn1:**

Sl. No	Current flowing through the resistances (Ammeter reading) I		Potential difference between the ends of resistances (Voltmeter reading) E		$R_s = \frac{E}{I}$	
	Unit	div	ampere	div		volt
1.						
2.						
3.						
4.						
5.						

To find the value of effective resistance when  $R_1$  and  $R_2$  are in series

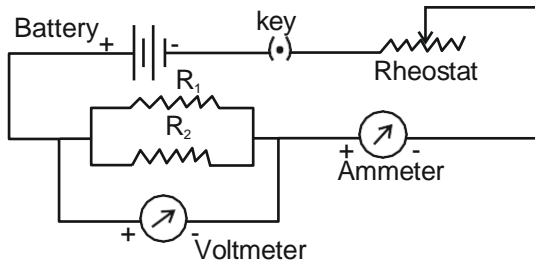
Average value of effective resistance when  $R_1$  and  $R_2$

are connected in series  $R_s =$  ohm

1)  $E =$  volt;  $I =$  ampere;  $R_s = \frac{E}{I}$

**Calculation:**

2)



**Circuit diagram for parallel connection**

Resistance  $R_1 =$  ohm;  $R_1 \times R_2 =$  Resistance  $R_2 =$  ohm

$$R_p = \frac{R_1 \times R_2}{R_1 + R_2}$$

Effective resistance



$$= \text{ ohm}$$

**Tabular column II:**

Sl. No	Current flowing through the resistances (Ammeter reading) I		Potential difference between the ends of resistances (Voltmeter reading) E		$R_s = \frac{E}{I}$ ohm
	div	ampere	div	volt	
1.					
2.					
3.					
4.					
5.					

To find the value of effective resistance when  $R_1$  and  $R_2$  are in parallel

Average value of effective resistance when  $R_1$  and  $R_2$

$$\text{are connected in parallel } R_p = \text{ ohm}$$

$$1) E = \text{ volt}; \quad I = \text{ ampere}; \quad R_p = \frac{E}{I} = \text{ ohm}$$

**Calculation:**

2)

## METRE BRIDGE

**Experiment No.:**

**Date:**

**Aim:**

To determine the unknown resistance of the given wire.

**Apparatus required:**

Metre bridge, given wire, variable resistance box, Leclanche cell,

plugkey, galvanometer, high resistance, jockey, connecting wires, etc.

$$X = \frac{X_1 + X_2}{2}$$

**Formulae:**

$$x = R \frac{l_1}{l_2} \text{ when the resistance box } R \text{ is in the right gap } (G_2)$$

$$x = R \frac{l_4}{l_3} \text{ when the resistance box } R \text{ is in the left gap } (G_1)$$

The resistance of the given wire is where

R = value of resistance introduced in the resistance box

$l_1, l_2, l_3$  and  $l_4$  are balancing lengths.

**Description:**

A metre bridge consists of a uniform wire AC of length one metre. It is stretched on a wooden board. The ends of the wire are fixed to two thick L shaped copper strips at A and C. Binding screws are provided at the ends of the strips. A third copper strip is fixed on the wooden board between the first two strips so as to form two gaps  $G_1$  and  $G_2$ . This strip is provided with three binding screws. A metre scale is fixed on the wooden board parallel to the length of the wire to measure the balancing lengths. A jockey J is provided by means of which contact can be made at any point on the wire.

**Procedure:**

The unknown resistance X (given wire) which is to be determined is connected in the gap  $G_1$  and a resistance box R is connected in the gap  $G_2$ .

A Leclanche cell and a key K are connected between the points A and C. A galvanometer, high resistance and the jockey are connected as shown in figure.

As a variable resistance (say  $2.0 \Omega$ ) is introduced in the resistance box R. Now the key is closed and the jockey is pressed at a point near the end A of metre bridge wire and then at a point near the end C. If the deflections in the galvanometer are in opposite directions, then the connections are correct.

The jockey is gently slid from A end on the bridgewire AC till the galvanometer shows null deflection and the balancing point J is found. The balancing length AJ is measured as  $l_1$  and the remaining length JC is noted as  $l_2$ , i.e.  $l_2 = (100 - l_1)$ . The experiment is repeated by increasing the values introduced in the resistance box in steps of  $0.2\Omega$ . The readings are tabulated. The resistance of given coil of the wire is calculated using the given formula. The average value of the  $X_1$  is found.

