INSERVICE TRAINING PROGRAMME IN PHYSICS
FOR THE PGTs of NVS

2.6.2003 to 22.6.2003

REPORT

Coordinator

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REGIONAL INSTITUTE OF EDUCATION, MYSORE-6

(NATIONAL COUNCIL OF EDUCATIONAL RESEARCH AND TRAINING)

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PREFACE

The 21-day training programme in Physics for the PGTs of Navodaya Vidyalaya Samiti, New Delhi was held in the Physics Section (DESM) of Regional Institute of Education, Mysore from 2^{nd} to 22^{nd} June 2003.

The programme was arranged at the request of NVS, New Delhi. The main objective of the programme was to enrich the content level of the teachers as per the revised curriculum.

The present volume contains a detailed report as well as the enrichment material provided to the teachers. Every effort has been made to make the material as explanatory as possible so that the teachers could find it useful in the classroom transactions.

The enrichment material on Communication Technology has been provided to the participants in the form of a CD.

I am glad to place on record the enthusiasm shown by the participating teachers in all the sessions and we thank them for the same.

My thanks are due to the authorities of Navodaya Vidyalya Samiti, New Delhi for having provided the funds for the programme and also for deputing the teachers for the course.

I am also grateful to Prof J S Rajput, Director, NCERT for having selected this Institute as one of the venues for this programme.

I express my heartfelt thanks to Prof G Ravindra, Principal, RIE, Mysore for extending cooperation for the conduct of the programme.

I wish to sincerely thank all the resource persons and guest lecturers who have actually contributed and shared their experiences with the participants.

My thanks are due to my colleagues in the Physics Department for their active participation and valuable guidance during the planning and implementation of the programme.

I wish to place on record the cooperation extended by my colleagues in other sections and departments during the conduct of the programme.

Lastly, I express my thanks to the administrative staff of the Institute and to the laboratory staff of the Physics Section who have spared no efforts in making the programme a grand success.

(R Narayanan)
Academic Coordinator

ABOUT THE TRAINING PROGRAMME

Knowledge in the present day world is developing at a fast pace. There is a tremendous growth of knowledge in the field of Science and Technology.

The present day teachers have to keep themselves up-to-date with the expanding knowledge. The inservice teachers need periodic refresher courses to fulfill this objective. In order to improve the capabilities of the teachers in content and pedagogy, the Navodaya Vidyalaya Samiti arranges inservice training of teachers at various levels in the form of orientation programmes and refresher courses.

The present programme was oriented toward the Post Graduate Teachers (PGTs) in Physics. It was held at the Regional Institute of Education, Mysore from 2nd to 22nd June 2003 (21 days). The programme was planned and implemented by the Physics Section of the Department of Education in Science and Mathematics of the Institute. In addition to the Physics faculty, faculty members from the Department of Education also worked as resource persons. Guest lectures and popular talks were also arranged using the expertise of external resource persons of eminence.

The main objectives of the training programme was to

i) enrich the content competency of the teachers so that they can execute the revised curriculum with greater confidence,

ii) provide a first hand experience in setting up, performing and interpreting the results of certain laboratory experiments and projects,

iii) make the teachers aware of recent thrust areas in the field of education so as to improve their professional competence, and

iv) make them familiar with certain skills and strategies required for effective teaching in the present day classrooms.

The programme consisted of two lecture sessions (1½ hours each) per day in the morning and a laboratory session (2 hours) in the afternoon followed by discussion/seminar (1 hour).

The laboratory and discussion session was attended by all the Physics faculty.

The topics on which lecture sessions were conducted were decided after an interactive session with the participants. Broadly the topics covered were from the following areas :

Mechanics
Waves and Oscillations
Electromagnetism
Current electricity
Electronics
Digital Electronics
Communication Systems
Solid State Physics
Nuclear Physics

The level of discussion was kept higher than the requirement at the plus two stage. During the laboratory sessions, the participants were encouraged to set up the experiments, suggest innovative investigatory projects and implement them wherever possible. The laboratory session was followed by presentation and discussion.

As a prelude to the laboratory sessions, two interactive discussions were presented on: (i) Problem solving approach to teaching of Physics and ii) Errors and significant figures.

The participants were able to get "hands-on" experience in computers and the use of multimedia. In addition to the content coverage lecture/discussion sessions were provided in the following professional areas.

a) Value Education
b) Evaluation and objective based Assessment
c) Creativity in teaching and learning
d) Needs and Problems of adolescents
e) Managerial skills for teachers
f) Skills of improving pupil participation
g) Motivation and positive attitude

In all, a variety of experiences were provided to the participants in order to enhance content enrichment and professional competence. It is hoped that the programme has sufficiently motivated the teachers.

(R Narayanan)
Academic Coordinator

LIST OF RESOURCE PERSONS

I. CONTENT AREA

RIE Faculty

Dr S S Raghavan
Professor in Physics

Dr P R Lalitha
Reader in Physics

Dr R Narayanan
Reader in Physics

Dr M N Bapat
Reader in Physics

Mr N R Nagaraja Rao
Sr. Lecturer in Physics

Guest Lecturers

Mr P Ramachandra Rao Reader in Physics (Retd) RIE, Mysore	**Electronics** **Error and Significant Figures** **Problem Solving Approach to Teaching of Physics**
Mr M A Chandrashekhar Reader in Physics (Retd) RIE, Bhubaneswar	**Digital Electronics**
Dr C R Nataraj Assistant Professor Electronics and Communications, SJCE, Mysore	**Communication Systems**

II. THRUST AREAS

Dr Kalpana Vengopal Lecturer in Education	**Creativity in Teaching and Learning** **Needs and Problems of Adolescents**
Dr N N Prahallada Reader in Education	**Value Education**
Dr V D Bhat Reader in Education	**Skills of increasing pupil participation** **Motivation and Positive Attitude**

Guest Lecturers

Dr C Gurumurthy Reader in Education Ramakrishna Institute of Moral and Spiritual Education Mysore	**Evaluation and Objective Based Assessment**
Prof K Shamanna Managerial Skills and Training Consultant and Fellow, Indian Society for Training and Management	**Managerial Skills for Teaching**
Prof G T Narayana Rao Editor, Kannada Encyclopaedia University of Mysore Mysore	**Evolution of Stars**

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CHAPTER 1
MOTION IN A PLANE

As we shall see la he motion of a particle subjected to a central force field is confined to a plane. Classic exampl e the motions of planets and satellites in the gravitational field of the Sun/Earth and the motion lectrons in an atom under the coulomb force of the (charged) nucleus.

To study motion i lane, it is convenient to employ plane polar coordinates (r, θ) defined by $x = r\cos\theta$; $y = r \sin$ (ig.1).

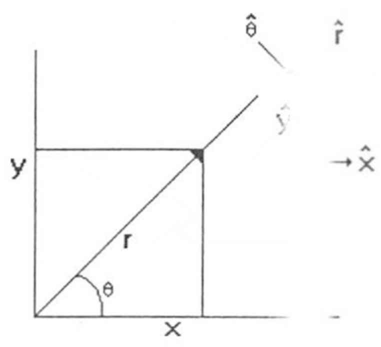


Fig.1: Plane Polar Coordinat

We define unit vectors $\hat{r}$ and $\hat{\theta}$ in the direction of increasing r and θ respectively. $\hat{r}$ and $\hat{\theta}$ are functions of θ and are related to $\hat{x}$ and $\hat{y}$ by

$$\hat{r} = \hat{x}\cos\theta + \hat{y}\sin\theta$$
$$\hat{\theta} = -\hat{x}\sin\theta + \hat{y}\cos\theta \qquad (1)$$

Taking derivatives with respect to θ,

$$\frac{d\hat{r}}{d\theta} = -\hat{x}\sin\theta + \hat{y}\cos\theta = \hat{\theta}$$

$$\frac{d\hat{\theta}}{d\theta} = -\hat{x}\cos\theta - \hat{y}\sin\theta = -\hat{r} \qquad (2)$$

The position vect r is given in terms of polar coordinates by

$$\mathbf{r} = r\,\hat{r}\,(\theta) \qquad (3)$$

We may describe t motion of a particle in polar coordinates by specifying r(t) and θ(t), thus determining the position v tor **r**(t).

The velocity vector

$$\mathbf{v} = \frac{d\,\mathbf{r}}{dt} = \frac{d}{dt}(r\,\hat{r})$$

$$= \frac{dr}{dt}\hat{r} + r\frac{d\hat{r}}{d\theta}\frac{d\theta}{dt}$$

$$= \dot{r}\hat{r} + r\dot{\theta}\hat{\theta} \qquad (4)$$

(using equation 2)

This means that the *radial* and *transverse* components of the velocity vector are:

$$\upsilon_r = \dot{r}, \qquad \upsilon_\theta = r\dot{\theta} \qquad (5)$$

The acceleration vector

$$\mathbf{a} = \frac{d\mathbf{v}}{dt} = \ddot{r}\hat{r} + \dot{r}\frac{d\hat{r}}{d\theta}\frac{d\theta}{dt} + \dot{r}\dot{\theta}\hat{\theta} + r\ddot{\theta}\hat{\theta} + r\dot{\theta}\frac{d\hat{\theta}}{d\theta}\frac{d\theta}{dt}$$

$$= (\ddot{r} - r\dot{\theta}^2)\hat{r} + (r\ddot{\theta} + 2\dot{r}\dot{\theta})\hat{\theta} \qquad (6)$$

(using 2)

Thus, the radial and transverse components of the *acceleration* vector are:

$$a_r = \ddot{r} - r\dot{\theta}^2$$

$$a_\theta = r\ddot{\theta} + 2\dot{r}\dot{\theta} \qquad (7)$$

If r is a constant, the motion is *circular* and $a_r = r\dot{\theta}^2 = -\upsilon_\theta^2 / r$ which is the familiar *centripetal* acceleration.

Momentum and Energy

The *linear momentum* of a particle is defined as $\mathbf{p} = m\mathbf{v}$ (8)

Newton's second law in vector form is

$$\frac{d\mathbf{p}}{dt} = \mathbf{F} \tag{9}$$

The change in momentum between times t_1 and t_2 is

$$\mathbf{p}_2 - \mathbf{p}_1 = \int_{t_1}^{t_2} \mathbf{F}dt \tag{10}$$

This is known as the *impulse* of the force.

In terms of Cartesian components,

$$\frac{d\mathbf{p}_x}{dt} = m\frac{d\vartheta_x}{dt} = F_x$$

Multiplying by ϑ_x on both sides,

$$m\vartheta_x \frac{d\vartheta_x}{dt} = F_x\,\vartheta_x$$

i.e., $\frac{d}{dt}\left(\frac{1}{2}m\vartheta_x^2\right) = F_x\,\vartheta_x$

Similarly, $\frac{d}{dt}\left(\frac{1}{2}m\vartheta_y^2\right) = F_y\vartheta_y$ and $\frac{d}{dt}\left(\frac{1}{2}m\vartheta_z^2\right) = F_z\vartheta_z$

$$\therefore \frac{d}{dt}\left[\frac{1}{2}m\left(\vartheta_x^2 + \vartheta_y^2 + \vartheta_z^2\right)\right] = F_x\,\vartheta_x + F_y\,\vartheta_y + F_z\,\vartheta_z$$

$$= \mathbf{F}\,.\,\mathbf{v}$$

i.e. $\frac{dT}{dt} = \vec{\mathbf{F}}\,.\,\mathbf{v}$ (11)

where T is the kinetic energy of the body.

Now, $\frac{d}{dt}(\mathrm{v}^2) = \frac{d}{dt}(\vec{\mathbf{v}}\,.\,\vec{\mathbf{v}}) = 2\vec{\mathbf{v}}\,.\,\frac{d\vec{\mathbf{v}}}{dt}$

$$\therefore \mathbf{F}\,.\,\mathbf{v} = m\frac{d\mathbf{v}}{dt}\,.\,\mathbf{v} = \frac{1}{2}m\frac{d(\mathrm{v}^2)}{dt} = \frac{d}{dt}\left(\frac{1}{2}m\mathrm{v}^2\right)$$

Multiplying (11) by dt and integrating,

$\int dT = \int \mathbf{F}\,.\,\mathbf{v}\,dt$

$$\therefore \quad T_2 - T_1 = \int_{t_1}^{t_2} \mathbf{F} \cdot \mathbf{v} \, dt \qquad (12)$$

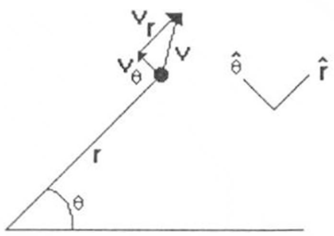
Since $\mathbf{v}\, dt = d\mathbf{r}$

$$T_2 - T_1 = \int_{r_1}^{r_2} \mathbf{F} \cdot d\mathbf{r} \qquad (13)$$

is the work done in going from point 1 (r_1) to point 2 (r_2).

The integral is to be taken along the path followed by the particle from point 1 to point 2.

Angular Momentum



The angular momentum of the particle of mass m about O is

$$L = r\, m\, v_\theta = m\, r^2\, \dot\theta \qquad (14)$$

$$F_r = m a_r = m\ddot{r} - m r \dot\theta^2 \qquad (15)$$

$$F_\theta = m a_\theta = m r \ddot\theta + 2 m \dot{r} \dot\theta \qquad (16)$$

and $$\mathbf{F} = \hat{r} F_r + \hat\theta F_\theta \qquad (17)$$

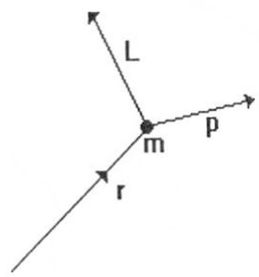
Fig.2. v_θ and v_r

From (14), $$\frac{dL}{dt} = 2 m r \dot{r} \dot\theta + m r^2 \ddot\theta$$

From (16), $$\frac{dL}{dt} = \frac{d}{dt}(m\, r^2\, \dot\theta) = r\, F_\theta = N \qquad (18)$$

N is the *torque* exerted by the force F about the point O.

From (18), $$L_2 - L_1 = m\, r_2^2\, \dot\theta_2 - m\, r_1^2\, \dot\theta_1 = \int_{t_1}^{t_2} r\, F_\theta\, dt$$



In vector language,

$$\mathbf{L} = \mathbf{r} \times \mathbf{p}$$
$$= m\,(\mathbf{r} \times \mathbf{v}) \qquad (19)$$

(**L** is perpendicular to **both r** and **p**).

Fig.3 Angular Momentum

Now, $\dfrac{d\mathbf{L}}{dt} = \dfrac{d}{dt}[\mathbf{r} \times m\mathbf{v}]$

$$= \mathbf{r} \times \frac{d}{dt}(m\mathbf{v}) + \frac{d\mathbf{r}}{dt} \times (m\mathbf{v})$$

$$= \mathbf{r} \times \frac{d}{dt}(m\mathbf{v}) + \mathbf{v} \times (m\mathbf{v})$$

$$= r \times \left(m \frac{d\mathbf{v}}{dt} \right) \qquad since \quad \mathbf{v} \times \mathbf{v} = 0$$

$$= \mathbf{r} \times \mathbf{F}$$

Thus, $\dfrac{d\mathbf{L}}{dt} = \mathbf{r} \times \mathbf{F} = \mathbf{N}$, the torque vector.

$$\therefore \quad \mathbf{L}_2 - \mathbf{L}_1 = \int_{t_1}^{t_2} \mathbf{N}\, dt \qquad (20)$$

Potential Energy

If the force **F** on a particle is a function of its position **r** = (x,y,z), then the work done by the force when the particle moves from $\mathbf{r}_1$ to $\mathbf{r}_2$ is given by $\int_{r_1}^{r_2} \mathbf{F}(r)\,.\,d\mathbf{r}$.

We define *potential energy* V (**r**) = V (x,y,z) as the work done by the force on the particle when it moves from **r** to some standard point $\mathbf{r}_s$.

$$\therefore \quad V(\mathbf{r}) = -\int_{r_s}^{r} \mathbf{F}(r) \cdot d\mathbf{r} \qquad (21)$$

Such a definition implies that V is a function only of (x,y,z) whereas the integral on the right hand side depends on the path of integration.

Let $\mathbf{F} = \mathbf{F}(x, y, z)$

The change in V when the particle moves from **r** to **r** − d**r** is given by

$$dv = -\mathbf{F} \cdot d\mathbf{r} \qquad (22)$$

Recall that $du = d\mathbf{r} \cdot \nabla u$ where u is a scalar function and ∇u is the gradient of u (grad u).

Hence, we may write $\mathbf{F} = -\nabla V$ (= - grad V) (23)

from which $F_x = -\dfrac{\partial V}{\partial x}, \quad F_y = -\dfrac{\partial V}{\partial y}, \quad F_z = -\dfrac{\partial V}{\partial z}$

Now, $\nabla \times \nabla = 0$

$$\therefore \quad \nabla \times \nabla V = \text{curl (grad V)} = 0 \qquad (24)$$

i.e. $\nabla \times \mathbf{F} = \text{curl } \mathbf{F} = 0$ (25)
[using (23)].

Thus, curl **F** = 0 is a *necessary condition* to be satisfied by **F** (x, y, z) before a potential function can be defined.

Now, if we consider a closed path c in space, the work done by the force **F** (r) when the particle travels around the path is

$$\int_c \mathbf{F} \cdot d\mathbf{r} = \iint_s \hat{n} \cdot curl\ \mathbf{F}\ ds \qquad (26)$$

(By Stokes theorem)
= 0 since curl **F** = 0

Thus, $\int_{c} \mathbf{F} \cdot d\mathbf{r} = 0$ (27)

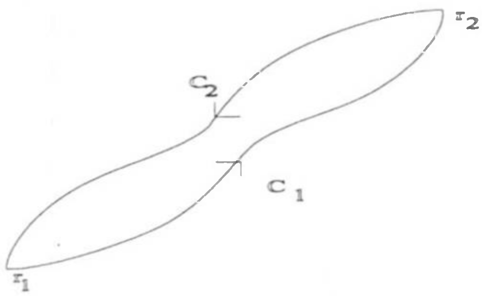


Fig. 4 Work done is independent of path

This means that the *work done in gong from* $\mathbf{r}_1$ *to* $\mathbf{r}$ *is independent of the path.*

i.e. $\int_{c_1} \mathbf{F} \cdot d\mathbf{r} + \int_{c_2} \mathbf{F} \cdot d\mathbf{r} = 0$

where c_1 and c_2 are two arbitrary paths as indicated in the figure.

Thus, curl $\mathbf{F} = 0$ is both *necessary and sufficient condition* for the existence of a potential function V ($\mathbf{r}$) when the force is a function of position $\mathbf{r}$ alone.

We can write $\int_{r_1}^{r_2} \mathbf{F} \cdot d\mathbf{r} = \int_{r_1}^{r_s} \mathbf{F} \cdot d\mathbf{r} + \int_{r_s}^{r_2} \mathbf{F} \cdot d\mathbf{r}$

$$= V(\mathbf{r}_1) - V(\mathbf{r}_2)$$

Recall that $T_2 - T_1 = \int_{r_1}^{r_2} \mathbf{F} \cdot d\mathbf{r}$

Hence, $T_1 + V(\mathbf{r}_1) = T_2 + V(\mathbf{r}_2)$ (28)

i.e., the total energy $E = T + V =$ constant.

Conservative Force

A force which is a function of position alone and whose ***curl*** vanishes is said to be a *conservative force*.

Central Force

A force which is directed always towards or away from a fixed centre and whose magnitude is a function only of the distance from the centre is called a *central force*.

Examples of a central force are gravitational and coulomb forces (both inverse square central forces) and the force (proportional to displacement) responsible for single harmonic motion.

In spherical coordinate, s $\mathbf{F} = \hat{r}\, F(r)$ with cartesian components

$$F_x = \frac{x}{r} F(r)$$

$$F_y = \frac{y}{r} F(r) \qquad F_z = \frac{z}{r} F(r)$$

$$r = \sqrt{x^2 + y^2 + z^2}$$

We can show that

$$Curl\mathbf{F} = \hat{z}\left(\frac{\delta F_y}{\delta x} - \frac{\delta F_x}{\delta y}\right) + \hat{y}\left(\frac{\delta F_x}{\delta z} - \frac{\delta F_z}{\delta x}\right) + \hat{x}\left(\frac{\delta F_z}{\delta y} - \frac{\delta F_y}{\delta z}\right) = 0$$

[Hint: Show that $\frac{\delta F_y}{\delta x} = \frac{\delta F_x}{\delta y}$, etc.)

This means, *a central force is a conservative force.*

Motion under a Central Force

Consider two interacting particles (such as the Sun and a planet or the nucleus and an electron in an atom). We regard one of them to be practically at rest with respect to the other. Since the force is central, $\mathbf{F}(r) = \hat{r}\, F(r)$

Torque $\mathbf{N} = \mathbf{r} \times \mathbf{F} = (\mathbf{r} \times \hat{r})\, F(r) = 0$

i.e. $\mathbf{L}$ = constant.

$$\therefore \quad \frac{d\mathbf{L}}{dt} = \mathbf{N} = 0$$

i.e. $\mathbf{L} = m(\mathbf{r} \times \mathbf{v})$ = constant.

Therefore, *both* **r** *and* **v** *must always lie in a fixed plane perpendicular to* **L**. In other words, *the motion is always planar.*

Since the force is radial only, $F_\theta = 0$ from (15) and (16)

$$m\ddot{r} - mr\dot{\theta}^2 = F(r) \tag{29}$$

$$mr\ddot{\theta} + 2m\dot{r}\dot{\theta} = 0 \tag{30}$$

These are the *equations of motion* for the particle. Since the force is conservative,

$$T + V = \frac{1}{2} m\dot{r}^2 + \frac{1}{2} mr^2\dot{\theta}^2 + v(r) = E, \quad \text{a constant.} \tag{31}$$

($\frac{1}{2} m\dot{r}^2$ - linear KE ; $\frac{1}{2} mr^2\dot{\theta}^2 = I\omega^2$ - rotational KE)

$$L \; (= I\omega) = mr^2\dot{\theta} \text{ is also a constant.}$$

Equation (29) can be written as

$$m\ddot{r} - \frac{L^2}{mr^3} = F(r)$$

or $$m\ddot{r} = F(r) + \frac{L^2}{mr^3} \tag{32}$$

This is exactly the form of an equation of motion in *one dimension* for a particle subjected
to the actual force F(r) plus a fictitious '*centrifugal force*' of magnitude

$$\frac{L^2}{mr^3} \; (= mw^2r = mv^2/r).$$

Inverse Square Central Force

$$\mathbf{F} = \frac{K}{r^2} \hat{r}$$

for which the potential function

$$V(r) = -\int_{\infty}^{r} F(r)\, dr = \frac{K}{r} \qquad (r_s = \infty)$$

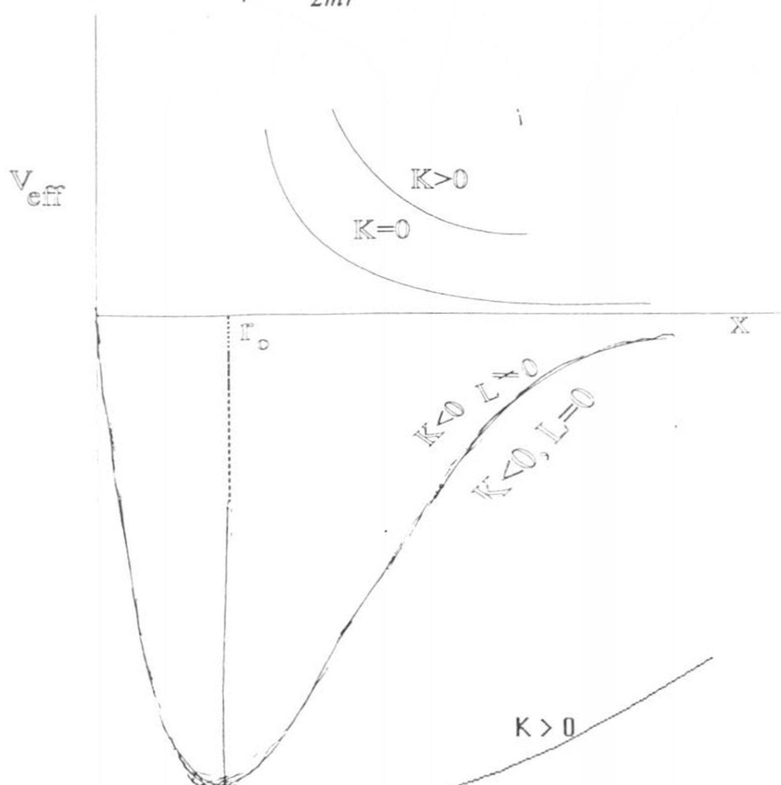
For gravitational force,
$K = -Gm_1m_2$

(Note: V is negative for an attractive force. At $r = \infty$, $V = 0$)

For the coulomb force, $K = q_1 q_2$
(The force is attractive or repulsive).

From (32) we can write,

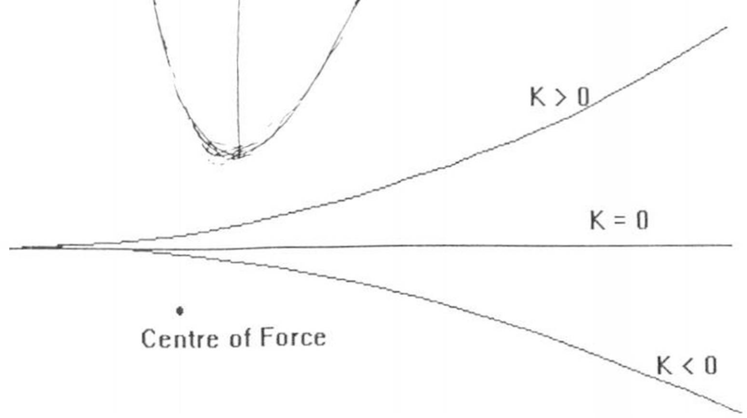
$$V_{eff}(r) = \frac{K}{r} + \frac{L^2}{2mr^2} \qquad (33)$$



where the second term on the RHS is the centrifugal potential.

K > 0 --- Repulsive potential
K = 0 --- No force (straight line path)
K < 0 --- Attractive potential.

Fig.5 Effective Potential



For $E = -\frac{1}{2}\frac{K^2 m}{L^2}$ the particle moves in a circle of radius

$$r_o = -\frac{L^2}{Km} \quad (K < 0).$$

Fig .6 Possible paths of particlesubjected to central force

For an *attractive force (K < 0)*, the path can be an *ellipse*, *parabola* or *hyperbola* (the *circle* is a special case of the ellipse) depending on whether E < 0, E = 0 or E > 0. Examples are planetary/satellite orbits, projectile paths and some cometary orbits.

For a *repulsive force (K > 0)*, we must have E> 0 and the orbit can only be a *hyperbola* (example: paths of charged particles scattered by a nucleus).

Kepler's II Law

$$L = mr^2\dot{\theta} = \text{constant.}$$

The area swept by the radius vector in time dt is $ds = \frac{1}{2} r^2 \, d\theta$

$$\therefore \quad \frac{ds}{dt} = \frac{1}{2} r^2\dot{\theta} = \frac{L}{2m},$$

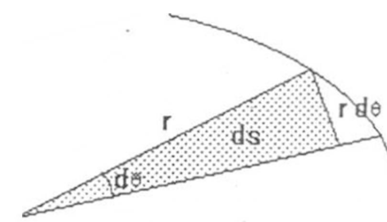


Fig.7 $ds = \frac{1}{2} r^2 d\theta$

a constant , i.e. the radius vector sweeps out equal areas in equal intervals of time. This is Kepler's second law. This explains why a çomet gains speed as it approaches the sun and loses speed as it moves away in its orbit.

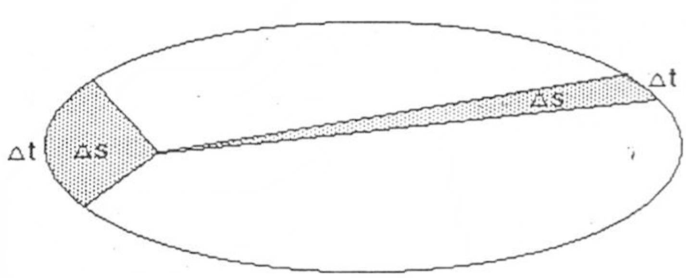


Fig.8 Illustration of Kepler's II Law

Earth Satellites

Artificial earth satellites are of recent origin (from 1957). However, the physical concept involved is traceable to an idea proposed more than 350 years ago by Newton himself. The idea is illustrated in Fig. No. 9. Observe that depending upon the tangential velocity of the object the path can be a parabola, a circle, an ellipse or a hyperbola. In the absence of any tangential velocity, the path is a straight line.

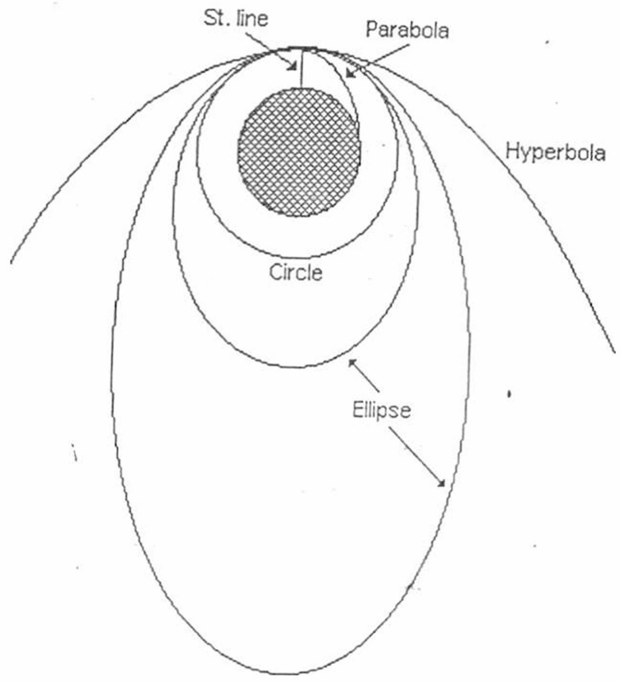


Fig.9 Paths of particle under earth's gravitational force

CHAPTER 2

ANGULAR MOMENTUM

A SPINNING TOP!

Almost everyone might have felt fascinated.in his childhood while playing with tops and asked an unanswered question, "what keeps the top standing?". A top is all grace and beauty, spinning about its axis, and at the same time revolving gently about a vertical line, a behaviour often called *precession.* One may feel as if it is going to fall. But somehow it manages to stay upright, somehow it manages to defy gravity.

Can it be that gravity acts differently on a spinning or a moving object?

The force of gravity remains the same irrespective of the motion of the object.

Another puzzle is the bicycle. Even a good acrobat may not be able to balance a bicycle that is standing still. Obviously, a bicycle has a self-balancing capacity which comes into effect only when it is in motion.

Can it be due to some force from the air on the revolving wheels of the bicycle much in the same way a fast spinning cricket ball swings in a curved path due to a side thrust from the air?

Air has no role to play in the stable motion of the bicycle wheels. They will behave in the same manner even in vacuum.

What then causes such strange behaviour of massive spinning objects? Yes, *massive spinning objects*. The catch lies there.

A heavy spinning object means large *angular momentum.* Every spinning object, a top, a bicycle, a spinning wheel, possess angular momentum. And angular momentum gives stability to objects against wavering motion, as in the case of a bicycle, against being toppled by gravity as in the case of a top, against wobbling of its axis, as in the case of a rifle bullet or an artillery shell which are invariably given spin at the time of firing in order to give them directional stability.

What is this angular momentum and how does it influence the behaviour of spinning objects?

Although a full understanding of angular momentum may not be easy, it is still possible to frame a simplified picture of this concept and present a lucid explanation of the laws governing the behaviour of spinning objects by analogy with linear momentum. Linear momentum is usually more familiarly known by a shorter name *momentum.*

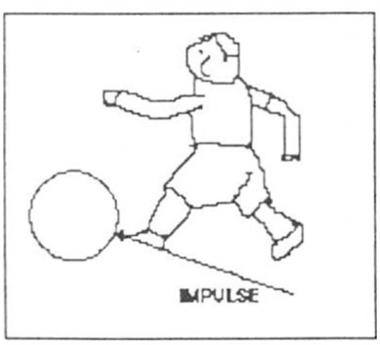
Let us therefore, briefly recall what we understand by momentum and how Newton's second law of motion explains its changing behaviour with time. Momentum **p** is often defined as mass times velocity.

$$\mathbf{p} = m\mathbf{v}$$

where m is the mass of the particle and **v** is its velocity. Now **v** is a vector and m is a scalar. When we take the product of a vector with a scalar, the resulting quantity is a vector whose direction is same as that of the original vector and whose magnitude is the product of the magnitudes of the vector and the scalar. Therefore, if we write v to mean the magnitude of **v**, then the magnitude of **p** is mv and the direction of **p** is the same as that of **v**. We, therefore, say that *momentum and velocity are parallel vectors.*

Now consider Newton's second law of motion. It is customarily written in the form

$$\frac{d\mathbf{p}}{dt} = \mathbf{F}$$



That is, the rate of change of momentum $d\mathbf{p}/dt$, is equal to the external force **F** acting on the particle. We shall often find it convenient to rewrite the above equation in the following form :

$$d\mathbf{p} = \mathbf{F}dt,$$

which means that the change $d\mathbf{p}$ in the momentum of a particle over a very small time period dt is equal to the product of the force **F** multiplied by the time interval dt.

A force acting over a small time interval is called an impulse. An impulse is a sort of kick. When you kick a football, you apply an impulse away from you.

In the same way the quantity $\mathbf{F}dt$ is an impulse, a very tiny impulse, given to the particle within the tiny time interval dt.

Continual application of a force can be looked upon as a train of such tiny impulse, or tiny

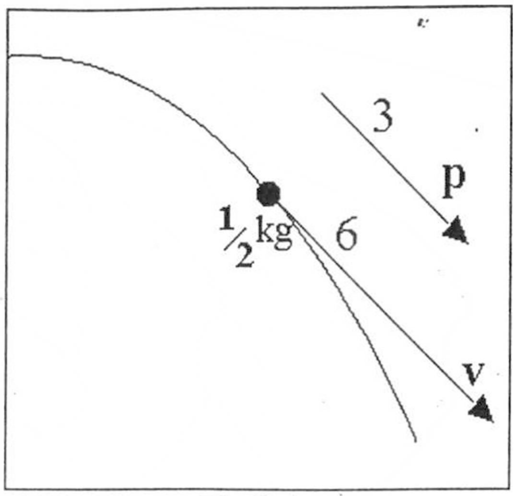


Fig 3.4

kicks, each impulse, or kick corresponding to a tiny time interval *dt*.

What Newton's second law tells us is that each tiny impulse **F***dt* changes the momentum by the vector *d***p** which equals the impulse **F** *dt*.

That means, if we consider an appreciable time span *T*, the tiny impulses `**F** *dt*'s over the tiny intervals `*dt*'s spread over the time *T* add up to an appreciable change in the momentum vector.

Consider a projectile of mass ½ kg moving in a parabola due to the action of gravity. At a certain point A it has a velocity **v** of magnitude 6m/sec. Its momentum is ½ x 6 = 3 units.

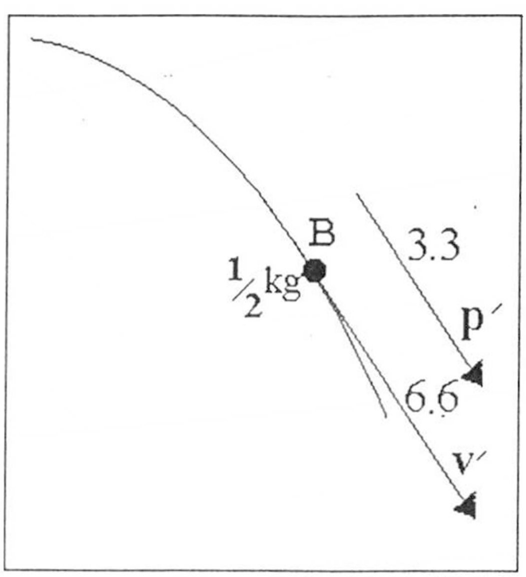
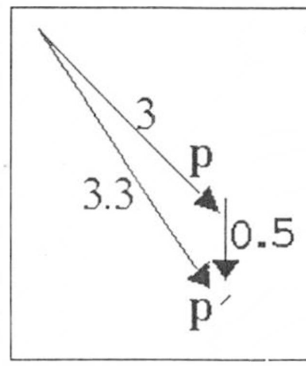


Fig 3.5

After a tiny time interval, say 1/10 the of a second, it moves to a point B where its velocity is 6.6 m/sec and momentum is 3.3 units. The velocity and momentum are shown as **v**′ and **p**′.

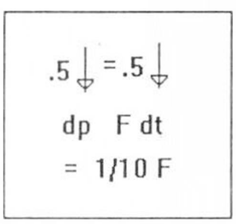
Question: What is the change in the momentum of the particle in the time interval *dt*? To this question, one may be tempted to suggest an obvious answer "0.3 units", an answer which is quite wrong.

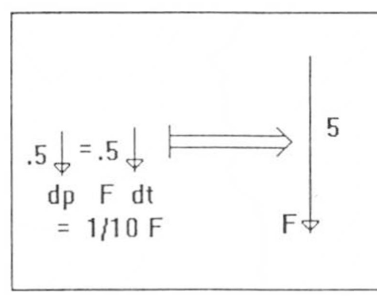
To obtain the answer correctly first look at the two momentum vectors **p** and **p**′ simultaneously and compare them by redrawing these vectors from the same origin O.



You know that vectors are added by the parallelogram rule or the triangle rule. According to this rule, the difference between the vectors **p** and **p**′ is the vector *d***p** whose magnitude is 0.5 units, and *not* 0.3 units, and whose direction is downward.

In other words, if you add to the original vector **p** of magnitude 3 another vector *d***p** of magnitude 0.5, as in the present case, you get the vector **p**′ whose magnitude is 3.3, which is less than the magnitude 3.5, you would have obtained had you added the magnitudes algebraically. This is an important peculiarity of vector addition.





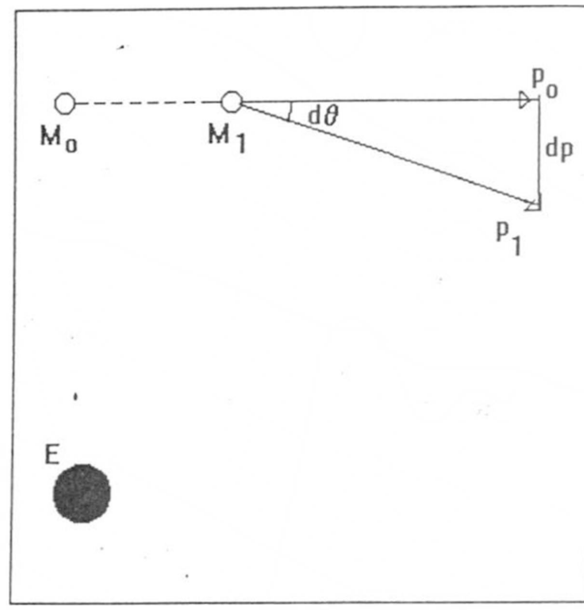
According to Newton's second law this change *d***p** in momentum accruing over the time interval *dt* is equal to the impulse **F** *dt* imparted to the particle by the external force.

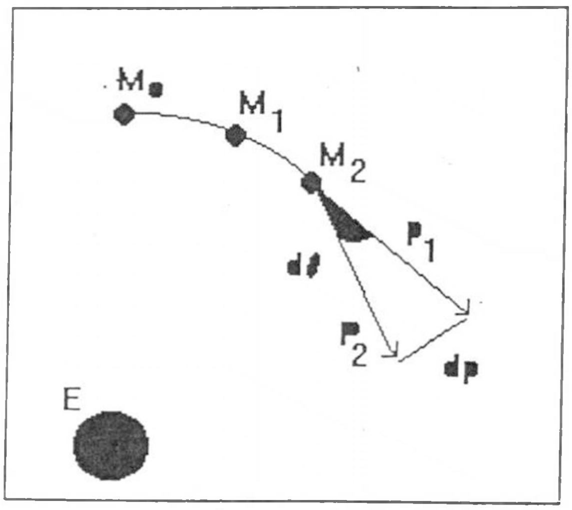
Since *d***p** is 0.5 units downward, and dt is 1/10 sec, the equation *d***p** = **F***dt* implies that the gravitational force on the projectile must be 5 newtons. We know that this is true if we take the acceleration due to gravity to be approximately 10m/sec^2.

Consider another example. The moon is moving in nearly a circular orbit due to the force of gravity directed towards the earth. Why doesn't the moon fall into the earth?

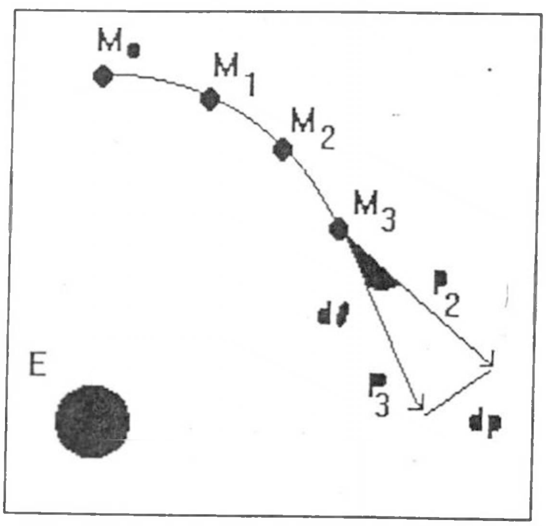
Again the answer can be provided by considering the change of the momentum vector due to a train of impulses directed toward the earth. Suppose E is the location of the earth, and M_o the location of the moon at some instant *t* and p_o is its momentum at that instant.

Because of the momentum $\mathbf{p}_0$ the moon moves to the point M_1 in the interval dt.. But when at M_1 its momentum is no longer $\mathbf{p}$, but has changed to $\mathbf{p}_1$ due to the impulse $\mathbf{F}dt$ which directed towards the earth contributes a momentum $d\mathbf{p}$ in the same direction. The direction of $\mathbf{p}_1$ differs from that of $\mathbf{p}_0$ by the an angle $d\theta$, so that the direction of motion has been deflected by the angle $d\theta$ in going from M_0 to M_1.

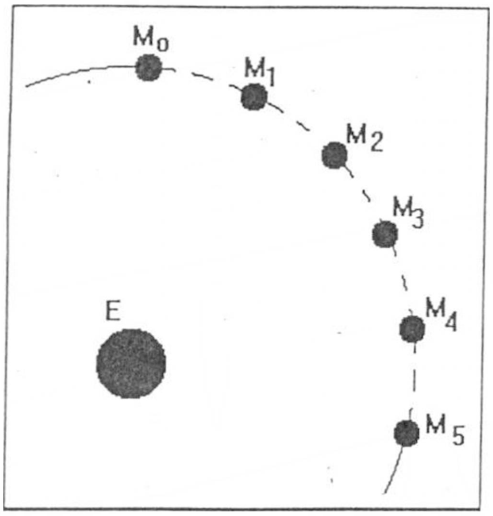


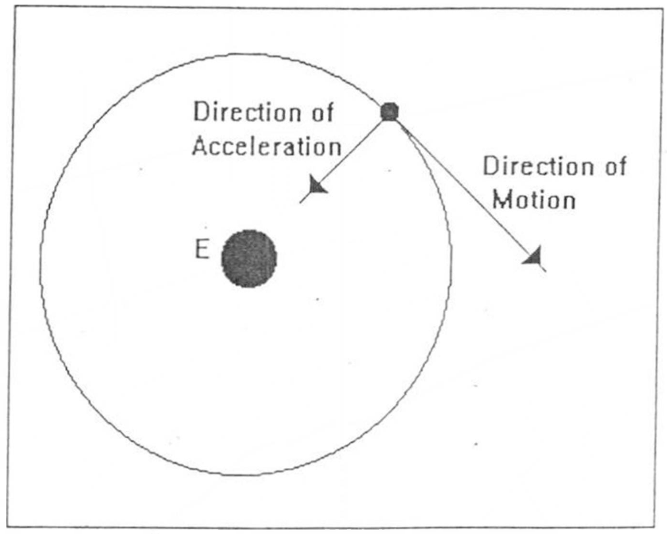


The moon moves from M_1 to M_2 in another interval dt after which its momentum changes to $\mathbf{p}_2$ due to another impulse directed towards the earth, deflecting the motion by another small angle $d\theta$.



In this way, the moon gets continually deflected by small angles dθs due to a train of small impulses, each impulse directed towards the earth.





As a result the moon moves in a circle.

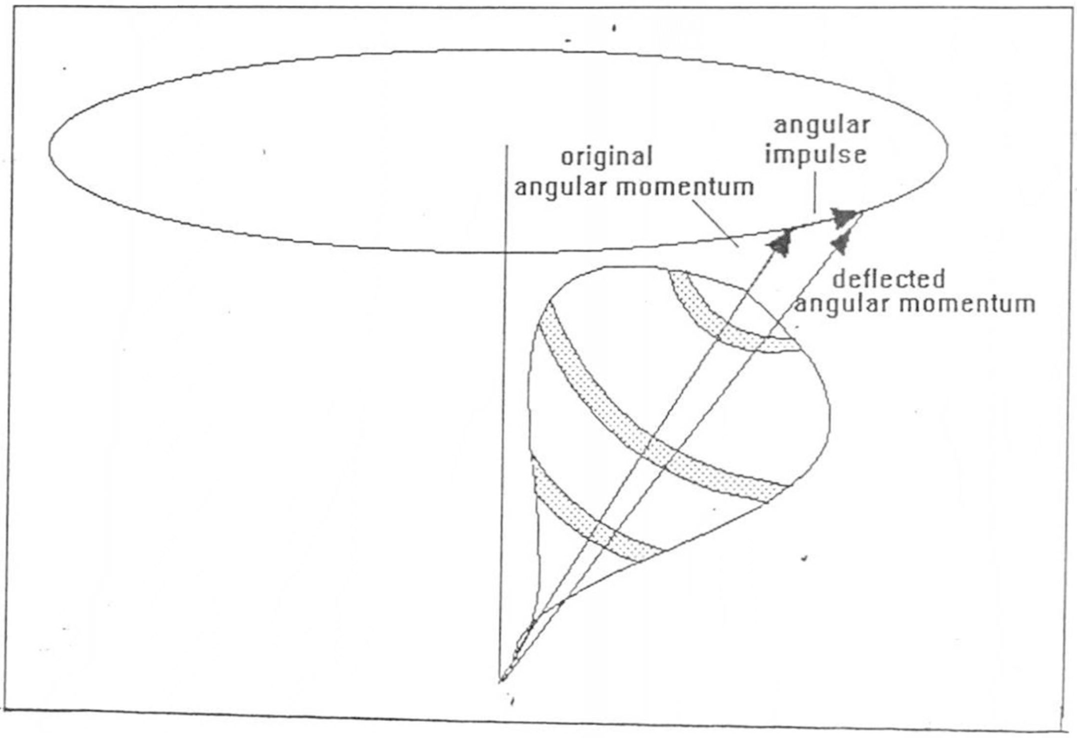
The moon has always been feeling an urge to fall into the earth. In fact she has been continually accelerating towards this cherished centre through millennia. But she has been incapacitated from reaching this goal because of her own momentum, much in the same way a running bull is incapacitated from stopping instantly but is carried away by its own momentum. If the moon didn't have an initial momentum she would have simply dropped onto earth like an apple a long time ago. Because of her momentum, she appears to be defying gravity.

It will be seen that there is a complete analogy between the motion of the moon, or any artificial satellite, and the motion of a top. The moon or any satellite is carried away in its circular or elliptical orbit because of its momentum. Even though it is always trying to fall on earth due to the pull of gravity, it never succeeds. This is happening, as we have seen, because every tiny impulse contributed by the gravity force is added perpendicularly to the existing momentum of the satellite according to the triangle rule of vector addition.

Exactly in the same way, the top is carried away in a beautiful precessional motion, its axis gliding along the surface of a cone. The top would have gladly fallen on the ground if it didn't have spin. A spinning top, on the other hand, has an angular momentum, which prevents it from falling. Like King Tantalus of Greek mythology who is kept chin deep in water but never allowed to drink it, the top is thirsting to fall but not allowed to do so by its own angular momentum.

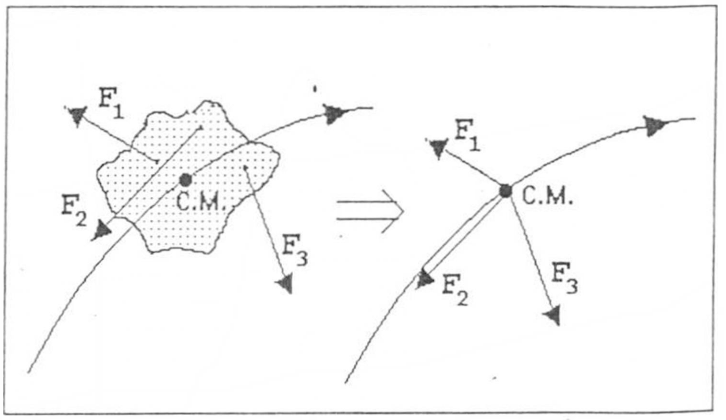
This similarity between the satellite motion and the top motion is due to a parallelism between the law of linear momentum (which is same as Newton's second law of motion), and the law of angular momentum. This latter law is written as :

$$\frac{d\,\mathbf{L}}{dt} = \mathbf{N}$$



where **L** represents the angular momentum of any object, which may be solid, liquid or gas, and **N** is the torque of all external forces. Both **L** and **N** are *measured with respect to the centre of mass of the object.* Without going into finer points, we shall identify the centre of mass of any terrestrial object, to be abbreviated as CM, to be the same as the centre of gravity.

In order to absorb the meaning of the above equation, we need a few remarks. Newton's second law of motion, as you know, is applicable to point particles. It does not automatically apply to an extended body. However, if we imagine the extended body to be made up of a very large number, say *K*, of tiny parts, each part being approximated as a point, then Newton's law applies to each tiny part. The force on any one them, say *A*, is the external force $\mathbf{f}_e$ and the force of interaction $\mathbf{f}_i$ from all other parts *B, C*,....composing this body. However, according to Newton's third law of motion, the force the part *A* exerts on another part *B*, is equal and opposite to the force that *B* exerts on *A*, so that all the forces of interaction within the body pair out into equal and opposite forces. Applying Newton's second law to each component part and then adding up the effects over the whole body, it can be shown that the gross motion of the extended object neatly separates out into two modes of motion. They are :

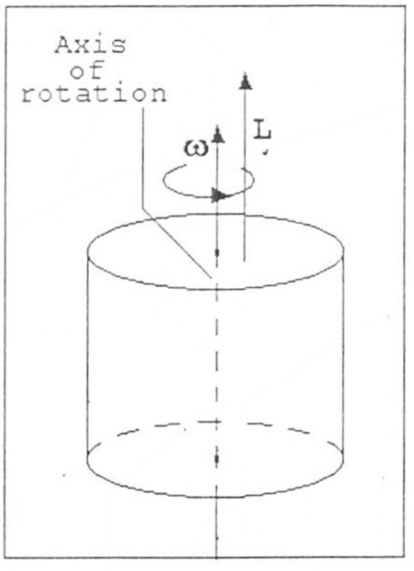


a) The *linear motion of the CM*, governed by the equation $d\mathbf{L}/dt = \mathbf{F}$ where **F** is the *vector sum of all the external forces* acting on the object.

The above equation is equivalent to the following more familiar form :

$$M\,\mathbf{a} = \mathbf{F}$$

where *M* is the total mass of the object and **a** is the *linear acceleration of its CM*. In this mode of motion, the extended body is imagined to be concentrated into a mass point at the CM.

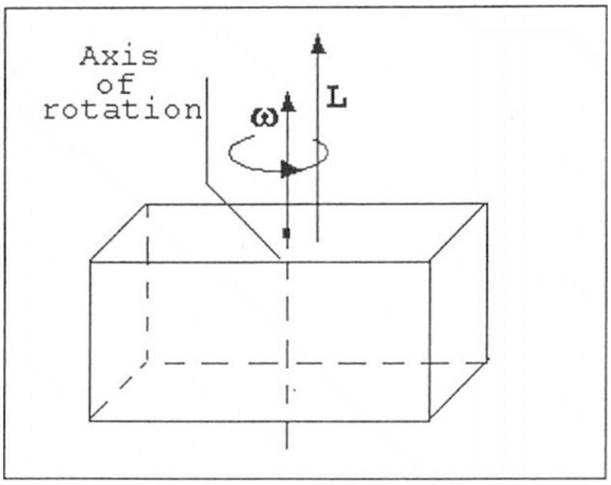


b) The *rotational motion about the CM*, governed by the law of angular momentum:

$$\frac{d\mathbf{L}}{dt} = \mathbf{N}$$

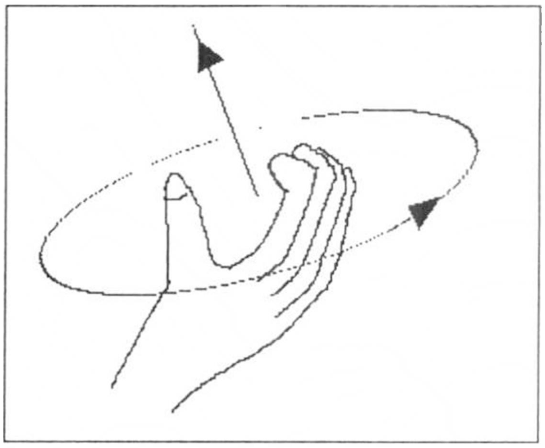
where **L** is the angular momentum about the CM and **N** is the torque of all external forces about the CM. In this mode of motion, the extended body is imagined to be revolving about the CM, which is now imagined to be a point fixed in space.

Before trying to illustrate the above equation, let us have a cursory understanding of angular momentum. A full understanding of this rather complex concept requires a lengthy



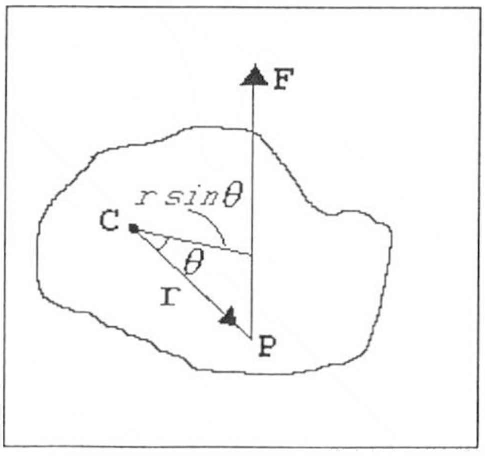
discussion. Therefore, we shall consider only the simplest case, namely a *rigid body which has an axis of symmetry* and which is revolving about this symmetry axis with an angular speed of ω radians per second. For this special case, the angular velocity vector ω and the angular momentum vector **L** are *parallel.*

The direction of the angular velocity vector is given by the right hand thumb rule (to be referred to as RHR). According to this rule, the direction of the angular velocity is along the axis about which the body is turning, its sense being such that if you curl the right hand fingers in the direction of turning then the thumb will give the direction of the ω vector.



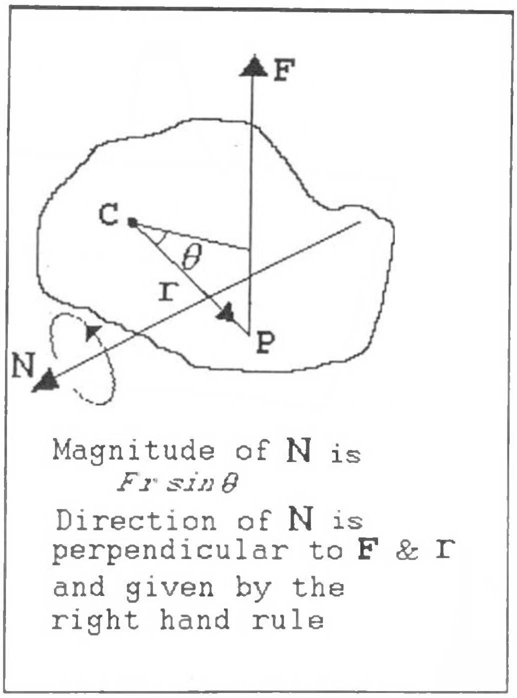
In this simple case, i.e. when a rigid body is turning about an axis XX which is also an axis of symmetry, the angular momentum of the body is given as

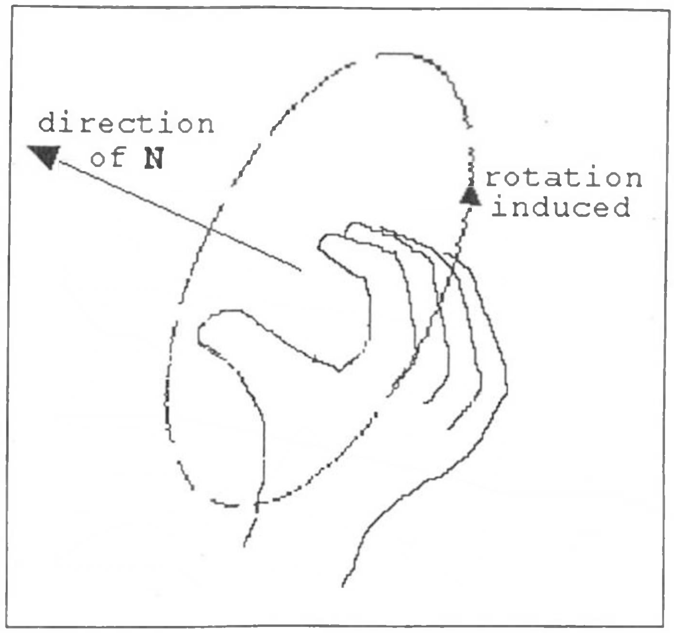
$$\mathbf{L} = I_{XX}\,\omega$$



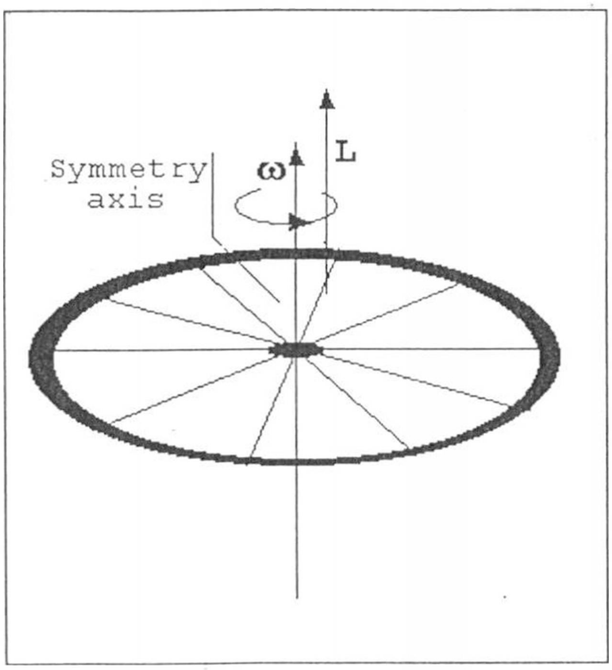
where I_{xx} is the moment of inertia of the body about the axis *XX* passing through the CM. Note that this parallelship between **L** and ω does not hold always. As in the above case, it holds when the axis is an axis of symmetry. Even when the rigid body does not have an axis of symmetry, there always exist at least three axes, perpendicular to one another, such that **L** will be parallel to ω, only when the body turns about any one of them. When this parallelship between L and ω holds, we call the axis of revolution a principal axis. Every axis of symmetry is a principal axis.

The torque **N** of an external force **F** about the CM is also a vector and is also given by the RHR. Let *C* represent the CM and let **F** be a force acting through the point *P* which is at a distance *r* from *C*. The perpendicular distance of the line of the force **F** from *C* is $r\,sin\theta$, where θ is the angle between **r** and **F**.





The torque vector is then perpendicular the plane defined by the vectors **r** and **F**, having a magnitude



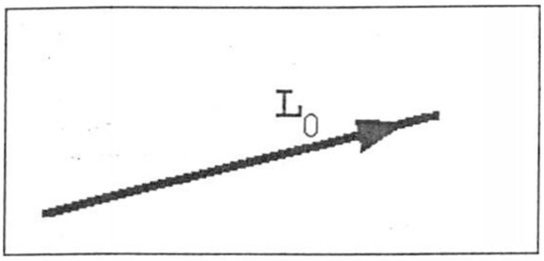
$N = Fr\ sin\theta$, its sense being such that if you curl the right hand fingers from **r** to **F**, then the thumb points in the direction of **N**.

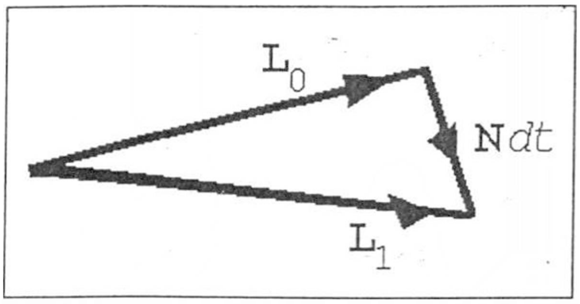
We shall now explain the motion of a top with the help of the law of angular momentum as just stated. For our purpose, we shall find it convenient to rewrite this law in the form of the following differential :

$$d\,\mathbf{L} = \mathbf{N}\,dt$$

which by analogy with linear momentum, says that the change in the angular momentum over time dt equals the angular impulse **N** *dt* imparted by the external forces over this time.

Now consider a spinning wheel. It is revolving about an axis of symmetry with an

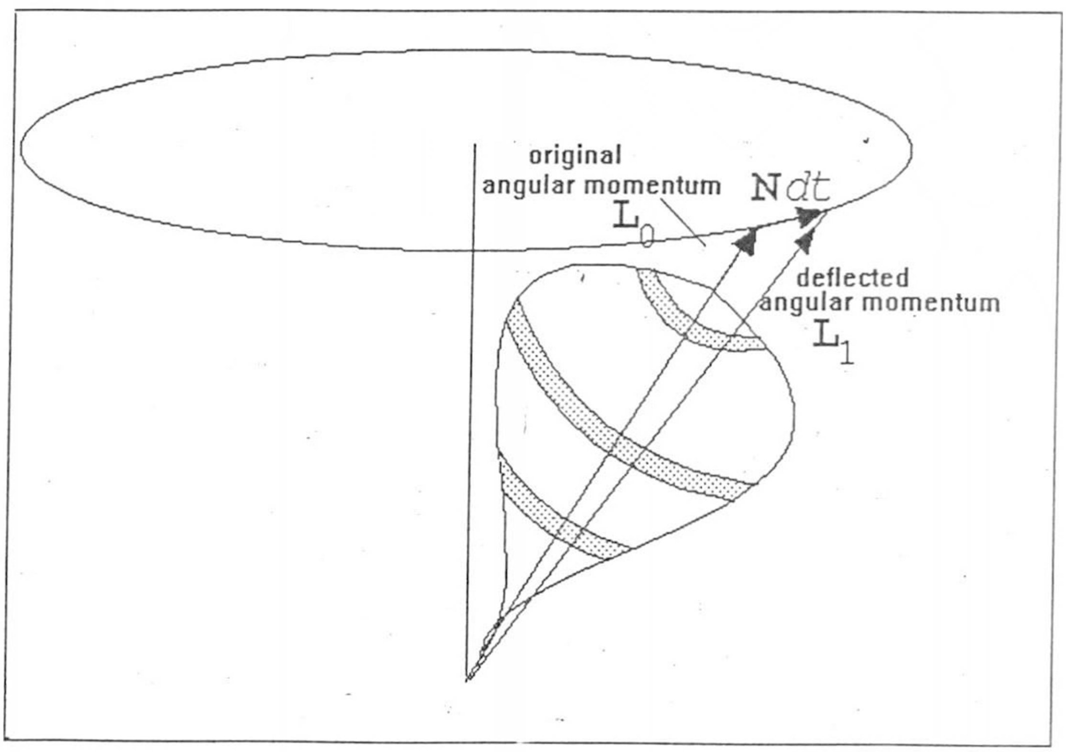




angular velocity ω. Therefore, its angular momentum is $\mathbf{L} = I\omega$, where I is the momentum of inertia about this axis.

Let this wheel have the angular momentum $\mathbf{L}_0$ at some instant t_0. The direction of the vector $\mathbf{L}_0$ therefore also represents the direction of the axis at t_0.

Let there be a torque vector **N** acting on the body over a small time interval dt so that an angular impulse **N** dt is added to the original angular momentum $\mathbf{L}_0$. Then the angular momentum of the body after this interval dt is the vector $\mathbf{L}_1$, which is obtained by adding to the original angular momentum $\mathbf{L}_0$ the small angular impulse $d\mathbf{L} = \mathbf{N}\,dt$ imparted over the interval dt. Since the new angular momentum vector $\mathbf{L}_1$ has a different direction than $\mathbf{L}_0$, and since the axis is following the direction of the angular momentum, the visible effect is that the direction of the axis has been deflected in time dt.



The strange behaviour of the top can now be easily explained in the light of the above mechanics. When the top is spinning (and also precessing) its angular momentum is approximately along the axis of the top (there is an extra component of angular momentum from the precessional motion, which we are ignoring for convenience). The top is experiencing two forces. The gravity force and the ground reaction force. Of these two, the gravity force *mg* is passing through the CM and therefore, does not produce any torque. The ground reaction **R** is mostly a vertical force but may also include a horizontal component if the ground is not smooth. This force **R** causes a torque **N** which is a horizontal vector perpendicular to the plane defined by the axis of the top and the reaction **R**. The resulting angular impulse **N***dt* over every small interval of time *dt* will then deflect the angular momentum vector perpendicular to itself. As a consequence, the angular momentum vector, and along with it, the axis of the top, will continuously precess along the surface of a cone.

Note that this is analogous to the way the gravitational impulse deflected the linear momentum of a satellite, sending it along a circular orbit. We have improvised a number of bicycle wheel devices to bring home to the learner the meaning, in fact the "feel" itself, of what we understand by angular momentum. Hold these wheels by hand, try to twist them, turn them, cuddle them while they are spinning. It will be an enlightening experience, an experience that can never come from any amount of book reading. For a while, you will disbelieve your eyes when you observe the wheel turning sideways in response to your efforts to twist it downwards. Then all of a sudden an enlightenment will dawn, and you will visualise the law of angular momentum before your eyes, and such strange phenomena like the motion of a spinning top will appear as natural as the fall of an apple. Only then you will have established your faith in the physics you have just learnt.

CHAPTER 3
WAVE MOTION

Wave Motion

Particle and wave are two major concepts in classical physics. When we wish to study the motion of material bodies, we often make use of the concepts of a particle. It suggests a tiny concentration of matter capable of transporting momentum and energy. A wave, on the other hand, suggests a broad distribution of energy, filling the space through which it passes. A wave transmits energy from one place to another without the actual movement of material particles between those places.

Suppose you intend to get in touch with a friend at a distant place. You can do it, either by sending a letter or by using the telephone. When you send a letter, a material object (the letter) moves from one place to another carrying the information. In the second case, a wave carries this information from you to the friend. There is no movement of any material object. When you make a telephone call, a sound wave carries your message from your vocal cords to the telephone; at the telephone system it is converted to an electromagnetic wave, which may pass through a copper wire or optical fibre (or may even move through space via a communication satellite). At the receiving end, your friend's telephone converts the electromagnetic wave into an audible wave (sound wave) and delivers the information to his ear. Thus the information is transported without the transport of any material medium.

Though we have talked about the electromagnetic waves above, the subject matter at present will be of a general nature on waves, especially waves in material media or elastic media.

Waves in General

A flag fluttering in breeze, ripple waves in water, sound waves in air and other media and seismic waves are some of the examples of mechanical waves. Mechanical waves are governed by Newton's laws, and they need a mechanical medium for propagation. Electromagnetic waves, on the other hand, are of a different nature and do not need any medium to pass through. All electromagnetic waves, irrespective of their wavelength, travel through vacuum with the same speed c, given by

$$c = 299,\ 792,\ 458 \text{ m/s}$$

Under special conditions, we come across another category of waves - matter waves - which are governed by the laws of quantum mechanics. An example is an energetic beam of electrons exhibiting wave characteristics under certain conditions.

Mechanical Waves

Many of the aspects of wave motion can be understood by considering the sinusoidal waves in a long stretched string. We assume that the string is infinitely long so that there is no effect of an echo due to reflection. Wave motion can be studied either by monitoring the waveform (shape of the wave)as it moves along the string, say from left to right, or we can concentrate on a specified element of the string and observe its motion. The waves in the string are transverse in nature since the displacement of any element is in the y direction perpendicular to the direction of travel of the wave which is the x direction.

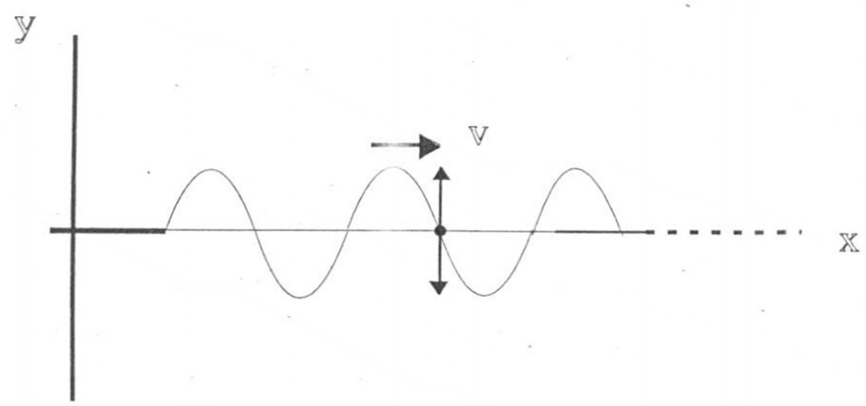


Fig.1 A typical element of the string moves up and down as the disturbance travels along

In contrast to this, consider the motion of a sound wave, set up in a long air-filled pipe by using an oscillating piston. In this case, the displacements of small elements of air are back and forth and hence parallel to the direction of propagation of the wave. Such wave motions are termed longitudinal. Sound waves are always longitudinal in nature.

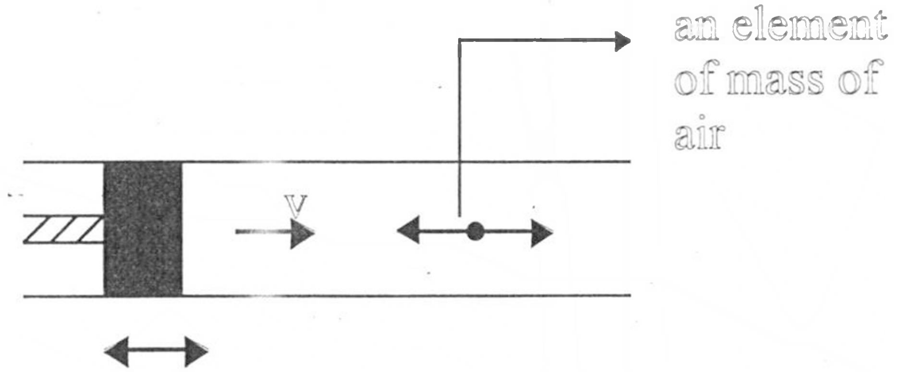


Fig.2 The to and fro motion of piston causes a to and fro motion of the element of mass in the parallel direction.

Transverse waves in a string

Consider a long string stretched by a tension T. Let a small portion of the string be given a sudden lateral displacement. Two things happen now. The displaced part of the string will tend to return to the original position i.e. restoring forces will act on the string. Secondly, the displaced portion of the string will exert lateral forces tending to displace adjacent parts of the string. This results in a pulse travelling out in each directions from the original undisplayed part.

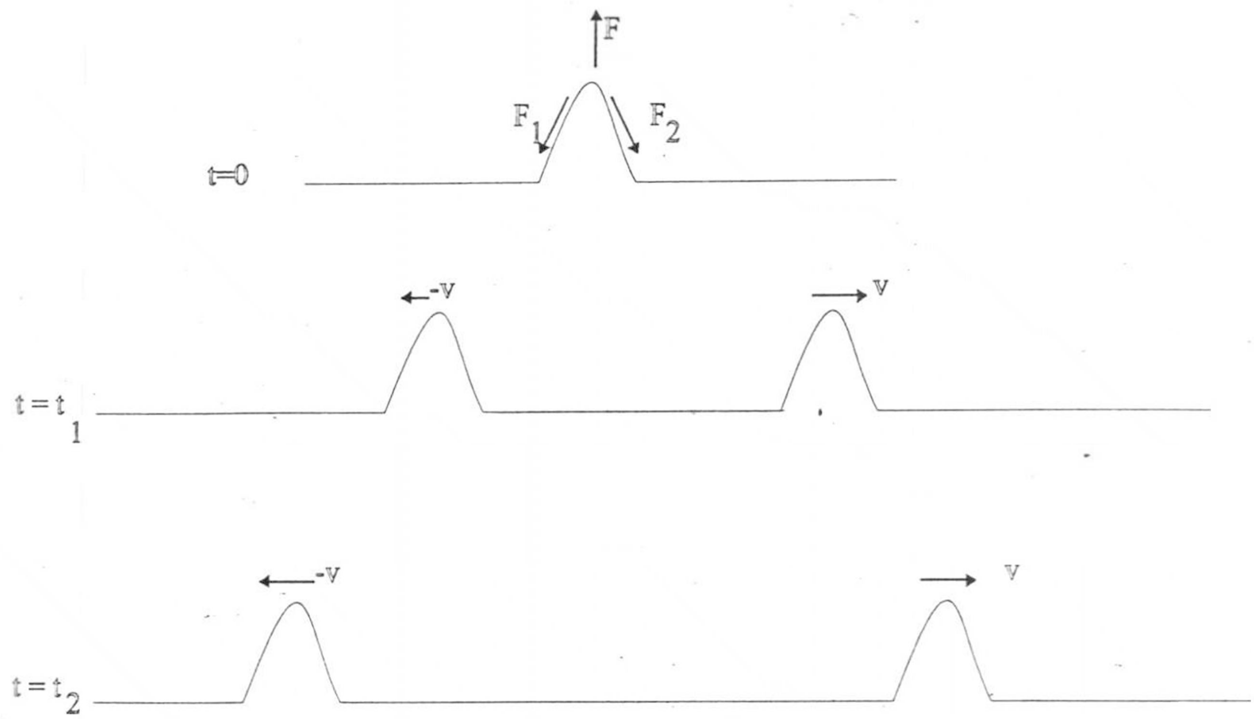


Fig. 3 A pulse travels along the string in both directions from the source

The speed with which a pulse travels along the string is characteristic of the condition of the string.

What are the conditions for a wave pulse (or a wave) to travel through the string? Or in other words, why an elastic medium, such as a string, when disturbed vertically, produces a pulse travelling laterally along the medium? These reasons can be listed as below. 1. There must be forces acting along the string which causes the displacement of the medium (string) as the pulse passes through. 2. The string must have elasticity, i.e. it does not tear apart under the stress created by the passing wave; at the same time, it should not be too rigid to yield to a pulse. 3. The string must have inertia so that when it reaches the equilibrium position, it continues to move and go beyond the equilibrium position. In figure 4, we have shown the forces acting on the cord corresponding to various stages. The forgoing reasons indicate that such forces must be acting so that the string is stretched and contracted to keep the pulse moving along. The segment of the string nearer the source passes its energy to the string segment adjacent to it by doing work on it (stretching the segment, for example). Once it has given up this energy, the crest formed by the wave pulse collapses and the string attempts to come to its original position, but due to its inertia it overshoots the equilibrium position creating a downward crest which is then passed on to the next segment of the string.

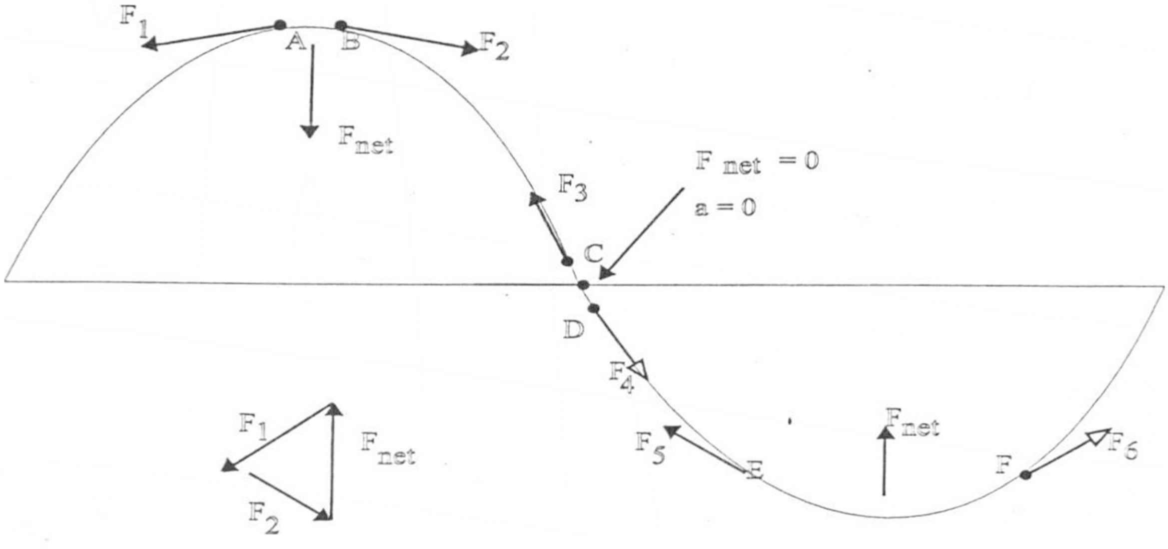


Fig. 4 Forces acting on a section of the string
Inset Shows the vector addition of F_1 and F_2 to get F_{net}

The figure illustrates the forces acting on the segments. When the string segment AB is pulled upward forces $\mathbf{F_1}$ and $\mathbf{F_2}$ (restoring forces) counterbalance it such that the upward motion of the string is stopped. The net downward force, $\mathbf{F_{net}}$, on the segment acts to accelerate it downward according to Newton's second law of motion. In the segment CD the sum of the forces $\mathbf{F_3}$ and $\mathbf{F_4}$ cancel out and no net force acts on this element and no acceleration at that point. The upward net force on EF can also be explained similarly.

To show that energy is transmitted along the string

Consider the forces acting at any point on the string. For example, $\mathbf{F_5}$ is the force acting at point E on the segment EF. The vertical component of this force is $F_5 \sin\theta$ which is acting upward and hence pulls a particle at point E upward with a velocity $V = s/t$. Let the particle move a distance $s = Vt$. This means that an amount of work $W = Fs = (F_5 \sin\theta)(Vt)$ is done on the point E by the adjacent segment of the string. This work is done by each particle in the string on the adjacent particle. In this manner, the string passes along the energy transmitted to it by the wave.

To derive an expression for the speed of the pulse in terms of the tension T and the characteristics of the string (medium).

Let us consider the forces acting on a small section of the pulse.

$r\Delta\theta$ = length of the arc of the string AB

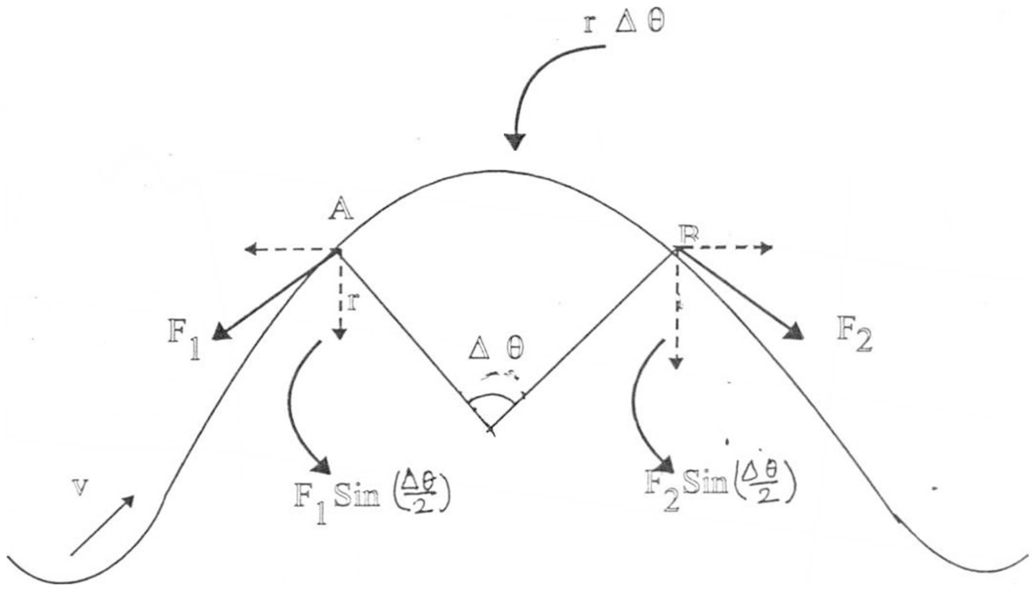


Fig.5 rΔθ = length of the arc of the string AB

Since we are interested in the relative speed of the pulse with respect to the string, we imagine that the string is moving over the top portion of a stationary hump (which has the shape of a pulse) with the relative speed v. We consider the top portion of the string as an arc of a circle of radius r such that the length of the segment AB $= r\,\Delta\theta$. From figure, we see that the net downward force on this segment is

$$F_{net} = F_1 \; Sin\left(\frac{\Delta\theta}{2}\right) + F_2 \; Sin\left(\frac{\Delta\theta}{2}\right)$$

Since $F_1 = F_2 = T$ (the tension in the cord,

$$F_{net} = 2\,T\,Sin\left(\frac{\Delta\theta}{2}\right)$$

$$\approx 2\,T\left(\frac{\Delta\theta}{2}\right) \quad \text{Since } Sin\left(\frac{\Delta\theta}{2}\right) \approx \frac{\Delta\theta}{2} \text{ for small } \Delta\theta.$$

$$\therefore\; F_{net} = T\,\Delta\theta \qquad (1)$$

The centripetal acceleration in the segment AB is $a = \frac{v^2}{r}$ and hence the force $F = ma = m\,\frac{v^2}{r}$.

Let μ be the mass per unit length of the string. Then, $\mu = \frac{m}{l}$, since $l = r\,\Delta\theta$, $m = \mu r\,\Delta\theta$.

$$F = \mu \, r \, \Delta \theta \, \frac{\upsilon^2}{r} \qquad (2)$$

Since these two forces must be equal, we have,

$$F = T \, \Delta \theta = \mu \, r \, \Delta \theta \, \frac{\upsilon^2}{r}$$

or, $\upsilon^2 = \frac{T}{\mu}$ or $\upsilon = \sqrt{\frac{T}{\mu}}$ (3)

Each pulse will travel along the string at constant speed which depends only on the tension T and the mass per unit length of the string. The assumption that $\Delta\theta$ is small means that the result holds good only for small transverse pulses (but the pulse may be of any shape).

Example

One end of the string is fastened to a stop and the other end hangs over a pulley with a 2.0 kg mass attached. What is the speed of the transverse wave in the string?
$T = M g = (2.0 \text{ kg})(9.8 \text{ m/s}^2) = 19.6 \text{ N}$

$$\mu = \frac{m}{l} = \frac{3.0 \ X \ 10^{-3} \ kg}{4.0 \ m} = 7.5 \ X \ 10^{-4} \ \frac{kg}{m}$$

$$\upsilon = \sqrt{\frac{T}{\mu}} = \sqrt{\frac{19.6 \ kg \ \frac{m}{s^2}}{7.5 \ X \ 10^{-4} \ \frac{kg}{m}}} = \sqrt{2.6 \ X \ 10^4 \ \frac{m^2}{s^2}} = 160 \ m \ s^{-1}$$

Travelling Waves

Consider a long string stretched in the x-direction. A transverse pulse generated at one end at t=0 is travelling along the string. Let the shape of the string at x at t=0 be represented by $y = f(x)$.

$$y = f(x), \quad t = 0$$

If we neglect the internal losses, such a wave would travel along the string without any change in shape. If v is the magnitude of the wave velocity the wave travels a distance vt in time t. Therefore, the equation of the curve at time t is given by

$$y = f(x - vt), \qquad t = t \qquad (4)$$

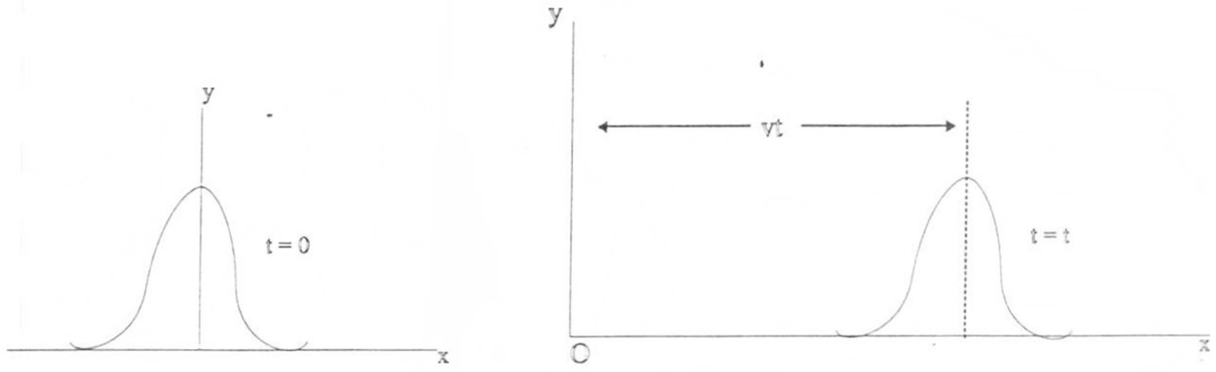


Fig. 6

This expression ensures that the waveform at t=t at $x=vt$ is the same as the waveform at x=0 at t=0. Equation (4) therefore gives the general equation representing a wave of any shape travelling to the right. If the wave were travelling to the left, we could represent it as $y = f(x+vt)$.

Let us now consider a particular part of the wave (phase) as time goes on. For this, we can look at a particular value of y, say the top portion. This means, we are looking at how x changes with t as $x-vt$ remains at some particular value. For this, x should increase as t increases so that $x-vt$ remains constant. To find the velocity of a particular phase of the wave, we can use the condition,

$$x - vt = \text{const.}$$

Differentiating w.r.to time t, $dx/dt = v$

v is called the phase velocity of the wave.

The wave equation can also be interpreted further. For a particular value of time t, the equation gives y as a function of x. This gives the actual shape of the pulse. The same result holds good for longitudinal waves also. An example for longitudinal wave is a long tube containing gas through which a pressure change is passing through. (See Fig.2)

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Let us now consider a particular waveform (which is also the simplest and the most important one). We represent this waveform at time t=0 by the relation,

$$y = A \sin \frac{2\pi x}{\lambda}$$

The shape of this curve is shown.

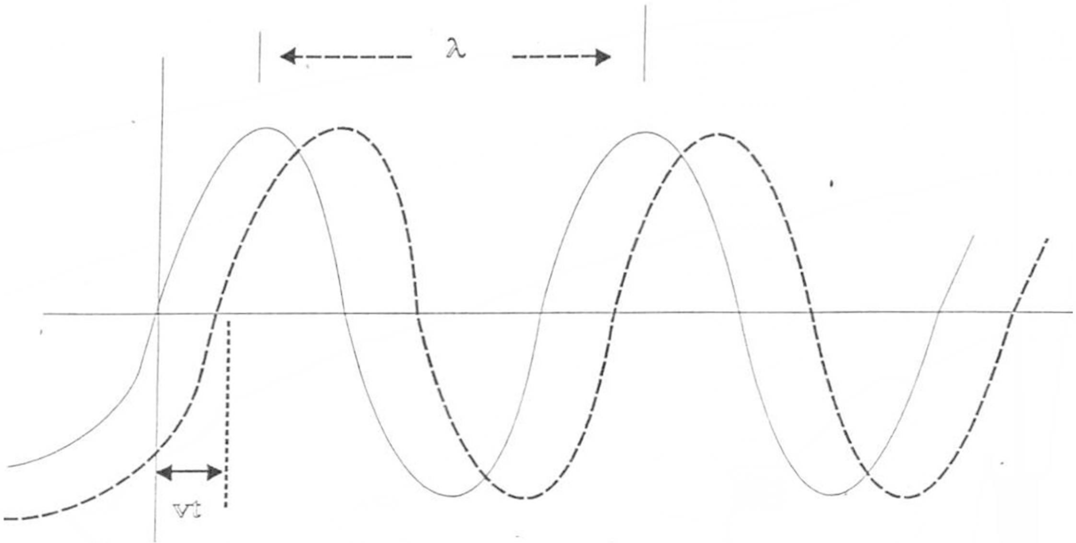


Fig. 7

A is called the amplitude of the wave. It corresponds to the maximum displacement. The value of the displacement y is the same at x as it is at x + λ , x + 2 λ etc. This distance λ is called the wavelength of the wave train . It represents the distance between two adjacent points in the wave having the same phase. Let us assume that the wave is travelling to the right with a phase velocity V. In time t, the wave would have travelled a distance Vt. Hence the equation of the wave at time t is given by

$$y = A \, Sin \frac{2\pi}{\lambda} (x - \upsilon t) \qquad (5)$$

The time required for the wave to travel a distance of one wavelength λ is called the period T.

∴ $\lambda = \upsilon t$.

Using this, we get

$$y = A \, Sin \frac{2\pi}{\lambda} (x - \upsilon t)$$

$$= A \ Sin \ 2\pi \left(\frac{x}{\lambda} - \frac{v}{\lambda} \ t \right)$$

From this equation, we see that at any given time t, y has the same value at $x + \lambda$, $x + 2\lambda$ etc. as it has at x.

$$= A \ Sin \ 2\pi \left(\frac{x}{\lambda} - \frac{t}{T} \right) \qquad (6)$$

Also any given position y has the same value at t+T, t+2T etc. as it has at t.

Let us now reduce eqn.(6) into other usable forms. We define two quantities wave number k and angular frequency ω, using the relations,

$$k = \frac{2\pi}{\lambda} \text{ and } \omega = \frac{2\pi}{T}$$

In terms of these quantities, a sine wave travelling to the right is represented by

$$y = A \sin (kx - \omega t) \qquad (7)$$

Similarly, a wave travelling to the left is given by

$$y = A \sin (kx + \omega t)$$

Using the relation $\lambda = VT$, Substituting for λ and T we get

$$\upsilon = \frac{\lambda}{T} = \left(\frac{2\pi}{k} \right) \left(\frac{\omega}{2\pi} \right) = \frac{\omega}{k} \qquad (8)$$

In our forgoing treatment we have assumed that the displacement y=0 at x=0 at t=0. But in practice, this need not be true. To take into account this, we write the general equation of a travelling wave as

$y = A \sin (kx - \omega t - \Phi)$ where Φ is called the phase constant.

Sometimes, we may be interested in the displacement at a particular point, say, $x = \pi/k$. This is obtained by cutting $x = \pi/k$ in eqn.(7).

We get $y = A \sin (\pi - \omega t - \Phi)$

$= A \sin (\omega t + \Phi)$ (9)

This represents a simple harmonic motion. Thus any particular element of the string undergoes simple harmonic motion about its equilibrium position when a wave train travels along the string.

CHAPTER 4

TRANSMISSION OF ENERGY

In all travelling waves, energy travels through the medium in the direction in which the wave travels. Each particle of the medium has energy of vibration, and passes energy on to succeeding particles.
Consider a portion of a string at some position x at time t.

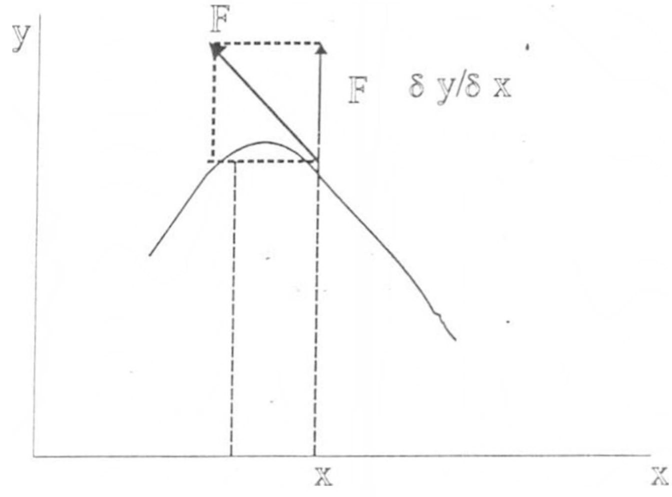


Fig.8

F is the tension acting on the string at position x. The transverse component of this force is $F_{trans} = -F \frac{\delta y}{\delta x} \delta x$. The minus sign is due to the fact that this force is exerted by the element to the left of x on the element to the right of x. Since $\frac{\delta y}{\delta x}$ is negative, F_{trans} is positive and is along the direction of increasing y. The transverse velocity of the particle at x is $\frac{\delta y}{\delta t}$ and it may be negative or positive. Since power is given by the product of force and velocity, the power expended by the force at x, or the energy passing through the position x per unit time in the positive direction is

$$P = \left(-F \frac{\delta y}{\delta x} \right) \left(\frac{\delta y}{\delta t} \right) \qquad (10)$$

Assuming sinusoidal wave, y = A sin (kx-wt) for a wave propagating in the positive x -direction.