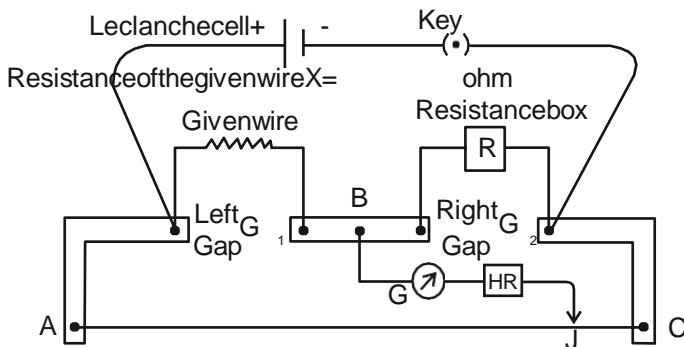
The given wire X and resistance box R are interchanged i.e. the unknown resistance X (given wire) is connected in the gap  $G_2$  and the resistance box R is connected in the gap  $G_1$ .

A suitable value of resistance (say  $2\Omega$ ) is introduced in the resistance box R. Now the key is closed and the jockey is pressed at a point near the end A of metre bridgewire and then at a point near the end C. If the deflections in the galvanometer are in opposite directions, the connections are correct.

The jockey is gently slid from A end on the bridgewire AC till the galvanometer shows null deflection and the balancing point J is found. The balancing length AJ is measured as  $l_3$  and the remaining length JC is noted as  $l_4$ , i.e.  $l_4 = (100 - l_3)$ . The experiment is repeated by increasing the values introduced in the resistance box in steps of  $0.2\Omega$ . The readings are tabulated. The resistance of given wire is calculated using the given formula. The average value of  $X_2$  is found.

Using the values of  $X_1$ ,  $resis$  and  $X_2$ , the average value of unknown tance is determined.

**Result:**



**Tabularcolumn1:**

Sl. No.	Known Resistance R	Balancing lengths				$x_1 = R \frac{l_1}{l_2}$
		$l_1$		$l_2 = 100 - l_1$	$l_2 = 1.0 - l_1$	
Unit	ohm	cm	m	cm	m	ohm
1.						
2.						
3.						
4.						
5.						

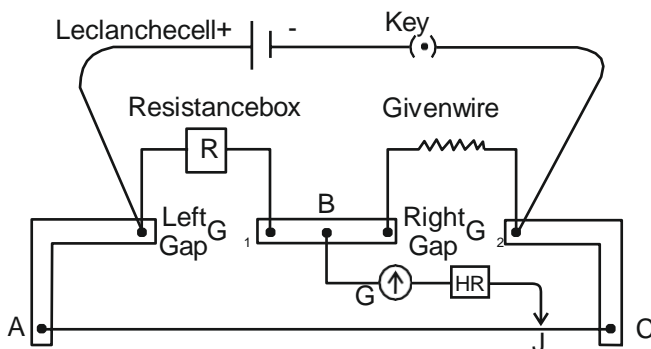
To find the unknown value of resistance  $X_1$  when  $R$  is in the gap  $G_2$ . The average value of  $X_1 =$                   ohm

**Calculation:**

$$x_1 = R \frac{l_1}{l_2}$$

1)  $R =$           ohm;  $l_1 =$           m;  $l_2 =$           m

$$x_1 = R \frac{l_1}{l_2}$$



2)  $R = \quad \text{ohm}; l_1 = \quad \text{m}; l_2 = \quad \text{m}$

**Tabularcolumn1:**

Sl. No.	Known Resistance R	Balancing lengths				$x_2 = R \frac{l_4}{l_3}$
		$l_3$		$l_4 = 100 - l_3$	$l_4 = 1.0 - l_3$	
Unit	ohm	cm	m	cm	m	ohm
1.						
2.						
3.						
4.						
5.						

To find the unknown value of resistance X, when R is in the gap G<sub>2</sub>

The average value of  $X_2 = \quad \text{ohm}$

**Calculation:**

$$x_2 = R \frac{l_4}{l_3}$$

1)  $R = \quad \text{ohm}; \quad l_3 = \quad \text{m}; \quad l_4 = \quad \text{m}$

$$x_2 = R \frac{l_4}{l_3}$$

Average resistance of the wire  $x = \frac{x_1 + x_2}{2}$

2)  $R = \quad \text{ohm}; \quad l_3 = \quad \text{m}; \quad l_4 = \quad \text{m}$   
 $= \quad \text{ohm}$

## POTENTIOMETER

**Experiment No.:**

**Date:**

**Aim:**

To compare the electromotive forces (e.m.f.s) of the given two cells, using a potentiometer.

**Apparatus required:**

Potentiometer, given two cells (Leclanche cell and Daniel cell), battery eliminator, rheostat, plug key, two way key, galvanometer, high resistance, jockey, connecting wires, etc.