Assuming sinusoidal wave, y = A sin (kx-wt) for a wave propagating in the positive x -direction.
∴ The magnitude of the slope at x is

$$\frac{\delta y}{\delta x} = k A \cos(kx - \omega t),$$ t = constant, and the transverse force is,

$$= -F k A \cos(kx - \omega t) \qquad (11)$$

The transverse velocity of the particle at x is ,

$$u = \frac{\delta y}{\delta t} = -\omega A \cos(kx - \omega t) \qquad (x = \text{constant}), \qquad (12)$$

∴ Power transmitted through x is

P = (transverse force) (transverse velocity) $(P = F \cdot \upsilon)$

$= [-F kA \cos(kx - \omega t)] [-\omega A \cos(kx-\omega t)]$

$$= A^2 k\omega F \cos^2(kx - \omega t) \qquad (13)$$

This equation shows that the power or rate of flow of energy is not constant. (The power input itself oscillates). The energy passing through the string is stored in each element of the string as a combination of kinetic energy of motion and the potential energy of deformation. The situation is similar to that obtained in an alternating current circuit, consisting of an inductance L and a capacitor C. As the power input oscillates, the energy is stored in the inductor and the capacitor alternately. In both cases, loss of energy occurs. In the string, it is due to internal friction and viscous effects; in the circuit, it is due to the resistive elements in the circuit. The power input to the string is found by taking the average over one period of motion. The average power delivered

$$\overline{P} = \frac{1}{T}\int_t^{t+T} P\,dt$$ where T is the period

$$= \frac{1}{T}\int_t^{t+T} A^2 k \omega F \cos^2(kx - \omega t)\,dt$$

$= \frac{1}{2} A^2 k\omega F;$ the average of $\cos^2$ over a period reduces to ½.

Using $k = \frac{\omega}{v}$, $\overline{P} = \frac{1}{2} A^2 \omega^2 \frac{F}{v}$, Where $\omega = 2\pi\gamma$

$$= 2\pi^2 A^2 \gamma^2 \frac{F}{\upsilon} \qquad (14)$$

We find that the average power does not depend on x or t. In the case of a string, the speed υ is related to F and μ as $\upsilon = \sqrt{\left(\frac{F}{\mu}\right)}$

$$P = 2\pi^2 A^2 \gamma^2 \mu \upsilon$$

Thus the rate of transfer of energy depends on the square of the wave amplitude and the square of the frequency. This fact holds good for all types of waves.

Principle of Superposition

Standing Waves (Stationary Waves)

If two sinusoidal waves of same amplitude and frequency travel in opposite directions through a medium, the two waves will be superposed in such a manner that stationary waves or standing waves are produced. Consider two such waves represented by

$y_1 = A \sin(kx - \omega t)$ and

$y_2 = A \sin(kx + \omega t)$

The resultant wave is given (by the principle of super-position)

$y = y_1 + y_2 = A [\,\text{Sin}(kx - \omega t) + \text{Sin}(kx + \omega t)\,]$

$$\left[Sin\ C + Sin\ D = 2\ Sin \frac{C + D}{2}\ Cos \frac{D - C}{2} \right]$$

$$Y = 2A\ \text{Sin}\ kx\ \text{Cos}\ \omega t \qquad (15)$$

This equation represents a standing wave. A particle at any point x executes simple harmonic motion as time goes on and all particles vibrate with the same frequency. Whereas in a travelling wave each particle of the string vibrates with the same amplitude, in a standing wave, the amplitude is not the same for different particles but varies with the location x of the particle. From eqn.15, we see that the amplitude is given by 2A sin kx which itself is a sinusoidal function of position. The amplitude has a maximum value of 2A at positions where

$$k x = \frac{\pi}{2}, \frac{3\pi}{2}, \frac{5\pi}{2}, \text{etc.}$$

i.e. $x = \frac{\pi}{2k}, \frac{3\pi}{2k}, \frac{5\pi}{2k}$

Since $k = \frac{2\pi}{\lambda}$, this leads to $x = \frac{\lambda}{4}, \frac{3\lambda}{4}, \frac{5\lambda}{4}$ etc.

These points are called the antinodes and the distance between two consecutive antinodes is equal to $\lambda/2$, one-half wavelength. At $kx = \pi, 2\pi, 3\pi$, etc, we find the amplitude to be zero i.e. at $x = \lambda/2, \lambda, 3\lambda/2$ etc. The points of zero amplitude are called nodes and nodes are also separated by one half wavelength. Fig.9 illustrates the nodes and antinodes in a standing wave pattern.

What happens if the amplitudes are different ?

We have seen that $P = 2\pi^2 A^2 \gamma^2 \mu \upsilon$ (rate of transfer of energy).

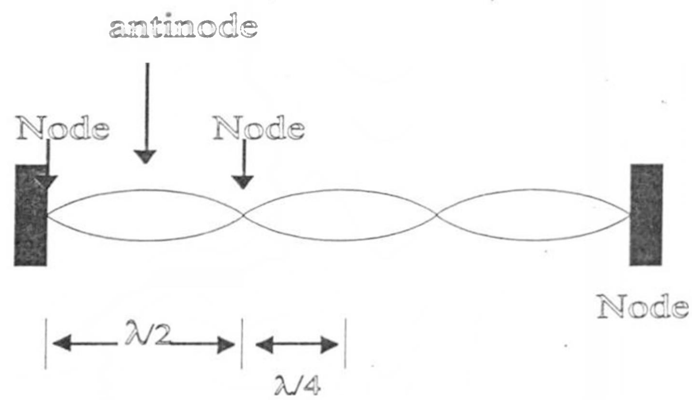


Fig.9

Do standing waves transport energy ?

Within a stationary wave, there is no flow of energy through the medium. Since the standing waves are produced by combining two waves of equal amplitude and frequency in opposite directions, energy transfer in one direction by one wave is equal to the energy transfer by the other wave in the opposite direction. The energy *alternates between vibrational kinetic energy and elastic potential energy and cannot be transmitted through the nodes*. We can also regard the standing wave pattern *as an oscillation of the string as a whole*, each particle oscillating with simple harmonic motion of angular frequency ω and an amplitude that depends on its location. We can imagine a vibrating string as a system of coupled oscillators where each part of the string has inertia and elasticity. Hence the vibrating string can be thought of as a collection of coupled oscillators.

What is the comparison with a spring mass system? A spring mass system has only one natural frequency of vibration whereas the vibrating string has a large number of natural frequencies.

Can we recall the standing wave pattern as a wave motion ? Yes - we can describe it as the superposition of two travelling waves, travelling in opposite directions.

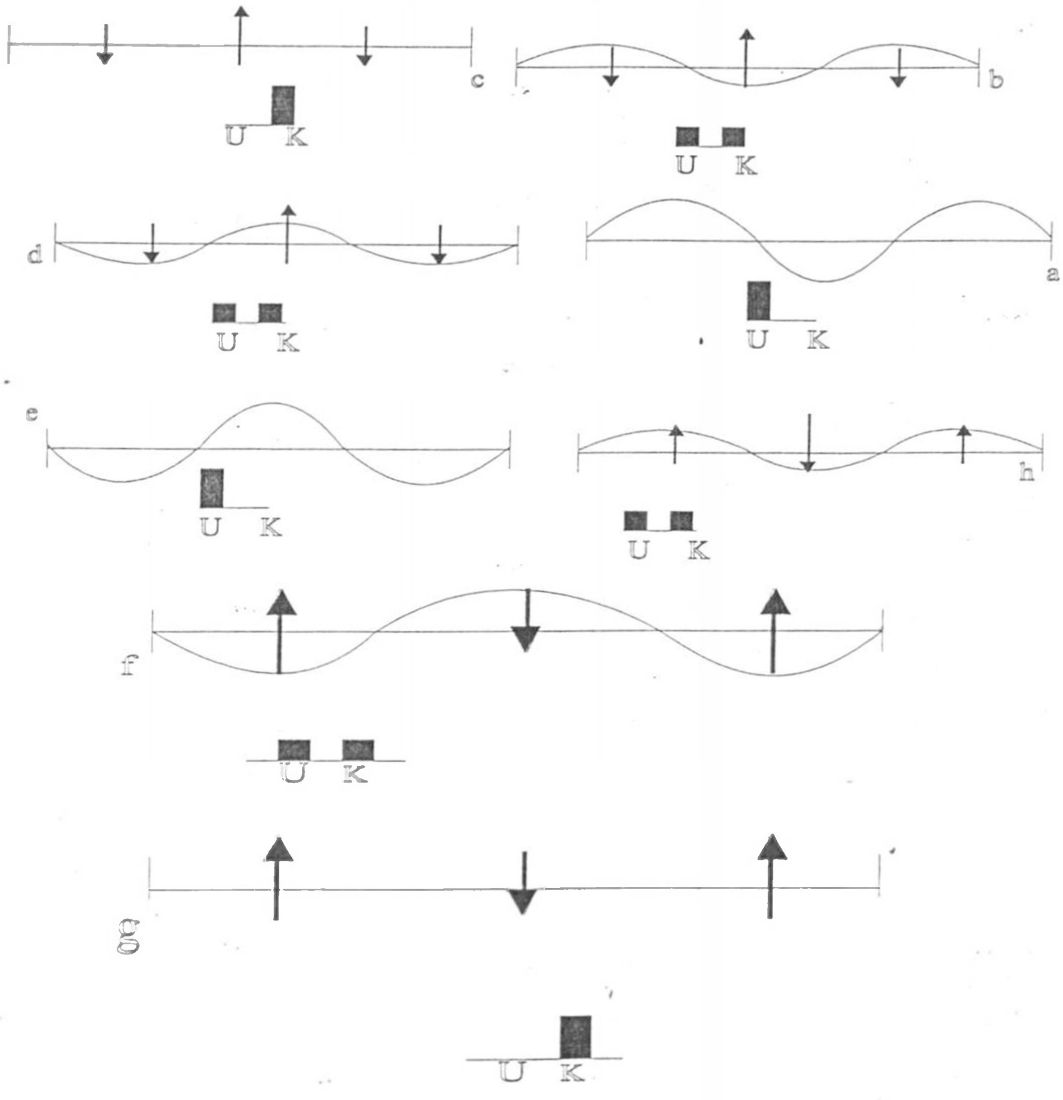


Fig.10

Figure 10 shows how the energy associated with the oscillating string shifts back and forth between kinetic energy of motion K and potential energy of deformation U during one cycle. A similar description can be given for the vibrating spring mass system.

a → all pot. E string momentarily at rest.
arrows → velocities of the string particles at the positions shown
c → string not displaced - but the particles have their maximum speed, energy is all kinetic.

The cycle is completed when the initial condition a is reached.

CHAPTER

RESONANCE

Whenever a body capable of oscillating is acted upon by a periodically varying force having a frequency equal to one of the natural frequencies of oscillations of the body, the body is set into vibration with a relatively large amplitude. This phenomenon is called "resonance" and the body is said to "resonate" with the applied impulses. The phenomenon of resonance can be understood by analysing the forced oscillations. Consider an *ideal mass-spring system* which has a natural frequency of oscillation given by $\omega = 2\pi\gamma = \sqrt{\frac{k}{m}}$, where ω is the angular frequency, k the force constant or the spring constant of the system and m is the mass attached to the spring. If there is friction represented by a frictional force bυ (where υ is the speed), the natural frequency of the spring-mass system is given by

$$\omega = 2\pi\gamma = \sqrt{\frac{k}{m} - \left(\frac{b}{2m}\right)^2}$$

What happens if the system is subjected to an oscillatory external force? The resulting oscillations are called forced oscillations. These forced oscillations have the frequency of the external force and not the natural, frequency of the body. But the response of the body depends on the relation between the forced and natural frequency. If the external force is such that it supplies a succession of small impulses at the proper frequency, the system can be set into oscillations of large amplitude. A child when swinging pumps at proper intervals and builds up a large amplitude. The problem of forced oscillation is useful in acoustic systems, alternating current circuits, atomic physics and also in mechanics.

Let us consider the equation of motion of a forced oscillator. We assume that the external driving force is given by $F = F_o \cos \omega t$ and let x be the displacement. The equation is given by

$$\frac{d^2x}{dt^2} + \frac{kx}{m} = \frac{F}{m} = \frac{F_o}{m}\cos\omega t \qquad (26)$$

We may assume F and x to be complex quantities for the purpose of mathematical analysis i.e., both x and F have a real part as well as an imaginary part. In the end, when the solving process is complete, we just take the real part of the solution. This approach using complex numbers makes the mathematical analysis very simple. For example, we may write $x = x_r + j\,x_i$ and $F = F_r + j\,F_i$. We may substitute for x and F and separate out the real and imaginary parts (two complex numbers are equal only when their real and imaginary parts are separately equal).

Let us now try to solve eqn. (26) by writing it as

$$\frac{d^2 x}{d t^2} + \frac{kx}{m} = \frac{\hat{F}}{m} e^{j\omega t} \qquad (27)$$

where $\hat{F} e^{j\omega t}$ is a complex number. The solution is expected to yield x also as a complex number. When the equation is applied to the case of a forced oscillator, $\hat{F} e^{j\omega t}$ is the driving force having some amplitude, phase and frequency, the frequency being that of the applied force. Let us assume that our solution x is also a complex quantity $x = \hat{x} e^{j\omega t}$. We know when an exponential function is differentiated, we can simply write it as the function multiplied by the simple exponent,

i.e. $\frac{d}{dt}(e^{j\omega t}) = j\omega e^{j\omega t}$

Or $\frac{d}{dt}(\hat{x} e^{j\omega t}) = \hat{x} j\omega e^{j\omega t}$

For a second derivative we multiply the right side again by jω i.e. we get

$\frac{d^2}{dt^2}(\hat{x} e^{j\omega t}) = -\hat{x}\omega^2 e^{j\omega t}$ since $j^2 = -1$.

∴ Our equation becomes

$$-\hat{x} w^2 e^{jwt} + \frac{k\hat{x}}{m} e^{jwt} = \frac{\hat{F}}{m} e^{jwt}$$

or, $\left(-\omega^2 + \frac{k}{m}\right)\hat{x} = \frac{\hat{F}}{m}$

or, $\hat{x} = \frac{\hat{F}/m}{\frac{k}{m} - \omega^2}$. $\qquad$ *Put* $\frac{k}{m} = \omega_o^2$ where ω_o is the natural frequency of oscillation

$$\therefore \quad \hat{x} = \frac{\hat{F}}{m(\omega_o^2 - \omega^2)} \qquad (28)$$

Since $m(\omega_o^2 - \omega^2)$ is a real number, the phase angles of F and x are the same. If $\omega^2 > \omega_o^2$, the phase angles are 180° apart. (When $\omega_o^2 > \omega^2$, $\omega_o^2 - \omega^2$, is +ve ∴ phase same).

We get what we want about resonance from equation (28), which is the solution of forced oscillation. The magnitude of x increases enormously when ω is nearly equal to ω_o and this condition is known as resonance (At resonance we have $\omega = \omega_o$ and x goes to infinity). That is, we *get a strong response when the driving force is applied at the right frequency.* For example, if we have an oscillating pendulum and we give a gentle push each time it comes to one side, we can build up a large amplitude of oscillation. This is what we see normally when a child playing in a swing pumps at proper intervals and builds up a large amplitude. Such forced oscillations are useful in acoustic systems, alternating current circuits, atomic physics and also in mechanics.

Let us now consider a more practical case of a forced oscillator. Our equation (28) tells us that if the frequency ω were exactly equal to ω_o, we would have an infinite response. But in practice no such infinite response occurs because other things like friction or damping limits the response. In our earlier analysis, we had ignored this parameter. Now we add a term to equation (27) to take account of the friction.

What must be the form of this frictional term depends on the problem at hand. However, in many circumstances, *the frictional force is proportional to the speed of the moving object.* An example is the frictional force or viscous force experienced by an object moving slowly in oil or thick liquid. There is no force when the body is not moving, but the faster it moves, the faster the oil has to go past the object and the greater will be the resistance. So we can assume that the resistance term or frictional force term is proportional to the velocity :

$F_f = -c\,\frac{dx}{dt}$. Why negative ? It is a frictional force. For convenience in mathematical analysis, we write the constant c as m times γ. Therefore, our equation of motion is,

$$m\frac{d^2x}{dt^2} + c\frac{dx}{dt} + kx = F \qquad (29)$$

Using $c = \gamma m$ and $\omega_o^2 = k/m$ or $k = \omega_o^2 m$, we get

$$m\frac{d^2x}{dt^2} + \gamma m\frac{dx}{dt} + \omega_o^2 m x = F$$

or,

$$\frac{d^2x}{dt^2} + \gamma\frac{dx}{dt} + \omega_o^2 x = \frac{F}{m} \qquad (30)$$

This equation is in a convenient form for solving.

A *very small value of γ corresponds to very small amount of friction; a large γ amounts to considerable friction.* To solve this equation, we can use the complex number method. We write F as the real part of $\hat{F}\, e^{j\omega t}$ and x as the real part of $\hat{x}\, e^{j\omega t}$. Substituting these into our equation, we find that

$$[\ (j\,\omega)^2\,\hat{x} + \gamma\,(j\,\omega)\,\hat{x} + \omega_o^2\,\hat{x}]\ \ e^{j\omega t} = \left(\frac{\hat{F}}{m}\right) e^{j\omega t} \qquad (31)$$

Dividing both sides by e^{iwt},

$$\hat{x} = \frac{\hat{F}}{m}\ \frac{1}{\omega_o^2 - \omega^2 + j\,\gamma\,w} \qquad (32)$$

Rewriting $\dfrac{1}{[\ m\ (\omega_o^2 - \omega^2 + j\,\gamma\,w)\ ]}$ as R

we have $R = \dfrac{1}{m\ (\omega_o^2 - \omega^2 + j\,\gamma\,\omega\ )}$ and $\hat{x} = \hat{F}\,R$ (33)

Since the factor R is complex, we may write it as $\rho\, e^{j\theta}$. The significance of this can be brought out as follows. Let $\hat{F} = F_o\, e^{j\Delta}$. Therefore, the actual force F

$$F = \hat{F}\, e^{j\omega t}$$

$$= F_o\, e^{j\Delta}\, e^{j\omega t}$$

$$= F_o\, e^{j(\omega t\ -\ \Delta)}$$

is the real part of $F_o\, e^{j\Delta}\, e^{jwt}$ i.e. $F_o \cos(\omega t + \Delta)$. Similarly from (33) we get,

$$\hat{x} = \hat{F}\,R = F_o\, e^{j\Delta}\,\rho\, e^{j\,\theta} = \rho\, F_o\, e^{j\,(\,\theta\,+\,\Delta)}$$

Since the displacement x is the real part of $x\ \ e^{iwt}$, it is given by the real part of

$$\hat{F}\,R\, e^{j\omega t} = F_o\, e^{j\Delta}\,\rho\, e^{j\theta}\, e^{j\omega t}$$

i.e. x is the real part of $\rho\ F_o\ e^{j(\theta+\Delta)}\ e^{j\omega t}$.

Since ρ *and* F_o *are real* this is given by

$$x = \rho\ F_o\ Cos\ (\omega\ t + \Delta + \theta) \qquad (34)$$

Equation (34) can now be interpreted as follows. *The amplitude of the response is the magnitude of the force F multiplied by a certain factor* ρ. It also tells us that x is not oscillating in phase with the force which has the phase Δ. The phase of x is further shifted by an amount θ. Thus ρ and θ can be interpreted to represent the size of the response and the phase of the response.

To get a physical idea of ρ :
To square a complex number, we multiply it by its complex conjugate.

$$R = \rho\ e^{j\theta} = \frac{1}{m\ (\omega_o^2 - \omega^2 + j\ \gamma\ \omega)}$$

$$\therefore\ R^2 = R\ R^* = \rho^2 = \frac{1}{m^2\ (\omega_o^2 - \omega^2 + j\ \gamma\ \omega)\ (\omega_o^2 - \omega^2 - j\ \gamma\ \omega)}$$

$$\rho^2 = \frac{1}{m^2 - [\ (\ \omega_o^2 - \omega^2\ )^2 + \gamma^2\ \omega^2\ ]} \qquad (35)$$

Also, $\frac{1}{R} = \frac{1}{\rho e^{j\theta}} = \frac{1}{\rho}\ e^{-j\theta} = m\ \ (\ \omega_o^2 - \omega^2 + j\ \gamma\ \omega)$

$$\therefore,\ \frac{1}{\rho}\ (cos\ \theta - j\ \ sin\ \theta) = m\ (\ \omega_o^2 - \omega^2\ ) + j\ m\ \gamma\ \omega$$

From this we find, $1\ /\ \rho\ \cos\theta = m\ (\omega_o^2 - \omega^2)$

$$-\frac{1}{\rho}\sin\theta = m\gamma\omega$$

and hence, $tan\ \ \theta = \frac{-\ \gamma\ \omega}{\omega_o^2 - \omega^2}$ (36)

This indicates that θ is negative i.e., the displacement x lags behind the force F by an amount θ. Eqn. (35) can be represented graphically by plotting ρ^2 against frequency ω. (See Fig. 15 a).

ρ^2 which is proportional to the square of the amplitude is also proportional to the energy. For small values of γ, the frictional force constant, the response tends to infinity when ω equals ω_o.

But due to the presence of the term $\frac{1}{\gamma^2 \omega^2}$, the response remains finite. Fig. 15b shows a plot of phase shift θ against frequency.

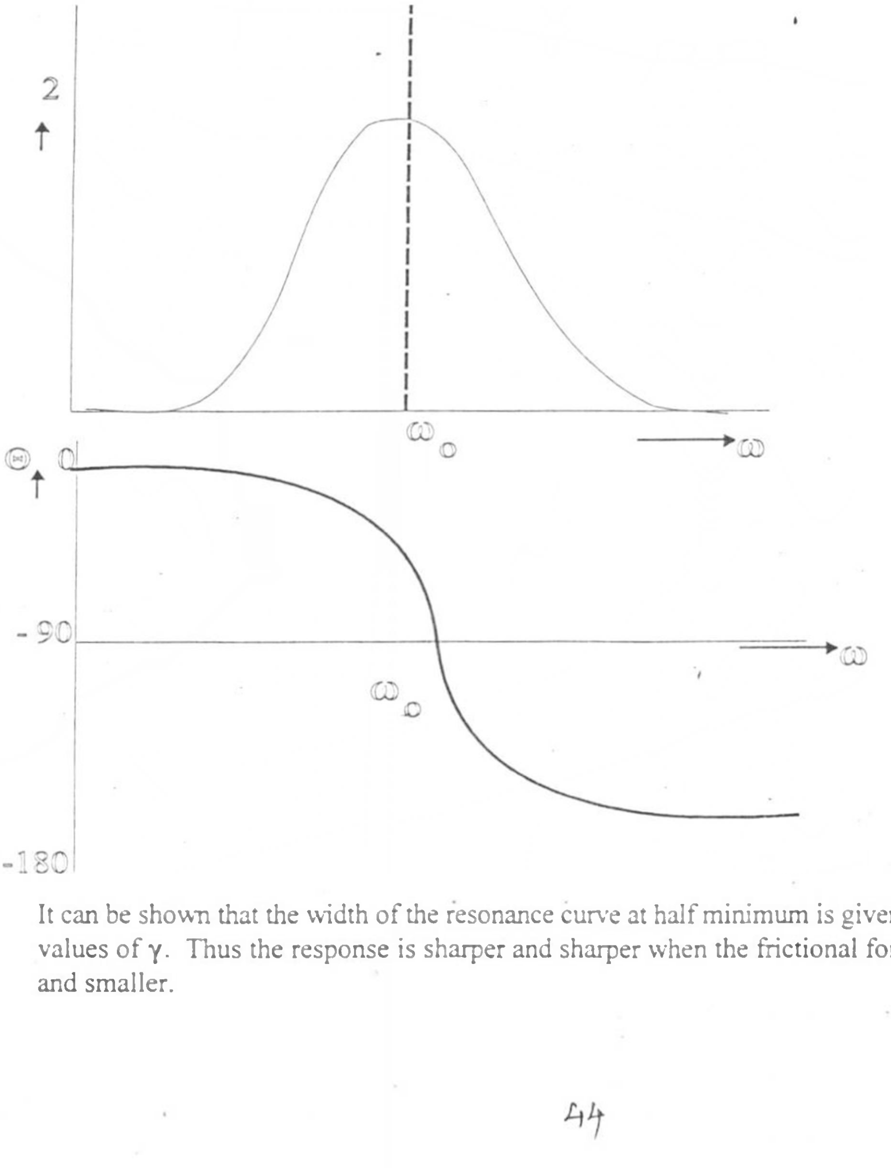


Fig.15a

Fig.15b

It can be shown that the width of the resonance curve at half minimum is given by $\Delta\omega = \gamma$ for small values of γ. Thus the response is sharper and sharper when the frictional forces are made smaller and smaller.

CHAPTER

DIFFRACTION

D.1. What is diffraction ?

Place a source of light on one side of a metal sheet with a hole pierced in it. If a opaque screen is kept on the other side, the area in front of the hole gets illuminated. If the hole is made smaller the area of illumination also becomes smaller. What happens if the hole is made very small? Surprisingly the area of illumination instead of becoming still smaller, starts getting bigger. Of course the intensity will be very small and so it is better observed in a dark room. This phenomenon which is contrary to expectation is due to a phenomenon called diffraction. When a beam of light passes through a narrow opening it spreads out to a certain extent into the region of geometrical shadow and this is due to diffraction. Light suffers deviation from its straight path while passing through narrow openings and while passing close to edges of objects. Some light bends into the geometrical shadow and its intensity there falls of rapidly. If the wavelength of light is smaller than the width of the obstacle of the opening, then the deviation is small. But if the wavelength is comparable, then the bending is appreciable. Diffraction also occurs when light goes over sharp edges of big objects.

If one observes the diffraction pattern formed by a narrow slit kept in front of a monochromatic light, dark and bright bands will be seen in the geometrical shadow. Unlike interference bands these bands are of unequal width.

Newton tried to explain diffraction on the basis of attraction and repulsive forces exerted by the edges on the corpuscles of light. Dr.Young tried to explain it on the basis of Huygen's theory as interference between incident light waves and light waves reflected at grating incidence. But they could not explain why the bands are not of equal width. Later Fresnel gave the correct explanation on the basis that diffraction is due to interference of secondary wavelets originating from various points of the wavefront which is allowed to pass through. These wavelets will have varying phase and amplitude and interference of these wavelets gives rise to diffraction bands.

Diffraction phenomena are divided into two categories.

1. Fresnel diffraction in which either the source or the screen or both are at finite distance from the aperture.
2. Fraunhotter diffraction, in which the source of light and the screen are effectively at infinite distance from the aperture.

D.2 Fresnel Diffraction

Fresnel made the following assumptions while explaining the diffraction phenomenon.

1. The wave front can be divided into large number of small zones called half-period zones. The net effect is combined effect of all these zones.

2. The intensity of the pattern is proportional to the amplitude of the disturbance at the opening.

3. The effect of any particular zone at a point is inversely proportional to the distance of the point from the zone.

4. The effect at a point will depend on the obliquity of the point. The obliquity factor is defined as $(1+\cos\theta)$ where θ is the angle the point makes with the forward direction (See Fig.1).

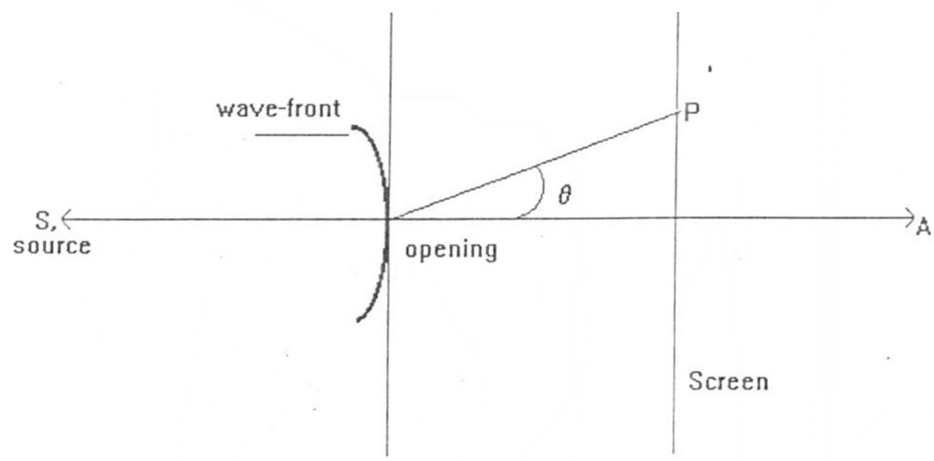


Fig.1 Fresnel Diffraction

Incidentally the dependence on obliquity factor explains why the intensity is zero behind the wave front where $\theta = 180°$.

Fresnel's explanation of propagation of light through an aperture

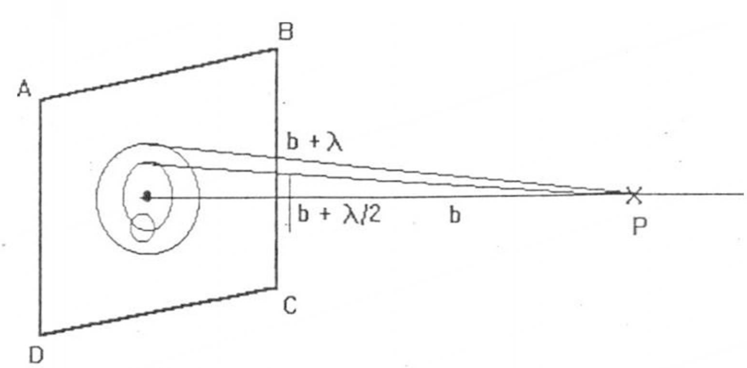


Fig.2 Fresnel's Half Period Zones

Let ABCD be a spherical wave front travelling in the forward direction and P an external point at a distance b from the wave front. Let the wavelength of light be λ and OP be perpendicular to the wavefront. Every point in this wave front can be thought as an origin of secondary wavelets. The secondary wavelets coming from ABCD produce the diffraction pattern on the screen. In order to find the resultant effect at P, we divide ABCD into zones, as follows. Around the point O construct series of circles on the wavefront which are at a distance $b + \lambda/2$, $b + 2\lambda/2$, $b + 3\lambda/2$ .... from P.

The area of the first zone will be approximately equal to $\Pi r_1^2 = \Pi b\lambda$. The area of the second zone will be equal to $\Pi r_2^2 - \Pi r_1^2$, which will also be equal to $\Pi b\lambda$. Observe that area of all the zones will be approximately equal to $\Pi b\lambda$ and the radius of the nth zone will be equal to $\sqrt{nb\lambda}$.

By Huygen's principle every point on the wavefront will be sending secondary wavelets in the same phase. But since their distance from P is different, they will reach P with different phases. Since nth zone is on an average distance $\lambda/2$ father from (n-1)the zone from P, the successive zones will produce resultants at P which differ by Π. This means that successive zones differ by half a period and that is why these zones are called half period zones.

If we represent by A_n the amplitude of the light from the nth zone, the successive values of A_n will have alternating signs because of their phase changing by Π. The resultant amplitude A of the whole wave can be written as

$$A = A_1 - A_2 + A_3 - A_4 \,.......(-1)^{n-1} A_n$$

The magnitude of the successive trains decreases slowly because (a) the amplitude decreases inversely with the average distance from P and (b) increasing obliquity.

The sum of the series can be evaluated as follows. Supposing n to be odd, then

$$A = \frac{A_1}{2} + \left(\frac{A_1}{2} - A_2 + \frac{A_3}{2}\right) + \left(\frac{A_2}{2} - A_4 + \frac{A_5}{2}\right) + \,..... + \frac{A_n}{2}$$

Since the amplitudes of any two adjacent zones are very nearly equal. We can write,

$$A = \frac{A_1}{2} + \frac{A_n}{2} \qquad (odd\ n)$$

It can be easily verified that if n is even then $A = \frac{A_1}{2} - \frac{A_n}{2}$. (even n)

If n is large enough then, $A_n << A_1$ and $A \approx A_1/2$

But if n is small, A will have different values as follows. If the geometry in such that there is only one zone, then $A = A_1$.
If there are only two zones, then $A \approx 0$.

If there are three zones, then $A \approx A_1 - A_2 + A_3 \approx \frac{A_1}{2} + \frac{A_3}{2}$.

This suggests why the intensity of a point P on the axis passes through maxima and minima as the screen is slowly moved away from the aperture.

Of course, it is possible to get the same effect by holding the screen in the same position and altering the size of the aperture.

Example 1: Consider the diffraction produced by a small circular aperture. Discuss the intensity of the diffraction pattern produced at a point P away from the axis.
Hint: Draw altitude PM where M is a point on the sheet having the aperture. Draw half period zones around M. Some zones will pass through the aperture. Resultant effect of these zones gives the intensity.

Example 2 : A circular aperture 1.2mm diameter is illuminated by monochromatic waves. A screen is steadily moved away from the aperture. When the screen is 30 cm. from the aperture, the centre of the patch becomes dark for the first time. Calculate the wavelength of light.
Hint: There will be only the first two zones in the aperture. Answer is 6000 A.

Example 3 : Consider a spherical wave front emitted from a point source of light and incident on a small opaque disc. P is any point on the principal axis on the opposite side. Explain why the diffraction pattern always consists of a central bright spot.
Hint: First few half period zones are cut off by the disc. Resultant of the remaining zones is never zero.

Example 4 : Discuss the diffraction pattern obtained when a thin wire is kept parallel to the slit kept in front of a source of monochromatic light.
Hint: Consider half-period zones of a cylindrical wave front, both above and below the obstacle. Diffraction bands are seen on both the sides.

Diffraction pattern due to cylindrical wire or straight edge etc. are obtained considering cylindrical wave fronts. The wave front is divided into zones and sub zones. . The resultant amplitude is obtained by adding the effect due to all the sub zones. The resultant of the amplitudes of zones and sub zones of wave fronts obtained vectorially gives Cornus Spiral. Its mathematical description gives Fresnel's integrals.

D.3 Fraunhofer Diffractions

Single Slit : Consider the diffraction pattern formed when light is incident on a single slit. In Fraunhoffer diffraction since both the source and the screen are effectively at infinite distance the light wave fronts will be plane and since the screen is finite and small, the obliquity factor and so the amplitude for all the zones will be the same. Consider the following geometrical arrangement (Fig.3).

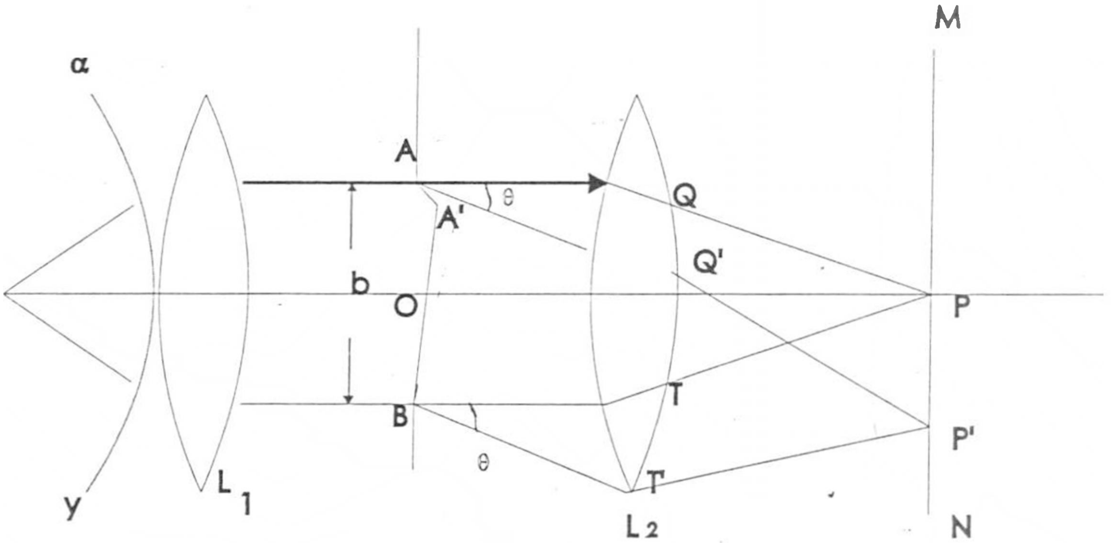


Fig.3 Fraunhoffer Diffraction

AB is a slit of width b spherical wave front XY is incident on a converging lens L_1. The emerging plane wave front passes through the slit. Each point on this wave front acts as source of light. The diffracted beam is converged by the lens L_2 and a pattern is formed on the screen MN.

Secondary waves travelling along the principal axis meet at P on the screen. They travel equal distances AQP and BTP. Secondary waves travelling at angle Θ will meet at P'. Let the incident wave front be divided into large number of strips. The magnitude of amplitude of vibrations for all the zones will be the same but not their directions. The path difference between rays from A and B is bsin Θ and the corresponding phase difference will be $2\alpha = 2\pi/\lambda$ b sin θ where λ is the wavelength of light. The resultant amplitude can be obtained by vector polygon method. Each strip contributes magnitude (a) to the resultant amplitude (Fig.4). The phase difference between successive strips is small.

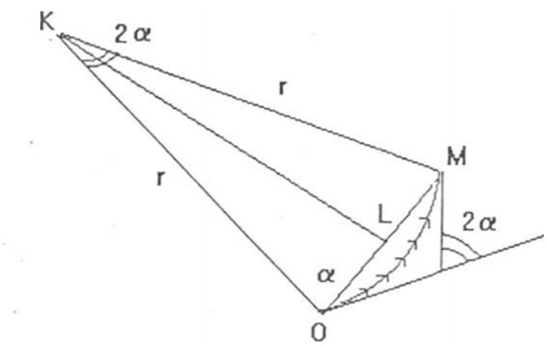


Fig.4 The resultant amplitude

If the number of strips is very large, then the polygon sides will form an arc OM = 2rα = ma.

The resultant amplitude is given by OM = 2r Sinα

$$= \frac{ma}{\alpha} Sin\ \alpha$$

$$= A_o \frac{Sin\alpha}{\alpha}$$

Where ma = Ao and $\alpha = \frac{\pi b Sin\Theta}{\lambda}$

The corresponding intensity will be given by $I = A_0^2 \frac{Sin^2\ \alpha}{\alpha^2}$.

The intensity distribution is given in Fig.5. The maxima and minima positions are as follows:

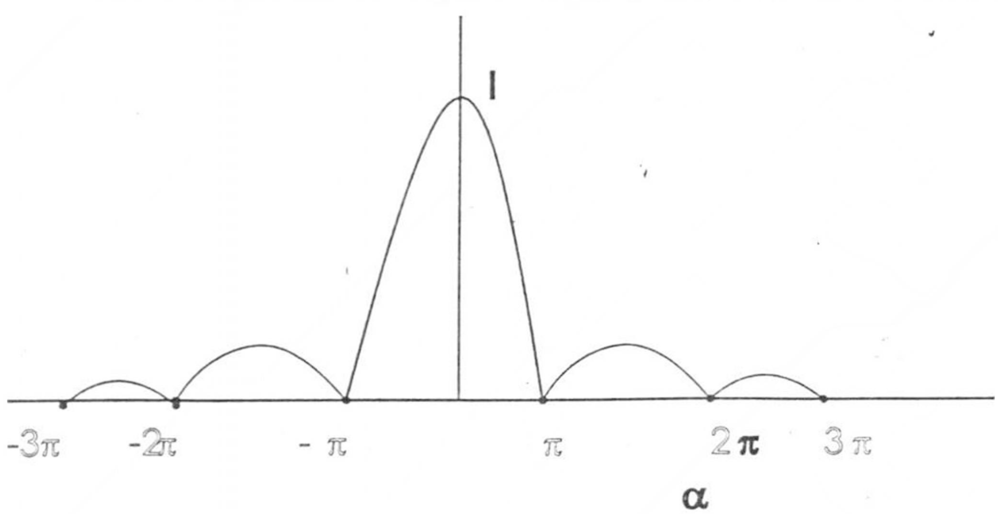


Fig.5 Intensity Variation

Central Maxima: At point P on the screen where $\theta = 0$, $\alpha = 0$.

Secondary maxima : Its direction is given by $Sin\alpha = 1$, $= \frac{\pi\ b\ Sin\ \theta}{\lambda} = \frac{2n+1}{2}\pi$.

Minima : Its direction is given by $Sin\alpha = 0$ or $\alpha = \frac{\pi\ b\ Sin\ \theta}{\lambda} = n\pi\ n \neq 0$.

It is easily verified that intensity of first secondary maxima is about 5% of that of the central maxima.

Example 5 : Calculate the ratio of the intensity of the third maxima to the intensity of the central maxima.

Example 6: Discuss the Fraunhotter diffraction at a circular aperture and show that the radius of the central maxima is equal to $f\lambda/d$ where d is the diameter of the aperture, λ is the wavelength and f is the focal length of the lens used to converge the diffracted rays.
Hint: Proceed as for single slit. The position of the first minima is given by $d\sin\theta = \lambda$ and if x is the radius, then $\sin\theta \approx \theta = x/f$.

Example 7: Discuss Fraunhofer diffraction pattern at a double slit.
Hint: Diffraction pattern is due to two phenomena i) interference of waves coming from corresponding points of two slits and ii) diffraction of secondary waves coming out of each slit.

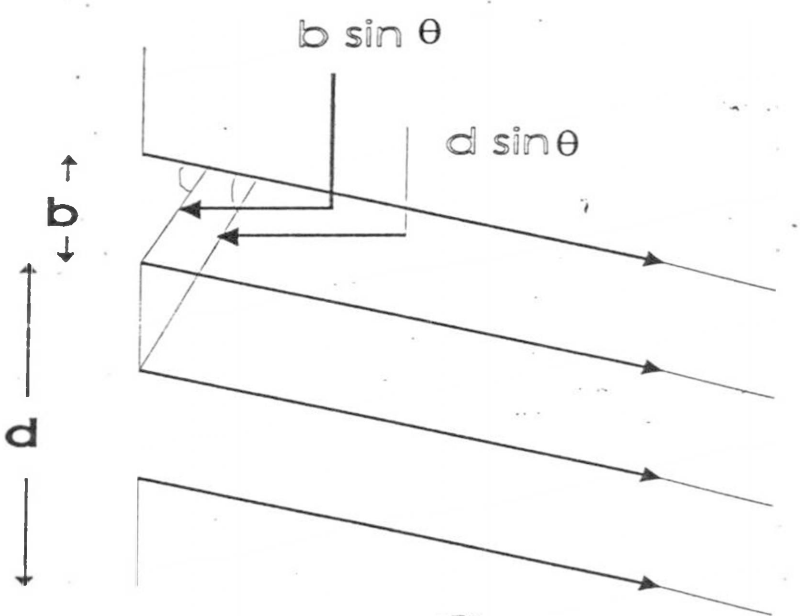


Fig.6
Double slit
diffraction

Due to interference, maxima is obtained where $d\sin\theta = n\lambda$ and minima when

$$d \sin \Theta = \frac{2n+1}{2} \lambda.$$

Due to diffraction, $b \sin \theta = n\lambda$ give minima, $n \neq 0$ and $b \sin \Theta = \left(\frac{2n+1}{2}\right) \lambda$ gives maxima.

The two may be combined to get the resultant intensity as $I = 4 A_o^2 \frac{\sin^2 \beta}{\beta} \cos^2 \Upsilon$ where

$$\beta = \frac{\pi b \sin \Theta}{\lambda} \qquad \Upsilon = \frac{\pi d \sin \Theta}{\lambda}.$$

Interestingly some angles where interference maxima are expected, minima is observed and this is because these positions correspond to diffraction minima also. These are commonly known as missing orders.

D4. Plane Diffraction Grating

It consists of very large number of narrow slits side by side. The common grating used in the class has about 6000 lines per centimetre. When wavefront is incident on the grating light is transmitted through the slits and is obstructed by the opaque portions. The diffraction pattern is due to both diffraction and interference phenomenon. The sharpness of the band increases and tends to become a line when number of slits is increased.

In the arrangement to obtain Fraunhofer diffraction, replace the single slit by a diffraction grating. A plane wave is incident on the grating. AB is the slit and BC the opaque portion of width b and a respectively.

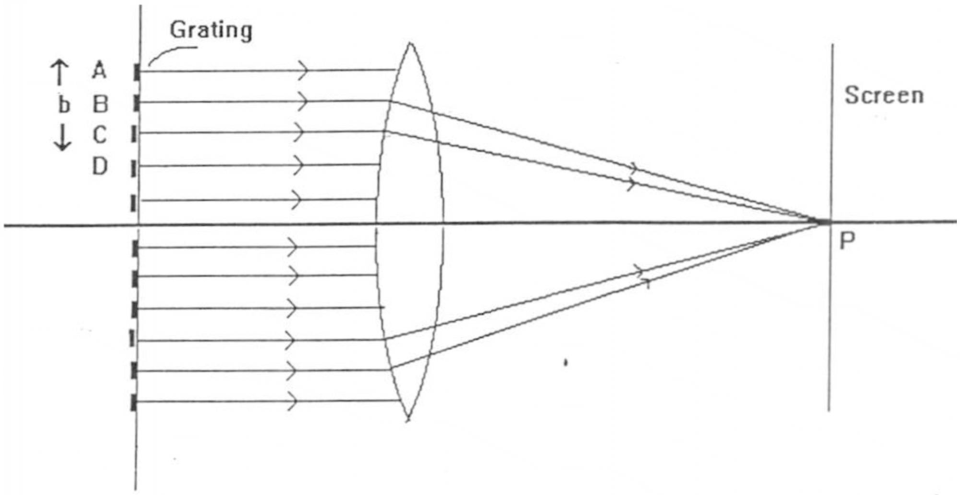


Fig.7 Diffraction grating

The screen is kept at the focal plane of the collecting lens. The point P where all the secondary waves reenforce gives the position of central bright maximum.

Next consider secondary waves travelling at an angle θ with the incident direction. Rotate the lens till its axis is parallel to the direction of secondary waves.

The secondary waves meet at P_1 on the sçreen. A and C, B and D are corresponding points. The intensity at P will depend on the path difference between the secondary waves coming from corresponding points.

∴ Path difference = AC sin θ

= (a+b) Sin θ

The intensity will be maximum if p.d. is integral multiple of λ. So position of maxima is given by $d\sin\theta = n\lambda$ Where $d = a+b$, $n = 0,1,2,....$

n is called the order of the diffraction pattern and d is called the grating space or the grating element.

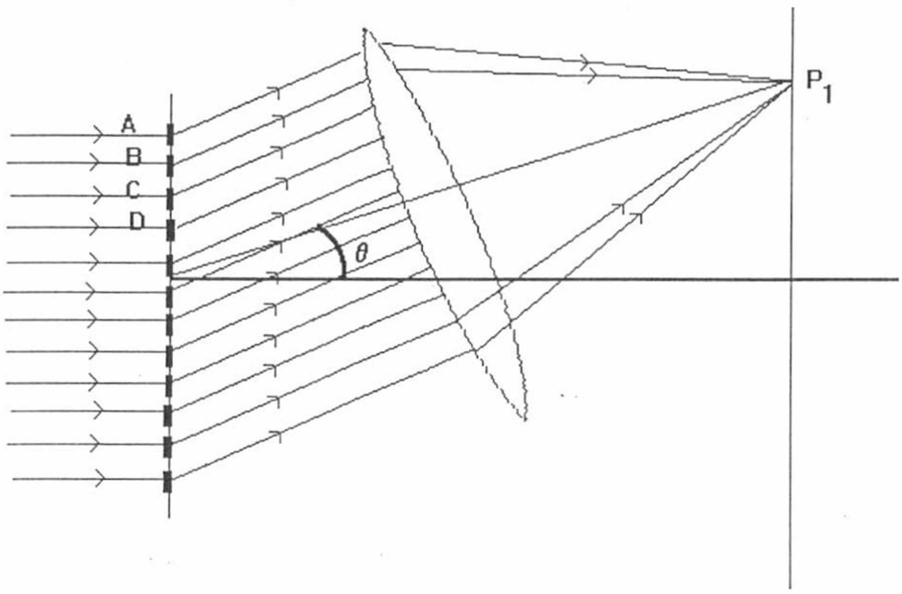


Fig.8 Formation of maxima

Example 8: If N is the total number of lines in the given grating surface show that there are N-1 minima between any two principal maxima.

Hint: Minima is obtained when all secondary waves from top half of the grating cancel the effect of those from lower half. Then the path difference between waves from extreme positions of the grating will be λ. This suggests that path difference for waves from A and C should be λ/N where N is the total number of grating element. Similarly, minima will be obtained when path difference

i) $2\lambda/N$, $3\lambda/N$ ......$(N-1)/N\ \lambda$

Additional Reading

1. A textbook of Optics - Subramanyam and Brijlal
2. Fundamentals of Optics - Jenkins and White.

CHAPTER

. POLARISATION

Light is electromagnetic wave. The electric vector E the magnetic vector B and the direction of propagation are mutually perpendicular to one another. Both E and B vectors vary sinusoidally in identical fashion. So it is sufficient to consider any one of them to describe optical phenomenon. Customarily E vector is considered.

Ordinarily a beam of light travelling in z-direction consists of millions of light waves. The electric vector of these waves vibrate in arbitrary directions in XY plane. Such a light is said to be unpolarised. However, if the electric vector of all the waves vibrate in one direction only, say Y direction, then the beam is said to be polarised in the XZ plane. In the diagram, it is polarised in a plane normal to the plane of the paper. The vibrations are represented by double headed segments in the YZ plane.

On the other hand, if the vibration are in X-direction only, they are represented by dots and the light is said to be polarised in YZ plane. There is no vibrations in the plane of polarisation.

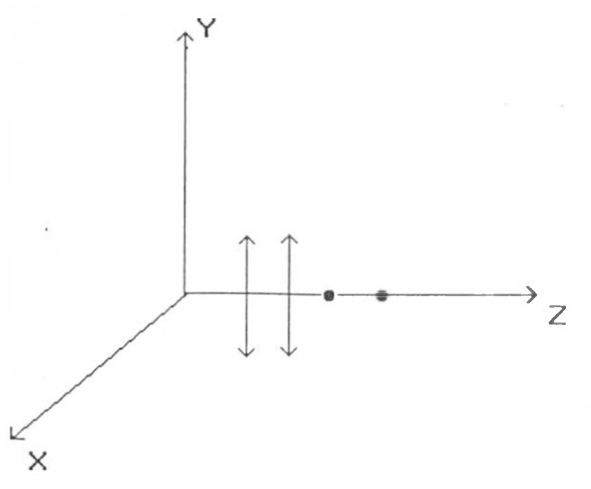


Fig 1 Vibrations in the plane of the paper (↕) and normal to the plane of the paper (.)

Analytical Treatment

A mathematical analysis of the vibrations of the E vector leads to further classification of the polarised light.

Consider a beam of light propagating in the z-direction. The E-vector of different waves will be vibrating in the XY plane in all possible orientations. These vectors can be resolved intocomponents along X and Y. Their resultant can be written as

$\Sigma X = E_x = A_x \cos wt$

and $\Sigma Y = E_y = A_y \cos (wt + \delta)$

where A_x and Ay are the amplitude of the resultant components along X and Y directions. Consider the following cases.

1. If $A_x = Ay$ and δ is $\pi/2$, then $E_x = A_x \cos \omega t$. $E_y = A_x \sin \omega t$

So, $E_x^2 + E_y^2 = A_x^2$

This represents circularly polarised light. The tip of the resultant vector will describe a circle as light propagates. In this case, if the resultant vectors are projected on a screen normal to direction of propagation the tips will describe a circle.

2. If $A_x \neq Ay$ and δ is $\pi/2$, then, $\frac{E_x^2}{A_x^2} + \frac{E_y^2}{A_y^2} = 1$

This represents elliptically polarised light , the tips of the resultant vector will describe an ellipse as the wave propagates.

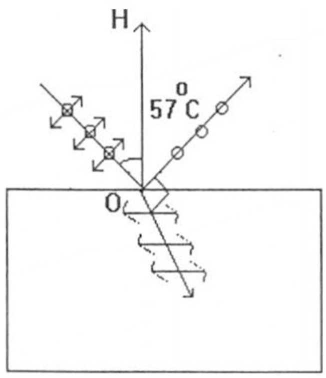
3. If $A_x \neq Ay$ and δ is $n\pi$ then $E_x = \pm A_x / Ay\ E_y$
4. If there is no definite phase relation between A_x and A_y , then the light is unpolarised.

This represents plane polarised light. The tips of the resultant vectors will describe a straight line.

Polarisation by Reflection

When unpolarised light falls on a glass plate, part of it is reflected and part of it is transmitted. Malus observed in 1808 that the reflected light is partially polarised. The degree of polarisation depends on the angle of incidence. The reflected beam is plane polarised when the incident angle is 57°.

Later in 1812, Brewster observed when the polarisation by reflection is maximum, then the reflected beam and the refracted beam are at right angles.



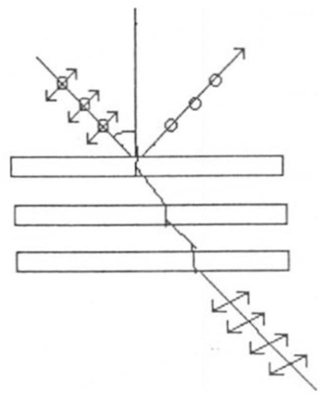


Fig.2 Brewster's Law

Fig.3 Polarisation by a pile of glass Plates

By Snell Law $\mu = \frac{\text{Sin } 57^\circ}{\text{Sin r}}$ and now $\angle 57^\circ + \angle r = 90^\circ$

So we get $\mu = \frac{\text{Sin } 57^\circ}{\text{Cos } 57^\circ} = \tan 57^\circ$

So the tangent of the polarising angle gives the refractive index of the material.

The refracted beam is also partially polarised and this polarisation can be increased by using large number of glass plates kept in parallel (Fig.3).

Example 1 : When plane polarised light falls on a quartz crystal, it is broken up into two beams of light whose E vectors are at right angles to one another. These beams then propagate through the crystal and interfere. If the refractive indices for them are 1.55 and 1.54, what should be the thickness of the crystal for the outcoming beam to be 1. Plane polarised, 2. Circularly polarised. Assume that their amplitude of vibrations are equal.

3. If in the above problem amplitude of vibrations are unequal, what changes do you expect?
Hint: To be plane polarised or circularly polarised, phase difference should be π or $\pi/2$ or the path difference should be λ/q or $\lambda/4$ respectively. Optical path difference = $\mu_1 t - \mu_2 t$.
So, $t(\mu_1 - \mu_2) = \lambda/2$ or $\lambda/4$ respectively.

Example 2 : Prove that when light falls on a plane parallel glass plate at its polarising angle, the refractive beam falls on the second face of the plate also at the polarising angle.

Law of Malus

Consider a beam of plane polarised light coming out of a polariser. Let the angle between this incident beam and the plane of transmission of the analyser be θ. The intensity of the light transmitted by the analyser varies as $\cos^2\theta$, (Fig.4).

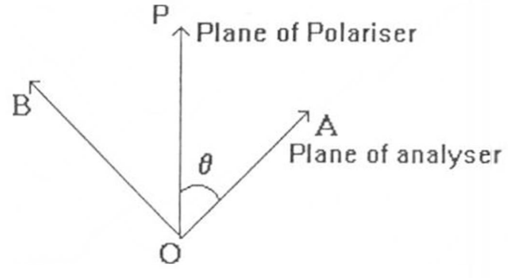


Fig.4 Malus law

Let OP be the plane of incident light of amplitude R. The amplitude can be resolved along the plane of the analyser OA and at right angles to it OB. The component along OA is transmitted by the analyser and that along OB is blocked. So, the amplitude of light passing through $A_1 = R\cos\theta$ or $I_1 = A_1^2 = R^2 \cos^2\theta = I_0 \cos^2\theta$. Malus law states $I_1 = I_0 \text{Cos}^2 \theta$.

Example 3: What is the angle between the analyser and polariser if the intensity of transmitted light is 25% of that of the incident light?

Polarisation by scattering

When sunlight falls on water molecules, dust particles, etc. the electric vector of the light acts on the positive and negative charges of these molecules. As a result, the positive and negative charges deviate along opposite direction and dipoles are formed. The electric vector changes sinusoidally and so the dipole moment also oscillates accordingly. So the dipole oscillates with the frequency of the electromagnetic waves. This effect is more in the direction of the E vector and less in other directions. So the intensity of emitted light will be more along one direction or the light is polarised.

Optical Activity

Consider a beam of light incident on a polariser. When the polariser and the analyser are crossed no light emerges out of the analyser. But if a sugar solution is kept between the analyser and the polariser, some light emerges out of the analyser. Sugar solution rotates the plane of polarisation. The amount of rotation depends upon the length and concentration of the sugar solution.

Fresnell Explanation of Rotation

A linearly polarised light can be considered as a resultant of two circularly polarised vibrations rotating in opposite directions.

If the two rotations travel with same velocity, then there is no optical rotation.

If the clockwise component travels faster then the rotation of the plane of polarisation is to the right and the optical material is called right handed or dextro rotatory. If the anticlockwise component travels faster, then the optical material is left handed or laevo rotatory. Quartz crystals are dextro rotatory whereas sugar solution is laevo rotatory. Calcite does not produce optical rotation. The amount of rotation of the plane of polarisation in sugar solution depends on (1) wavelength, (2) concentration of the solution, (3) length of the solution and (4) temperature.

Analytical Treatment

Consider the optical rotation in quartz crystal. The incident plane polarised light is broken up into two circularly polarised light.

The clockwise circularly polarised vibration has components

$x_1 = a \cos (wt + \delta)$

$y_1 = a \sin (wt + \delta)$

The anticlockwise circularly polarised vibration has components

$x_2 = -a \cos wt$

$y_2 = a \sin wt$

The resultant displacement is given by

$X = x_1 + x_2 = a \cos (wt + \delta) - a \cos wt$

$= 2a \sin \delta/2 \sin (wt + \delta/2)$

and $Y = y_1 + y_2 = a \sin (wt + \delta) + a \sin wt$

$= 2a \cos \delta/2 \sin (wt + \delta/2)$

Both X and Y have same phase = $(wt + \delta/q)$. They are at right angles and their amplitudes are different. So their resultant is plane polarised and it makes an angle $\delta/2$ with the original direction.

In the above analysis of δ is zero, then the resultant vibration will be given by X=0 and Y= 2 a sin wt and is the resultant polarised beam is not rotated.

Example: Using a polarimeter tube, 30cm long and containing sugar solution, the plane of the polarisation was rotated by 12°. If specific rotation of sugar solution is 60°, estimate the strength of the sugar solution.

$$S = \frac{10\theta}{lC} \quad \text{by definition}$$

$$\text{or } \frac{1}{C} = \frac{Sl}{10\theta} = \frac{60 \times 30}{10 \times 12}$$

or C = 0.067 g/cc

Some crystals are optically active, some are not. It is found that crystal whose lattice is same as its mirror image is not optically active. E.g. cubic crystal. Crystal whose lattice is not the same as its mirror image is optically active e.g. rhombic crystal.

CHAPTER

LASER

LASER is a short form for Light Amplification by Stimulated Emission of Radiation. It is a very special source of light having the following characteristics.

1. **Highly directional**: Electromagnetic waves propagate along a particular direction. The spread of the beam, even after travelling large distances, is extremely small.

2. **Highly intense** : The brightness of a given source of light is the power emitted per unit area of the surface per unit solid angle. An ordinary light emits light in all directions whereas laser is highly directional which makes it highly intense.

3. **Highly monochromatic**: Only electromagnetic waves of a particular frequency gets amplified and emitted. There is no spread in the specified frequency. If the wavelength is specified as 6000A it remains the same always. It will not be 6001 A or 5999 A.

4. **Highly Coherent**: It is coherent in space and coherent in time. Spatially coherent means that at any time light has some phase everywhere across the wavefront. Time coherence, implies that phase of all the waves even after they have traveled for a time t is the same.

Production of Laser

Two fundamental processes namely absorption and emission are involved in the production of laser as detailed below.

Absorption

Consider atoms and molecules having energy levels E_1 and E_2 . If they are irradiated by electromagnetic waves of energy $E_2 - E_1$, then they absorb the incident energy and get excited to higher level E_2. The excited states are unstable and so these molecules will return to ground state by emitting radiation of frequency $(E_2 - E_1)/h$. This emission may be spontaneous or stimulated.

Emission

In spontaneous emission, the excited atoms return to the ground state in about 10^{-8}s. The emitted radiation will not have correlation either in phase or in direction. In stimulated emission, the photon of energy $(E_2 - E_1)$ is incident on excited molecules. All the excited molecules will then return to its lower state by emitting frequency $(E_2 - E_1)/h$. The direction and phase of the emitted radiation will be same as that of incident radiation.

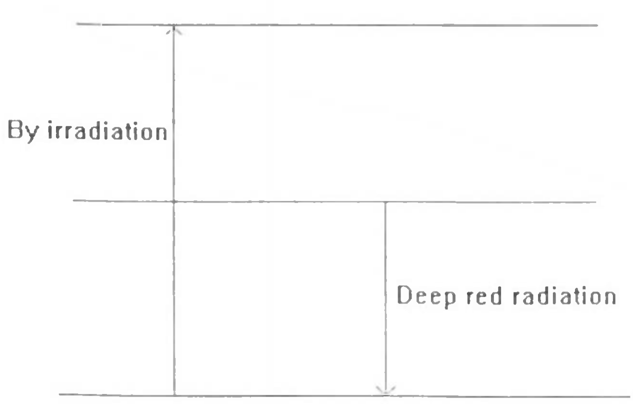
In normal conditions, number of molecules in lower energy level will be more than those in the upper level i.e. population of lower level is more than that in the upper level. So when external radiation falls on the molecules, the molecules in the ground level will get excited to higher level and molecules in the upper level will return to the ground level. So under normal conditions number of molecules going to upper level will be more than those returning to lower level. Initially absorption predominates emission and this will continue till the processes of absorption and emission compensate each other. So if the intensity of emitted radiation has to be more than the number of molecules in upper level should be more or there should be population inversion. So, in order to produce laser, it is essential to produce population inversion in the sample. Of course, it is also essential that the external radiation interacts with the sample. When radiation of frequency $(E_2 - E_1)/h$ falls on the molecules whose population is inverted, then the molecules emit radiation of same frequency. The intensity of this radiation is then increased manyfold by multiple reflections. In what follows a brief description of the working of a solid-state laser i.e. ruby laser and that of a gas-laser i.e. He-Ne is given.

Ruby Laser

Ruby consists of aluminium oxide in which a few of aluminium atoms are replaced by chromium atoms. The normal aluminium oxide is a colourless material. Substitution of chromium makes it red.

When ruby is irradiated by external radiation Cr gets excited to state E_1 from ground state G.

Cr gives some of its excitations energy to do Al by thermal transfer and all of them come to slightly lower level E_2. This state E_2 is a meta stable state of Cr. So Cr can stay in this state for longer time interval (10^{-5} s). In this interval more and more Cr ions will come to E_2 and thus population inversion is achieved.



E_1 Al + Cr only Cr is excited.

E_2 Cr loses energy to Al and comes to lower state E_2 (meta stable state)

Deep red radiation (E_2 - G)

G Al + Cr in ground state

Fig.1 Energy level in ruby crystal

From E_2, Cr gives out deep red radiation. Normally this radiation is spontaneous and not coherent. To make it more intense and coherent an external radiation of the same frequency is applied and the excited atoms are stimulated to give out intense and coherent radiation. Its working is briefly illustrated below.

Consider a ruby crystal cylindrical in shape (R) exposed to neon flash lamp and kept between a pair of parallel reflecting mirrors. The excited Cr atoms give up excess energy by spontaneous emission. The radiation is emitted in all directions and some of them will be normal to the reflecting mirrors. These later photons start the laser action.

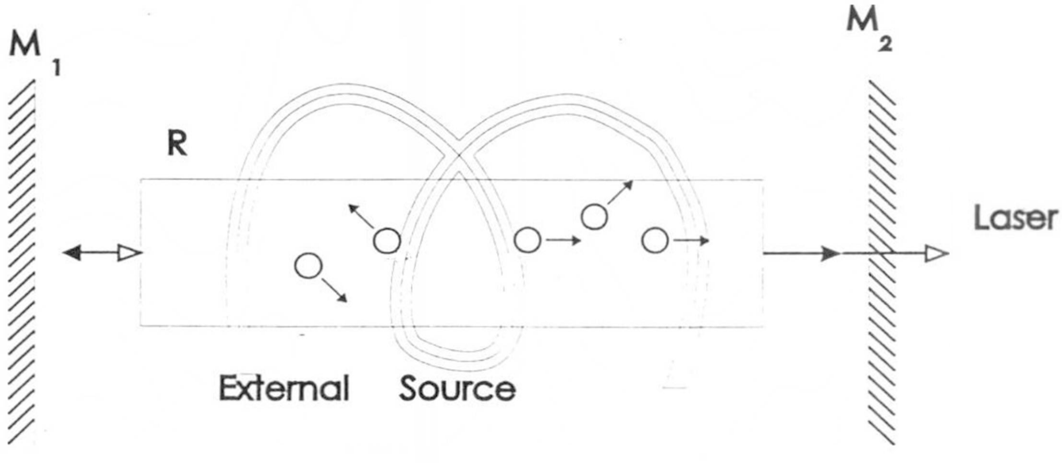


Fig.2 Working of a Ruby Laser

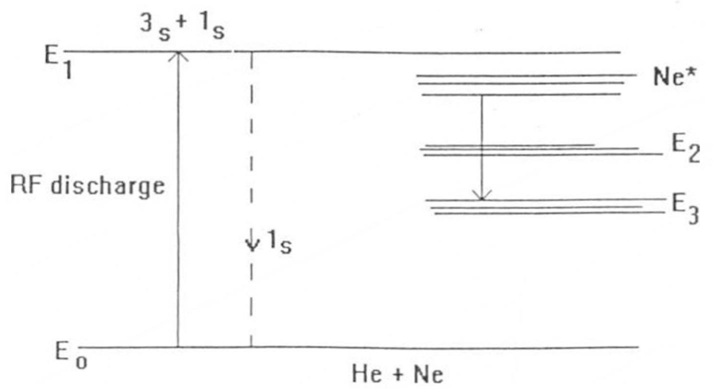
These photons travelling normal to the mirrors get reflected repeatedly at M_1 and M_2 and increase their number. The reflected light acts as a stimulant and successive reflections add to the intensity in a particular direction. The intensity of light emitted in other directions will be small or negligible. One of the mirrors is partially silvered which enables the light to come out in the form of laser beam.

In actual practice, mirrors are replaced by the polished ends of the ruby rod. These polished ends are coated with reflecting material.

Gas Lasers

It is essential that the gas molecules absorb external radiation and get their population inverted. Since the absorption band of gas molecules is narrow as compared to that of solids, it is more difficult to develop gas lasers than solid state lasers. Since the chances of absorption of external radiation is less, it is essential that the exciting radiation be very intense.

In a typical helium-neon laser, the sample consists of 90% He and 10% Ne. The external radiation from RF discharge, raises He from ground level E_0 to excitation level E_1. The excited state will have He both in singlet state and triplet state. The singlet state electrons readily come down to ground state which is also a singlet state. But the triplet state electrons are forbidden to come down to singlet states. These triplet state He exchange energy by collision with the ground state Ne atoms.

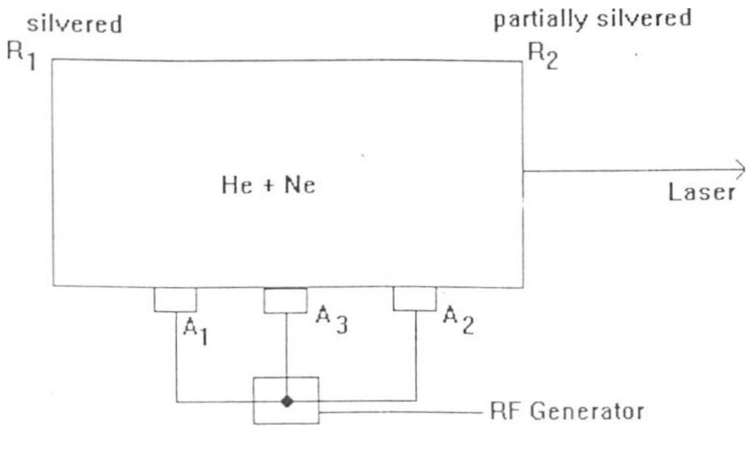


It so happens,

$He^* + Ne \rightarrow Ne^* + He$

that excitation level of Ne is very close to excitation level of He. Because triplet state of He is a metastable state, the population of excited Ne goes on increasing. Besides, this excited state E_1, Ne has lower excited states E_2, E_3, etc. By applying external radiation of frequency $(E_2 - E_3)/h$, laser is produced.

Fig.3 Energy levels of He - Ne mixture



In this case, the energy is provided by RF discharge and He acts only as an intermediary agent.

The construction of He+Ne laser is briefly described in Fig.4.

Fig.4 He-Ne Laser

Electrodes A_1 and A_2 are used to excite He. Electrode A_3 is used to stimulate excited Ne to produce laser. R_1 and R_2 are polished and silver plated. R_2 is partially silvered which allows laser beam to come out.

Laser Applications

1. **Scientific tool**

   It provides highly monochromatic and intense beam of light for use in research and technology. Raman spectra which needed many hours of exposure previously can now be obtained in few minutes using laser beam as incident light.

   Raman effect experiments can also be used to provide laser beams of slightly different frequency than that of incident laser beam.

2. **Hollowgraphy**

   In hollowgraphy where one observes three dimensional images, it is necessary that the light used to obtain the image and the light used to observe the image should have exactly same frequency and are coherent and intense. These conditions are met if lasers are used as light sources.

3. **Calibration**

   Since the wavelength and frequency are very well defined it is used in defining the standard of length.

4. **Biology**

   Individual cells can be destroyed by using lasers and the effect of these cells in the biological system can be easily studied.

5. **Medicine**

   Laser is of great help in surgical operations. Operation can be done without piercing the body by surgical instruments and this solves the problem of sterilisation and infection. Laser beam can be used both for cutting and for destroying the human tissues, tumors, etc.

6. **Communication**

   The laser beams are very intense and have practically no dispersion. This facilitates RADAR communication. They are also increasingly used in fibre optics communications.

7. **Industrial use**

   Lasers can make holes in metal blocks in few seconds. It is, therefore, very useful in industrial cutting. Lasers can be focussed to very small region and this helps in point welding in electronic circuits.

CHAPTER
GAUSS'S LAW

Charles Hugustin Coulomb (1736-1806) measured electrical attractions and repulsions quantitatively and deduced a law known as 'Coulomb's Law'. Two charges q_1 and q_2 separated by a distance 'r' found to exert a force resulting a twist in suspended torsion balance. The force of repulsion was found to be proportional to the product of two charges and inversely proportional to the square of the distance between them. This force acts along the line joining the two charges. It is interesting to note that the charge q_1 sets up an electric field in the space around itself and this electric field acts on the charge q_2 and the charge q_2 experiences a force. Thus the electric field acts as an intermediary role in the forces between charges. If the charge q_1 suddenly moves it will create a field disturbance which will be immediately communicated to the charge q_2. We will see later when we deal with electromagnetic waves that such electric disturbances will move at the speed of light.

In this chapter we will discuss electric field around a point charge and electric potential and apply the knowledge to capacitors and dielectrics. The electric field is generally represented by imaginary lines of force. The line of force and electric field vector E are such that (i) the tangent drawn to the line of force at any point gives the Direction of E at that point and (ii) the lines of force are drawn such that number of lines per unit cross sectional area perpendicular to the lines is proportional to the Magnitude of E. When the lines are close to each other, E is large and when they are apart E is small.

Karl Friedreich Gauss (1777-1855) a German scientist and mathematician made a number of scientific contributions to experimental and theoretical physics. His well known contribution is known as 'Gauss's Law' which is a statement of an important property of electrostatic fields.

1. Consider a field E of a single isolated point charge q. This charge is surrounded by a hypothetical closed surface of arbitrary shape. The field intensity E at every point on the surface is directed radically outward from the point charge and its magnitude is

$$E = k\left(\frac{q}{r^2}\right) \qquad (1)$$

over a small area ds of the surface. This area is so small it will have the same field in magnitude and direction. The component normal to the surface can be written as

$$E_n = E\cos\theta$$

Where θ is the angle between vector E and the outward normal to the surface.

$$E_n\, ds = E\cos\theta\, ds = kq\left(\frac{ds\cos\theta}{r^2}\right) \qquad (2)$$

This ($ds\cos\theta$) is the projection of the area ds at right angle to 'r' and the quotient ($ds\cos\theta / r^2$) equals to the solid angle $d\omega$ subtended at the charge q by the area ds. Refer fig. (1).

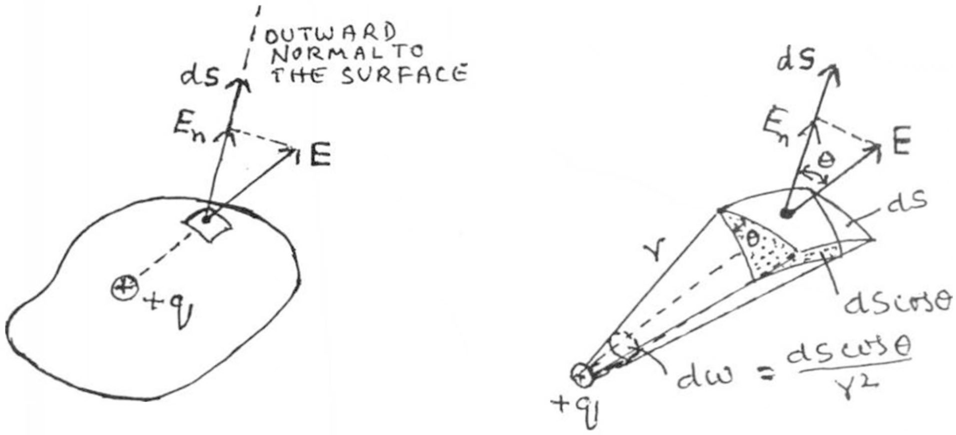


Fig.1 Electric field surrounded by a closed surface

$$E_n \; ds = kq \, d\omega \qquad (3)$$

Integrate both sides over the entire closed surface.

$$\oint E_n \, ds = kq \oint d\omega = 4 \pi kq \qquad (4)$$

The left hand side of this equation formed by multiplying the normal component of E at the surface by an element area of the surface and is called the *SURFACE INTEGRAL OF E OVER THE SURFACE. THE EQUATION POINTS OUT THE SURFACE INTEGRAL IS PROPORTIONAL TO THE ENCLOSED CHARGE q REGARDLESS OF THE SHAPE AND SIZE OF THE SURFACE OR THE LOCATION OF THE CHARGE.* If the point charge is negative the direction of E is inward and the angle would be greater than 90° and its cosine will be negative.

(2) If the charge is distributed inside a closed surface then it could be subdivided in imagination, into point charges q_1, q_2 and q_3.......etc. and the equation could be written for each point charge and sum up over all charges. The sum of the integral becomes the surface integral of the resultant field and the charge become Σq the algebraic sum of all charges inside the closed surface.

$$\oint E_n \; ds = 4 \pi \kappa \; \Sigma q \qquad (4)$$

THE SURFACE INTEGRAL OF THE COMPONENT OF E OVER ANY CLOSED SURFACE IN AN ELECTROSTATIC FIELD IS EQUAL TO 4πk TIMES THE NET CHARGE INSIDE THE SURFACE.

In $\oint E_n \; ds = 4 \pi k \; \Sigma q$ the constant k when expressed in terms of permitivity constant ϵ_o is given by

$4\pi k = 1/\epsilon_o$ and the value of $\epsilon_o = 8.8541878 \times 10^{-12} C^2 N^{-1} m^{-2} \approx 9 \times 10^{-12} \; C^2 N^{-1} m^{-2}$

The above equation can be rewritten as

$$\oint E_n \; ds = q / \epsilon_o \; \ldots\ldots\ldots\ldots\ldots\ldots \quad .(5)$$

The surface integral of E over a closed surface is called ELECTRIC FLUX $\phi_{E.}$

$$\phi_E = \oint E_n \; ds = q / \epsilon_o \ldots\ldots\ldots\ldots\ldots\ldots \quad (6)$$

THEREFORE THE ELECTRIC FLUX LINKED WITH A CLOSED SURFACE IS PROPORTIONAL TO THE CHARGE ENCLOSED BY THE CLOSED SURFACE.

$\epsilon_0 \oint \vec{E} . d\vec{s} = q$ is called the GAUSS'S LAW.

This law is one of the fundamental equations in electromagnetic theory and we will see more of it when we deal with Maxwell's equations.

3) Consider a case of two equal and opposite point charges as shown in fig (2). The dashed lines represent the intersections with the plane of the figure of the hypothetical closed Gaussian surfaces. The flux Φ_E is positive where the positive charge is enclosed by a closed surfac$_1$ S It is negative when negative charge is enclosed by the Gaussian surface S_2.

What would be the value of flux Φ_E at enclosed surfaces S_3 and S_4? Give reasons for your answers.

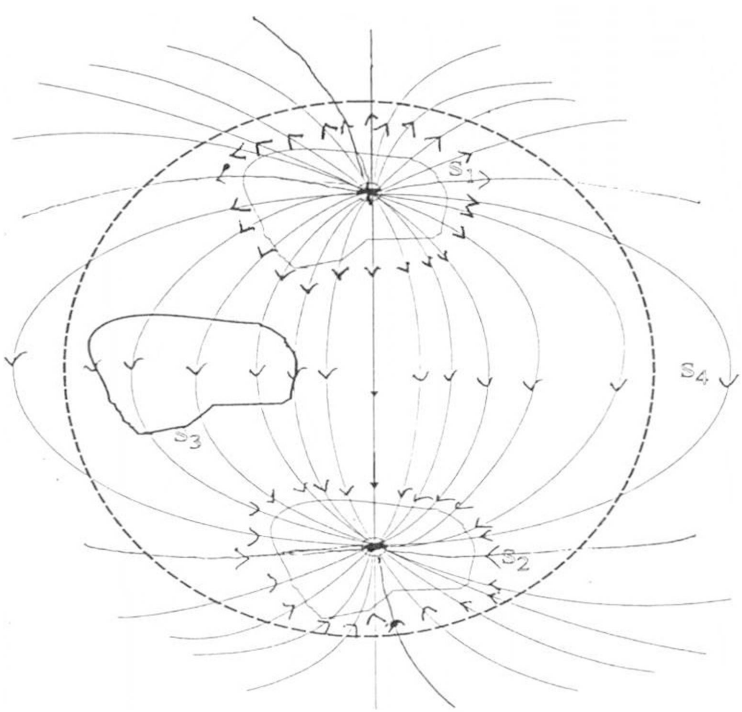


Fig.2 Electric flux in the case of two equal and opposite charges

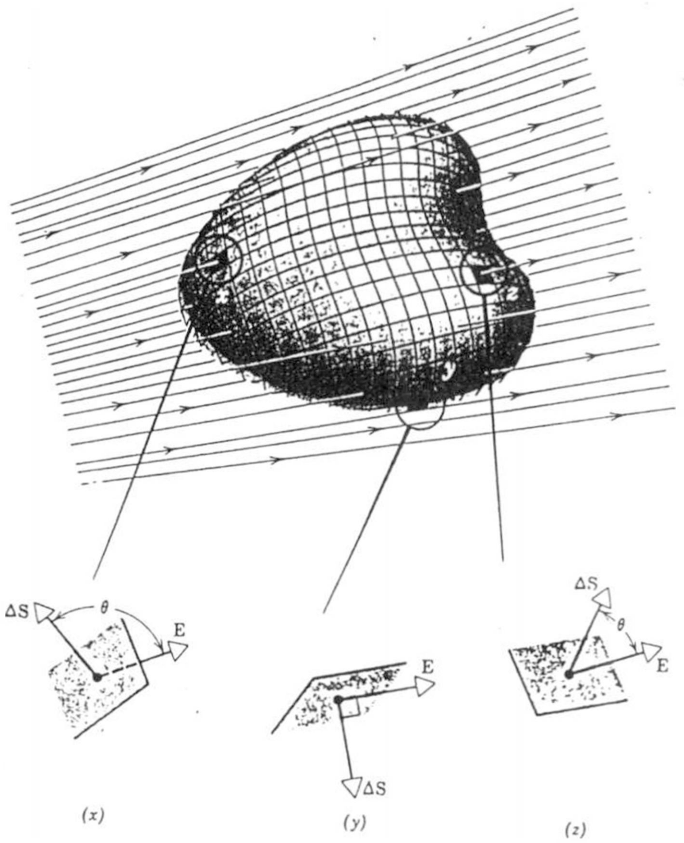


Fig.3 Flux in the case of a non uniform electric field

4) Now consider a case of non uniform electric field as given fig (3). Let the enclosed surface be divided into a large number of element squares ΔS as shown in this figure. These element squares are small enough to be considered as plane surfaces. Such element areas may be considered as vectors whose magnitude will be ΔS and the directions are taken as outward drawn normal to the surfaces as shown in fig (3b).

At every square element, one can construct an electric field **E.** Since the square is very small at all points on the square, $\vec{E}$ may be considered as constant.

The vectors Δ **S** and **E** that characterize each square make an angle with each other. Refer fig (3b). We have selected three such areas and magnified them. Note that when the angle $\theta > 90^0$, **E** points in and when $\theta = 90^0$ it is parallel to the surface and when $\theta < 90^0$ the field **E** points out.

In the above case one can write a semiquantitative definition of electric flux as

$$\phi_E \cong \Sigma\, \mathbf{E}.\Delta\, \mathbf{S} \text{ ....(7)}$$

which instructs to add the scalar quantities of **E**.Δ **S** for all element areas into which the surface has been divided. For all purpose the surfaces where θ is less than 90^0 it is positive and **E** is outward. If **E** is everywhere inward, θ will be more than 90^0 , then **E** .Δ **S** is negative and flux ϕ_E for the surface will also be negative.

The exact definition of flux is given by the differential limit of the equation.

$$\phi_E \cong \Sigma\, \mathbf{E}.\Delta \mathbf{S} = \Sigma |E| \Delta S \cos\theta \text{ ....(7a)}$$

Replace the sum over the surfaces by an integral over the entire surface.

$$\phi_E = \oint \mathbf{E}.\Delta\, \mathbf{S} \text{.................................(8)}$$

The surface integral suggests that the surface is considered as divided into very large number of infinitesimal elements of area ds and the scalar quantity of **E**.d**S** is to be evaluated for each element and the sum is to be taken for entire surface. The surface integral also shows that it is a closed surface.

The S.I. unit of electric flux is

$$|\phi_E| = \left(\frac{newton \times (meter)^2}{coulomb}\right) = \left(\frac{joule \times meter}{coulomb}\right)$$

5) Consider a case of spherically symmetric charge distribution as shown in fig (4). Let R be the radius of spherical distribution of charge and ρ the charge density which is the charge per unit volume (Cm^{-3}).

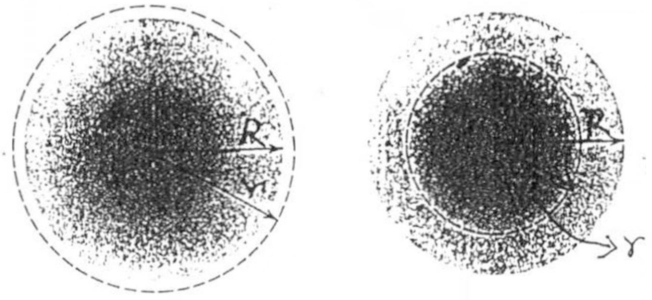


Fig.4 A spherically symmetric charge distribution

This charge density depends only on the distance of the point from the centre. Let us find an expression for E for points (i) outside and (ii) inside the charge distribution.

(i) Draw a Gaussian surface at a distance 'r' which is greater than R. From the equation (4) one can write

$$\epsilon_o \oint \mathbf{E} \cdot d\mathbf{S} = \epsilon_o \, E \, 4\pi r^2 = q$$

$$E = \frac{1}{4 \pi \epsilon_o} \left(\frac{q}{r^2} \right) \qquad (9)$$

where q is the total charge. Thus for points outside the charge distribution of spherically symmetric, the electric field E has a value that would have if the charge q is concentrated at its centre. (Compare this result with that of spherically symmetric distribution of mass and gravitational force at points outside it).

(ii) Now draw a Gaussian surface at a distance 'r' which is less than R the radius of spherically symmetric charge distribution. From the Gauss's law.

$$\epsilon_o \oint \mathbf{E} \cdot d\mathbf{S} = \epsilon_o \, (4 \pi r^2) = q^1 \qquad (10)$$

Here q' is the part of the charges contained within the sphere of radius r.
Therefore

$$E = \frac{1}{4 \pi \epsilon_o} \frac{q'}{r^2} \qquad (10a)$$

Note that the part of the charges outside the radius r do not contribute to the electric field E. (Compare this result with that of Gravitational field. Note that this result agrees well with that of spherical shell of matter which exerts no gravitational force on the body inside it).

A special case may be considered here where the distribution of charges inside the spherically symmetric charge is uniform. The charge density ρ will be constant throughout the charge distribution of radius R. It will be zero outside R.

$$\therefore \; q' = q \frac{(4/3) \pi r^3}{(4/3) \pi R^3} = q \left(\frac{r}{R} \right)^3$$

The expression for $E = \frac{1}{4 \pi \epsilon_o} \left(\frac{q r}{R^3} \right)$ (12)

What is the value of E at the centre of the spherically symmetric charge distribution ?

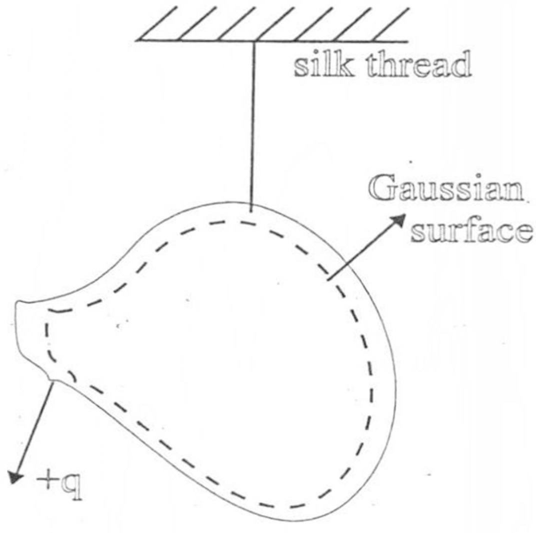


Fig.5 An insulated metallic conductor

6. Let us look into a case of an insulated charged conductor of any shape as shown in the fig. (5). It can be proved that any excess of charge placed on the insulated conductor resides on its entire surface. Draw a Gaussian surface very close to the inner side of the conductor.

Any excess of charge on the conductor will set up electric field inside the conductor. This field acts as charge carrier of the conductor and cause their movements. That means they set up inner electric currents. This will redistribute the charges in such a way that the electric fields will be automatically reduced in magnitude and eventually becomes zero. The internal currents also stops and the electrostatic conditions prevail. This redistribution of excess charges will take place in negligible time. At electrostatic equilibrium E is zero everywhere inside the conductor. Since the Gaussian surface is drawn just inside the conductor the excess of charges will also be zero at the Gaussian surface. Hence the electric flux on the Gaussian surface will also be zero. Therefore Gauss's law predicts that there will be no net charge inside a Gaussian surface. If the excess charge is not inside the Gaussian surface then it is logical to conclude that the excess of charges must be at the outer surface of the conductor.

If a charged metal ball is lowered with a silk thread inside an 'insulated' metal can, it will induce opposite charges on the metal can both inside and outside it. But when the charged body touches the insulated metal can the charge on the ball will immediately spread over the outer surface of the insulated metal can. There will be no charges on the inner side of the insulated metal can.

7. The Coulomb's law could be deduced from the Gauss's law. Consider an isolated point charge q. From symmetric condition the lines of force will be radial and uniformly distributed. The field **E** must be normal to a Gaussian surface drawn at a distance 'r' from the point charge. Hence the **E** and ds at any point on the Gaussian surface are directed radically outward. The angle between them will be zero and **E**.ds becomes Eds.

$$\therefore \quad \epsilon_0 \oint \mathbf{E} \cdot d\mathbf{S} = \epsilon_0 \oint E\, ds = q \qquad (13)$$

On this sphere E is constant and the above equation can be rewritten as

$$\epsilon_o \; E \oint ds = q$$
$$\epsilon_o \; E\,(4\,\pi\, r^2) = q$$

$$\therefore\; E = \left(\frac{1}{4\pi\epsilon_o}\right)\left(\frac{q}{r^2}\right) \qquad (14)$$

If a test charge q_o is brought to the Gaussian surface, the above electric field will act on the charge q_o and its magnitude will be

$$F = \left(\frac{1}{4\pi\epsilon_o}\right)\left(\frac{q q_o}{r^2}\right)$$

$F = Eq_o$
which is precisely Coulomb's Law.

(8) Draw necessary Gaussian surfaces for the following cases and determine the electric field E

(i) An infinite line charges with linear charge density λ, find E at a distance r from it.

(ii) An infinite sheet of charges on a non conducting sheet with surface charge density σ. Find E at a distance r from it.
Will the electric field change if the non conducting sheet is changed to a Conductor ? Find its value.

(iii) Two infinite plane parallel sheets having surface charge densities σ_1 and σ_2 are considered. Refer fig (6). What would be the electric fields in the space I, II and III ?

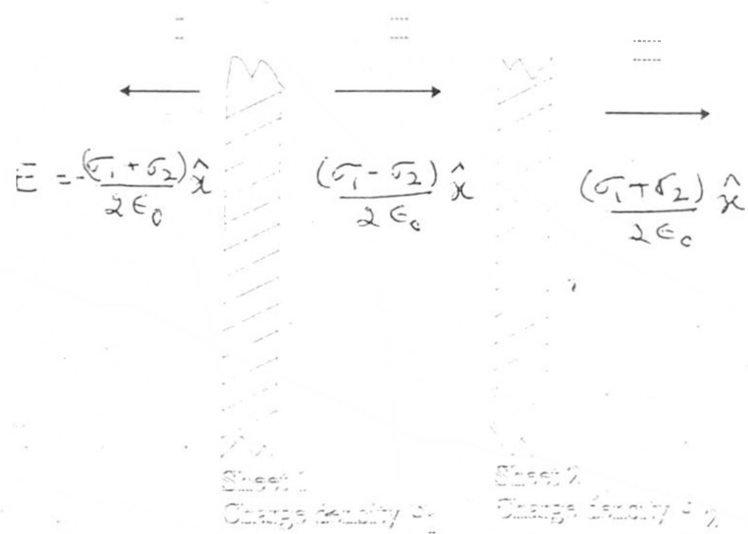


Fig.6 Field due to two infinite plane parallel sheets of charge density ($\sigma_1 > \sigma_2 > 0$). Only a section of finite part is shown.

9. When two plane parallel plates are given equal and opposite charges the field between and around them is shown in fig (7). While most of the charges accumulates at the opposing faces of the plates - the field is essentially uniform in space between them - there will be spreading or

'Fringing' of field at the edges of the plates. When the plates are made larger and distance between them is diminished the fringing effect is reduced and can be neglected entirely. This field between the plates is uniform and the charges are distributed uniformly over the opposing surfaces.

If the resulting fields at the two surfaces of the plats are E_1 and E_2, find the resultant field E at the space (i) to the left of the first plate (ii) in between the plates and (iii) to the right of the second plate.