**Formula:**

$$\frac{E_1}{E_2} = \frac{l_1}{l_2}$$

The ratio of e.m.f.s of two cells

where

$E_1$  = e.m.f. of first cell (Leclanche cell)

$E_2$  = e.m.f. of second cell (Daniel cell)

$l_1$  = balancing length when e.m.f. of Leclanche cell is projected in the primary circuit

$l_2$  = balancing length when e.m.f. of Daniel cell is projected in the primary circuit

**Description:**

The potentiometer consists of 4m, 6m or 10m long wire of uniform resistance kept stretched on a wooden board between two terminals. To make the apparatus compact the wire is arranged on the board in zigzag manner in segments, each of one metre in length. A jockey is also provided to make contact at any point on the wire. A metre scale is fixed parallel to the length of the wire using which readings can be taken directly.

The battery  $B_a$ , key  $K$ , rheostat  $R$  and the potentiometer wire  $AB$

are connected in series. This is called primary circuit.

$E_1$  and  $E_2$  are the e.m.f.s of the Leclanche cell and Daniel cell respectively which are to be compared. The positive terminals of Leclanche cell and Daniel cell are connected to the terminal A of potentiometer wire. The negative terminals are connected to the two outer terminals of two way key. Then the middle terminal of two way key is connected with jockey through galvanometer and high resistance. This is called secondary circuit.

**Procedure:**

The primary circuit is closed and the rheostat in the primary circuit is adjusted to send a suitable current. The top gap of the two way key is closed to have Leclanche cell of e.m.f.  $E_1$  included in the circuit. Then the jockey is pressed at the ends A and B of the potentiometer wire. If the deflections in the galvanometer are in opposite directions, then the connections are correct. Similarly the Daniel cell of e.m.f.  $E_2$  is included in the circuit and the circuit connections are checked.

The rheostat's movable slide is kept at one end. Now, the top gap of the two way key is closed to have Leclanche cell of e.m.f.  $E_1$  included in the circuit. The jockey is gently slid from A on the wire and the balancing point  $J_1$ , at which there is no deflection in the galvanometer, is obtained accurately. The balancing length  $AJ_1$  is measured as  $l_1$ .

Again, the bottom gap of the two way key is closed to have Daniel cell of e.m.f.  $E_2$  included in the circuit. The jockey is gently slid from A end of the wire and the balancing point  $J_2$  is obtained accurately. The balancing length  $AJ_2$  is measured as  $l_2$ .

The experiment is repeated by adjusting movable slide of the rheostat for varying current in the primary circuit. In each case the balancing lengths  $l_1$  and  $l_2$  are measured and readings are tabulated in the tabular column. Then, the ratio of e.m.f. of two cells is calculated using the formula.

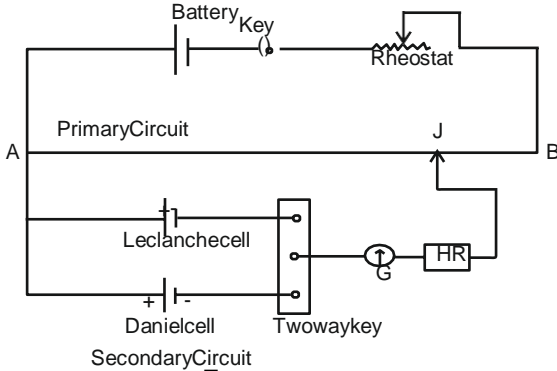
**Note:**

If value of e.m.f. of one of the two cells is known, the e.m.f. of the other cell can be calculated using the formula  $(E_1/E_2) = (l_1/l_2)$ . If  $E_1$  is known, the

e.m.f. of  $E_2$  can be calculated using  $E_2 = (E_1 X l_2) / l_1$

**Result:**

- i) The e.m.f. of given two cells are compared.
- ii) The ratio of e.m.f. of given two cells is found to be =



Tabular column: To find  $\frac{E_1}{E_2} =$

Sl. No.	Balancing lengths				$\frac{E_1 \cdot l_1}{E_2 \cdot l_2}$
	For Leclanche cell (of e.m.f. $E_1$ ) $l_1$		For Daniel cell (of e.m.f. $E_2$ ) $l_2$		
Unit	cm	m	cm	m	
1					
2					
3					
4					
5					
6					
7					

The average of  $\frac{E_1}{E_2} =$



**Calculation:**

1.  $I_1 =$  m;  $I_2 =$  m.

$$\frac{E_1 - I_1}{E_2 \quad I_2} =$$

2.  $I_1 =$  m;  $I_2 =$  m.

$$\frac{E_1 - I_1}{E_2 \quad I_2} =$$

## JOULE'S CALORIMETER

**Experiment No.:**

**Date:**

**Aim:**

To determine the specific heat capacity of water (liquid), using Joule's calorimeter.