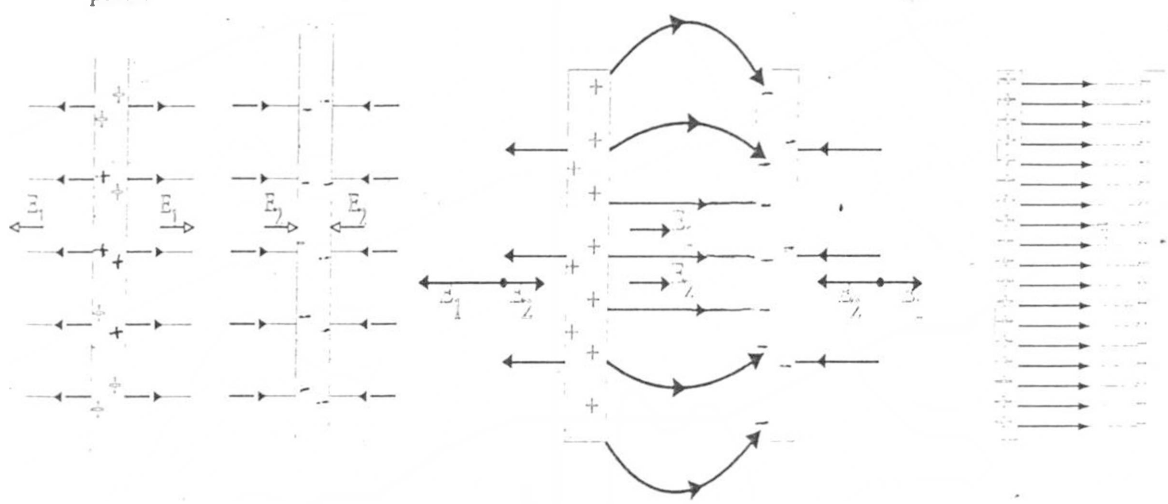


Fig.7 Electric field between oppositely charged parallel plates

EVALUATION

1. If the Gaussian surface encloses a dipole, what would be the electric flux ϕ_E for this surface?

2. An early model atom is considered to have a positive charged point nucleus of charge Ze surrounded by a uniform density of negative charges up to a radius R. The atom as a whole is neutral. Discuss the electric field at a distance 'r' from the nucleus. r > R.

3. In the Rutherford or nuclear atom model the positive charge,of the atom is concentrated in a small region (the nucleus) at the centre of the atom. For gold it is of about 6.9×10^{-15} m. What is the electric field at the nuclear surface ? Neglect the effect associated with the atomic electrons. Z for gold is 79.

4. A non conducting sphere of radius 'a' is placed at the centre of a spherical shell of inner radius 'b' and outer radius 'c'. A charge of +Q is distributed uniformly through the inner sphere and the charge density is $\rho\ Cm^{-3}$. The outer shell carries -Q charges. Find E_r (i) within the sphere $r < a$, (ii) between the sphere and shell (iii) inside the shell $b < r < c$ and (iv) outside the shell.

5. Charge is distributed uniformly throughout an infinitely long cylinder of radius R (i) show that E at a distance 'r' from the cylinder axis is given by $E = \frac{\rho r}{2\epsilon_o}$ when ρ is the charge density when $r < R$. (ii) what do you expect as the result if $r > R$?

CHAPTER
ELECTRIC POTENTIAL

We have seen that Electric Field is a *VECTOR* quantity. At a point its magnitude and direction can be shown with help of lines of force. One can visualize its nature and its variations in space in quantitative way. The electric field around a charged body can be described not only by electric field vectors but also by another scalar quantity known as *POTENTIAL*. In a static condition the potential contains just as much information as the Electric Field. Consider a test charge q_o being moved in a static field due to a point charge q. Draw lines of force around the charge. The lines of force will be radial. Suppose the test charge q_o is moved along a radial path I from A to B. (See fig. 8). As the lines of force also indicate the direction of electric field, **E** will be acting along the radial direction. A force of **E** q_o will be acting on the test charge due to the field. This force is towards the charge q_o. In order to move the charge q_o from A to B one has to do work against the electric field. Let the external agent do the work. In order to keep equilibrium this force must be equal and opposite to **E** q_o. Let us represent the potential at B as V_B and that at A is V_A. The difference between the two potentials is

$$V_B - V_A = (W_{AB})/q_o \qquad (1)$$

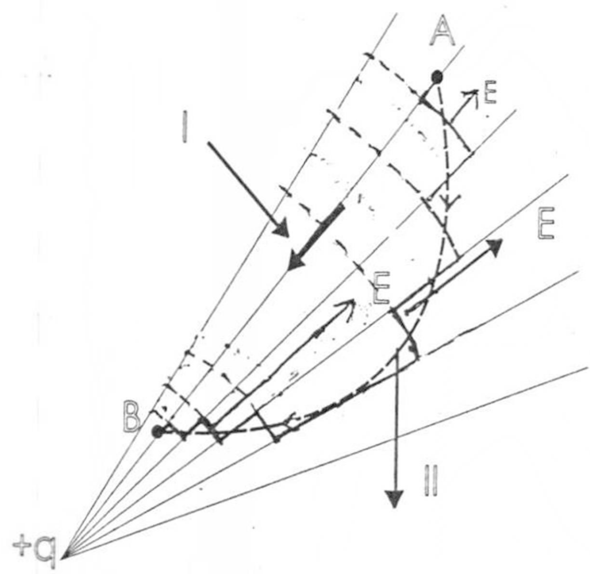


Fig.8 A test charge q_o is moved from A to B in the field of charge q along I & II paths

In order to define the POTENTIAL we select a condition that V_A becomes zero. That will happen only when the point A is far away from the charged body. That means at infinity distance. Dropping suffix B, one can write the potential at a point as

$$V = (W)/q_o \qquad (2)$$

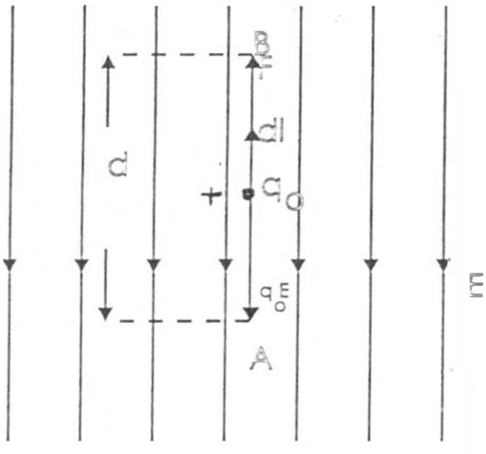
POTENTIAL AT A POINT IN AN ELECTRIC FIELD IS THE WORK TO BE DONE BY AN EXTERNAL AGENT BY BRINGING A UNIT TEST CHARGE FROM INFINITY TO THE POINT IN QUESTION.

The unit of potential is joule coulomb^{-1}, generally called volt (V).

Instead of following the radial path I now consider any arbitrary path II (see fig 8). It can be broken up into a large number of radial and arc segments. When these segments are small, it can be made arbitrary close to the actual path. It can be shown that no work will be done along the 'arc' as the force **F** and displacement **dl** are right angles to each other. Which means work is done only along radial segments. The sum of work done on the radial segments will be same as the work done in the previous radial path I. Since the second path is arbitrary, it is clear that work done is the same for all paths connecting A and B.

- What happens when $V_A = V_B$?

No work is required to move a test charge from A to B. Such surfaces are called *EQUIPOTENTIAL* surfaces. If a test charge is moved from one equipotential to another, work to be done by the external agent will not be zero. But it will be the same for *DIFFERENT PATHS* between those two equipotential surfaces.



From the symmetry, the equipotential surfaces for a spherical charge are a *FAMILY OF CONCENTRIC SPHERES.* And for a uniform field they are a *FAMILY OF PLANES AT RIGHT ANGLES TO FIELD.* In such cases the equipotential surfaces are at right angles to the lines of force and thus to electric field **E.** If they are not at right angles to each other then there will be a component lying on the surface and work has to be done.

Fig.9 A test charge q_o is moved from A to B in a uniform electric field E

NO WORK WILL BE DONE BETWEEN TWO POINTS ON THE SURFACE IF THE SURFACE IS EQUIPOTENTIAL AND E *MUST BE AT RIGHT ANGLES TO THE EQUIPOTENTIAL SURFACE.*

1) Consider a case where the two points A and B are in a uniform electric field **E**. Let A and B are at a distance d in the field direction. The force acting on a test charge q_o in the field is

$F = q_o E$ along the direction of **E.**

In order to move the test charge q_o from A to B one must counter act the above force by an external agent. This force must be equal but in the opposite direction.

If the displacement is dl, the work done W_{AB} by the external agent is the work contributed for all infinitesimal segments into which the path is divided.

$$W_{AB} = \int_A^B \vec{F} \,.\, d\,\vec{l} = -q_o \int_A^B \vec{E} \,.\, d\,\vec{l} \qquad (3)$$

$$V_B - V_A = \frac{W_{AB}}{q_o} = -\int_A^B \vec{E} \,.\, d\vec{l} \qquad (3\,a)$$

If point A is at infinity V_A will be zero and potential at B is

$$V = -\int_\infty^B \vec{E} \,.\, d\,\vec{l}$$

In the above case

$$V_B - V_A = \frac{W_{AB}}{q_o} = \mathbf{E}\ d \qquad (4)$$

This gives a new relation between the electric field and the potential difference between two points in the uniform field.

2. Now consider a non uniform electric field. Let the path in which the test charge q_o is moved by a curved path as shown in the figure 10. The force acting on the test charge is E q_o. To keep the test charge q_o without acceleration, the external agent must apply equal force but in the opposite direction. $\mathbf{F} = -\mathbf{E}\,q_o$. Therefore the work done by the agent in moving the test charge q_o from A to B is

$$W_{AB} = \int_A^B \vec{F} \,.\, d\vec{l} = -q_o \int_A^B \vec{E} \,.\, d\vec{l} \qquad (5)$$

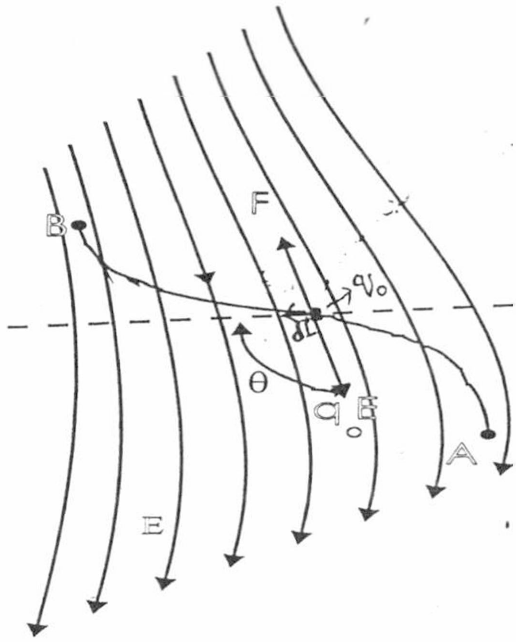
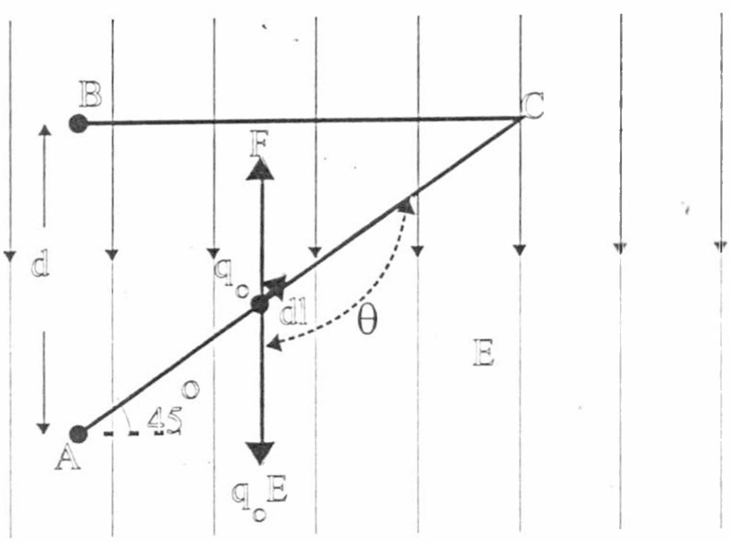


Fig.10 A test charge q_o is moved from A to B in a nonuniform electric field

Such an integral is called line integral. If the point A is at infinity the potential V at B is

$$V = -\int_{\infty}^{B} \vec{E} \cdot d\vec{l} \qquad (6)$$



Calculate the potential difference between A and B when the test charge q_o is taken along the path A to C and C to B is shown in the figure (11). The electric field E is uniform.

Fig.11 A test charge q_o is moved along the path ACB in a uniform field E

3. Consider a case of an isolated + point charge q. Let A, B and q lie on the same straight line. Assume that a test charge q_o is moved along the radial path without acceleration. In the fig. 12 **E** points to right and the direction of dl is to the left. Therefore.

$$E.\,d\,l = E \cos 180^\circ\, dl = -\,E.\,dl \qquad (7)$$

We are moving the test charge in a direction which reduces r where r is the measurement from q which is the origin.

$\therefore\ dl = -\,dr$ $\qquad\qquad\qquad \therefore\ E.\,dl = E\,dr$

substituting in 5

$$V_B - V_A = -\int_A^B \vec{E}\,.\,d\vec{l} = -\int_{r_A}^{r_B} E\,dr$$

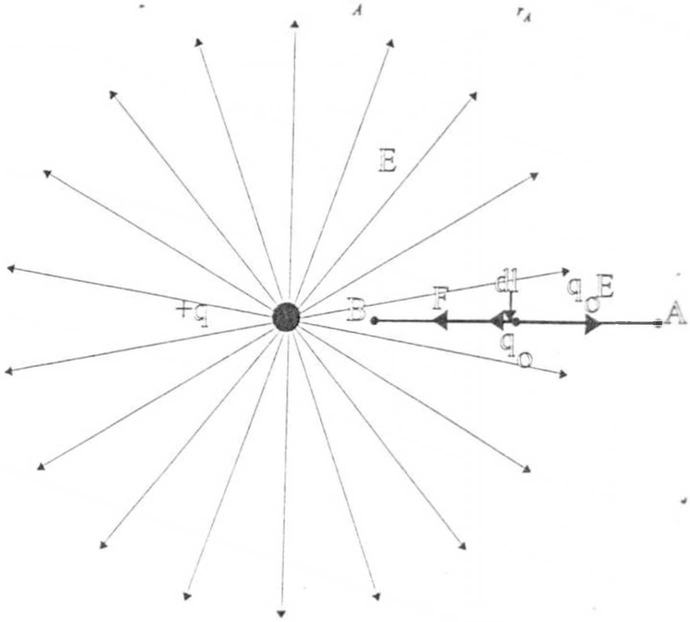


Fig.12 A test charge q_o is moved by external agent from A to B in the field set by as isolated charge q

Knowing the value of electric field at the site

$$E = \left(\frac{1}{4\pi\epsilon_o}\right)\left(\frac{q}{r^2}\right)$$

we obtain $$V_B - V_A = -\frac{q}{4\pi\epsilon_o}\int_{r_A}^{r_B}\frac{dr}{r^2} = \frac{q}{4\pi\epsilon_o}\left(\frac{1}{r_B} - \frac{1}{r_A}\right) \qquad (8)$$

choosing reference point A to be at infinity the potential at B will be

$$V = \frac{1}{4 \pi \epsilon_o} \left(\frac{q}{r} \right) \qquad (9)$$

THIS SHOWS THAT EQUIUPOTENTIAL SURFACES FOR AN ISOLATED POINT CHARGE ARE SPHERES, CONCENTRIC WITH THE POINT CHARGE.

Will the above relation hold good for external points to spherically symmetric charge distribution? Answer to this question is considered in the following problem.

What is the electric potential at the surface of a gold nucleus ? The radius of the nucleus is 6.6 x 10^{-15} m and atomic numbers of gold Z = 79.

Assume the nucleus to be spherically symmetrical and behaves electrically for an external points as if it is a point charge.

4) When there are group of charges, the potential is found out by calculating potential V_n due to each charge as if other charges are not present and then adding the quantities algebraically.

$$V = \sum V_n = \frac{1}{4 \pi \epsilon_o} \sum_n \frac{q_n}{r_n}$$

If the charge distribution is continuous rather than being a collection of points, the summation is replaced by an integral.

$V = \int dV = \frac{1}{4 \pi \epsilon_o} \int \frac{dq}{r}$ where dq is the differential element of charge distribution.

POTENTIAL ENERGY

In an electrostatic condition the charges q_1 and q_2 are separated by a distance r. If the separation distance is increased the external agent has to do work and it will be positive if the charges are of opposite sign and negative otherwise. *THE ENERGY REPRESENTED BY THIS WORK CAN BE THOUGHT OF AS STORED ENERGY CALLED ELECTRICAL POTENTIAL ENERGY.*

WE DEFINE THE ELECTRICAL POTENTIAL ENERGY OF A SYSTEM OF CHARGES AS THE WORK TO BE DONE TO ASSEMBLE THE SYSTEM OF CHARGES BY BRINGING THEM FROM AN INFINITE DISTANCE.

Let us imagine q_2 is removed to infinity. The electrical potential caused by q_1 at the original site of q_2 is

$$V = \frac{1}{4 \pi \epsilon_o} \frac{q_1}{r}$$

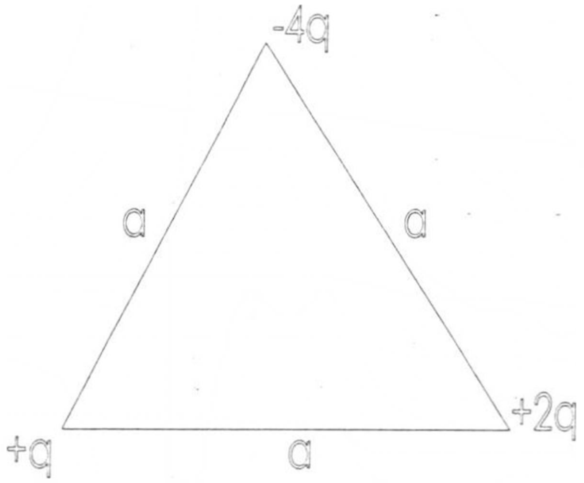
If q_2 is moved from infinity to its original position, work required by definition of electrical potential is $W = V q_2$. Combining the above two equations we get $U = \frac{1}{4 \pi \epsilon_o} \frac{q_1 q_2}{r_{12}}$ which is the potential energy of the system.

Two protons in a nucleus of U^{238} are 6.6×10^{-15} m apart. What is their mutual electric potential energy?

If a system contains three charged q_1 q_2 and q_3 separated by distances r_{12}, r_{13}, and r_{23} then the total energy of the configuration is the sum of the energies of each pair of particles.

$$U = U_{12} + U_{13} + U_{23}$$

It is considered that at infinity distance the potential is zero. A positive potential energy corresponds repulsive electric forces and a negative potential energy to attractive electrical forces. In the case of nucleus protons are held together by attraction and a non electrical forces, otherwise they would have moved apart. This force is called *NUCLEAR FORCE.*



Calculate the potential energy of the configuration of 3 charges as shown in the fig. 13.

Fig.13 The charges are held fixed by an external force

6) Consider an insulated conductor. We have seen earlier while dealing Gauss's Law that any excess of charge q placed on that conductor will move to its outer surface. We will discuss potential at different points of a charged conductor and also the electric field.

Consider two points A & B on the conductor. If they are not at the same potential, the charge carriers in the conductor near the lower potential would tend to move toward the higher potential. In a steady state there is no movement of charges and all points, both on the surface and inside it, must have the same potential. Since the surface of the conductor has static charge it will be an equipotential surface. The vector E for points on the surface must be at right angles to the surface. When excess charge is placed on the conductor it will spread over the surface on the conductor until E equals to zero for all points inside it. The same statement can be stated as the excess charge will move until all points of the conductor - both outer surface points and interior points are brought to the same potential. For V to be constant E has to be zero everywhere inside the conductor $E_l = \frac{dV}{dl}$.

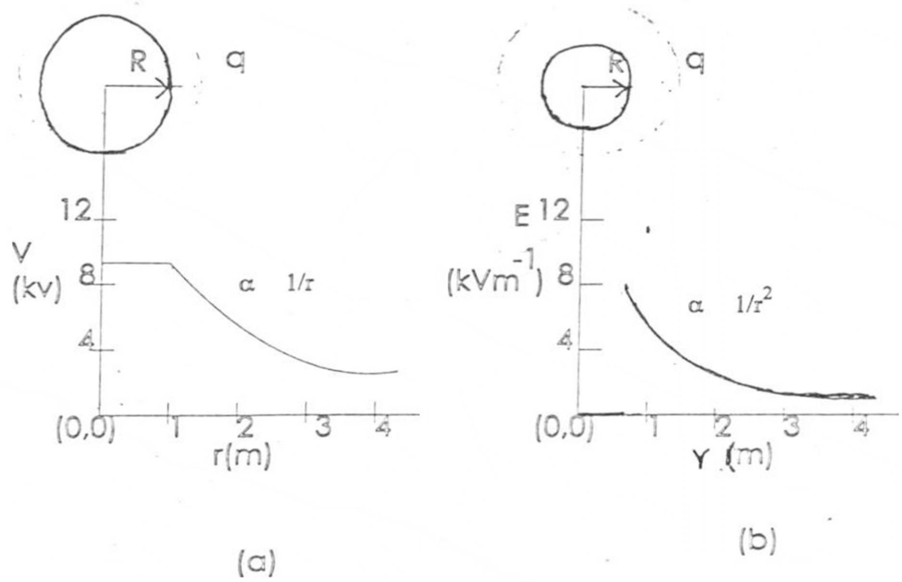


Fig.14 The potential (a) and electric field (b) for points near a conducting spherical shell of radius 1.0m and carrying a charge of 1.0×10^{-6} C

When we plot V against distance r for a charged conducting sphere, we obtain a graph as in the fig.14a. Inside the conductor the potential is constant and as the point under consideration is away from the outer surface of the conductor it decreases as $V \propto 1/r$.

When we plot a graph for variation of electric field E for various positions, we obtain a graph as in fig.14b. E is zero inside the conductor and is maximum for a point on the conductor. It decreases rapidly as the distance r increases.

$$E \propto \frac{1}{r^2}$$

It should be noted that the charge density tends to be high on isolated conducting surfaces whose radii of curvature are small and conversely. At sharp points, the charge desntiy is relatively high and similarly on the plane regions of the conductor it is relatively low.

7. The electric field E at points immediately above a charged surface is proportional to the charge density So it will reach very high value near the sharp points. One can see glow discharges from sharp points during thunderstorms. Lightning rod acts in this way to neutralize charged clouds and thus prevent lightning strokes.

Suppose two sphers R and r having different radii are charged and connected by a very long fine wire. The charges on them are Q and q respectively. Let V be the potential of the entire assembly. Find ratio of their surface charge densities σ_1 and σ_2 where σ_1 corresponds to the first sphere and σ_2 for the later sphere.

Which sphere has larger total charge and which has the greater change density? Refer fig.(15).

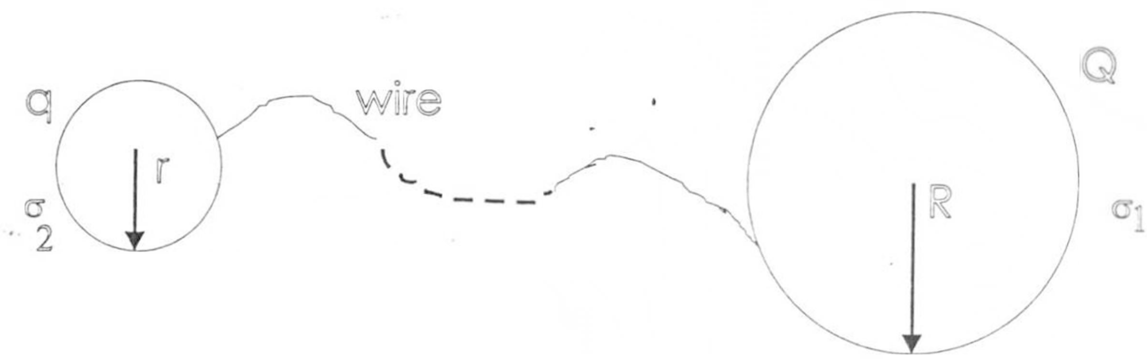


Fig.15 Charged spheres of different radii connected by a wire

We have studied the properties of electric fields in an electrostatic condition. The Gauss's law holds good and work done by carrying a test charge from one point to other is independent of the path. The concept of potential and equipotential surfaces are also well explained. We will discuss briefly the behaviour of conducting bodies in an electrostatic field. The very word 'static' means nothing changes with time. That means, there is no movement of charges and no current either in the interior or on the surface of the conductor. But a conductor has large number of free electrons. If there is any field in the interior of the conductor, the charges must move (a current flows) and the 'static' condition will be destroyed. Similarly, if there is any surface current, there must have a component of electric field tangential to the

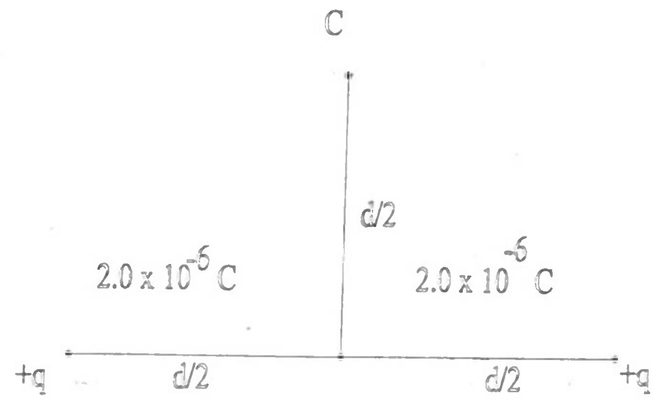


Fig. 16 Two charges 2.0×10^{-6} C are fixed in space at a distance of 2.0 cm apart

surface. So the 'static' condition forbids the presence of any electric field in the interior of the conductor and also no tangential component of field over the outer surface of the conductor. The field can have only normal component on the outer surface of the conductor. E is normal to the outer surface of the conductor.

EVALUATION

1. If V equals a constant value throughout a given region of space, what can you say about **E** in that region?

2. In the case of an isolated positive point charge q, is it necessary to assume that A, B and q lie on a straight line in order to prove $V = -\int_{\infty}^{B} \mathbf{E} \cdot d\mathbf{l}$ is true ? (Refer Fig.12)

3. Do electrons tend to go to regions of high potential or low potential ? Give reason.

4. An infinite charged sheet has a surface charged density $\sigma = 1.0 \times 10^{-7}$ Cm^{-2}. How far apart are the equipotential surfaces whose potentials differ by 5.0V ?

5. Two charges $q = +2.0 \times 10^{-6}$ C are fixed in space a distance d = 2.0 cm apart as in the figure (16).
   i) What is the electrical potential at point C ?
   ii) You bring a third charge $q = 2.0 \times 10^{-6}$ C very slowly move from infinity to C. How much work must you do?
   iii) What is the potential energy U of the configuration when the third charge is in place?

6. What is the charge desntiy σ on the surface of a conducting sphere of radius 0.15m whose potential is 200V ?

7. Two metal spheres are 3.0 cm in radius and carry charges $+1.0 \times 10^{-8}$ C and -3.0×10^{-8} C respectively, assumed to be uniformly distributed. If their centres are 2.0 m apart, calculate i) the potential of the point half way between their centres and ii) the potential of each sphere.

CHAPTER

AMPLIFIERS

Introduction

Very often, we have to amplify a small voltage signal from a transducer, such as a phonograph pick up, to a level which is suitable for the operation of another one, such as a loudspeaker. The arrangement is called an amplifier. We shall first study a 'black-box' representation of an amplifier. Then we shall proceed to discuss transistor as an amplifier and how the performance of the amplifier could be improved by providing negative 'feed back'.

Black-box representation of an amplifier

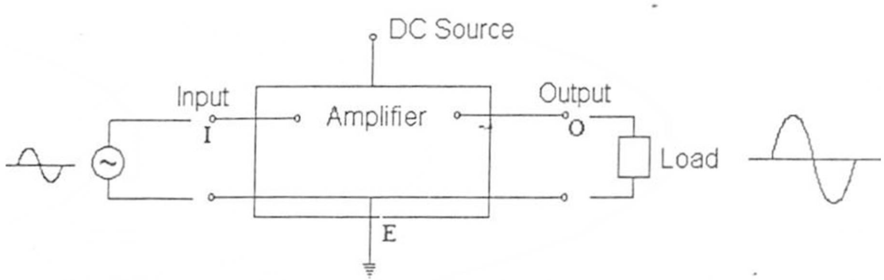


Fig.3.1 Black box representation of an amplifier

The above figure gives a black box representation of an amplifier. Input and output are alternating in nature at some fixed or variable frequency. The input is a low level voltage such as that obtained from a phonograph cartridge, or a tape-head (or a transducer such as a thermo couple, pressure gauge, etc).

The output is an enlarged version of the input and may be fed to a loudspeaker as in an audio amplifier. The amplifier has at least one active device, such as a transistor and may have a common connection E between input (I, E) and output (O, E) terminals.

In order to magnify the output, an amplifier needs a source of energy - a dry battery as in portable ones or a dc source resulting from a rectifier and filter combination. The active device basically converts the energy from the dc source into energy at the output of the amplifier that is proportional to the input signal. The ac input signal merely serves as a means of *controlling the dc to ac conversion* in the active device. This is usually accomplished with comparatively little input signal power.

The efficiency of conversion

$$\eta = \frac{\text{ac signal power delivered to the load}}{\text{dc input power}}$$

The maximum value of efficiency is in the range of 25% to 90%, depending upon the way in which the load is coupled to the active device (series fed, transformer coupled, push-pull etc) and the class of operation of the amplifier (class A, AB, B and C).

Transistor - as an amplifier

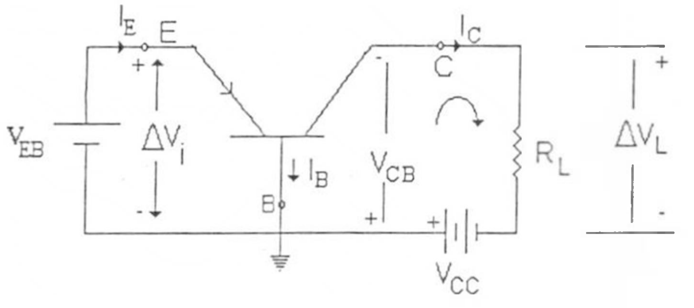


Fig.2 Transistor as an amplifier common base PNP

A small change in voltage between E and B produces a relatively large emitter current change ΔI_E. A fraction α of this current change is collected and passed through a load R connected in series with the collector supply voltage V_{cc}

$$\left(\alpha = \frac{\Delta I_C}{\Delta I_E} \right)_{V_{CB}}$$

The change in voltage across R_L is

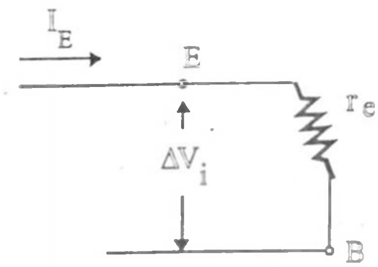
$$\Delta V_L = + (\Delta I_C) R_L = \alpha \, \Delta I_E R_L \qquad (1)$$

This could also be many times the change in input voltage ΔV_i. i.e. $\Delta V_L = A \, \Delta V_i$ (2)

where , $A = \frac{\Delta V_L}{\Delta V_i}$ the voltage amplification.

From (1) and (2)

$$A = \frac{\Delta V_L}{\Delta V_i} = \alpha \frac{\Delta I_E}{\Delta V_i} R_L$$



Now, $\Delta V_i = r_e \Delta V_E$ where r_e is the dynamic resistance of the emitter junction

$$A = \alpha \frac{R_L}{r_e}$$

$$V_i = I_E r_e \; ; \quad \Delta V_i = \Delta I_E r_e$$

V_i

The value of $r_e = \frac{0.026}{I_E} \Omega; \quad \alpha \propto 1$

Let $R_L = 3000\,\Omega$, $r_e = 40\,\Omega$. Then, A = 75

The above consideration gives physical explanation of why the transistor acts as an amplifier. The current in the low-resistance input circuit is transferred to the high-resistance output circuit. The word 'transistor' has originated as a contraction of '*transfer resistor*'.

Transistor provides power gain as well as voltage or current gain.

Graphical explanation

The amplification action can well be explained by considering the input (I_E - V_{EB}) and output (I_c = V_{CB}) characteristics. The fig 3 gives the common-base input characteristics of a PNP transistor. At an operating point Q, if the input voltage varies as a function of time as shown in (b), the emitter current varies as shown in (c).

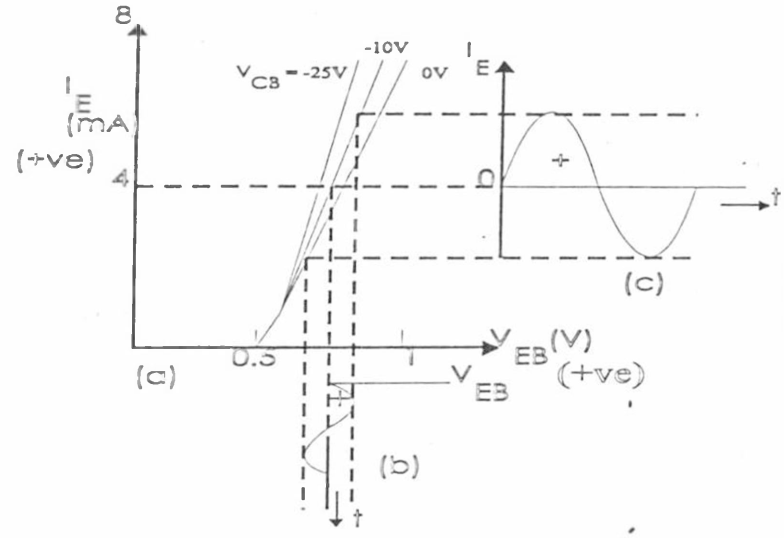


Fig.3 Common base input characterist ic (PNP)

To find the effect of this on the output, we have to consider the output characteristics which is shown in fig.4.

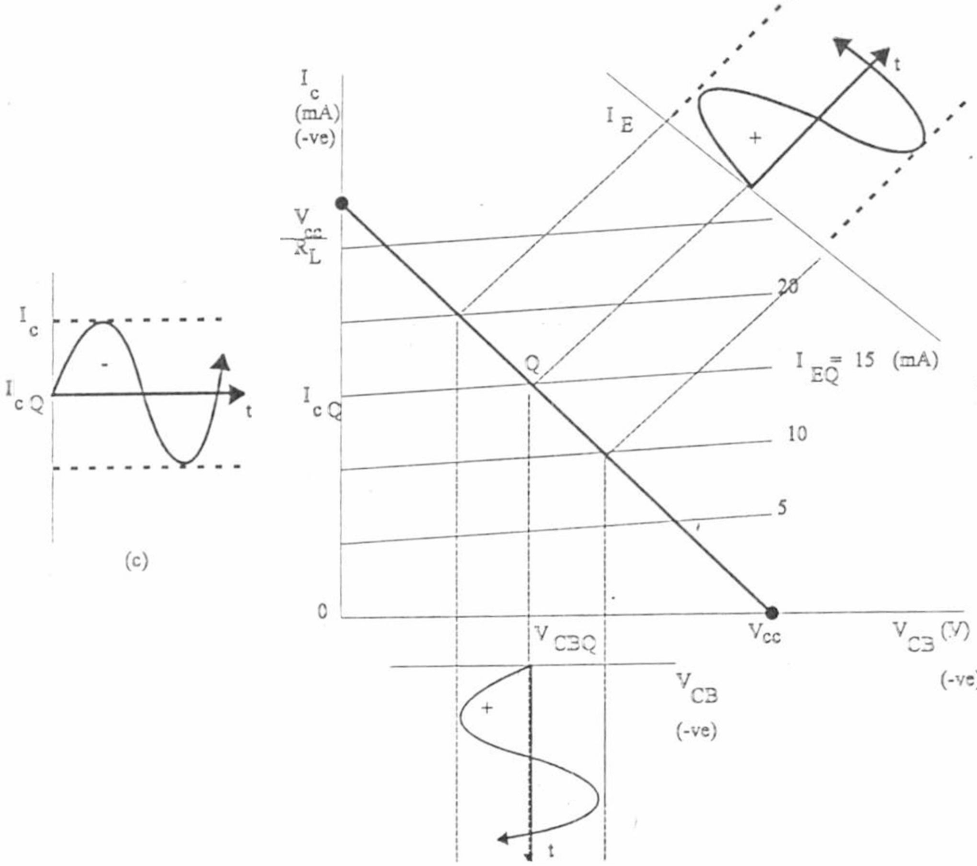


Fig.4 Load-line on output characteristic and graphical explanation of amplifier :tions

Next, draw the load-line AB on the graph. For this, apply Kirchhoff's voltage law to the output circuit.

$- V_{CB} - I_C R_L + V_{CC} = 0$

Rearrange $I_c = -\frac{1}{R} V_{CB} + \frac{1}{R_L} V_{cc}$

This straight line cuts V_{CB} axis at V_{CC} because when $I_C = 0$, $V_{CB} = V_{CC}$ and it cuts the I_C axis at V_{CC}/R_L because when $V_{CB} = 0$, $I_C = V_{CC}/R_L$.

If the input circuit current I_E is varied as a function of time, about the operating point Q (V_{CBQ}, I_{CQ}, I_{eQ}) as shown in fig 4b, the corresponding variation in I_C and V_{CB} are shown in fig. 4c and fig. 4d. On comparing fig.4(d) with fig.3(b), the output (collector) voltage is in phase with the input (emitter) voltage. This is described as phase inverse, As V_{BE} goes positive, I_E goes positive, I_C goes negative resulting in V_{CB} going positive.

The circuit for common base NPN transistor is given in fig.5.

The input and output characteristics for a NPN transistor will of course have the same genral appearance as those shown for a PNP in figures 3 and 4. Note that V_{EB} is negative, I_E is negative and V_{CB} and I_C are positive.

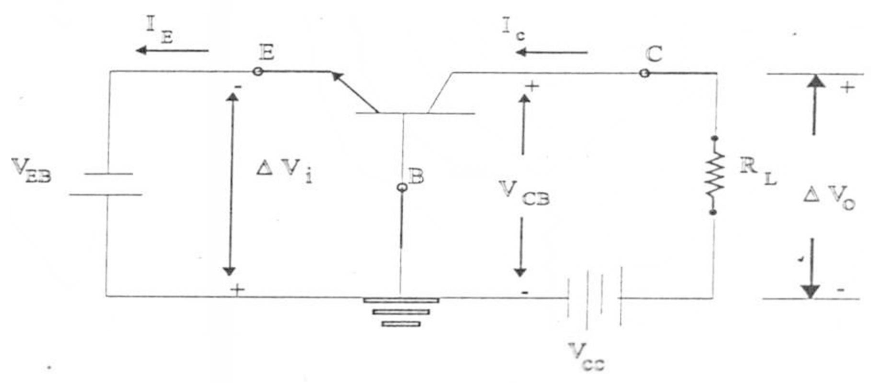


Fig.5 Transistor as an amplifier (common base NPN)

Similar explanation is possible for common emitter and common collector amplifiers. It should be noted that the output voltage suffers phase inversion in also common emitter configuration and not in *common collector* configuration, for which input characteristics are as in Fig.6(b) and output in 6(c). When V_{BC} goes negative, I_B goes negative, I_E goes positive and V_{EC} negative. No phase reversal.

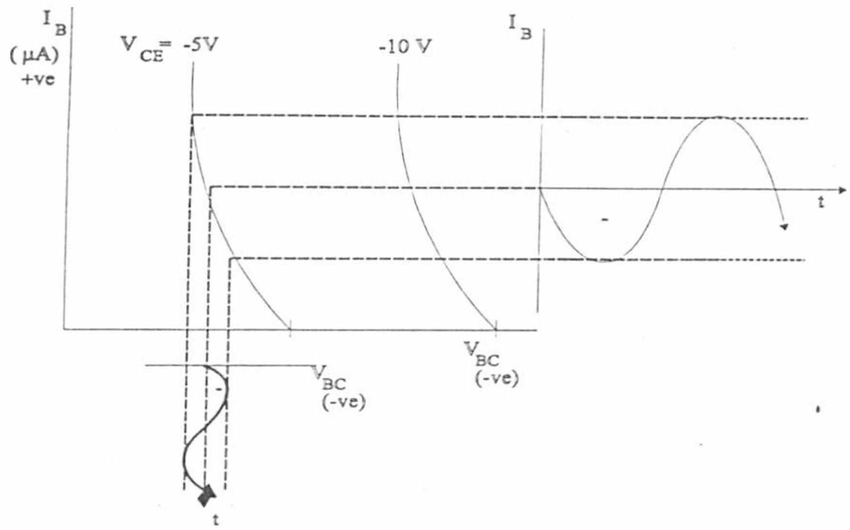
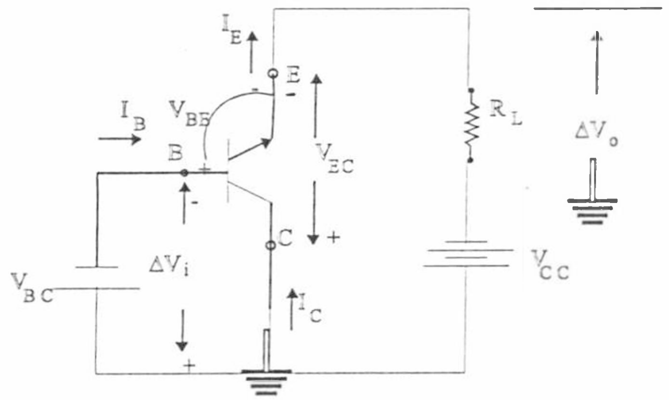
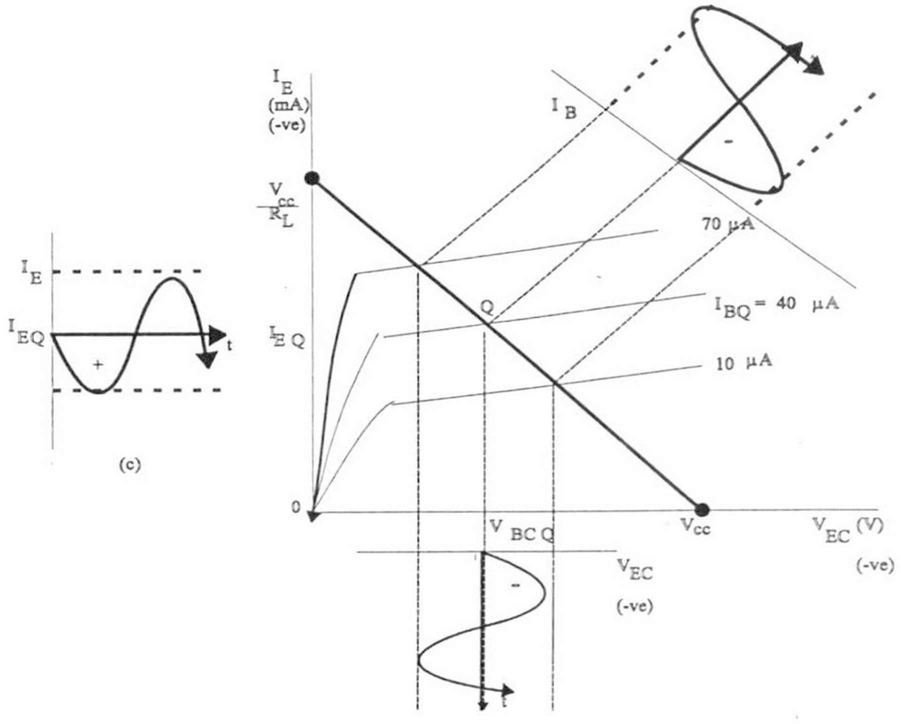


Fig.6 a) Common-collector amplifier (NPN)



b) Input characteristics of the same



c) Output characteristics

Coupling of amplifier to source and load

The coupling is generally done using capacitors C_i and C_o. The capacitor C_i blocks dc from going to the source and similarly C_o blocks dc from going to load R_L.

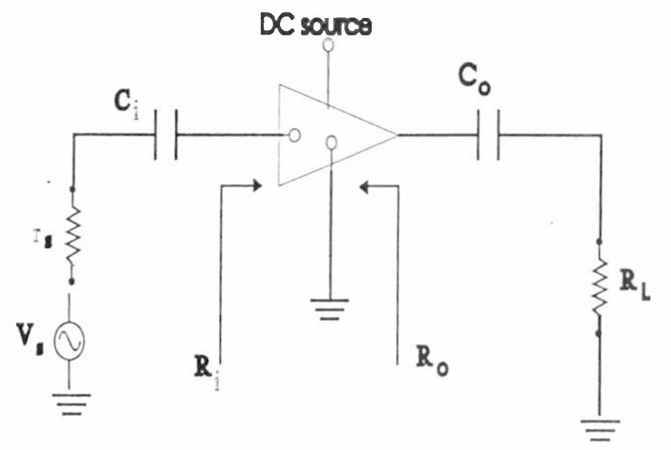


Fig.7 Coupling of an amplifier

Characteristics of a good amplifier

An amplifier should have

i) high input impedance R_i and low output impedance R_o (Refer fig.7)

ii) high fidelity, least distortion (different types) and constant gain over a wide range of frequencies, etc.

Comparison of characteristics of the three configurations and uses

Characteristics	**Common base**	**Common emitter**	**Common collector or emitter follower**
Output resistance R_o	Highest 1-2 M Ω	Moderate 50 kΩ	Lowest 100 Ω- 1 k Ω
Input resistance R_i	Lowest 20-50 Ω	Moderate 1-2 kΩ	Highest 150 kΩ - 600kΩ
Current gain A_I	Low ≈ 1(<1) 0.85 to 0.995	Large 20-200	Large 20-200
Voltage gain A_v	Hig ≈ $A_{V(CE)}$	High	Low <1 (0.99)
Power gain A_P	Moderate	Large	Small
Phase Change	No	180°	No
Uses	i) For impedance matching - to match a very low impedance source to drive a high impedence load.	Popular because of high voltage, current and power gains.	i) As a buffer amplifier between a high impedance source and a low impedance load
	ii) As a non inverting amplifier with a voltage gain >1 iii) as a constant current source (e.g. in sweep circuit to charge a capacitor).		ii) Level shifter in direct coupled circuits.
Stages in cascading of amplifier	Input - when a transducer requires near short circuit operations.	Intermediate	Input: When a transducer requires near open-circuit operation. Output: to derive a low impedance load (especially capacitive)

Feedback in amplifier

The amplifier performance can be improved by providing proper feedback (i.e. combining a portion of the output signal with the external input signal). The concept of feedback can be understood from the block diagram given below.

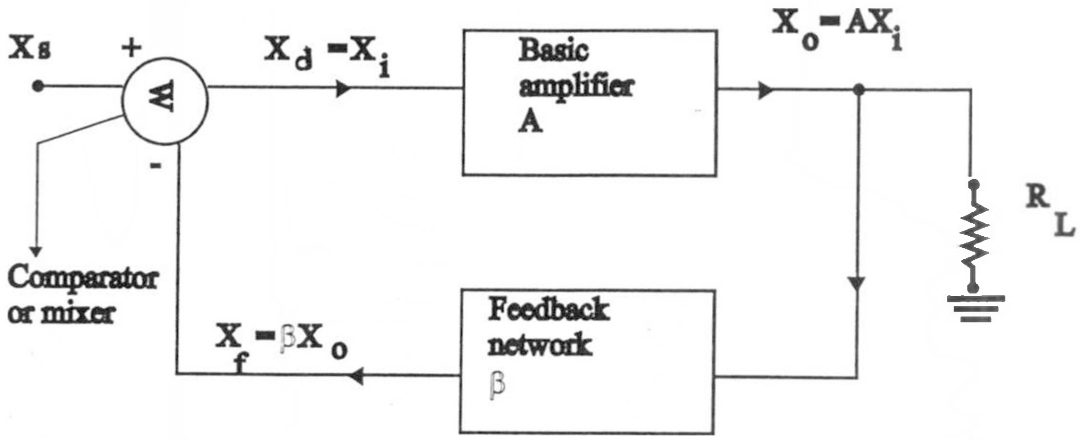


Fig.8 Block diagram of an amplifier with feedback

Input to the amplifier $X_i = X_d$, the difference signal $(X_S - X_f)$. The gain of the basic amplifier

$$A = \frac{X_o}{X_i} = \frac{X_o}{X_s - X_f}, \quad where\ X_f = \beta X_o$$

The gain of the amplifier with feedback

$$A_f = \frac{X_o}{X_s} = \frac{AX_i}{X_s} = \frac{A(X_s - X_f)}{X_s}$$

$$= A\,\frac{(X_s - \beta X_o)}{X_s}$$

$$= A - \beta A A_f$$

Rearranging,

$$A_f(1 + \beta A) = A$$

$$A_f = \frac{A}{1 + \beta A}$$

$(-\beta A)$ is called the loop gain :

Case 1: If $|A_f| < |A|$, the feedback is called negative or degenerative. The voltage feedback causes signal reduction because it is out of phase with the external signal, yielding a lower output. Note $|1 + \beta A| > 1$ i.e. the loop gain is negative.

Case 2 : If $|A_f| > |A|$ the feedback is termed positive, or regenerative. The phase of the voltage feedback is such as to increase the output, yielding a greater output. Note $|1 + \beta A| < 1$ i.e. the loop gain is positive.

Advantages of negative feedback

The negative feedback, though reduces the overall gain, has the following advantages:

i) The gain is stabilized against variations of 'parameters' of the active device (due to aging, temperature, replacement, etc.)

ii) The input resistance is increased and output resistance is reduced (for some circuits).

iii) The frequency response can be significantly improved (wide band width).

iv) The distortions are reduced.

v) The noise is suppressed.

CHAPTER

OSCILLATORS

Introduction

An oscillator is a device that generates a periodic ac output signal of desired frequency, without requiring any form of input signal. Oscillation can be described as a form of instability caused by feedback that regenerates or reinforces a signal that would otherwise die out due to energy loses. In order for the feedback to be regenerative, it must satisfy certain amplitude and phase relations.

Block diagram discussion

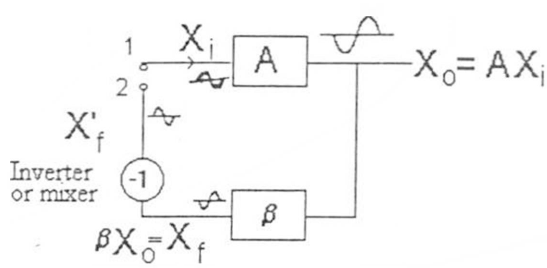


Fig.1 An amplifier with transfer gain A and feedback network β, not yet connected to form a closed loop

The fig. 1 shows an amplifier, a feedback network, and an input mixing circuit not yet connected to form a closed loop. The output signal of the amplifier $X_o = AX_i$ is due to the input X_i applied directly to the amplifier input terminal. The feedback voltage $X_f = \beta X_o = A\beta X_i$. This is simply inverted in the mixing circuit. The output of the mixing circuit $X_f^1 = - X_f = - A\beta X_i$

The loop gain = $\frac{X_f^1}{X_i} = - \frac{X_f}{X_i} = -\beta A$,

Supposing that X_f^1 is identically equal to externally applied input signal X_i the amplifier is not in a position to distinguish the source of the input signal applied to it. Therefore, if external source were removed and if terminal 2 were connected to terminal 1, the amplifier would continue to provide the same output X_o as before. The condition that $X_f^1 = X_i$ means that $- A\beta = 1$ or the loop *gain must equal* unity. This statement has two implications :

i) $|A\beta| = 1$ and

ii) the phase of $-A\beta$ is zero. The loop gain phase shift is zero (or, an integral multiple of 2π)

These conditions are called the Barkhausen criteria. The reactive elements in the amplifier and/or feedback circuits cause the gain magnitude and phase shift to change with frequency. In general, there is only one frequency at which the gain magnitude is unity and at which, simultaneously, the total phase shift is equivalent to zero. The designing an oscillator means selecting reactive components and incorporating them into circuitry so that the conditions are satisfied at a predetermined frequency.

Refer to the feedback formula $A_f = \dfrac{A}{1 + \beta A}$

For if $-\beta A = 1$, thus $A_f \rightarrow \infty$. This may be interpreted to mean that there exists an output voltage even in the absence of an externally applied signal voltage.

An oscillator must have an amplifier to supply energy (from the dc supply) to replenish resistive losses and thus sustain oscillation.

Practical considerations

Refer to fig.1. If $|\beta A| = 1$, the removal of external generator will result in a cessation of oscillations. (Fig.2a)

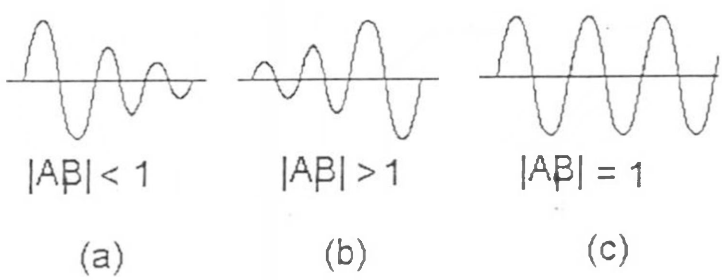


Fig.2 Gain magnitude and amplitude of oscillation

If $|A\beta| > 1$, the amplitude of oscillations will continue to increase, fig.2-b (limited by the onset of non-linearity of operation in the active devices associated with the amplifier).

If $|A\beta| = 1$, then, with the feedback signal connected to the input terminal, the removal of the external generator will make no difference. The amplitude of oscillation is a steady one as in fig.2-c.

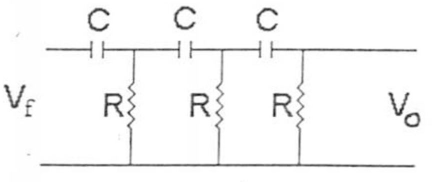
In practice, an oscillator in which the loop gain is exactly unity is an abstraction, completely unrealizable. The value of loop the magnitude of gain $|A\beta|$ is made somewhat larger (about 5%) than unity in order to ensure that, with incidental variations in active device and circuit parameters, $|\beta A|$ shall not fall below unity. After the output voltage reaches a desired level, the value of $|A\beta|$ decreases to unity and the output amplitude remains constant.

How does the oscillation start?

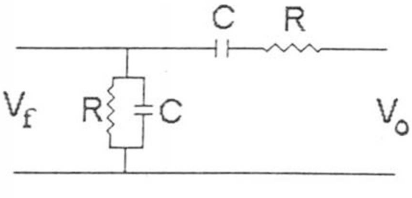
When the circuit is switched on every resistor in the circuit generates noise voltages due to the random motion of electrons in it. The noise signal is a complex signal, which can be viewed as made up of sinusoidal signals of frequencies over 10^{12}Hz. These signals are very small in amplitude. All are amplified and appear at the output terminals. A part of the amplified noise output passes through the feedback circuit. The Barkhaussen criteria are satisfied for only the predetermined frequency which goes for amplification again. With the magnitude of loop gain $|A\beta|$ slightly greater than unity, the oscillations build up at this frequency. When suitable level is reached, $|A\beta|$ decreases to unity and a steady output is obtained.

Feedback circuits

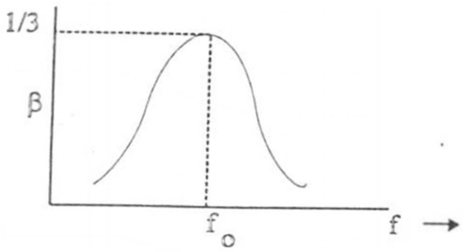
Some of the commonly used feedback circuits for oscillators are given below.



i) RC network - low frequency

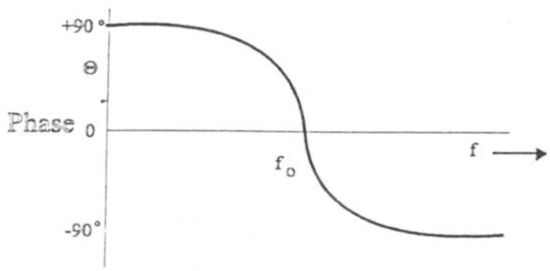


ii) RC network - Phase lead-lag network as in Wein-bridge oscillator - a very good one for audio frequency operation.



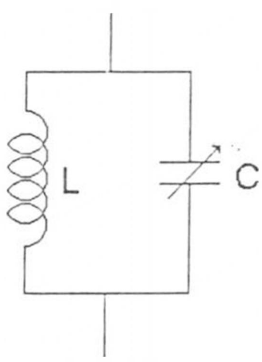
$$f = \frac{1}{2\pi RC}\ Hz$$

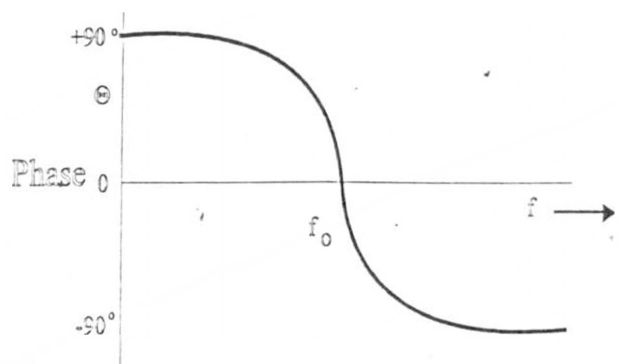
A = 3, β = 1/3

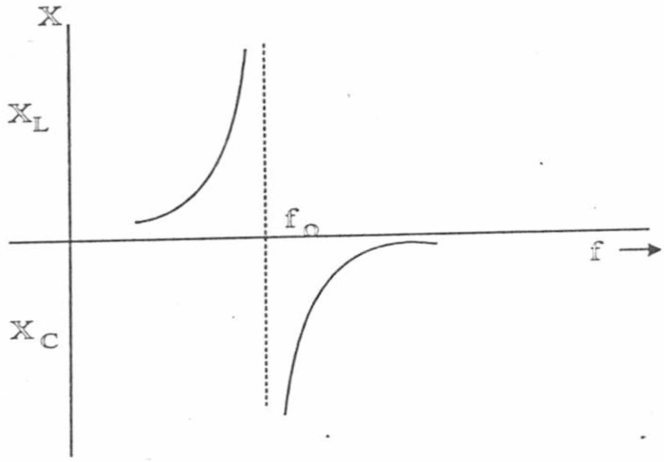


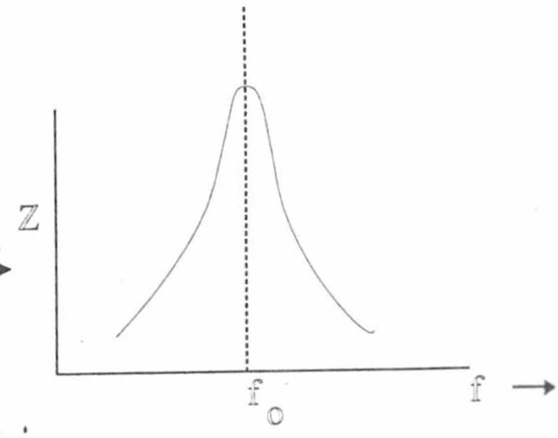
Amplitude stability is maintained by having a pair of components : resistance-sensistor / tungsten lamp (positive temperature coefficient or thermistor (negative temperature coefficient).- resistance forming one arm of the wheatstone bridge network, the other arm being the phase lead-lag components.

iii) LC network - for high frequency applications upto 500 MHz - generally tuned/resonant network

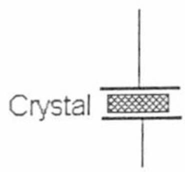


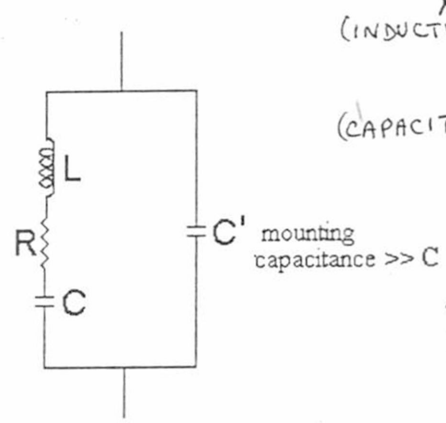


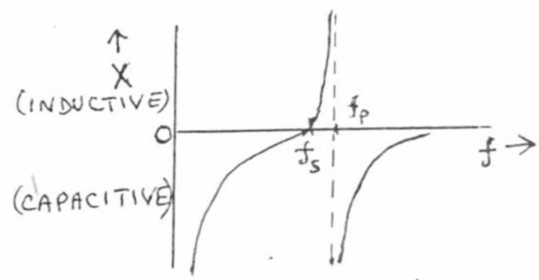




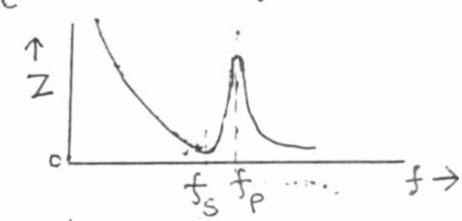
iv) Crystal oscillator



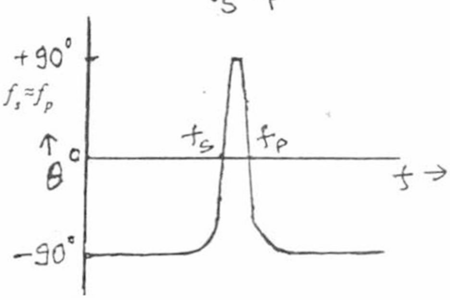




(R negligible)



$$f_s = \frac{1}{2\pi}\sqrt{\frac{1}{LC}} \qquad f_p = \frac{1}{2\pi}\sqrt{\frac{1}{L}\left(\frac{1}{C}+\frac{1}{C^1}\right)} \quad \text{and } f_s \approx f_p$$



Desired characteristics of an oscillator

i) amplitude stability

ii) frequency stability - a measure of its ability to maintain as nearly a fixed frequency as possible over as long a time interval as possible. A measure of frequency stability is $\frac{d\theta}{df}$

The larger the value of $\frac{d\theta}{df}$ the more stable is the oscillator frequency (θ is the phase of voltage with respect to current.

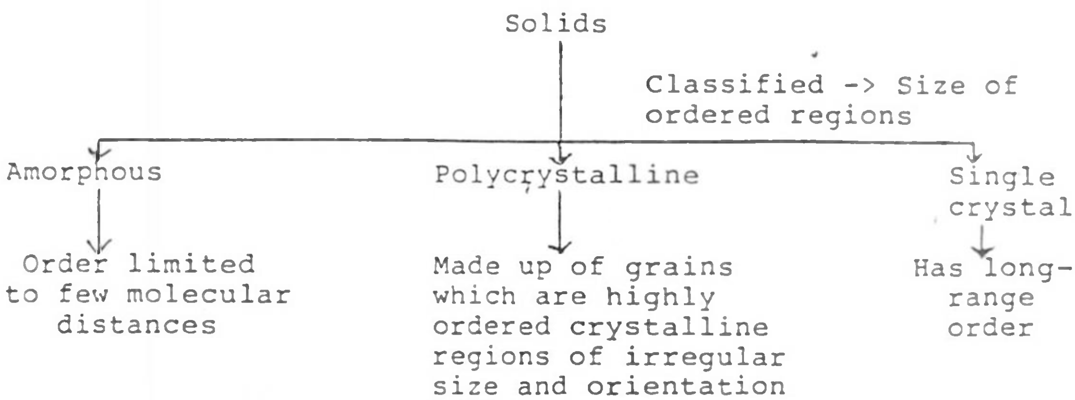
iii) perfect sinusoidal waveform - especially for audio frequency operations.

CHAPTER

SOLID STATE PHYSICS

Solid State Physics is a study of the physical properties of solids like electrical conductivity, dielectric properties, elastic properties, thermal properties, magnetic properties. It deals with properties common to a large number of compounds and the quantitative relation between the properties and the underlying structure.

It is sometimes also termed as 'Condensed Matter Physics'.



Many important properties of materials depend on the structure of crystals and the electron states within crystals - Band theory.

Aim of Crystal Physics

The aim of Crystal Physics is the interpretation of the macroscopic properties in terms of the microscopic particles of which the solid is composed.

Science of Crystallography

It is the study of geometric form and other physical properties of crystalline solids by using X-rays, electron beams, neutron beams, etc.

Lattice Points and Space Lattice

Crystal Structure: Atomic arrangement in a crystal is called crystal structure. In a perfect crystal, there is a regular arrangement of atoms (periodic) in 3-dim.

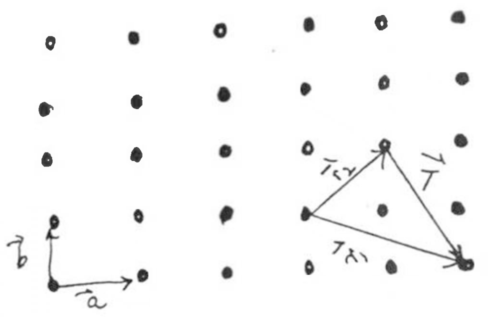
Periodicity may be different in different directions.

Location of atoms are specified by points called **Lattice points.** Totality of such points forms a crystal lattice or space lattice.

If all the atoms at the lattice points are identical, the lattice is called **Bravais lattice.**

3-dim space lattice is a finite array of points in 3-dim, in which every point has an identical environment as any other point in the array.

Consider a 2-d array of points. The environment about any two points is the same.



$\vec{a}$, $\vec{b}$ are the fundamental translation vectors.

Choose some origin and join it to two points A and B by vectors $\vec{r}_1$ and $\vec{r}_2$. If the difference T of the two vectors $\vec{r}_1$ and $\vec{r}_2$ satisfies the following relation

$$\vec{T} = n_1\vec{a} + n_2\vec{b}$$

where n_1 and n_2 are integers, then the array of points is a 2-d lattice.

For a 3-d lattice

$$\vec{T} = n_1\vec{a} + n_2\vec{b} + n_3\vec{c}$$

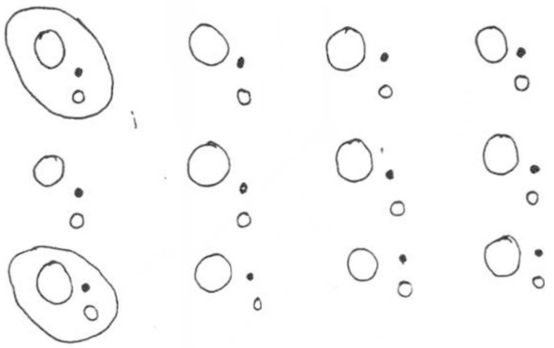
Crystal lattice refers to the geometry of a set of points in space whereas crystal structure refers to the actual ordering of its constituent ions, atoms and molecules in the space.

The Basis and Crystal Structure

Crystal structure is got by associating every lattice point with a unit assembly called Basis.

A **basis** is an assembly of atoms or molecules identical in composition orientation and arrangement. All lattice points are connected by a translation.

Lattice + Basis = Crystal Structure



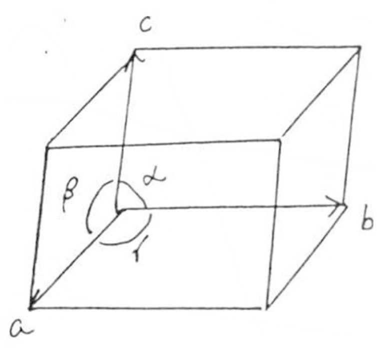
Unit Cells and Lattice parameters

In every crystal some fundamental grouping of particles (atoms) is repeated. Such a fundamental repeat entity is called a **unit cell**. This smallest unit repeated in 3-d gives rise to the crystal and constitutes the building block. Unit cells for most crystals are parallelepipeds or cubes having three sets of faces which are parallel.

A unit cell is chosen to represent the symmetry of the crystal structure, wherein all the atom positions in the crystal may be generated by translations of the unit cell integral distances along each of its edges.

More than a single unit cell may be chosen for a particular crystal structure. However, we generally use the unit cell having the highest geometrical symmetry.

A space lattice is a regular distribution of points in space, in such a manner that every point has identical surroundings. The lattice is made up of a repetition of unit cells, and a unit cell is completely described by the three vectors $\vec{a}$, $\vec{b}$, $\vec{c}$ when the length of the vectors and the angles between them (α, β, γ) are specified.



Lattice parameters of a unit cell

Taking any lattice point as the origin, all other points on the lattice can be obtained by a repeated operation of the lattice vector $\vec{a}$, $\vec{b}$, $\vec{c}$. The lattice vectors and the interfacial angles are called the lattice parameters of a unit cell. Hence if we know the values of these intercepts and the interfacial angles, we can easily determine the form and the actual size of the unit cell.

The vectors $\vec{a}, \vec{b}, \vec{c}$ may or may not be equal. Same is true of angles α, β, γ. They may or may not be right angles. The above conditions determine the seven crystal systems. If the atoms are at the corners only the seven crystal systems yield seven types of lattices. More space lattices can be constructed by placing atoms at the body centres of unit cells or at the centres of faces giving the body-centred and face-centred lattices. Bravais showed that the total number of different space lattice types (obeying the condition that every point has identical surroundings) is only fourteen. Hence the term "Braivais Lattice".

Unit Cell vs. Primitive Cell

Primitive Cell: It is a geometrical shape which, when repeatedly placed indefinitely in three dimensions will fill all space and is equivalent of one lattice point.

Primitive cell contains only one lattice point at the corners of the unit cell.

Unit cells may be primitive cells but all primitive cells need not be unit cells.

Crystal systems

There are thirty-two classes of crystal systems based on geometrical considerations (i.e. symmetry and internal structure). But, it is a common practice to divide all the crystal systems into seven groups or basic systems which are distinguished from one another by the angles between the three axes and intercepts of the faces along them. They are

- Cubic (Isometric)	Minerals -> galena,
- Tetragonal	garnet,
- Orthorhombic	zircon, rutile
- Monoclinic	barite
- Triclinic	gypsum plagioclase
- Trigonal (rhombohedral)	calatequartz
- Hexagonal	graphite, molybdenite

The seven crystal systems and their properties are given in the table.

Sl. No.	Crystal system	Unit cell parameters	Examples
1	Triclinic	$a \neq b \neq c; \alpha = \beta = \gamma = 90°$	$K_2Cr_2O_7$ $CuSO_4$ $5H_2O$
2	Monoclinic	$a \neq b \neq c; \alpha = \beta = 90° \neq \gamma$	$CaSO_4$ $2H_2O$ (Gypsum) $FeSO_4$ Na_2SO_4
3	Orthorhombic (rhombic)	$a \neq b \neq c; \alpha = \beta = \gamma = 90°$	KNO_3, $BaSO_4$
4	Tetragonal	$a = b \neq c; \alpha = \beta = \gamma = 90°$	TiO_2, SnO_2 $NiSO_4$
5	Cubic	$a = b = c; \alpha = \beta = \gamma = 90°$	Au, Cu, NaCl
6	Hexagonal	$a = b \neq c; \alpha = \beta = 90°, \gamma = 120°$	SiO_2, Zn, Mg
7	Rhombohedral (Trigonal)	$a = b = c; \alpha = \beta = \gamma \neq 90°$	As, Sb, Bi, Calcite

A crystalline substance can be looked upon as a closely packed aggregate of atoms or ions which are usually assembled to have a spherical shape. It has been observed that structure of many crystals can be profitably understood in terms of the packing of spheres in space.

Every one is familiar with a bunch of grapes or bananas, a pile of oranges, close packing of seeds in a pomegranate fruit,k a raft of soap bubbles and a beehive (Figs. 1-4).

In all these cases we observe that one piece is surrounded by six other identical units on one surface which are in close contact with it. This is how nature fills the space to the maximum extent wherever it is possible. Similarly it is possible to keep six marbles (or ping pong balls) in contact with only one ball in a single layer (Fig. 5). A second similar layer can be superimposed on this layer in such a way that each sphere is in contact with three spheres of the adjacent layer as shown (Fig. 5). A third layer can be added now in two ways. In one way it is possible to keep the spheres directly above the first layer as in Fig. 6a. The other way is to keep the spheres over the holes in the first layer not occupied by the second layer (7a). The first arrangement is called hexagonal close packing (HCP) and the second is known as (CCP) cubic close packing.

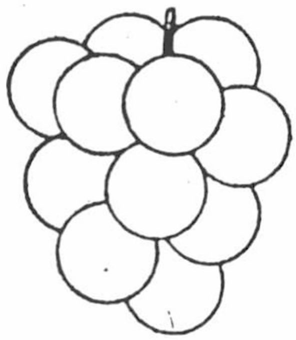


FIG.1 – GRAPES

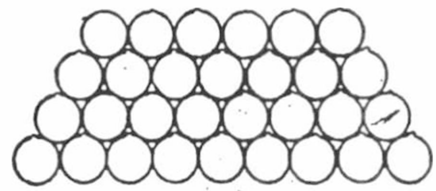


FIG.2 – ORANGES

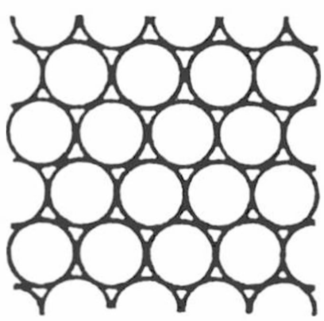


FIG.3 – SOAP BUBBLES

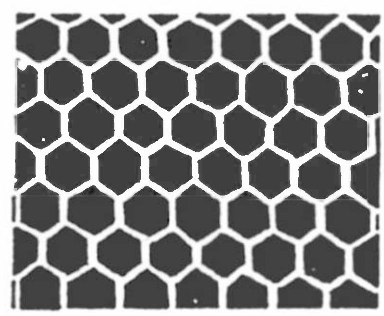


FIG. 4 – BEEHIVE

PACKING ARRANGEMENT

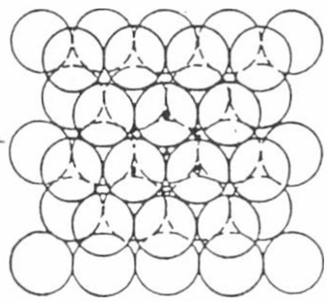


FIG 5

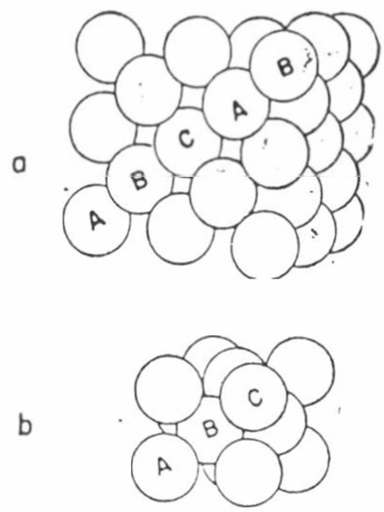


FIG 7 FCC

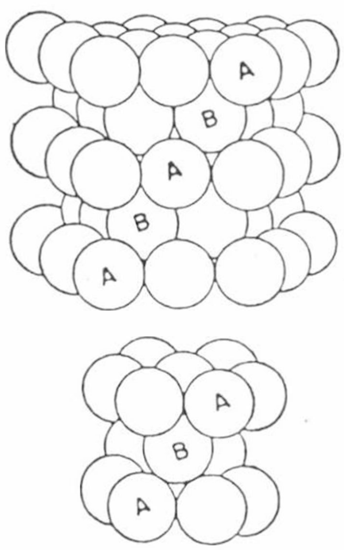


FIG 6 HCP

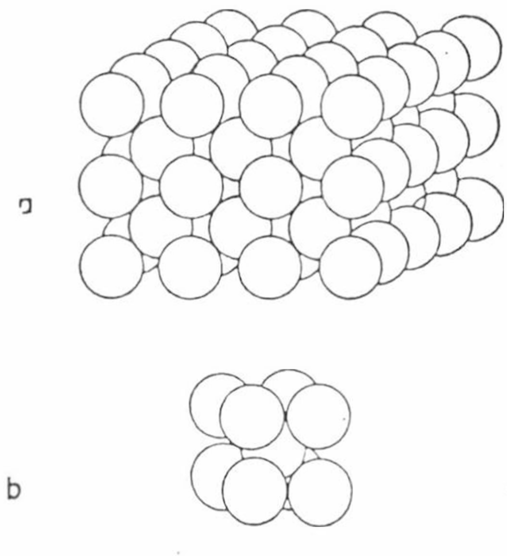


FIG 9 BCC

Let us designate the layers as A, B and C, we observe that in HCP, the sequence of layers will be repeated. We observe that in HCP, the sequence of layers will be repeated in terms of AB, AB, AB,... or BC, BC, BC or AC, AC, AC ... In the case of CCP the sequence will be ABC, ABC, ABC, ... Thus the structure repeats itself after two layers in HCP and after three layers in CCP.

We can count 12 spheres in contact with one sphere in both the packing arrangements (6b, 7b). In one plane one sphere is surrounded by six other spheres in HCP with three other spheres on both sides in a triangular way. In the case of CCP the lower triangle is rotated through an angle of 60°. The number of immediate neighbours which a sphere can have is represented by the **coordination number**: 12 for HCP and CCP. Some important crystal structure terms are defined below:

Coordination number (N): Number of equidistant nearest neighbours that an atom has in the given structure. Greater the coordination number the more closely packed up the structure.

Nearest neighbour distance (2r): The distance between the centres of two nearest neighbouring atoms is called **nearest neighbour distance.** It will be 2r if r is the radius of the atom.

Atomic radius (r): Atomic radius 'r' is defined as half the distance between nearest neighbours in a crystal of pure element.

Atomic packing factor: The fraction of the space occupied by atoms in a unit cell is called **atomic packing factor** (APF); or simply packing factor, i.e. it is the ratio of the volume of the atoms occupying the unit cell to the volume of the unit cell relating to that structure.

Cubic (simple) structure

Coordination no. = 6

No. of atoms/unit cell = 1

Nearest neighbour distance 2r = a

Lattice parameter a = 2r

No. of lattice points = 1

Volume of all atoms $v = 1 \times 4/3\pi r^3$

Volume of unit cell $V = a^3 = (2r)^3$

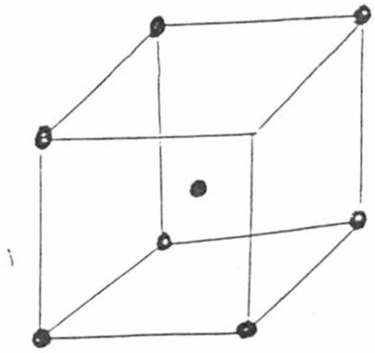
Packing fraction (atomic)

$$= \frac{\text{Volume of atoms}}{\text{Volume of unit cell}} = \frac{v}{V}$$

$$= \frac{4}{3}\,\frac{r^3}{a^3} = \frac{4}{3}\,\frac{r^3}{8r^3} = \frac{\;}{6} = 52\%$$

BCC

BCC Lattice



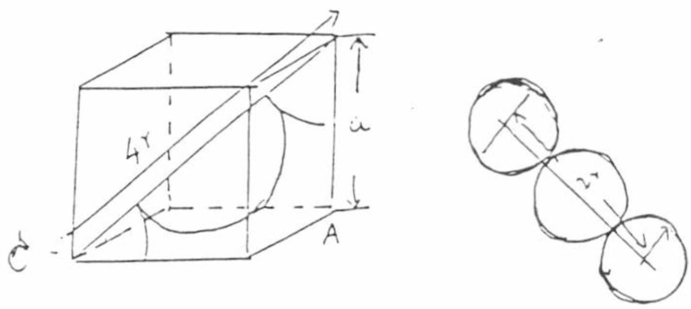
$AC^2 = a^2 + a^2 = 2a^2$

$FC^2 = (AC)^2 + (AF)^2$

$= 2a^2 + a^2 = 3a^2$

$(4r)^2 = 3a^2$

$a = \frac{4r}{\sqrt{3}}$ $2r = \frac{a\sqrt{3}}{2}$



Coordination no. 8; $2r = a\sqrt{3}/2$

Lattice constant $a = 4r/\sqrt{3}$

Number of atoms/unit cell = 2

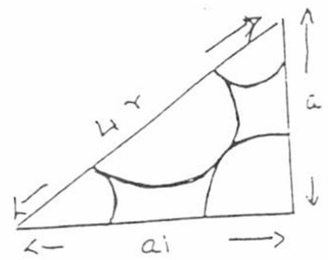
v = Volume of atoms in unit cell = $2 \times (4/3)\pi r^3$

Volume of unit cell $V = a^3$

APF = 68% eg. metallic crystals

Similarly for a face centred cubic.

FCC We can calculate the APF.



$N' = 12$, $2r = \frac{a\sqrt{2}}{2}$

$n = 4$ $\quad v = 4 \times \frac{4}{3}\pi r^3$

$V = a^3 = \frac{64r^3}{2\sqrt{2}}$

$P.F = 74\%$

eg Cu, Al, Ag

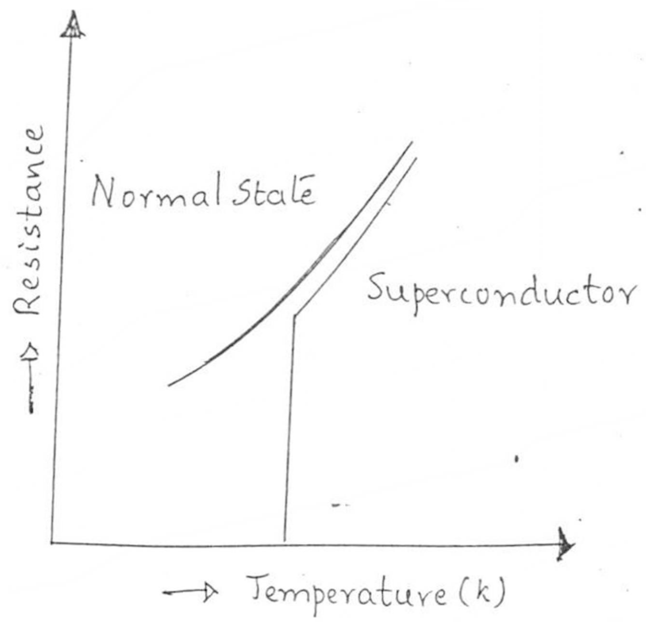
Mg, Co, Zn, Ti, Se have HCP structure. Most of the remaining one-third of the metals do

CHAPTER
SUPERCONDUCTIVITY

Superconductors, materials that offer no resistance to the flow of electricity, are one of the last great frontiers of scientific discovery. Super-conductivity has been observed in certain metals, alloys and ceramics. Not only have the limits of superconductivity not yet been reached, but also the theories that are used to explain superconductor behaviour seem to be constantly under review. The exotic phenomenon of superconductivity was first observed in mercury by the Dutch Physicist Kammerlingh Onnes of University of Leiden in the year 1911. When he cooled it to the temperature of liquid helium, 4 degrees Kelvin, its *resistance* suddenly disappeared. The Kelvin scale of temperature represents the "absolute" scale of temperature. The sudden transition to a state of no resistance was not confined to the pure metal but occurred even if the mercury was quite impure. The new state at which the electrical properties became quite unlike those previously known he called the "superconducting state".

Transition Temperature

The temperature at which a superconductor loses resistance is called its superconducting transition temperature or critical temperature Tc. This is different for different materials and is a characteristic of the given material or compound. In general the transition temperature



Temperature dependence of The resistance of a normal and superconducting metal.

is not very sensitive to small amounts of impurities, but the superconductivity of a few metals such as iridium, molybdenum, which in the pure states have very low transition temperatures, may be destroyed by the presence of minute quantities of magnetic impurities. Such elements hence exhibit superconductivity only if they are pure. Substances with regular lattice can only become superconductors. Imperfections in the lattice can render superconductivity impossible for substances with imperfect lattice have a finite resistance even at very low temperatures close to absolute zero. Ferromagnetics are not superconductors.

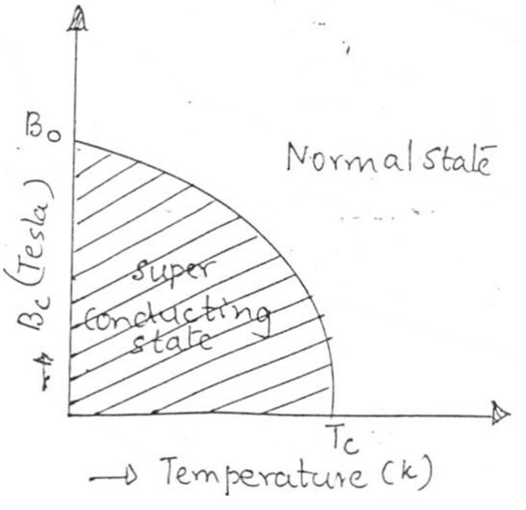
Critical field and Current Density

Superconductivity will disappear if the temperature of the specimen is raised above its transition temperature or if a sufficiently strong magnetic field or current density is made to flow through the superconductor. The applied field necessary to restore the normal resistivity is called the critical field B_c. Furthermore, superconductivity vanishes if the current flowing through the specimen exceeds a certain limit called the critical current I_c. Both B_c and I_c depend on temperature and on each other. Experimentally it is found that the critical magnetic field, at zero current depends on temperature as follows.

$$B_c = B_o \left[1 - \left(\frac{T}{Tc}\right)^2\right]$$

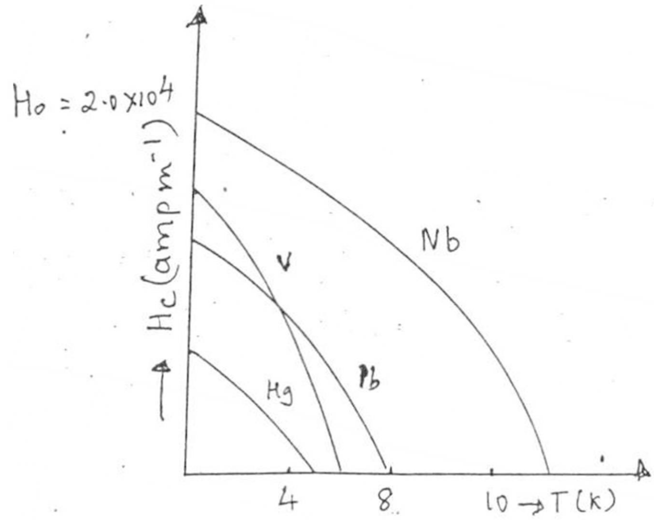
where B_o is the critical field at 0°K. Thus the field has its maximum value B_o at T = 0°K.

Critical Field.



The critical magnetic field at which superconductivity disappears.

Effects of Magnetic Field.



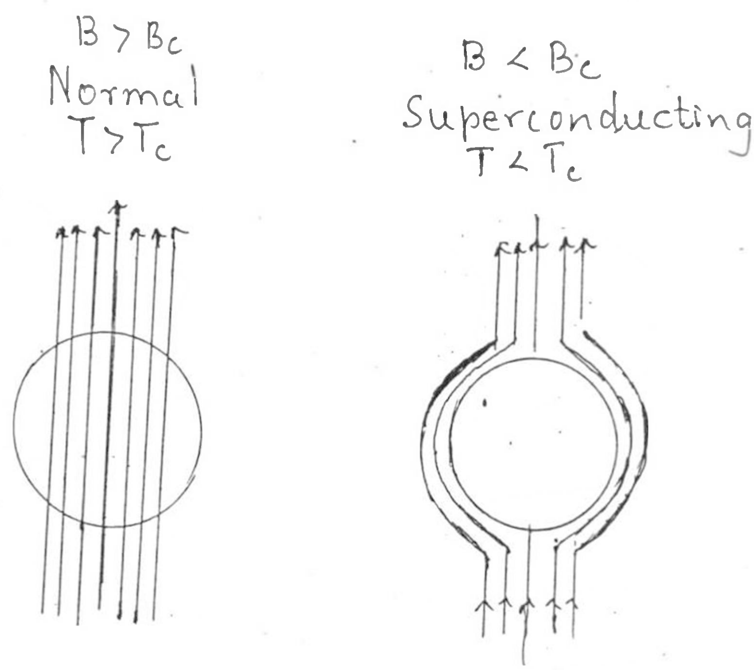
Variation of critical field as a function of temperature

The critical field B_c also depends on the material. We find that these materials are superconductors only for values of T and B below their respective curves and are normal conductors for values of T and B above these curves.

Magnetic Properties

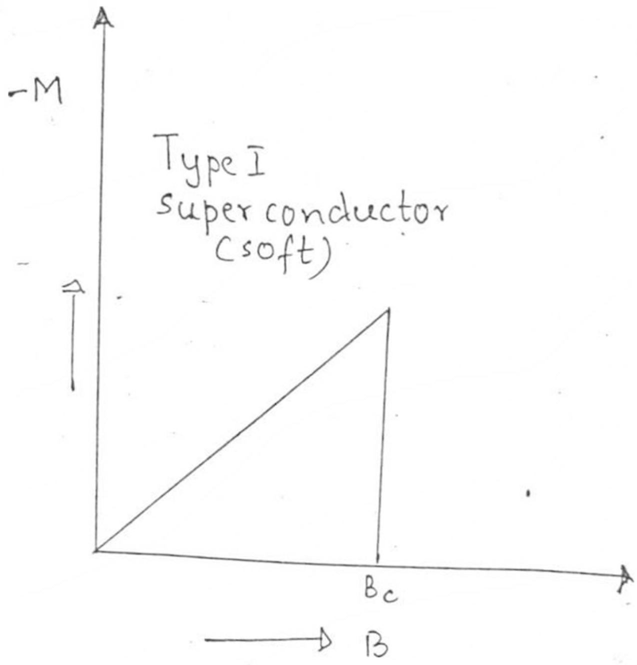
The magnetic properties of superconductors are as remarkable as their electrical properties. The ideal magnetic behaviour of superconductors falls into two classes. Type I and type II. Below the critical temperature and for $B < B_c$ the material is perfectly diamagnetic i.e. the field does not penetrate the superconductor. This behaviour illustrated in the figure below is called the Meissner Effect.

Meissner Effect.

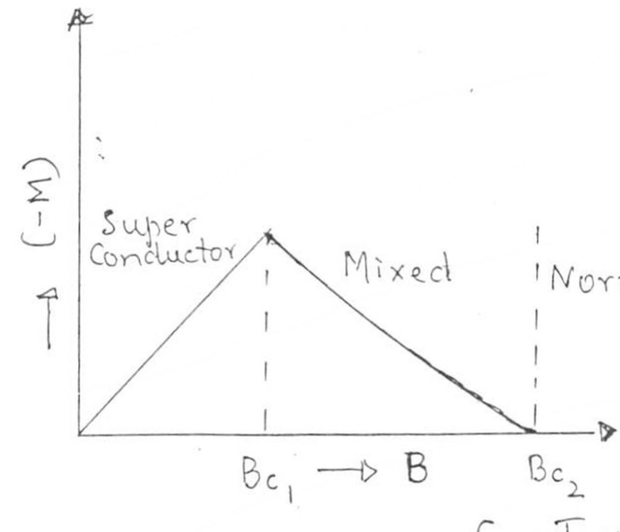


Magnetic field configuration for a spherical superconductor in an applied field.

As can be seen from the figure that when the specimen becomes superconducting the field is concentrated at the sides of the specimen, but not at the top or bottom. When this happens the specimen must exist as a mixture of the normal and superconducting states called the intermediate state. Consequently such specimens are either superconducting (if $B < B_c$) or normal if $B > B_c$.



Magnetisation v.s applied field for a perfect superconductor.



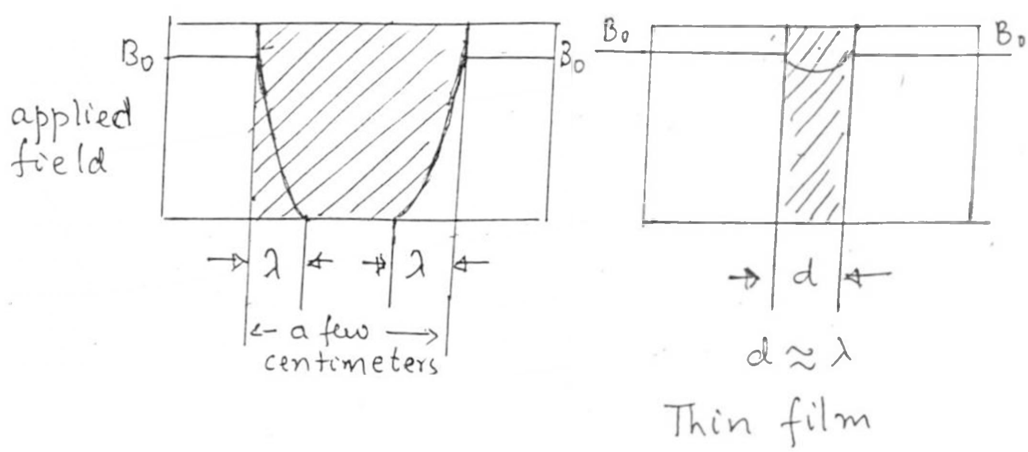
Magnetisation curve for Type II superconductors.