**Apparatus required:**

Joule's calorimeter, battery eliminator, ammeter, voltmeter, rheostat, plug key, stop watch, thermometer, digital balance, liquid (water), connecting wires, etc.

**Formula:**

$$S = \frac{EIt}{(m_2 - m_1)(T_2 - T_1)} - \frac{m_1 S_c}{(m_2 - m_1)}$$

The specific heat capacity of the liquid

where

E = potential difference across the heating coil (Voltmeter reading)

I = current flowing through the heating coil (Ammeter reading)

t = time for which the current flows to increase the temperature of water

$m_1$  = mass of empty and dry calorimeter with stirrer

$m_2$  = mass of calorimeter with stirrer and water

$S_c$  = specific heat capacity of the material of the calorimeter

$T_1$  = initial temperature of the liquid

and  $T_2$  = final temperature of the liquid

### **Description:**

The Joule's calorimeter consists of a brightly polished copper or aluminium vessel closed by an ebonite lid with two holes. A heating coil, connected to two binding screws provided on the ebonite lid is kept within the calorimeter. Through holes in the lid a sensitive thermometer and a stirrer can be inserted. The calorimeter is placed inside a wooden box stuffed with cotton or wool.

### **Procedure:**

The mass of the clean, dry calorimeter with stirrer ( $m_1$ ) is determined using a digital balance. The calorimeter is filled to two-third of its volume with the given liquid (water) and the mass ( $m_2$ ) is once again determined.

The calorimeter is kept in the wooden box and the heating coil is completely immersed in the liquid. Connections are given as shown in the circuit diagram.

The circuit is closed for a moment and the rheostat is adjusted for a suitable current of 0.8 A or 1.0 A. The ammeter and voltmeter readings are noted as  $I$  and  $E$  respectively. The circuit is kept open. Now, the liquid is stirred well and the initial temperature ( $T_1$ ) of the liquid is determined using a sensitive thermometer.

The circuit is again closed and simultaneously a stopwatch is started. The liquid is stirred slowly and uniformly. The temperature of the liquid is noted for every minute in the tabular column. When the temperature of the liquid is increased by  $5^\circ\text{C}$ , the key is removed and the stopwatch is stopped simultaneously.

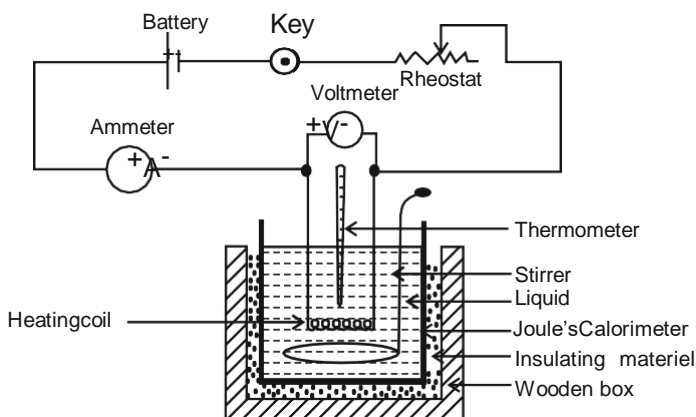
The time ( $t$  second) for which the current was flowing through the circuit is noted from the stopwatch. The specific heat capacity of the material of the calorimeter ( $S_c$ ) is taken from tables.

Then, the specific heat capacity of the water ( $S$ ) is calculated using the given formula.

**Result:**

The specific heat capacity of water is found to be  $S =$

$\text{J kg}^{-1} \text{K}^{-1}$



**Observation:**

Mass of the empty calorimeter with stirrer	$(m_1) =$	g
	$=$	kg
Mass of calorimeter with stirrer + water	$(m_2) =$	g
	$=$	kg
Current flowing through the heating coil (Ammeter reading)	$(I) =$	A
Potential difference across the coil (Voltmeter reading)	$(E) =$	V
Initial temperature of water	$(T_1) =$	$^{\circ}\text{C}$
Final temperature of water	$(T_2) =$	$^{\circ}\text{C}$
Time for which current is passed to increase the temperature of water	$(t) =$	minutes
	$=$	second
Specific heat capacity of the material of the calorimeter (copper)	$(S_c) =$	$\text{J kg}^{-1} \text{K}^{-1}$

**Tabular column:**

To record the time taken for 5°C raise in temperature of water

Time in minutes	Temp. of water in °C

Time taken for 5° craise in temp. of water (t) =                      s

**Calculation:**

The specific heat capacity of water

$$S = \frac{Elt}{(m_2 - m_1)(T_2 - T_1)} - \frac{m_1 S_c}{(m_2 - m_1)}$$

$$= \quad \quad \quad \text{J kg}^{-1} \text{K}^{-1}$$

## COPPER VOLTAMETER

**Experiment No.:**

**Date:**

**Aim:**

To determine the electrochemical equivalent (e.c.e) of copper, using copper voltameter.