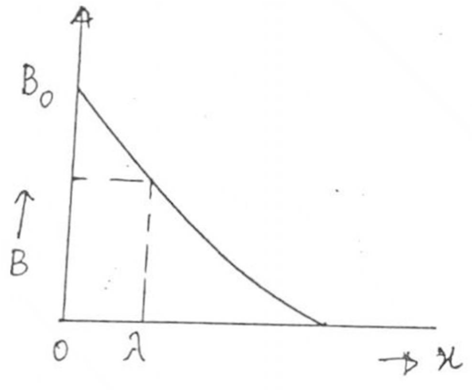
Type II superconductors behave differently. For applied fields below B_{c1} (called the lower critical field), the material is diamagnetic, field is completely excluded. At B_c the field begins to penetrate the specimen and the penetration increases until B_{c2} (called the upper critical field). At this field the magnetization vanishes and the specimen becomes normal. The magnetization of a type II superconductor vanishes gradually as the field is increased, rather than suddenly as in type I superconductors. Type I superconductors are also called Soft Superconductors, while type II are called hard superconductors.

Penetration Depth and thin films

The applied field does not suddenly drop to zero at the surface of a Type I superconductor, rather it decays exponentially. As a consequence, the field is fairly large over a distance from the surface. The penetration depth, λ, ranges from 300 to 5000 A depending on the material.

Penetration Depth





Decay of field.

In the above figure is depicted the penetration of the magnetic field into a bulk specimen and into a thin film whose thickness is less than the penetration depth. In ordinary specimens whose dimensions are much larger than 5000 A, the major fraction of the volume is not penetrated by the field.

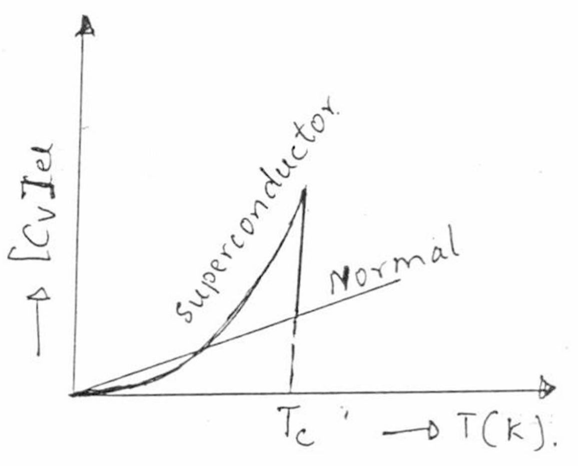
Persistent Currents

If a superconductor is in the form of a ring a current can be induced in it by electromagnetic induction. The resistivity of a superconductor can be measured by observing the induced current as a function of time. If the material is in the normal state, the current damps out quickly because of the resistance of the ring. But if the ring has zero resistance, the current once set up, flows indefinitely without decrease in value. In a typical experiment, a lead ring could carry an induced current of several hundred amperes for over a year without any change. Such currents are called "persistent currents'. Physicists found that the upper limit for the resistivity of a superconducting lead ring was about 10^{-25} Ωm. The fact that this is about 10^{-17} as large as the value at room temperature does indeed justify taking $\rho = 0$ for the superconducting state.

Thermal Properties

Thermal properties such as specific heat capacity and thermal conductivity of a substance change abruptly, when it passes over into the superconducting state.

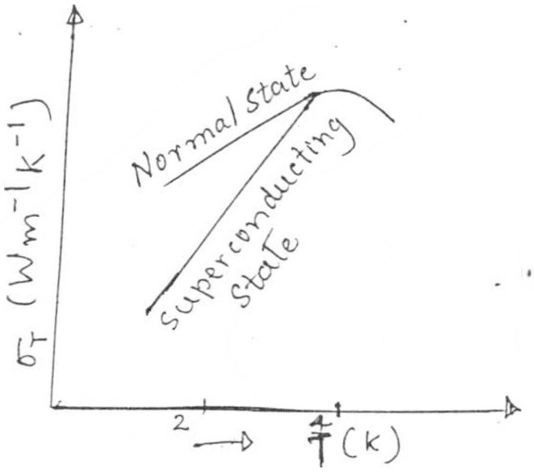
Specific Heat



Temperature variation of electron contribution to the specific heat of a conductor in the normal & super conducting states

$$C_v = [C_v]_{el} + [C_v]_{La}$$
$$= \gamma T + \beta T^3$$

Thermal Conductivity



Thermal conductivity of a specimen of tin in the normal and superconducting states.

According to the thermodynamic theory of superconductivity, the superconducting and normal states are two phases of a substance, each being converted into the other at definite values of the state variables the temperature T and the magnetic field intensity B. The conversion of a superconductor to the normal state by the action of a magnetic field i.e. at $T < T_c$ is a first order phase transition. The same convention in the absence of a magnetic field is a seond order phase transition.

Causes of Superconductivity

Though a number of a theories were proposed, the first widely-accepted theoretical understanding of superconductivity was advanced in 1957 by Ameican physicists John Bardeen, Leon Cooper and John Schreeffer. Their themes of superconductivity became known as BCS theory which fetched them a Nobel Prize in 1972. They used the idea advanced by

Cooper (1956) that pairs of electrons could condense into a lower energy phase provided that there was some attraction, however weak, between them. Accordingly, they were able to show that such an attraction does exist between electron pairs due to electron-phonon interactions. Phonon is a quantum of acoustic energy. According to the BCS theory the electrons responsible for superconductivity are coupled together in pairs called Cooper pairs. These electrons have opposite spins and equal and opposite momenta. The attractive force between the electrons of a pair extends over a relatively long distance of the order of 10^{-4} cm, called the distance of correlation.

The binding energy, 2Δ, between the two electrons is temperature dependent and becomes zero if the temperature approaches T_c. At absolute zero, all electrons are paired. For $T_c > T > 0$ some electrons are paired and others are excited. For $T > T_c$ there are no electron pairs. The excited electrons behave normally in every respect. The paired electrons are responsible for superconductivity. Since there are electron pairs for $T < T_c$, the material is a superconductor for $T < T_c$.

The mathematically complex BCS theory was successful in explaining superconductivity at temperatures close to absolute zero for elements and simple alloys. However, at higher temperatures and with different superconductor systems, the BCS theory has subsequently become inadequate to fully explain how superconductivity is occurring.

Another significant theoretical advancement came in 1962 when Brian D Josephson, a graduate student at Cambridge University, predicted that electrical current would flow between two superconducting materials – even when they are separated by a non-superconductor or insulator. His prediction was later confirmed and won him a share of the 1973 Nobel Prize in Physics. This tunneling phenomenon is today known as the "Josephson Effect".

High Temperature Superconductors

The year 1986 saw a breakthrough in the discovery made in the field of superconductivity. Alex Miller and Georg Bednorz, created a brittle ceramic compound that superconducted at the highest temperature then known 30K. What made this discovery remarkable was that ceramics are normally insulators. They don't conduct electricity well at all.

Researchers had not therefore considered them as possible high temperature superconductor candidates. The Lauthanum Bariem, Copper and Oxygen compound that Miller and Bordnoiz synthesised, behaved in a way which was not yet understood. Tiny amount of this superconducting copper oxide were found to be actually super-conducting at 58K, due to a small amount of lead having been added as a calibration standard-making the discovery more noteworthy. Muller and Berdnoz's discovery led to a lot of activity in the field of superconductivity. In an attempt to cook up ceramics of every imaginable combination leading to higher and higher T_cs by substituting Yttrium for Lanthanum an incredible 92 T_c was achieved today referred to as YBCO. This temperature is warmer than liquid nitrogen temperature which is a commonly available coolant. Additional milestones have been achieved eversince by using exotic and often toxic elements in the base *perovskite* ceramic. The latest world record T_c of 138 K is held by a molecule of Mercury, Thallium, Barium, Calcium, Copper and Oxygen, created in 1995. Under extreme pressure its T_c can be coaxed up even higher – approximately 25 to 30 degrees more at 300,000 atmospheres.

Though a lot of advancements in superconductor T_cs have been achieved in recent years, other discoveries of equal importance have been made. Researchers in 1997 discovered that at a temperature very near absolute zero an alloy of gold and indium was both a superconductor and a natural magnet. Conventional wisdom held that a material with such properties could not exist! Recent years have also seen the discovery of the first high-temp superconductor that does *NOT* contain any copper and the discovery of the first *plastic superconductor*!

Fullerenes also called buckyballs exist on a molecular level when 60 carbon atoms join in a closed sphere. When doped with more alkali metals the fullerene becomes a "fulleride" and will often superconduct. Fullerenes like ceramic superconductors are a fairly recent discovery. They are technically a part of the larger family of organic conductors. Organic conductor family includes : molecular salts, polymers and pure carbon systems. The molecular salts within this family are large organic molecules that exhibit superconductive properties at very low temperatures. They are also therefore referred to as molecular superconductors. About 50 organic superconductors have been found with T_cs ranging from 0.4 K to 12 K (at ambient pressure). Since these

T_cs are in the range of type I superconductors, engineers have yet to find a practical application for them. Their unusual properties have made them the focus of intense research. These properties include giant magneto resistance, rapid oscillations, quantum hall effect. Organic superconductors are composed of an electron donor (the planar organic molecule) and an electron acceptor (a non-organic anion). A few examples of organic superconductors are:

$(TMTSF)_2\ ClO_4$
[Tetramethyltetra seleniafulvene + acceptor]

$(BETS)_2GaCl_4$
[boro(ethylenedioxy)tetrathiafulvene + acceptor]

'Borocarbides' are another system of superconductors, which also contain ferromagnetic transition metals like iron cobalt or nickel disprove the fact that they cannot form superconductors. Boron and Carbon act as mitigator to his unwritten rule. In addition, when combined with elements that have unusual magnetic properties (like Holmium) some borocarbides exhibit. What is known as 're-entrant' behaviour? Below T_c where they should remain superconductive, there is a discordant temperature at which they briefly retreat to a 'normal' non-superconductive state. Most borocarbides contain a rare earth element. A few of the unique compounds are listed below.

Compound	T_c
YPd_2B_2C	23K
YNi_2B_2C	15.5K
$HoNi_2B_2C$	7.5K

Other superconductors are Ruthenates and only one polymer *Polythiophene* which has been successfully waxed into the super-conducting state. Such startling discoveries are forcing scientists to continually re-examine longstanding theories on superconductivity and to consider heretofore-unimagined combination of elements.

Uses of Superconductors

Magnetic excitation is an application where superconductors perform extremely well. Transport vehicles such as trains can be made to 'float' on strong superconducting magnets, virtually eliminating friction between the train and its tracks. Conventional magnets waste much of the electrical energy and are also physically much larger than the superconducting magnets. Use of MAGLEV vehicles however has not caught up in spite of the technology having been proven. The world's only MAGLEV train to be used commercially was in Birmingham, England which of course closed down in 1997 after 11 years of operation.

Superconductors perform a life caring function in the field of biomagnetism Magnetic Resonance Imaging (MRI). By impinging a strong superconductor – derived magnetic field into the body, hydrogen atoms that exist in the body's water and fat molecules are forced to accept energy from the magnetic field. They then release this energy at a frequency that can be detected and displayed by a computer.

SQUID (Superconducting Quantum Interference Device) is used in magneto-encelography. SQUID's are capable of sensing a change in magnetic field upto 100 billion times weaker than the force that moves the needle on a compass. With this technology, the body can be probed to certain depths without the need for the strong magnetic fields associated with MRIs.

Electric generators with superconducting wire are far more efficient that conventional generators wound with copper wire. Their efficiency is 99% and their size is half that of the conventional ones so that they are lucrative ventures for power utilities.

An idealised application for supercoductors is to employ them in the transmission of commercial power to cities. However, due to the high cost and impracticality of cooling miles of superconducting wire to cryogenic temperatures, this has only happened with short test runs. Most recently this month, workers pulled out nine cables from underground conducts at a Detroit Power Station, to be replaced by the first higher temperature superconductor cables in a working power grid.

In the electronic industry, ultra high performance filters are now being built. Since superconducting wire has near *zero resistance*, even at high frequencies, many more filter stages can be employed to achieve a desired frequency response. This translates into an ability to pass desired frequencies and block undesirable frequencies in applications such as cellular telephone systems.

Superconductors have also found widespread applications in the military. HTSC *SQUIDS* are being used by the US Navy to detect mines and submarines. *Significantly smaller motors* are being built for NAVY ships using superconducting wire and tape.

Among emerging technologies ultrasensitive, ultrafast *super-conducting light detectors* are being adapted to telescope due to their ability to detect a single photon of light. Superconductors may even play a role in Internet communications soon. Internet data traffic is doubling every 100 days and superconductor technology is being called upon to meet this *super* need.

Assuming a linear growth rate it is expected that the world wide market for superconductor products is to be nearly doubled between year 2010 and 2020. Should new superconductors with higher transition temperatures be discovered, growth and development in this exciting field would explode virtually overnight.

References

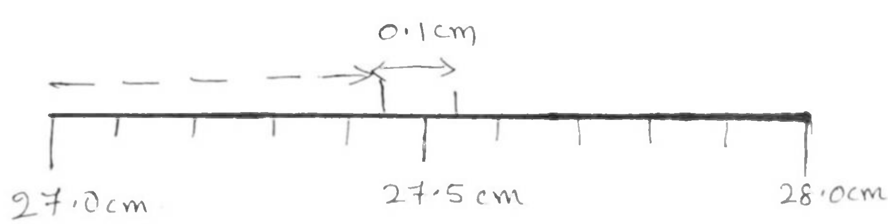
Charles Kittel, Solid State Physics, Asia Publishing House, New Delhi, 1965, pp.451-476.

Pillai S.O., Solid State Physics, Wiley Eastern Ltd., New Delhi, 1994, pp.350-357.

Website http://www.superconductor.org on the Internet.

III. Significant figures

Every physical quantity must have a unit, to tell what was counted, an order of magnitude and a statement about its reliability, which for the present we can indicate in a rough way by writing only the correct number of significant figures. The digits that are certain and one more are called significant figures. For example, in the statement that the length of the textbook is 27.5 cm, the digits 2 and 7 are certain and there is uncertainty of 0.1 cm in digit 5 because conventionally a length ranging from 27.45 to 27.5 is written as 27.5. The reading is known only to the nearest tenth of a centimeter. It has three significant figures. The error implied in this is 1 part in 275, i.e. $\frac{0.1}{27.5}\times 100 \cong 0.4\%$.



The speed of light in vacuum is 2.99792458×10^8 ms^{-1}. There are nine significant figures. The greater the number of significant figures the greater is the accuracy of our measurements. In our calculations speed of light can be taken as 3.00×10^8 ms^{-1} (to three significant figures).

Scientific notation power – of – 10 notation

A measured quantity is written in the form $m \times 10^n$ where $1 < m < 10$, and n is an integer, positive or negative. All digits in 'm' are significant.

For example, a measured quantity 0.00780 m is indicated as 7.8×10^{-3} m (two significant figures) if the measurer is certain of digit 7 and as 7.80×10^{-3} m (three significant figures) if he is certain of digit 8.

Propagation of uncertainty through arithmetic operation

The uncertainty in derived quantities is fixed by uncertainties in the measurements that are to be combined. Results are not improved by carrying out simple arithmetic operations to many figures. "No chain is stronger than its weakest link". We have an analogous situation with regard to measurements and their use in calculations. How to detect the weakest link and to judge how weak it is?

Addition and subtraction

An example : 36.34m + 0.0386 m + 4.133 m. On examination of these numbers, we find that 36.34 m is known only to the nearest one-hundredth of a metre, the other two numbers are known respectively to the nearest one-ten thousandth and one-thousandth of a metre. Therefore the weakest link is the first number, its accuracy is 0.01m. Therefore, it is enough if we know the sum to one-hundredth of a metre. Hence, we follow this step : round off the other numbers to one-hundredth's place and add.

The sum = (36.34 + 0.04 + 4.13) m = 40.51 m

The same rule is to be followed in case of subtraction. Note that the subtraction reduces the number of significant figures.

Multiplication and Division

In this case, the answer will have same significant figures as the factor having the least significant figures.

(1) For example, density = $\frac{33.35\,g}{15.42\,cm \times 5.53\,cm \times 2.7\,cm}$

Here the weakest link is 2.7 cm. The factor 2.7 has the least significant figure of two. So the density is expressed to two significant figures.

The step adopted is, density = $\frac{33.4\,g}{15.4 \times 5.53 \times 2.7\,cm^3} = \frac{33.4}{230}$

$= 0.1452$

$= 0.15\ g\ cm^{-3}$

(2) Multiplication involving constants such as 2, 1/4 , etc.

Area of a surface is $2.30 \times 10^{-2}\ m^2$. What is twice its area?

It is $2 \times 2.30 \times 10^{-2}\ m^2$ (three, significant figures and not one, since numbers like 2, ¼ etc. have unlimited accuracy).

(3) Multiplication involving constants like π, G, etc.

The diameter of a wire is 0.57 mm. What is its area of cross-section ?

The formula $A = \frac{\pi D^2}{4}$ is preferred to $A = \pi r^2$ (why?).

Further π = 3.14159265.

The measured quantity has two significant figures. Express π to 3 significant figures, i.e. one more – to take care of rounding off errors.

$$A = \frac{3.14 \times (0.57 \times 10^{-3})^2}{4} m^2 .$$

$$= 0.255 \times 10^{-6} \text{ m}^2$$

$$= 2.6 \times 10^{-7} \text{ m}^2$$

Note: These rules are not rigid ones. There are many examples which do not conform exactly.

Advantages

1. It provides an easy introduction to the existence of uncertainty in measurements.
2. It helps to avoid misleading numbers and unnecessary calculations when measured quantities are subjected to arithmetic operations.

Disadvantages

1. Significant figures furnish only a rough estimate of uncertainty or accuracy.
2. There is no single rule for all the four operations.
3. Significant figure omit reference to the accumulation of uncertainty as data are combined.

For example, the implied error in the final value of density referred to earlier i.e. $\frac{0.01}{0.15} \times 100 \cong 7\%$, whereas, the implied errors in l, b, h and m are 0.1%, 0.2%, 4% and 0.03% respectively. The error indicated in the final answer is on the higher side; the accumulated error (maximum possible error) in this measurement having, $(0.1 + 0.2 + 4 + 0.03)\% \approx 4.3\%$.

IV. Quantities to be measured with greater accuracy in an experiment

When measured quantities are substituted in a formula in order to calculate a desired physical quantity, the individual errors influence the uncertainty/ error in the final result. For example, in the simple pendulum experiment, to determine 'g', the acceleration due to gravity, the two measurements made are the length 'l' and the period 'T'. In this, which one of them is to be measured to a greater accuracy? Should we use a vernier calipers to determine the radius of the bob? The length l' from the point of suspension to the surface of the bob is measured using a metre scale to an accuracy of one millimeter. By measuring the diameter to an accuracy of 1/10 of a millimeter (the usual L.C. of the vernier caliper) and adding half that value to l', will not improve the accuracy. (Refer: – addition – significant figures). Therefore, it is enough to read the main scale of the vernier.

The other quantity T, occurs as squared in the formula for 'g'. If one commits an error of x% in T, its contribution to the error in g is 2x%.

[Note: The error in 'T' is reduced by measuring time 't' for a large number of oscillations say 20. If one is using a stop-clock (L.C. = 1s) and t = 40s then $T = \frac{40 \pm 1}{20}$ = (2.00 ± 0.05) S. The period is measured to an accuracy of 0.05 S.

Error is $\frac{0.05}{2.00} \times 100 = 2.5\%$. Its contribution to the error in g is 5.0%]. The error in

'l' is $\frac{0.1}{100.0} \times 100 = 0.1\%$ where 100.0 is the length of the pendulum set-up. Hence, the period 'T' has to be measured with greater accuracy].

In general, those quantities which have higher powers (exponents) in a formula are to be measured with greater care because their contribution to the error in the final result is (power (or exponent) × error in the quantity).

In the experiment to measure the resistivity of a wire, the quantity to be measured with less error is the diameter of the wire, rather than the length of the wire.

The resistance unplugged can be considered as a constant for expressing the final result in terms of significant figures.

In the determination of viscosity of a liquid by the Poiseuille's method, the quantity which is to be measured with greater care is the radius of the capillary tube because it appears in the formula as $(\text{radius})^4$.

Should we use the physical balance for weighing by the method of oscillation in the following cases ?

(i)	Calorimetry	- determination of specific heat
(ii)	Calorimetry	- determination of latent heat of steam/ice.
(iii)	Faraday's law of electrolysis	- determination of e.c.e. of copper.

Error – Its effect on procedure of an experiment

Suppose that we are to determine 'g' using a simple pendulum. One can set up a pendulum of a given length 'l' and measure the period T by the usual method. Substitute these values in the formula $g = 4\pi^2 \frac{l}{T^2}$ and calculate 'g'. The value of 'g' so obtained may be higher or lower because of the errors that crept in. How can one proceed to get a better result – errors being ironed out? The procedure should influence the value of 'g' on both sides. That means many sets of readings are to be taken and the average is to be found out. The formula tells us that $\frac{l}{T^2}$ is a constant. So, one can set up pendulum of different lengths and measure the corresponding periods in the usual way. Calculate $\frac{l}{T^2}$ in each case and substitute its mean value in the formula to determine g'. That is $g = 4\pi^2 \left(\frac{l}{T^2}\right)_{mean}$

The relation $\frac{l}{T^2}$ = constant suggests that the graph of T^2 on x-axis and 'l' on y-axis is a straight line. Its slope gives the value of $\left(\frac{l}{T^2}\right)$ and hence $g = 4\pi^2$ (slope).

The value of g determined this way is a better value as the best-fit line drawn further "irons" out the errors. Note that the formula for $g = \frac{4\pi^2}{slope}$ if 'l' is plotted on x-axis and T^2 on y-axis ($l - T^2$ graph).

Thus, the attempt to reduce the errors, determines the procedure that is adopted (many sets of readings, graphical analysis and the average of the relation between the variable quantities in the formula). Also note that we do not calculate 'g' in each set and then take the average.

Some aspects to be kept in mind while collecting and recording data

1. The values selected for independent variable must be convenient ones for (i) plotting, and (ii) calculation. For example, in simple pendulum experiment, let the lengths be a whole number like l = 60.0 cm and not as 60.2 cm.
2. Among the values of the independent variables, let there be the values which are multiples of initial value (some times 1.25, 1.50,... etc. of the initial value). e.g. l = 90.0 cm, 120.0cm etc. This enables to find out the proportionality: when one quantity is doubled is the other doubled?
3. While tabulating, let the independent variables be arranged in an order (increasing/ decreasing). This enables the experimenter to see the proportionality.
4. Let the number of trials be at least five (in case of graphical analysis).
5. Let the unit be written only on top of the tabular column.
6. Let the measured quantities be entered to the required number of decimal places, in accordance with the least count. E.g. l = 60.0 cm and not 60 cm when the least count is 0.1 cm.
7. While establishing relations, the physical quantities can be measured in any convenient and arbitrary unit and not necessarily in S.I. units. E.g. (1) When a body moves with uniform velocity, unit for time can be in terms of distance, (2) range of the projectile can be a measure of the speed and hence momentum (if masses of objects are the same).
8. Train ~~the~~ students to record the measured values in ink in the tabular column. If wrong, let them cancel it with a neat single stroke.

V. Data analysis – Graphical method

The data tabulated can be analysed by arithmetic and graphical methods. The latter is a convenient one because

(i) the comprehension of the relation is easy and quick, and

(ii) the graph drawn properly makes the uncertainties to affect on either side – thus 'ironing' out the errors. The value calculated, thus, is a *better* one.

Mathematically, a graph is a curve or other line representing relation of the elements in a equation or function ($y = mx$, $y = mx + c$, $y = ax^2$, $y = ax + bx^2$ etc).

In data analysis, it is a line or diagram showing how one quantity depends on or changes with another. The following are the features of importance :

(i) the nature
(ii) the intercept
(iii) the area under the curve, and
(iv) the slope

1. Plotting a graph : Choice of quantities to be plotted on x and y-axes

We need not follow strictly the con~~nectivity~~vention – independent variable on x-axis and dependent variable on y-axis. Let us take them in such a way that we get useful quantities in less number of steps. For example, if we plot the displacement on y-axis and time on x-axis, the slope gives the mean speed. If the quantities are interchanged, the reciprocal of the slope gives the mean speed. In this case, an additional step is involved which could be avoided.

2. Marking of x- and y-axes

Let x- and y-axes are marked on the edges of the ruled portion itself.

3. *Selection of scale*

The graph plotted must cover as large a portion of the graph sheet as possible. This is done by taking as large a scale as possible. However, while selecting the scale, *convenience* has to be looked into. For this, let the scale be

1 cm = 1, 2, 5,10, 20, 50, 100....etc. units

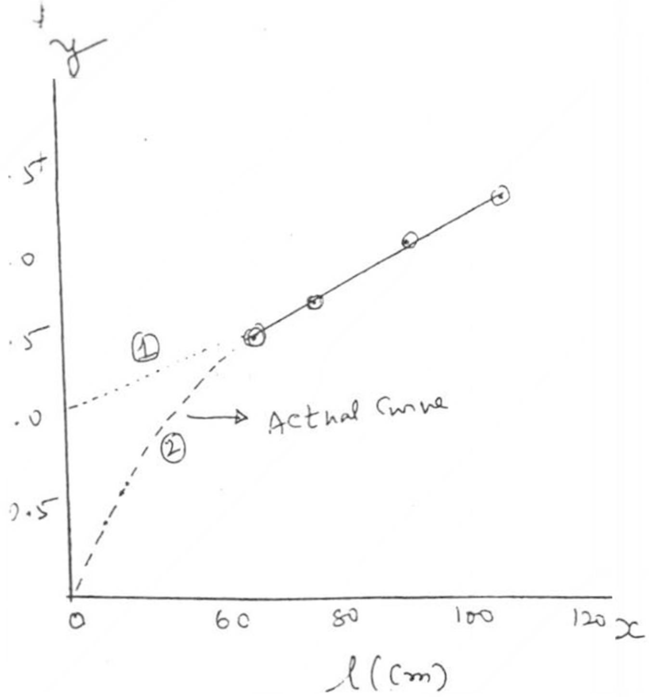
= 0.1, 0.2, 0.5.......

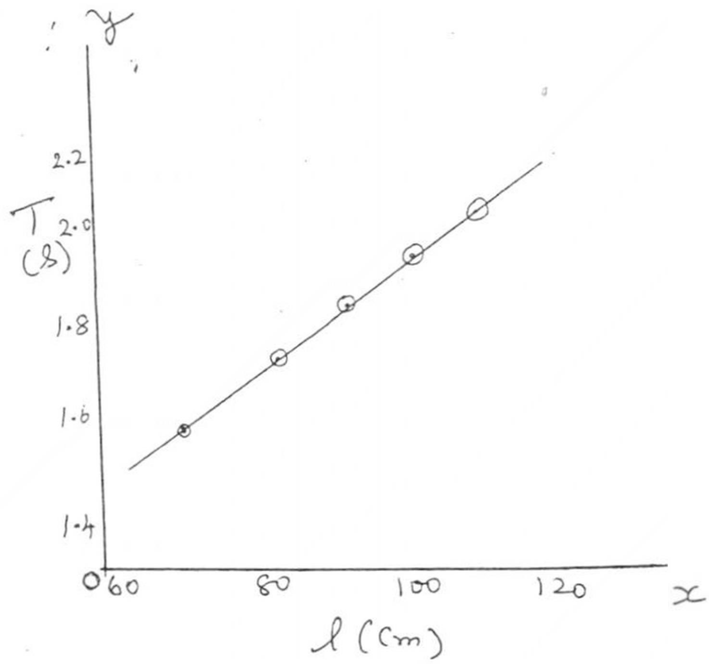
= 0.01, 0.02, 0.05,..... and so on.

For marking the graduations express the values in scientific notation: $m \times 10^n$ where $1 < m < 10$.

4. *Choice of the origin*

(i) When we are interested in the nature of the relationship among the two quantities the origin has to be (zero, zero). For example, in the simple pendulum experiment, suppose we plot the graph between l and T, taking the values over which the measurements are made, the plotted points appear to be on a straight line. We may draw a wrong conclusion that T is directly proportional to l.





Exact relation is obtained when we plot the graph with (0,0) as origin. The graph appears to be a straight line. However, if you extrapolate the graph, you find the graph cutting time axis at a finite point, corresponding to a pendulum of zero length. This is absurd, which means our assumption that l – T graph is a straight line is wrong. If the students are to determine the relation, you can instruct them to modify one of the quantities and draw graphs such as $l - T^2$, $\sqrt{l} - T$ etc. Let them try $l - T^2$ since squaring is easier than taking the square root. [In case, you are interested in $\sqrt{l} - T$ graph, then the independent variable can be assigned values 49.0, 64.0, 81.0, 100.0 and 121.0 cm]. This graph is a straight line, passing through the origin. Hence $T^2 \propto l$, or $T \propto \sqrt{l}$.

Similar cases will be observed in the following experiments.

(a) oscillations of a liquid column

(b) oscillations of a spring-mass system.

In this case, T^2 - M graph will not pass through the origin. There will be an intercept on M-axis, which represents the corrections for the oscillating mass due to the mass of the spring. This value is about one-third the mass of the spring.

(ii) If one is interested in the value of the slope alone, then it is enough to accommodate the range of values measured in each of the axes, taking as large a scale as possible, but a *convenient* one.

(iii) Suppose that the quantity is measured using a precision instrument, say a traveling microscope of least count 0.001 cm and if the least count (value of one division in the axis) of the axis on which they are represented is 0.01 cm, then round off the measured values to 1/100th place (i.e. multiples of L.C on the axis) and plot.

5. ***Drawing a line or curve***

Draw a small circle around the plotted points. In case of a straight line, draw using a *transparent scale*, a best-fit line, which contains as many points as possible on the line, and the other points being scattered equally on either side. If (0,0) is a point (as in the simple pendulum experiment), the best-fit line must be drawn from it. In the case of a curve, just do not join the successive points by a curve, but draw a 'free-hand' smooth curve again with as many points as possible on the curve and the rest equally scattered on either side of it. A flexible stick or a tongue cleaner can be used.

6. ***Drawing inference***

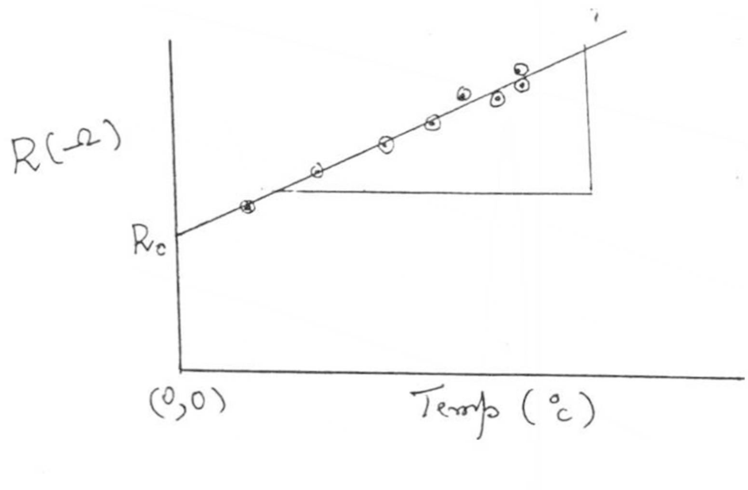
From the graph, first tell how the line is and then state the relation.

(i) A straight line passing through the origin → there is a direct proportionality between the elements of the graph i.e. $y \propto x$.

[Theoretically if the graph is to pass through the origin and with the experimental data, it does not exactly pass through the origin, then give an account for the discrepancy].

(ii) We prefer linear graph. When two quantities measured indicate inverse relationship, then plot one quantity against the reciprocal of the other. For example, in Boyle's law, $P = \left(\frac{1}{V}\right)$ graph. Note P-V graph is a rectangular hyperbola.

(iii) If the graph is a linear one, but has an intercept, then measure the intercept and indicate what does that represent. For example, in the temperature coefficient



of resistance, the intercept on resistance axis gives the resistance of the wire at 0° C.

$R_t = (R_o \alpha) t + R_o$

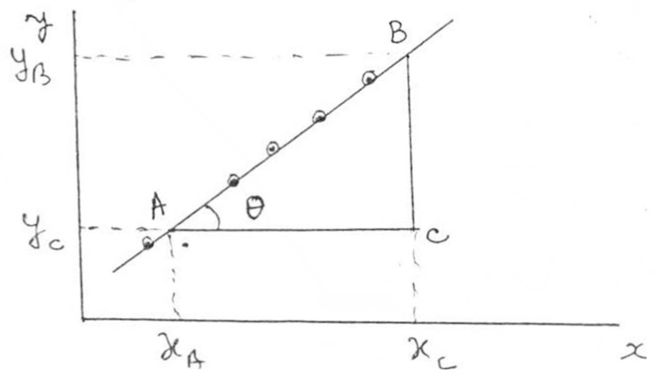
Slope = αR_o

$$\alpha = \frac{slope}{R_o}$$

$$\alpha = \frac{slope}{intercept}$$

7. ***Determining the slope of the graph***

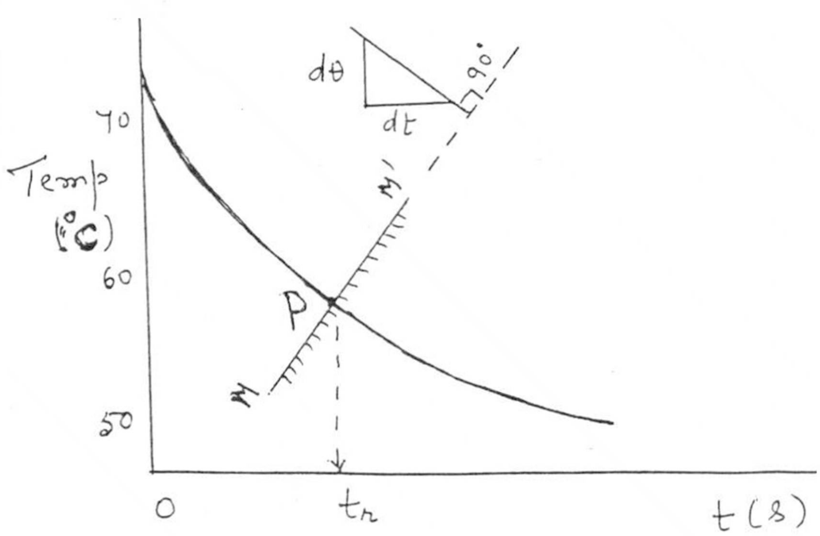
While determining the slope, don't select the plotted points for the triangle to be drawn. Instead, go along the graph and identify the two points on the graph, as far apart as possible where the graph passes through the intersection of the x- and y- lines.



The slope has to be calculated as the ratio of the values which the lines BC and AC represent and NOT of their geometrical lengths. Record the values of BC and AC in terms of appropriate significant numbers, taking L.C on the axis into account. Identify what quantity does the slope represent.

$$\tan\theta = slope = \frac{BC}{AC} = \frac{y_B - y_C}{x_C - x_A}.$$

8. ***Drawing a tangent to a curve at a given point***



Consider the temperature – time curve in the experiment: Newton's law of cooling. We wish to find the rate of cooling at a given temperature t_r. One method is to draw a tangent to the curve and determine its slope. (This gives the *instantaneous* rate of cooling). Place a plane mirror strip MM' across the curve corresponding to the given point P. Rotate the mirror about this point, till the portion of the curve in front of the mirror and its image through the mirror appears continuous. Trace the mirror surface line MM'. Draw a normal to it at that point or at any point on an extended line MM'. (This avoids the crowding of line MM' normal and triangles drawn to determine the slope for them). Determine the slope, as usual.

Instantaneous velocity at a given instant can be measured by the same method on displacement – time graph.

I.C.LOGIC GATES.

AIM: To construct NAND,NOR,NOT,AND,OR and EX-OR logic gates and study their characteristics.

INTRODUCTION: The building blocks of modern digital computers are the logic circuits. Logic circuits are designed using logic gates. the OR,AND and NOT are called the basic gates. The NAND and NOR gates are called universal gates because the basic gates can be realised from them. Hence we can realise the action of all the gates using NAND or NOR logic gates. ICs 7400 and 7402 are comercially available digital ICs with four 2-input NAND gates & four 2-input NOR gates respectively. These two ICs are available as 14-pin dual-in-line packages.

EXPERIMENTAL PROCEDURE: A) USE OF NAND LOGIC IC.

NAND logic: The circuit shown in fig. 1 is asembled. The two inputs to the gate are A and B.The output of the gate is observed for variuos input conditions as shown in the Truth Table 1 and the table is completed.

AND logic: The circuit shown in fig.2 is assembled.The output of the gate is observed for various input conditions and the Truth Table 2 is completed.

OR logic: The circuit shown in fig.3 is assembled.The output of the gate is observed and the Truth Table 3 is completed.

NOT logic : The circuit shown in fig. 4 is assembled. The output of the gate is recorded and the Truth Table 4 is completed.

NOR logic: The circuit shown in fig. 5 is assembled. The output of the gate is observed and the Truth Table 5 is completed.

EX-OR logic: The circuit shown in fig. 6 is assembled. The output of the logic gate is recorded and the Truth Table 6 is completed.

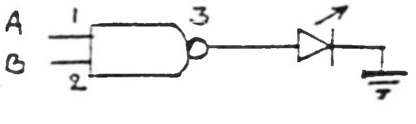
B) USE OF NOR LOGIC IC:

Circuit connections are made for NOR logic (fig.7),AND logic (fig.8), OR logic (fig.9),NOT logic (fig.10),NAND logic (fig.11) and EX-OR logic (fig.12) in that order. For each gate the output is recorded and the corresponding Truth Table is completed.

[Signal binary '1' to a gate means connecting the input terminal to Vcc or +ve terminal of the DC source.Signal '0' means connecting the gate terminal to GND or -ve terminal of the DC source. The output of the gate is recorded with the help of the LED.If the LED lights up the output is '1'. If it remains OFF the output is '0'.For all the circuits to function,pin 14 of the IC is connected to +5 V and pin 7 to GND of the DC source.]

Fig.1

7400-NAND LOGIC.

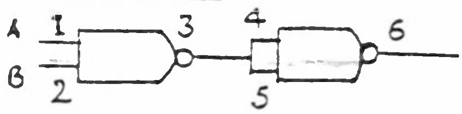


Truth Table 1

Input Level A	Input Level B	Output Level $\overline{A \cdot B}$
0	0	
0	1	
1	0	
1	1	

Fig.2

7400-AND logic

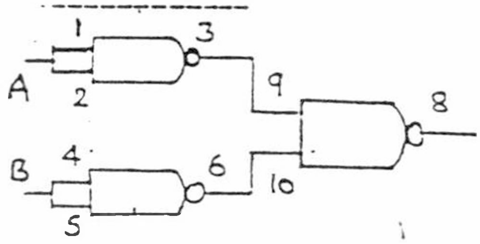


Truth Table 2

Input Level A	Input Level B	Output Level · A.B
0	0	
0	1	
1	0	
1	1	

Fig.3

7400- OR logic

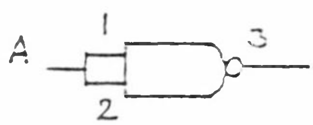


Truth Table 3

Input Level A	Input Level B	Output Level A + B
0	0	
0	1	
1	0	
1	1	

Fig.4

7400-NOT logic

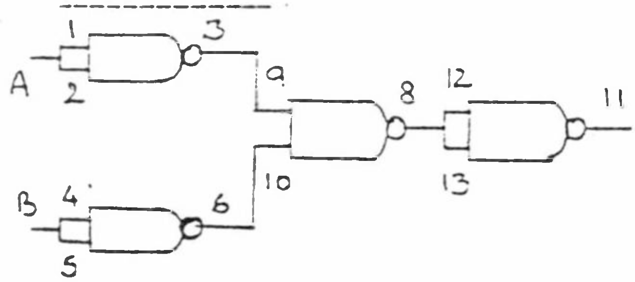


Truth Table 4

Input Level A	Output Level $\overline{A}$
0	
1	

Fig.5

7400-NOR logic

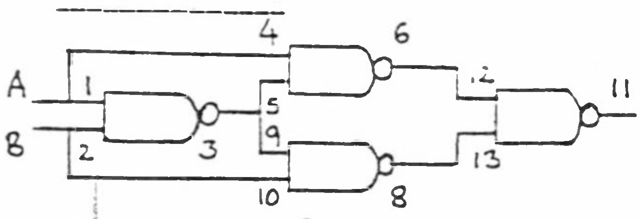


Truth Table 5

Input Level A	Input Level B	Output Level $(\overline{A + B})$
0	0	
0	1	
1	0	
1	1	

Fig.6

7400-EX-OR logic

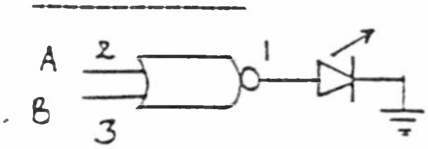


Truth Table 6

Input level A	Input level B	Output Level $A \oplus B$
0	0	
0	1	
1	0	
1	1	

Fig.7

7402-NOR logic

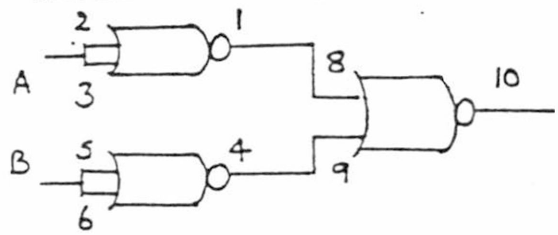


Truth table 7

Input Level A	Input Level B	Output Level $(\overline{A + B})$
0	0	
0	1	
1	0	
1	1	

Fig.8

7402-AND logic

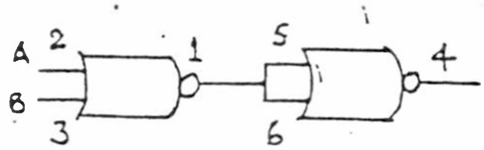


Truth Table 8

Input Level A	Input Level B	Output Level A . B
0	0	
0	1	
1	0	
1	1	

Fig.9

7402- OR logic

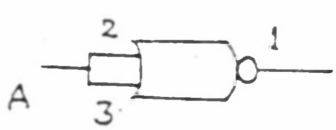


Truth Table 9

Input Level A	Input Level B	Output Level A + B
0	0	
0	1	
1	0	
1	1	

Fig.10

7402-NOT logic

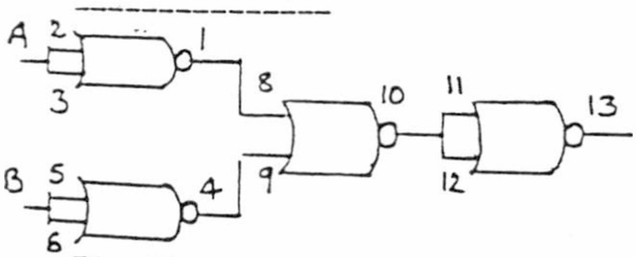


Truth Table 10

Input Level A	Output Level $\overline{A}$
0	
1	

Fig.11

7402-NAND logic

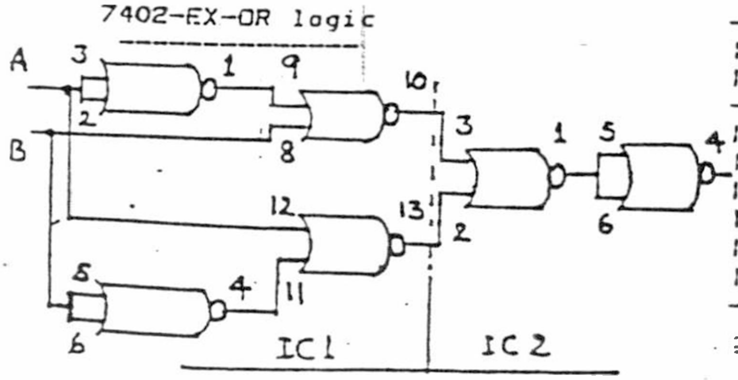


Truth table 11

Input Level A	Input Level B	Output Level $\overline{A \cdot B}$
0	0	
0	1	
1	0	
1	1	

Fig.12

7402-EX-OR logic



Truth Table 12

Input Level A	Input Level B	Output Level A ⊕ B
0	0	
0	1	
1	0	
1	1	

LOGIC GATES USING SEMICONDUCTOR DEVICES

Aim :

To design and study the characteristics of logic gates using semiconductor devices.

Apparatus :

A few diodes (BY 127), a few transistors (AC 127), a few resistors, LEDs, a 5v dc power supply.

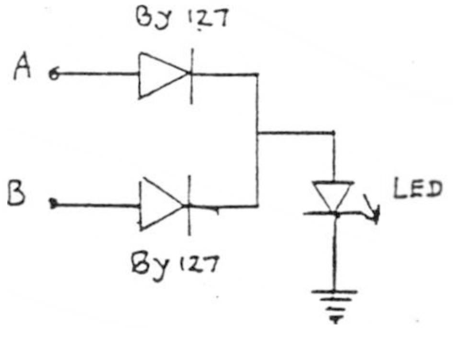
Theory :

Logic gates are fundamental building blocks of a digital computers. A computers ability to solve a problem depends on its ability to make decisions as it progresses through the steps in the problem. Circuits that make decisions are called logic circuits. These decisions are of Yes, No variety, that is of the two-state types logic circuits can be in one of two positions - ON or OFF, HIGH or LOW. Information and logical conditions are represented by a dc level on a signal line. Logic circuits analyse a combination of line levels at their input and produce a desired output when the input combination is correct for that particular circuit. Some of the Logic circuits are OR, AND, NOT, NAND and NOR. These can be designed using semiconductor devices. The functioning and the output of a logic gate is given by the Truth table.

Experimental Procedure :

The following electrical connections are made one after another and the output is noted for the various input conditions for each logic. The truth table is then drawn. Signal input is represented in binary form. Signal input A is binary 1 when A is connected to +5V of the dc source. A is binary 0 when it is connected to GND (-ve of dc source). The output is binary 1 when the LED lights up and it is binary 0, when LED does not lights up.