**Apparatus required:**

Copper voltameter, battery eliminator, plug key, rheostat, ammeter, stopwatch, digital balance, connecting wires, emery sheet, etc.

**Formula:**

The electrochemical equivalent of copper

$$Z = \frac{m}{It} = \frac{m_2 - m_1}{It}$$

where

$m$  = mass of copper deposited on the cathode

$m_1$  = mass of the cathode before passing current

$m_2$  = mass of the cathode after passing current

$I$  = current flowing through the electrolyte (Ammeter reading)

$t$  = time of passage of current

### **Description:**

The copper voltameter consists of a glass vessel, containing a strong solution of copper sulphate. Three plates of copper are dipped in the solution. The two outer plates are connected together and they form the anode. The middle plate serves as the cathode.

### **Procedure:**

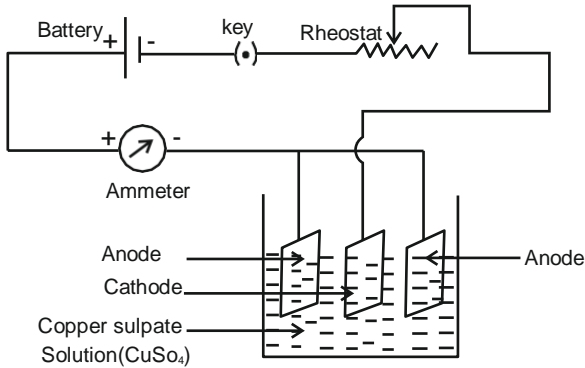
The copper voltameter is connected in series with the battery, plug key, rheostat and ammeter. It should be noted that the negative of the battery is connected to the central cathode plate. After making all the connections, the circuit is closed for a moment and the rheostat is adjusted for a suitable current, say 1 ampere, in the ammeter. The circuit is kept open, the cathode plate is removed and cleaned with emery sheet. It is then washed in water, dried and its mass  $m_1$  is determined accurately using a digital balance.

The cathode plate is then fixed in its position and the circuit is closed, simultaneously starting a stop watch. The current is passed for 30 minutes and during the passage of current the ammeter value should be maintained as the same.

After 30 minutes the cathode is taken out, rinsed with water, dried and its mass  $m_2$  is found accurately. The time of passage of current is noted. The mass of copper deposited is determined and hence e.c.e. of copper is calculated using the given formula.

### **Result:**

The electrochemical equivalent of copper  $Z =$   $\quad\quad\quad$   $\text{kgC}^{-1}$



**Observation:**

Mass of the cathode before passing current $m_1$	Ma	=	
			g
			kg
Mass of the cathode after passing current $m_2$		=	g
			kg
Mass of the copper deposited at cathode $m$		=	$(m_2 - m_1)$
			kg
Current flowing through the electrolyte (Ammeter reading)	}		I = A
Time of passage of current		t =	minutes
		=	second

**Recording of time of passage of current (in minutes):**

1	2	3	4	5	6	7	8	9	10	11	12	13	14	15
16	17	18	19	20	21	22	23	24	25	26	27	28	29	30

**Calculation:**

The electrochemical equivalent of copper

$$Z = \frac{m}{It} = \frac{m_2 - m_1}{It}$$

$$=$$

$$= \text{kgC}^{-1}$$

# CHARACTERISTICS OF P-N JUNCTION DIODE

**Experiment No.:**

**Date:**

**Aim:**

To draw the voltage-current characteristics of a P-N Junction diode in forward bias and to find the 'dynamic forward resistance' and 'knee voltage' from the graph

**Apparatus required:**

P-N junction diode (Germanium or Silicon), D.C. power supply (0 – 10 V), D.C. milliammeter (0 – 15 mA), D.C. Voltmeter (0 – 1.5 V), resistance (500), etc.

**Formula:**

The dynamic forward resistance  $r = \frac{\Delta V}{\Delta i}$

$\Delta V$  = a small change in potential difference (p.d) applied to the diode

$\Delta i$  = the corresponding change in the current flowing through the diode

Knee voltage ( $V_k$ ): It is the approximate voltage above which the diode current increases sharply.

**Theory:**

When the positive terminal of the battery is connected to P-type and negative terminal to N-type of a semiconductor diode, current flows easily through the diode. Now it is said to be in forward bias.

When the P-type material is connected to negative and N-type to positive of the battery, the diode offers a high impedance path and little current flows through the diode. Now it is said to be in reverse bias.

This property of P-N Junction diode, allowing the flow of current in one direction easily is made use of in many electronic circuits, especially in solid state half-wave and full-wave rectifiers.

### Procedure:

The anode and cathode of a P-N junction diode can be identified by seeing a colour band or spot on the cathode. Also the cathode may be having pointed edge in some type.

For forward biasing, the connections are given neatly as shown in the diagram. Suitable ranges of meters, resistance and power supply can be selected.

The current flowing through the diode is noted for various values of p.d. applied across the diode and the readings are noted in the tabular column. With these readings, a neat graph is drawn by taking forward voltage applied across the diode in X-axis and forward current flowing through the diode (in mA) along Y-axis.

This curve is called as the forward characteristics of the diode. From this graph, the approximate voltage above which the diode current sharply increases, can be determined, which gives the "Knee Voltage" of the diode.

For germanium diode, the knee voltage is 0.3 volt and for silicon diode the knee voltage is 0.7 volt, which is almost equal to the barrier potential.

In the upper region of the curve, two perpendicular lines OA and AB are drawn to find the dynamic forward resistance, as shown in the model graph. In this OA gives the value of small change in p.d. applied across the diode ( $\Delta v$ ) and AB gives the corresponding change in the current flowing through the diode ( $\Delta i$ ). These values are determined.

Then the dynamic forward resistance of the diode is determined by using the given formula.

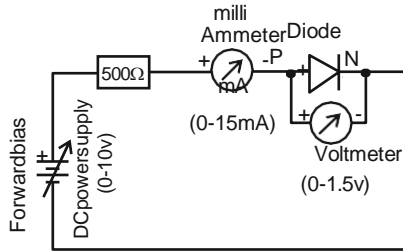
### Result:

The forward characteristics of a P-N junction diode is drawn.

From the graph,

- (i) The dynamic forward resistance =                  ohm
- (ii) The knee voltage                   $V_k =$                   volt





Circuit diagram for drawing the forward characteristics of P-N junction diode.

**Tabular column:**

Readings for drawing the forward characteristics of P-N junction diode

Sl. No.	P.d applied across the diode (v)		Current flowing through the diode (i)	
	div	volt	div	milliampere (mA)
1.				
2.				
3.				
4.				
5.				
6.				
7.				
8.				
9.				

**Calculation:**

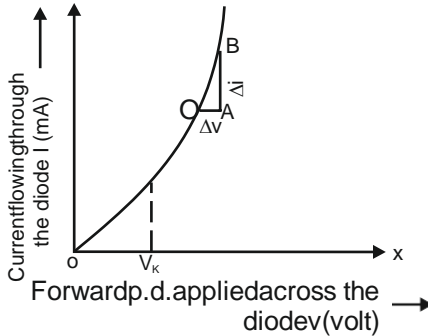
From graph,

$\Delta V = \quad \text{div}; \Delta i = \quad \text{div}$

For calculating  $\Delta V$  in volt

In X-axis 10 div =  $\quad$  volt; 1 div =  $\quad$  volt

$\therefore \Delta V = \quad \text{div} = \quad \times \quad = \quad \text{volt}$



For calculating  $\Delta i$  in mA

$$\begin{aligned} \text{In Y-axis } 10 \text{ div} &= \text{mA}; 1 \text{ div} = \text{mA} \\ \therefore \Delta i &= \text{div} = x = \text{mA} \\ &= \text{mA} \times 10^{-3} \text{ A} \end{aligned}$$

(i) The dynamic forward resistance is  $r_f = \frac{\Delta V}{\Delta i}$

(ii) The knee voltage is  $V_k = \text{div} = \text{ohm}$

$$\begin{aligned} &= x \\ &= \text{volt} \end{aligned}$$

## LOGIC GATES

**Experiment No.**

**Date:**

**Aim:**

To find the output condition for different combinations of the inputs for NOT gate and two inputs AND, OR, NAND and NOR logic gates, using IC chips.

**Apparatus required:**

Logic gate apparatus kit with chip mounting base, integrated circuit chips (IC chips) for the above logic gates.

**Description:**

The logic gate kit consists of required DC power supply, light emitting diodes (LEDs), to indicate the input and the output signal, toggle switches to provide various combinations of input and breadboard to mount the desired IC chip.

In the kit when the LED glows, the corresponding condition is "ON", and the binary is 1. If the LED does not glow the corresponding condition is "OFF", and the binary is 0.

**Procedure:****(a) NOT gate:**

It is the simplest form of the logic gates. It has only one input and one output. The function of this gate is to produce one logic level from the opposite logic level. i.e. If the input condition is "ON", the output condition will be "OFF" and if the input condition is "OFF", the output condition will be "ON".

The NOT gate is available in IC chip number 7404. The pin configuration for the power supply, ground, input and output of the IC chip is given in the figure and the Boolean expression is also given.

The 7404 IC is mounted on the base provided in the kit. Suitable input level is given and the corresponding output level is noted in the truth table of NOT gate.

The function of this gate is inverting or complementing the input signal. So, this gate is also called as a 'Complementer or Inverter gate'.

**(b) AND gate:**

This gate has two or more inputs and only one output. The function of this gate is to produce a true condition on its output, if and only if all its input conditions are true. i.e., the output of the gate is in "ON" condition if and only if all its inputs are in "ON" condition.

The two inputs AND gate is available in IC chip number 7408. The pin configuration for the power supply, ground, input and output of the IC chip is given in the figure and the Boolean expression is also given.

The 7408 IC is mounted on the base provided in the kit. Suitable combinations of input levels are given and the corresponding output levels are noted in the truth table of AND gate.

This gate is also called as a 'Multiplier gate'.

**(c) ORgate:**

This gate has two or more inputs and only one output. The function of this gate is to produce the true condition on its output, when one or more of its input conditions are true. i.e., The output of the gate is in "ON" condition if any one of its inputs is in "ON" condition.

The two inputs OR gate is available in IC chip number 7432. The pin configuration for the power supply, ground, input and output of the IC chip is given in the figure and the Boolean expression is also given.

The 7432 IC is mounted on the base provided in the kit. Suitable combinations of input levels are given and the corresponding output levels are noted in the truth table of OR gate.

This gate is also called as a 'Adder gate'.

These above three gates are called 'Basic Gates'.

**(d) NANDgate:**

It is a combination of AND and NOT gates. The output of NAND gate is complementation of that of an AND gate.

The two inputs NAND gate is available in IC chip number 7400. The pin configuration for the power supply, ground, input and output of the IC chip is given in the figure and the Boolean expression is also given.

The 7400 IC is mounted on the base provided in the kit. Suitable combinations of input levels are given and the corresponding output levels are noted in the truth table of NAND gate.

**e) NORgate:**

This gate is a combination of NOT and OR gates. The output of the NOR gate is complementation of that of an OR gate.

The two inputs NOR gate is available in IC chip number 7402. The pin configuration for the power supply, ground, input and output of the IC chip is given in the figure and the Boolean expression is also given.

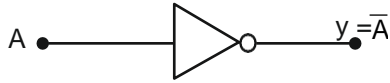
The 7402 IC is mounted on the base provided in the kit. Suitable combinations of input levels are given and the corresponding output levels are noted in the truth table of NOR gate.

These above two gates are known as 'Universal gates'.

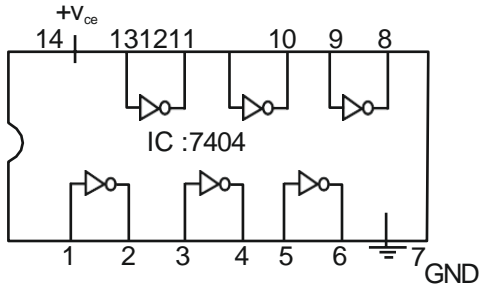
**Result:**

Thus the output condition for different combinations of the inputs for NOT gate and two inputs AND, OR, NAND and NOR logic gates are noted in the truth table.

**a) NOTgate:**



**Symbolic representation**



**NOTgateICchip**

IC No: 7404

Pinconfiguration:

Inputpins :1,3,5,9,11,13

Outputpins:2,4,6,8,10,12

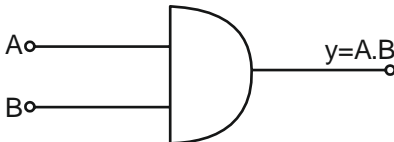
Groundpin :7

Powerpin(+V<sub>cc</sub>):14

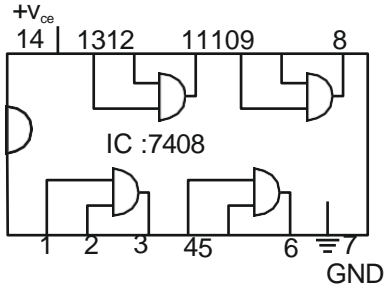
**TruthtableforNOTgate**

INPUT	OUTPUT
A	Y= A
0	
1	

**b) ANDgate:**



**Symbolic representation**



**TwoinputANDgateIC chip**

IC No: 7408

Pinconfiguration:

Inputpins : 1,2,4,5,9,10,12,13

Outputpins :3,6,8,11

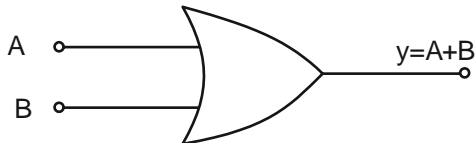
Groundpin 7

Powerpin( $+V_{cc}$ ):14

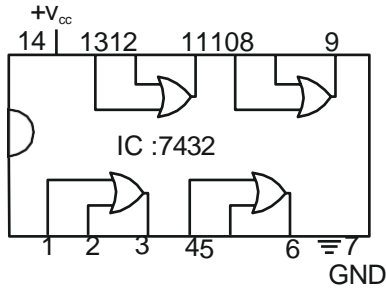
**TruthtableforANDgate**

INPUTS		OUTPUT $Y = A.B$
A	B	
0	0	
0	1	
1	0	
1	1	

**c) ORgate:**



**Symbolic representation**



**TwoinputORgateICchip**  
IC No: 7432

Pinconfiguration:

Inputpins:1,2,4,5,9,10,12,13

Outputpins:3,6,8,11

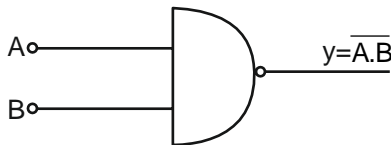
Groundpin:7

Powerpin(+V<sub>cc</sub>):14

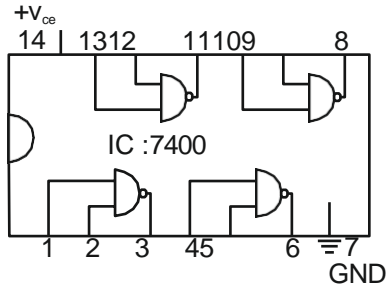
**TruthtableforORgate**

INPUTS		OUTPUT Y=A+B
A	B	
0	0	
0	1	
1	0	
1	1	

**d) NANDgate:**



**Symbolic representation**



### TwoinputNANDgateIC chip

IC No: 7400

Pinconfiguration:

Inputpins:1,2,4,5,9,10,12,13

Outputpins:3,6,8,11

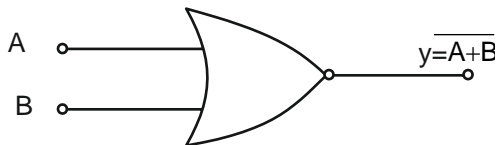
Groundpin:7

Powerpin(+V<sub>cc</sub>):14

### TruthtableforNANDgate

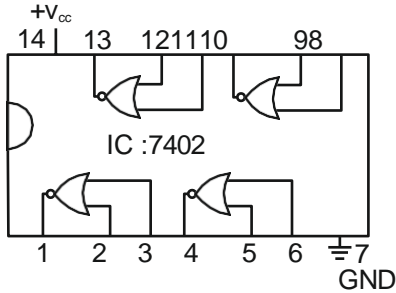
INPUTS		output $y = \overline{A \cdot B}$
A	B	
0	0	
0	1	
1	0	
1	1	

### e) NORgate:



Symbolic representation





**TwoinputNORgateICchip**

IC No: 7402

Pinconfiguration:

Inputpins:2,3,5,6,8,9,11,12

Outputpins:1,4,10,13

Groundpin:7

Powerpin(+V<sub>cc</sub>):14

**TruthtableforNORgate**

INPUTS		output
A	B	$y = \overline{A+B}$
0	0	
0	1	
1	0	
1	1	

## MODEL QUESTION PAPER

1. Determine the refractive index of the given transparent liquid using traveling microscope.
2. Draw the V-I characteristics of the solar cell.
3. Measure the angle of the prism and the angle of minimum deviation using spectrometer and then calculate the refractive index of glass.
4. Verify the laws of resistances by connecting the two given standard resistances (i) in series and (ii) in parallel, using Ohm's law.
5. Determine the unknown resistance of the given wire, using metre bridge.
6. Compare the electro motive forces (e.m.f.s) of the given two cells, using potentiometer.
7. Determine the specific heat capacity of water, using Joule's calorimeter.
8. Determine the electro chemical equivalent (e.c.e.) of copper using Copper Voltmeter.
9. Draw the voltage-current characteristics of a P-N junction diode in forward bias and then find the 'dynamic forward resistance' & 'knee voltage' from the graph.
10. Find the output conditions for different combinations of the input for NOT gate (7404) and 2 inputs AND, (7408) OR, (7432) NAND (7400) & NOR (7402) logic gates using IC chips.