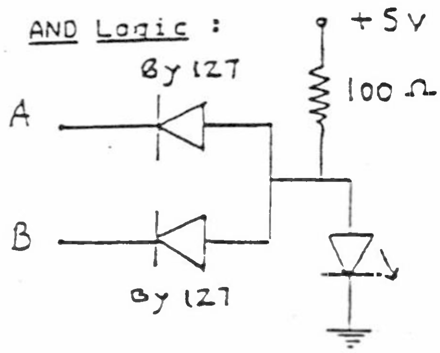
OR Logic :



Truth Table 1

Input dc level		Input Binary level		Output Binary level
A	B	A	B	(A + B)
GND	GND	0	0	
GND	5V	0	1	
5V	GND	1	0	
5V	5V	1	1	

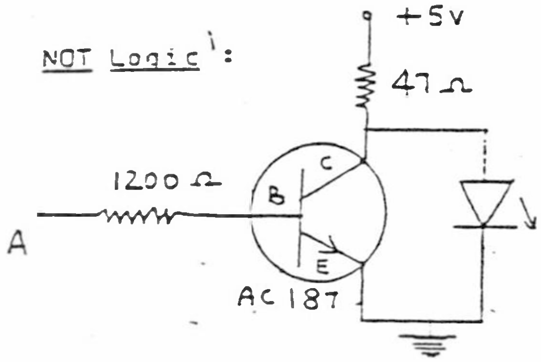
AND Logic :



Truth Table 2

Input Level		Output level
A	B	(A . B)
0	0	
0	1	
1	0	
1	1	

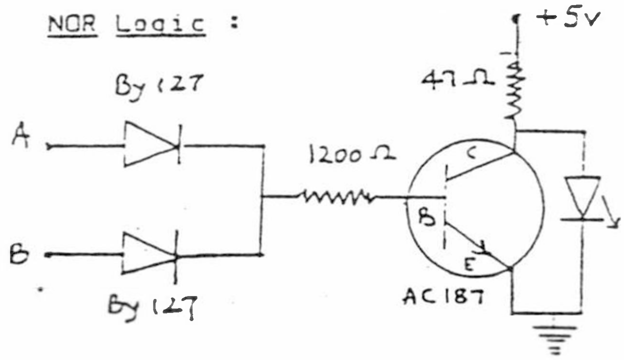
NOT Logic :



Truth Table 3

Input	Output
A	$\overline{A}$
0	
1	

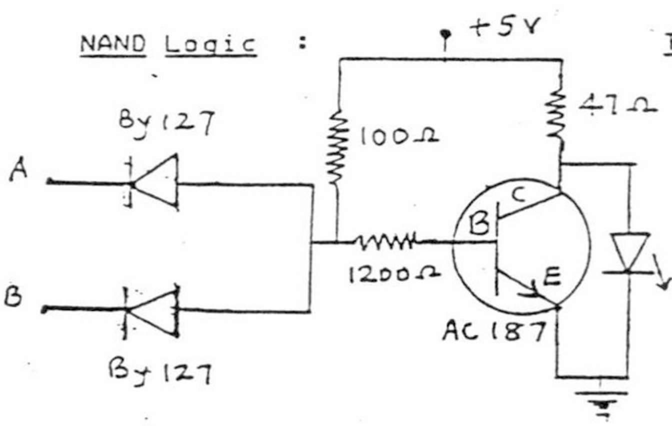
NOR Logic :



Truth Table 4

Input Level		Output level
A	B	($\overline{A + B}$)
0	0	
0	1	
1	0	
1	1	

NAND Logic :



Truth Table 5

Input Level		Output level
A	B	($\overline{A . B}$)
0	0	
0	1	
1	0	
1	1	

SE3.TXT

HALF ADDER AND FULL ADDER CIRCUITS.

AIM: To design half and full adder circuits and study their characteristics.

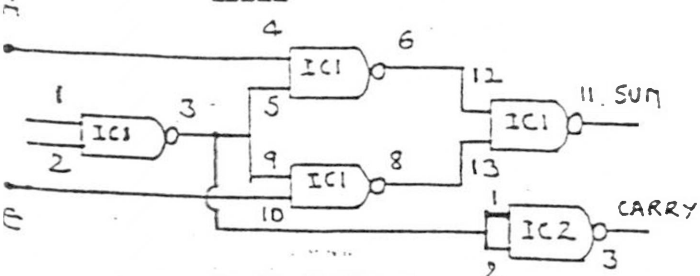
INTRODUCTION: A digital computer contains circuits which can perform arithmetic operations like addition ,subtraction,multi ication and division.The basic operationas are addition and subtrac as multiplication is repeated addition and division is repeated sub ction. Even subtraction can be achieved by using adders. Hence the computer can be built using adders only for arithmetic operations.

The simplest binary adder is the half adder capable of adding two bits at a time providing a sum output and a carry output necessary. A half adder can be realised using NAND logic or alterna using the combinational logic of EX-OR and AND gates. The half adde two inputs and two outputs.

The full adder has 3 inputs and two outputs. It is use add three binary digits at a time. The simplest way to connect a ful adder circuit is two use two half adders and an OR gate.

EXPERIMENTAL PROCEDURE: The circuit shown in fig. 1 is assembled. For various input conditions shown in the Truth the output is recorded.The Truth Table (1) is completed. The circui shown in fig.2 is now assembled and the above procedure is repeated. The full adder circuit shown in fig.3 is assembled. For the differer input conditions shown in Truth Table (2) the output is recorded anc table is completed.The results are verified by actual binary additic [A or B or C '0' means connecting the corresponding pin to GND c -ve terminal of the DC source. A or B or C '1'means connecting the pin to +5V or +ve terminal of the source.]

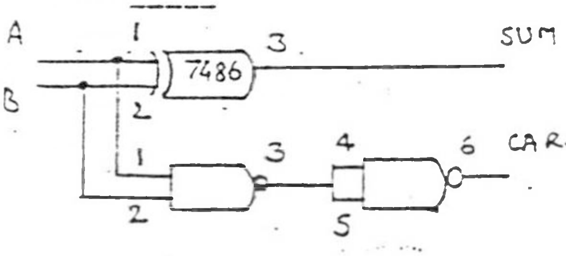
Fig.1



Truth Table 1

Input Level A	Input Level B	Output Level Sum (A + B)	Output Level Carry (A.B)
0	0		
0	1		
1	0		
1	1		

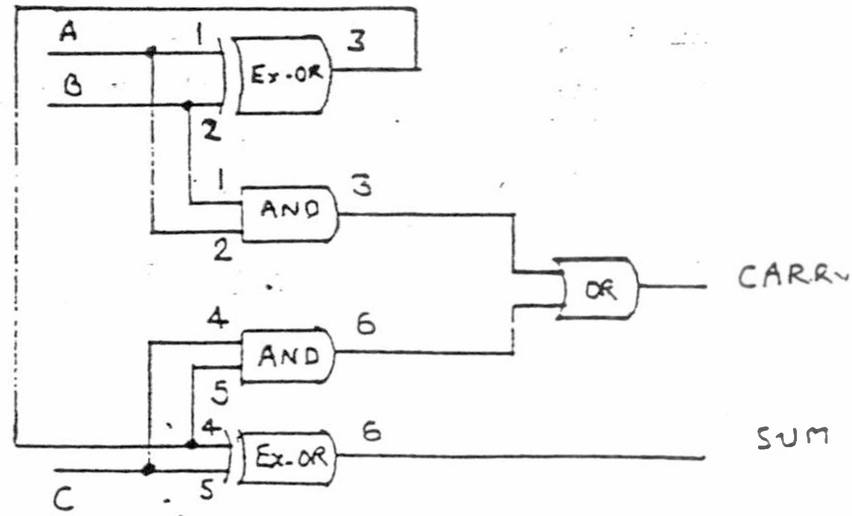
Fig.2



Truth Table 1

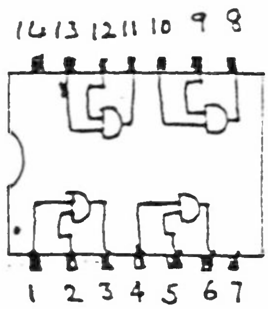
Input Level A	Input Level B	Output Level Sum (A + B)	Output Level Carry (A.B)
0	0		
0	1		
1	0		
1	1		

Fig.3

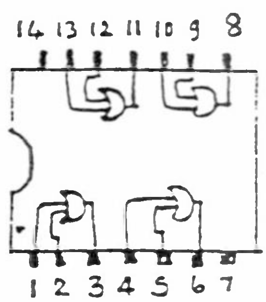


Truth table 2

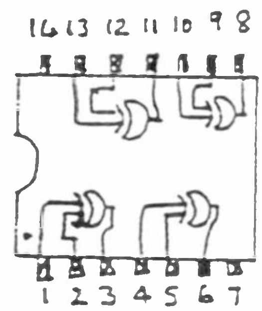
Input Level A	Input Level B	Input Level C	Output level Sum	Output level Carry
0	0	0		
0	0	1		
0	1	0		
0	1	1		
1	0	0		
1	0	1		
1	1	0		
1	1	1		



IC 7408
(AND)



IC 7432
(OR)



IC 7486
(Ex-OR)

[PIN 14 TO +5V AND PIN 7 TO GND FOR ALL THE ABOVE ICs.]

PARALLEL ADDER AND SUBTRACTOR

AIM :

To construct and use parallel adder, half subtractor and full subtractor.

APPARATUS :

Trainer board with IC-7486, Adder Subtractor trainer board, 5V dc power supply, LED indicator.

PARALLEL ADDER :

Parallel adders are employed in digital computers. In parallel addition all the binary words to be added are applied to the inputs simultaneously. Parallel adders are faster. 7483 IC is a commercially available 4-bit parallel adder IC. Two four bit binary words can be added at a time with the 7483 IC.

Experimental Procedure :

The circuit shown in fig.1 is rigged up. The inputs for circuit shown in fig.1 are two binary words A4, A3, A2, A1 and B4, B3, B2, B1. Different input words are fed and the outputs recorded in table 1 and the truth table is completed.

SUBTRACTORS :

The rules of binary subtraction are 0-0 = 0, 1-0 = 1, 0-1 = -1 and 1-1 = 0. We refer 0-1 = -1 as being a difference 1 and a borrow of 1.

HALF SUBTRACTOR :

HALF SUBTRACTOR subtracts two binary bits according to the binary subtraction rules and produces a borrow & a difference. A Half subtractor can be constructed using EX-OR, NOT and AND gates as shown in fig.2. The circuit is assembled and for various input conditions the output is recorded and the truth table-2 is completed.

FULL SUBTRACTOR :

A full subtractor can be constructed using two half subtractors. The circuit as shown in fig.3 is rigged up. For various input conditions the output is recorded and the truth table-3 is completed.

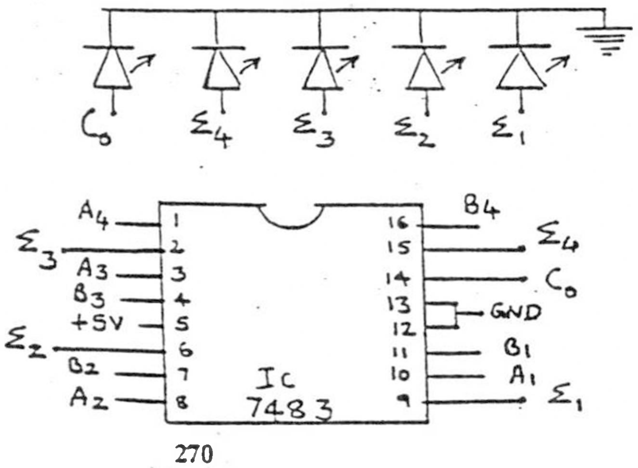


Table - 1

Tr. No.	INPUTS								OUTPUTS				
	A_4	A_3	A_2	A_1	B_4	B_3	B_2	B_1	C_o	Σ_4	Σ_3	Σ_2	Σ_1
1													
2													
3													
4													
5													
6													
7													
8													

Table - 2

A	B	Difference	Borrow
0	0		
0	1		
1	0		
1	1		

Table - 3

A	B	B_{in}	Difference	Borrow
0	0	0		
0	0	1		
0	1	0		
0	1	1		
1	0	0		
1	0	1		
1	1	0		
1	1	1		

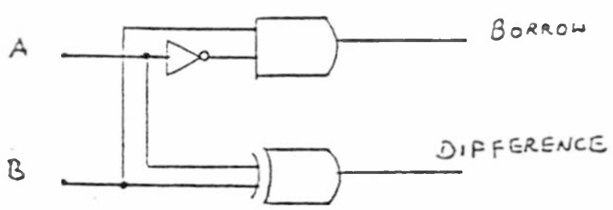


Fig. 2.

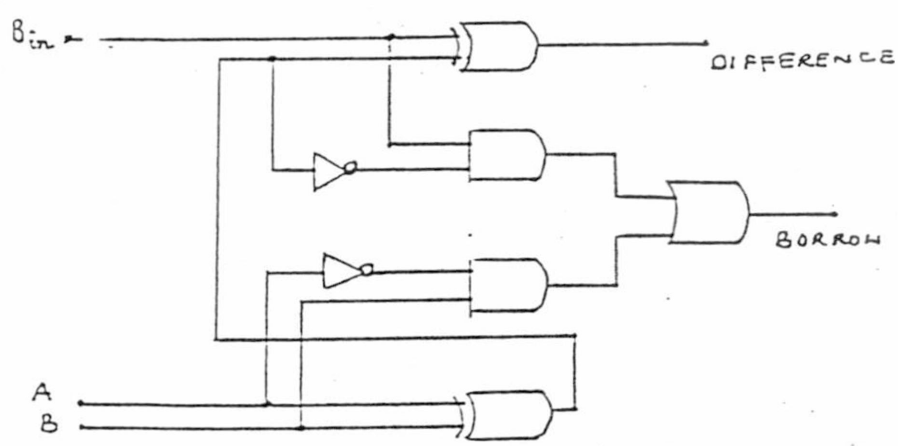


Fig. 3.

TE14.T

ADOLESCENCE

Need and Importance

The term adolescence comes from the Latin word *adolescere*, meaning "to grow" or "to grow to maturity". People of the earlier civilizations considered the child became an adult when capable of reproduction. Today the term adolescence has a broader meaning. It includes mental, emotional and social maturity along with physical maturity.

Psychologically, adolescence is the age when the individual becomes integrated into the adult society. This integration into adult society has many affective aspects, more or less linked with puberty. It also includes profound intellectual changes.

The Adolescent Years

It is customary to regard adolescence as beginning when children become sexually mature and ending when they reach the age of legal maturity. Research studies have revealed that changes in the individual are more rapid in the earlier year of adolescence than the latter. The period adolescence may be divided into two - early (12 - 15 years) and late (16 - 20 years).

Adolescence is an important period

The period of adolescence is important due to both immediate effects and long term effects. The immediate effects are the onset of puberty and accompanying rapid physiological and mental developments. These give rise to long term effects, like the need for necessary mental adjustments and the need for establishing new attitudes, values and interests, thereby entering adulthood normally.

Adolescence is a transition period

It is a passage from one stage of development to another. It means that what has happened before will leave its mark on what happens now and in the future. While moving from childhood to adulthood, they must "put away childish things". There is confusion about the roles the individual is expected to play. He/She is neither a child nor an adult.

Adolescence is a problem age

In childhood most of their needs and problems are taken care of by parents and teachers, as a result many adolescents are inexperienced in coping with problems alone. Secondly, because adolescents want to feel independent they demand the right of coping with their own problems, rebuffing attempts on the part of parents and teachers to help them. Very often, they are unable to solve problems not due to individual incapacity but because all their energies are engaged in battling the rapid physiological and mental changes that are problematic.

Adolescence is a time of search for identity

Throughout late childhood, known as the "gang age", they conform to group standards. Gradually in early adolescence, they crave for identity and are no longer satisfied to be like their peers in every respect, as they were earlier. One of the ways they try to establish themselves as individuals in late adolescence is by using status symbols in the form of owning a vehicle, clothes and other readily observable material possessions. They hope, in this way, to attract attention to themselves and to be recognized as individuals while, at the same time, maintaining their identity with the peer group.

Adolescence is a dreaded age

Many popular beliefs and popular stereotypes have definite evaluative connotations and, unfortunately, many of them are negative. The beliefs of adults that teenagers are sloppy, unreliable individuals, inclined towards destructiveness and anti-social behaviour influence the self-concepts and attitudes of adolescents towards themselves. This makes their transition difficult as they are unable to seek help from parents or teachers as all adults have poor opinion about them.

Adolescence is the threshold of Adulthood

As adolescents approach legal maturity, they are anxious to shed the stereotype of teenagers and to create the impression that they are near adults. Dressing and acting like adults, they discover, are not always enough. So, they begin to concentrate on behaviour that is associated with adult status - smoking, drinking, using drugs and engaging in sex, for example. They believe that this behaviour will create the image they desire.

DEVELOPMENTAL TASKS OF ADOLESCENCE

Havighurst's Developmental Tasks :

Adolescence

- Achieving new and more mature relations with age-mates of both sexes.
- Achieving a masculine or feminine social role
- Accepting one's physique and using one's body effectively.
- Desiring, accepting and achieving socially responsible behaviour.
- Achieving emotional independence from parents and other adults.

All these tasks are preparing them for adulthood.

PHYSICAL CHANGES

Few adolescents experience body - cathexis or satisfaction with their bodies. However, they do experience more dissatisfaction with some part of their bodies than with other parts. This failure to experience body - cathexis is one of the causes of unfavourable self-concept and lack of self-esteem during adolescence.

Concerns about Physical Changes

- Awareness of social reactions to different body builds, especially of the endomorphic type leading to obesity.
- Menstruation, in the case of girls. Physical discomfort, cramps, weight gain, mood swings, depression, etc.
- Menstruation as "the curse" for girls; as boys do not experience any such form of physical discomfort, it colours of the attitude of girls and encourages them to behave as if they are martyrs.
- Acne and other skin eruptions, marring their chances for physical attractiveness.
- Physical attractiveness and its role in social relationships with their peer group and the opposite sex.

Emotionality during Adolescence

Traditionally, adolescence has been thought of as a period of "storm and stress" - a time of heightened emotional tension resulting from the physical and glandular changes that are taking place. Adolescent emotionality can be attributed mainly to the fact that boys and girls come under social pressures and face new conditions for which they received little or no preparation during childhood. Not all adolescents go through a period of exaggerated storm and stress, but most do so due to the necessity of making adjustments to new patterns of behaviour and to new social expectations. For example, problems related to romancing, worry of the future as their schooling comes to an end.

Emotional Patterns in Adolescence

During early adolescence they are often irritable, excited and explode easily but as they grow into later adolescence instead of having temper tantrums they express anger by sulking and refusing to speak.

Emotional Maturity

Indications of emotionality are that boys and girls

- do not 'blow up' emotionally in the presence of others but wait for a convenient time and place to let off emotional steam in a socially acceptable manner.
- assess a situation critically before reacting to it in an emotionally unthinking manner.
- are stable in their emotional responses and do not swing from one mood to another.

SOCIAL CHANGES DURING ADOLESCENCE

Social Adjustments

One of the most difficult developmental tasks of adolescence relates to social adjustments. These adjustments must be made to members of the opposite sex in a relationship that never existed before and to adults outside the family and school environments.

Increased peer-group influence

Since adolescents spend most of their time outside the home with the peer group, it is understandable that peers have a greater influence on their attitudes, speech, interests, appearance and behaviour than the family has. The peer group is their real world, providing them a stage upon which to try out themselves. The peer group offers them a world in which he may socialize in a climate where the values are not set by adults but by others of his own age. As adolescence progresses peer group influences give place to close, personal friendships.

New values in selection of friends

Adolescents want as friends those whose interests and values are similar to theirs, who understand them and make them feel secure, and in whom they can confide problems and discuss matters they feel they cannot share with parents or teachers.

Interest in the opposite sex becomes increasingly stronger as adolescence progresses.

SOME ADOLESCENT INTERESTS

Recreational Interests

- Relaxing
- Games and sports
- Travelling
- Hobbies
- Dancing
- Reading
- Movies
- Radio and Records
- Television
- Daydreaming

Social Interests

- Parties
- Drinking
- Drugs
- Conversation
- Helping others
- World affairs
- Criticism and reform

Personal interests

- Interest in appearance
- Interest in clothes
- Interest in achievements
- Interest in independence
- Interest in money

Educational Interests

Typically, young adolescents complain about school in general and about restrictions, homework, required courses, food in the hostel, the way the school is run and their teachers. The attitudes of older adolescents towards education are greatly influenced by their vocational interests.

Factors influencing adolescent attitudes towards education

- Peer attitudes - whether they are college oriented or work oriented.
- Parental attitudes - whether parents consider education a stepping stone to upward social mobility or only a necessity because it is required by law.
- Grades, which indicate academic success or failure.
- The relevance or practical value of various courses.
- Attitudes towards teachers, administrators, and academic and disciplinary policies, success in extra-curricular activities.
- Degree of social acceptance among classmates.

Why do adolescents dislike school ?

Adolescents dislike school when

- parents have unrealistically high aspirations for their academic, social or athletic achievements.
- They find little acceptance among their classmates.
- They mature early and are conspicuously large among classmates.

Vocational Interests

Boys and girls as adolescents are fascinated by the world of work. As early adolescents they are fanciful about their vocation, but as they near adulthood they become more realistic and focused. They are eager to earn money and believe that this is the final scene to the play of attaining independence. This is the 'exploratory stage' and are on the look out for vocational information.

Sex Interest and Sex Behaviour

The first developmental task relating to sex adolescents must master is forming new and more mature relationships with members of the opposite sex. Now that they are sexually mature, new interest begins to develop when sexual maturation is complete, is romantic in nature and is accompanied by a strong desire to win the approval of members of the opposite sex. Because of their growing interest in sex, adolescent boys and girls seek more and more information about it. Few adolescents feel they can learn all they want to know about sex from their parents. Consequently they take advantage of whatever sources of information are available to them - discussion with friends, sex books, experimentation through masturbation, petting or intercourse.

Changing social trends in sexual behaviour

- Broader outlook of parents
- Co-educational institutions
- Role of school counselors
- Importance of providing sex education for adolescents
- Easy availability of information on sex through media

FAMILY RELATIONSHIPS DURING ADOLESCENCE

Deterioration in the relationship between parents and adolescents is usually due to fault on both sides. The so-called 'generation-gap' between adolescents and their parents is partly the result of radical changes in values and standards that occur due to a rapidly changing culture and better educational and social opportunities available to the younger generation. Thus it is more a 'cultural gap' rather than 'generation-gap' as differences are not entirely due to chronological age differences.

Common causes of family friction during adolescence

- Standards of Behaviour
- Methods of discipline
- Relationships with siblings
- Feeling victimized
- Hypercritical attitudes
- Family size
- Immature behaviour
- Rebellion against relatives
- "Latchkey problems"

ADOLESCENT PERSONALITY

Conditions influencing the Adolescent's Self-Concept

- Age of maturing
- Appearance
- Sex appropriateness
- Names and nicknames
- Family relationships
- Peers
- Creativity
- Level of aspiration

HAZARDS OF ADOLESCENCE

Physical hazards

- Mortality
- Suicide
- Physical defects

Psychological hazards

- Poor foundations
- Late maturing
- Prolonged illness
- Role change
- Prolonged dependency
- Social discrimination
- Sexual rejection
- Family relationship

Common danger signals of Adolescent maladjustments

- Irresponsibility - shown in neglect of studies in favour of fun and social approval.
- Overly aggressive - cocksure attitude
- Feeling of insecurity - which cause the adolescent to conform to group standards in a slavishly conventional manner.
- Feeling of martyrdom
- Excessive daydreaming
- Regression to earlier levels of behaviour

- Use of defense mechanisms such as rationalization, projection, fantasizing and displacement.

INTELLECTUAL DEVELOPMENT IN ADOLESCENCE

- Transition from concrete operational thought to formal operations - hypothesizing, analytical thought process, inductive thought process, accommodate new experiences, concept mapping.
- More abstract, liberal and knowledgeable
- Trying to understand purpose, need, meaning, able to think abstractly decline in authoritarian views, increase in political knowledge.
- Engaged in establishing a personal value system - emulate behaviour, modeling, critical about contradictory values.

ROLE OF THE TEACHER AND EDUCATIONAL INSTITUTION IN THE DEVELOPMENT OF ADOLESCENTS

As a teacher

- aid in the adolescent's search for identity
- provide for sufficient career information and vocational guidance
- be patient with the mood swings of adolescents
- try to be friendly with them as one of them to gain popularity
- be trustworthy to win their confidence
- provide a positive picture of the world of work.
- make them feel that everyone has gone through this period including yourself
- do not label them
- provide for a counsellor who is easily approachable
- be more appreciative of their uniqueness
- provide for courses that will help them know about their physical self
- provide for sex education

- do not be authoritative unnecessarily, democratic ways of approach are more appreciated.
- help students make short-term goals if they are underachievers.
- recognize their worth and respect them.
- do not make gender differences in your class transaction.
- be patient with disruptive behaviour
- dialogue is appreciated to come to concensus
- allow boys and girls to interact in the class
- offer guidance to students who find it difficult to get along with others.
- help them accept their physical body and use it effectively
- guide them in achieving correct sex-roles
- help them to become emotionally independent
- provide positive and realistic views of marriage and family life
- develop skills and concepts necessary for civic competence in order to become prospective voters
- channelise their energies in achieving satisfaction (cathexis) in extra curricular activities and service programs.
- teach them to develop skills for problem solving.
- help them to develop positive attitudes in life.
- help them abstain from eve-teasing and gender-abuse.
- motivate them to raise their level of aspiration
- aid in developing their interests, talents and hobbies
- educate them about hazardous status symbols like rash driving, pre-marital sex, drugs, alcohol and smoking.
- counsel to improve family relations of the adolescent
- help them to make a realistic assessment of their strengths and weaknesses
- help to develop a positive self-esteem
- be a good role-model yourself.

SIGNIFICANCE OF VALUE EDUCATION

The problem of value education of the young is assuming increasing prominence in educational discussions during recent times. Parents, teachers and society at large have been concerned about values and value education of children. National policy on Education (NPE) 1986 and revised NPE 1992 has given all importance to the promotion of Value Education in Schools. Education is expected to play a major role in promoting national development in all its ramifications. At the same time, it should bring harmonious development of all the faculties towards adequate preparation for life. The present situation in India demands a system of education, which, apart from strengthening national unity, must strengthen social solidarity through meaningful and constructive value education.

The worldwide resurgence of interest in value education has been explained as the natural response of the modern industrialized societies to the serious erosion of moral values in all aspects of life and the crisis of values experienced in modern times.

It is now commonplace to say that sweeping political, economic and social changes have overtaken human civilization during the past few centuries and these have been largely responsible for the predicament of modern man. The factors such as personal greed, meanness, selfishness, indifference to others' interests and laziness also have brought about large-scale corruption in almost all spheres of life – personal and public, economic and political, moral and religious. We can achieve a better moral standard in our democratic way of national life if we become more industrialized and thus overcome mass poverty and the general feeling of insecurity which gives rise to greed.

We are witnessing a tremendous value crisis throughout the world today. A lackadaisical attitude towards value and its institutions is ubiquitous everywhere around the globe. As the vitality of human belief in values is dying out in every land, the younger generation has started to pooh-pooh the unique religious epics of antiquity and religious institutions, giving room for corrosion of godliness and erosion of spiritual and moral values. As a result, the mind of man has been laciniated and divided into small fractions and fragments which makes the value content of human life a diminishing factor in modern times.

The reappearance of barbaric qualities of selfishness, clashes and conflagration and other destructive forces which are burning the society, give clear indication of the degenerating process of human society. Now, there is an urgent need for a great effort to revive and reform the values of human life and to rejuvenate the foundation of the new civilization.

Concerted efforts and continuous dependence on good books and institutions will give students sterling and inspiring qualities of concentration, infinite love, justice, honesty, purity, selfishness, wisdom, faithfulness, humility, forgiveness, mercy, trustworthiness, respect for others, obedience, sincerity and a host of other virtues which are *sine qua non* to build the equipment of life. This should be the central theme of value education. Whatever be the cause of the present value crisis, there is no gain – saying the fact that the weakening of moral values in our social life is creating serious social and ethical conflicts. It is this changing context – the declining moral standards in personal and public life on the one hand, and the national ideological commitment to the values of democracy, socialism, secularism and modernization on the other – that constituted the driving force behind the recommendations stressing the importance of value education in educational institutions.

While there is general dissatisfaction with the fall in moral standards of both young and the old and disenchantment with the disregard to moral values witnessed in personal and public life, there has been no concerted attempt on the part of the society to address itself squarely to the problem of value education. Unfortunately, education is becoming day by day more or less materialistic and the value traditions are being slowly given up. A modern Indian is being educated mainly with the bread and butter aim of education; as a result most of our graduates run after money, power and comforts, without caring for any type of value.

The degeneration in the present day life, the demoralization of public and private life, the utter disregard for values, etc. are all traceable due to the fact that moral, religious and spiritual education has not been given due place in our educational system.

The Education Commission of 1964-66 says that "a serious defect in the school curriculum is the absence of provision for education in social, moral and spiritual values". In the life of the majority of Indians, religion is a great motivating force and is intimately bound up with the formation of character and the inculcation of ethical values.

A national system of education that is related to life, needs and inspiration of the people cannot afford to ignore this purposeful force. Value crisis of the present day life is baffling the minds of educators and the educands as well. The effect of the value crisis on present day life is witnessed in the following :

- The democratic ideology that has been accepted by our country is yet to be actualized in the form of social and economic democracy as to realize democratic values guaranteed by the Constitution of India.

- The individual is becoming a prey to the contradictory values and is being converted as a consequence into an extreme radical, a reactionary, a skeptic or cynic.
- The present Indian educational system is reflecting more or less borrowed ideologies and philosophies and the national values are relegated to the back.
- The teacher-educators and teachers are not being clearly oriented to the national values and ideas, ideal and ideologies that they have to inculcate in the students. Hence, they are not in a position to play their role as value educators.
- The student community is drowned in neck-deep poverty, ignorance and unhealthy surroundings. Hence, they are not in a position to comprehend the real values of our contemporary India.
- Our curriculum does not reflect human values and the value system, hence our schools and colleges have become examination centers and not value centers.

The problem with value education, it appears, is that while everybody is convinced of its importance, it is not clear as to what it precisely means and what it involves. In our educational reconstruction, the problem of an integrated perspective on values is pivotal, for its solution alone can provide organic unity for all the multifarious activities of a school or college curriculum programme. An integrated education can provide for integrated growth of personality and integrated education is not possible without integration of values.

In value education, as in any other areas of education, what is asked of the teacher is a total commitment to the development of rational autonomy in both thought and action.

It should be noted that the most important aspect of value education consists not in unwilling adherence to a set of rules and regulations but in the building and strengthening of positive sentiments for people and ideals. Value education should prepare individuals for participation in social life and acceptance of social rules. What is more important in value education is that schools should provide a healthy climate for sharing responsibilities, community life and relationships.

The new National Curriculum Framework for School Education (NCFCE) prepared by NCERT gives uppermost importance to Value Education in schools. NCERT has been contributing richly to the area of Value Education by way of organizing inservice education courses for key level persons, preparation of instructional materials, etc. The RIE, Mysore under the Coordinatorship of Dr Prahallada has brought out a 686 page material titled **"TREASURE TROVE OF VALUES"** which consists of Anecdotes, Fables, Stories, Legends, Biographies and Folk Tales related to values which will be of great use at primary stage.

Also, 115 page Package on Value Education has been brought out by RIEM consisting of importance of Value Education, approaches to Value Education, Lesson Planning in Value Education. The package will be useful for the teachers for the inculcation of values at primary school stage.

Regional Nodal Centre on Value Education at RIEM

The NCERT, New Delhi has been identified by the MHRD (Department of Education). Government of India as the nodal center for strengthening value education in the country at school level. Subsequently, a National Resource Centre for Value Education (NRCVE) has been set up in order to plan and implement programmes on value oriented education. NCERT, New Delhi has launched a National Programme for Strengthening Value Education. This programme has been visualized as a national level initiative to sensitize parents, teachers, teacher educators, educational administrators, policy makers, community agencies etc. about the need for promotion of value oriented education. The focus of the programme is on generating awareness, material development, teachers training, development of school programmes, promotion of research and innovations in the area of education of human values and development of a framework of value education for the school system.

In this context, a Regional Nodal Centre (RNC) has been set up at the RIE, Mysore from September 2002 which will be responsible for linkages networking, monitoring and follow up etc. at the State, District and grassroot level for implementation of value education programmes. The Centre will take up the responsibility of organizing National Consultation and Regional Workshop on Value Education with focus on strategies of awareness generation, material development and teacher's training. The RNC comprises of representatives drawn from SCERTs, IASEs, CTEs, DIETs, NGOs, School Boards, Bureau of Textbooks and eminent professionals/educationists from the southern states.

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CHAPTER 28

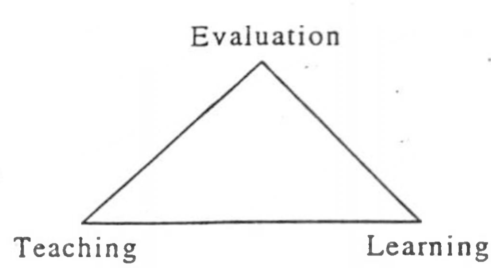
EVALUATION

C. GURUMURTHY

All our educational programmes emerge out of aims and objectives set out in the curriculum. The learning experiences provided to the pupils are directed towards realising these objectives. The teacher wishes to bring about a change in behaviour in the pupil in a desired direction. If after the instruction the student exhibits such changes in behaviour, then the objective is said to be realised.

A casual observation by the teacher in the normal setting of a school may not help him to judge whether the objective are realised or not. A systematic procedure in a controlled environment is necessary to arrive at the above judgement. Hence objective based evaluation forms an integral part of instruction. The results of evaluation is made use of to improve teacher's own instruction and thereby pupil's learning.

Thus evaluation, teaching and learning are interdependent on each other.



What is Evaluation ?

Before arriving at an acceptable definition of evaluation let us examine the meaning of the terms 'value' and 'valuation'. By value we mean, price, worthiness, procures, estimation, set of principles, habits, customs, character, etc. Valuation, means estimating the performance of the learner in certain situations and marking on the basis of predetermined set standards.

Evaluation could then be defined as valuation plus judging the worthiness of learning outcomes. It involves a systematic process and identification of objectives in advance.

Purposes and Functions

Evaluation can be made use of for varied purposes.

- to adapt instruction to the differing needs of individuals.
- to identify the hardspots of a pupil in a given subject and suggest suitable remedies.
- for selection like within the school for higher education and outside by the employer.
- to offer personal guidance for scholastic career, placement, solving immediate problems of pupils, etc.

In short we can say that it is important for diagnosis, prediction, selection, grading and offering guidance.

Categories

The various categories of evaluation are

- purpose - specific category
- mode - specific category
- process - specific category

Evaluation is usually done through tests and examinations. An objective based test contains various types of test items. What then is an item ?

An item is referred to as a learning activity presented in the form of a specification of a task to the tester.

Features of an item

An item

- is based on a learning content/learning activity
- presents the learner with a task
- expects a response from a learner
- expects the response to undergo process of evaluation

Mechanics of an item

(i) The task that an item specifies should in the process of learner-response to it, demand and reflect only those specific aspects or skills or bits of learning that are being listed.

(ii) It should specify precisely

- what the learner is to do

- the conditions under which it is to be done and
- to what level (standard to be accomplished)

(iii) The medium used to present the task specification should be such that there may not be any gap in its communication to the tester.

Items can be of different categories and formats.

Item categories

Category 1

(i) supply type

(ii) open-ended

(iii) subjective items

Category 2

(i) selection type

(ii) closed end

(iii) objective type

Item formats

Selection type

1. Constant alternative

True-False

Yes-No

Agree-Disagree

Right-Wrong

Modified True, False

Agree/Disagree/Don't know

Always/Never/Something

2. Multiple choice

Simple selection type

Multiple selection type

Reason-Assertion type/Multiple facet type

Sequencing type/Rearrangement type

Matching type

Linked type

Negative multiple choice type

Analogy type

Complex multiple choice type

3. Supply type

Simple questions

Completion type

Short-answer type

Long-answer type

Problem-solving questions

Standardisation of a test

It is important to standardise a list before using it. A number of steps involved in the process. They are

i. Preparation of a blue print

ii. Test construction

iii. Pilot study

iv. Item analysis

Preparation of a blue print is necessary in order to decide the extent of the content to be tested, the objectives on which the test will be based and the type of items. It helps the teacher decide the weightage to be given to each of these aspects (refer sample blue print). Keeping the blue-print in view the test is later constructed.

During the pilot study the test can be administered to a selected population of students. Item analysis may be done and suitably modified before standardising the same.

Item Analysis

Item analysis refers to the process of examining students response to each test item. We use the facilitative value and discriminative index while deciding about whether the items are to be retained, modified or rejected.

(a) Facilitative Value

This indicates the difficulty level of each item and is given by

$$FV = \frac{R}{N} \times 100$$

R = Total number of right responses in both the groups.

N = Total number of students in both the groups.

Note: Both the groups mean High Ability Group (HAG) and Low Ability Group (LAG).

Discriminative Index

This indicates the extent to which the question discriminates a higher scorer from a low scorer in the same test.

$$DI = \frac{N(HAG) - N(LAG)}{n}$$

N(HAG) = Total number of right responses in (HAG)

N(LAG) = Total number of right responses in (LAG)

n = Total number of students in either group (HAG)/(LAG)

Interpretation

After determining the facilitative value and discriminative index, they can be used to draw up a table which indicates the range of scores, what it implies and what should be done with the item as shown below.

Facilitative Value

From	To	What it means	What is to be done with the item
0	25	The item is too difficult	Modify the item
25	75	The item is within the suitable range of facility	Retain the item
75	100	The item is too easy	Reject the item

Discriminative Value

From	To	What it means	What is to be done with the item
-1.00	+0.20	HAG is not doing better than LAG	Modify
+0.20	+0.80	Item is satisfactory	Retain
+0.80	+1.00	Item is very good	Retain

The standardised test item performs its function effectively and the standardised test discriminates between good and poor learners. It also helps a teacher to choose a suitable technique of teaching for the classroom teaching-learning, depending on the learners abilities. Further facilitative value provides a basis for comparison and can also help in defining and maintaining standards in schools. The discriminative indices help identify topics to be addressed to all learners and topics to which learners of lower ability are to be registered.

Strategies for Increasing Positive Student Behaviours

Guidelines for Effective Praise

One of the most powerful strategies is providing praise for appropriate behavior. The planning of how and when to use praise rests with the teacher.

1. **Define the appropriate behaviour while giving praise.**
Praise should be specific for the positive behaviour that the student displays. This means any comments about behaviour should focus on what the student did right. The praise should include exactly what part of the student's behaviour is acceptable.

   Situation: The teacher would like to see seatwork done quietly.
   Example: "That is great that you did your seat work so quietly today."
   Non-example: "You didn't disturb others today."

2. **Praise should be given immediately.**
The sooner an approving comment is made about appropriate behaviour, the more likely the student will repeat the desired behaviour.

3. **The statements used as praise should vary.**
Individual statements that one uses should be varied. When students hear the same praise statement used over and over, it looses its value for the student.

4. **Praise should not be given continuously or without reason.**
If praise is given too frequently or without stating what the student is doing that is "good", then praise looses its value to the student.

5. **Be sincere with your praise.**
Students will notice if you do not mean what you say. Nonverbal cues like facial expressions and posture will alert the students that your praise is not sincere. The praise will not be effective if the student perceives that it is not sincere. Smiles communicate that the praise given is genuine.

6. **Be consistent when praising the target.**
It is important to be consistent with the behaviours that you praise. Students learn more quickly when they are always praised for desirable behaviours. Consistency between teachers is important in order to avoid confusion about behavioural expectations.

7. **Praise should be developmentally appropriate.**

   Statements to younger or developmentally delayed students should be in language that is at their level so they clearly understand what behaviour is seen as appropriate. However, if older students perceive they are being "talked down to", it is likely that the praise will be discounted.

Why Praise Works

1. Praise is readily available as reinforcement for positive behaviours.
2. Praise can be administered immediately after the desired behaviour.
3. Praise can be used over and over again if praise statements are varied.
4. Praise may be used in combination with other strategies to increase behaviours.
5. Praise can be tailored to a variety of behaviours by being specific about the activity.
6. Praise works if the relationship between the student and the person giving the praise is a positive relationship.

Why Punishment Does Not Work

- Punishment is a less effective means of dealing with unacceptable student behavior. Punishment gives attention to the wrong behavior. When the teacher gives attention to inappropriate behaviours, frequently the behavior increases. The student may repeat the behavior just to get attention. For some students, attention of any kind is desirable.
- Punishment can damage the student's relationship with the teacher. If a student is punished for behaviour that is unacceptable, he or she may become uncooperative at other times. The student may not try or work for the teacher when requested to do so.
- A student's self-esteem can suffer if the only attention from teachers is in the form of punishment. The negative feelings that come from only experiencing punishment can result in an attitude that he or she can do nothing right. With the use of punishment, there is not an opportunity for the student to be recognized for the behavior that is acceptable.
- Punishment can discourage both unacceptable and acceptable behaviors. If a student is frequently met with negative responses for behaviour, the student may decrease both positive and negative behaviors. If positive behaviors decrease, the student will not have the opportunity to learn or practice acceptable behaviours. Punishment does not encourage a student to take social risks.

Non Verbal Social Approval used to Increase Positive Behaviors

Praise is one form of social approval., Other social means of communicating that the behavior is appropriate may include nods, smiles or a "thumbs up" sign. Where developmentally appropriate, a pat on the back or a "high five" can be used to signal the student that their behavior is appropriate. Just as with praise, these other forms of social approval should be given as soon as possible after the positive behavior is observed.

Rules and Instructions as a Means of Increasing Positive Behaviors

Rules and instructions can help the student increase positive behaviors in a number of ways.

1. Rules and instructions can provide a guideline for what behaviours are appropriate. Students may not know what is expected of them. Learning what positive behaviours are can help speed up the identification of acceptable behaviours.
2. Giving clearly stated instructions or having rules displayed enhances communication about expected behaviour.
3. Rules and instructions can be used effectively with praise or other strategies to increase positive behaviours.
4. Restating the rules or instructions just prior to an activity will remind or cue the student about the behaviour that is expected.

Example: It is the first day of school and you have playground duty for 5th and 6th graders. You and the teachers arrange a time to meet with the students in their classrooms to go over the school rules on playground behavior.

Modeling
For some students an explanation of desirable behaviors is not enough. Demonstration is another way of making expectations clear.
Example: The students you are working with become unacceptably loud. You start talking to them in a very low voice. This would demonstrate to them what voice level you want them to use.

Build a Positive Relationship with the Student
Working to establish a relationship with students is an important strategy in effective behavior management. Investing time to get to know students is a good first step in establishing a positive relationship with them. A positive relationship sets the groundwork for all the other strategies. Students are more likely to listen and respond to rules, requests and reinforcement if they know their interactions with the teacher will be positive. Ways to be positive include:

1. Demonstrating to students their importance (i.e., by learning their names, actively listening to them, remembering things said by them.)
2. Praising continuation of appropriate behaviours.
3. Showing interest in helping students.
4. Explaining the reasons for having rules.
5. Encouraging students to participate in activities.

Students respond better to adults who take a personal interest in them. Personal knowledge of each student is one way to strengthen and improve these relationships. It.provides the opportunity to model interpersonal behaviors.

Please Remember

- Behaviour Management should be viewed as an opportunity for teaching and not as an opportunity for punishment.
- Consider the impact on the student's best interests.
- Avoid embarrassing students.
- Suggestions should be in the form of constructive criticism.
- Constructive criticism should occur in private.
- Never engage in a power struggle. Strive for win/win.
- Thank students when they are trying to improve.
- Do not touch a student who is upset.
- Keep other teachers/ the H.M. informed.
- Documentation should be objective and free of emotion.

Encouraging Participation

- ➢ Integrate Discussion Into Your Teaching
- ➢ Respond to Student Questions
- ➢ Help Students Prepare for Discussion

1. *Explain the purpose of discussion*
2. *Create an appropriate physical setting for discussion*

3. *Identify discussion questions/issues in advance*
4. *Use an assignment as a basis for discussion*
5. *Begin with common experiences*
6. *Divide the class into smaller groups*
7. *Prompt discussion through the use of key phrases*
8. *Try brainstorming techniques*

➢ Sustain and Focus Discussion

1. *Encourage heated debates*
2. *Intercede if the discussion breaks down*
3. *Keep notes during discussion*
4. *Assign students responsibility for summarizing major points*

➢ Get and Use Feedback

1. *Increase your eye contact with students*
2. *Ask students if they understand what you are saying*
3. *Call on students to paraphrase or to summarize*
4. *Begin your lesson with a series of questions*
5. *Ask questions during teaching*
6. *Give students problems to solve during class time*
7. *Reserve the last 10 minutes of class for questions*
8. *Give frequent assignments*
9. *Give frequent quizzes*
10. *Ask students to define, associate or apply concepts*
11. *Periodically borrow students' lecture notes*
12. *Encourage students to form study groups*

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ssroom Management Techniques

ned Ignoring

:times the most effective way to deal with student misbehaviour is to ignore it.

n to Ignore Behavior:

1. When the inappropriate behaviour is unintentional or not likely to reoccur.
2. When the goal of misbehaviour is to gain teacher attention.
3. When you want a behaviour to decrease.
4. Do not intervene when there is nothing you can do.

n to Intervene:

1. When there is physical danger or harm to you, others or the child.
2. When a student disrupts the classroom.
3. When there are violations of classroom rules or school policy.
4. When there is interference with learning.

5. When the inappropriate behaviour will spread to other students.

Providing Cues to Students

An important aspect of behaviour management is developing ways to communicate with students that provide reminders that support your expectations. It's simply a way to let the student know that you want their attention, or you're aware of the behaviour, and that you want it changed. These cues can also be used to reinforce positive behaviour patterns as well (i.e., reminders to continue the quality of interaction during an activity). These techniques may be non-verbal, including eye contact, physical gestures (i.e., raising your hand in silence), tapping or snapping of your fingers, coughing or clearing your throat, facial expressions (i.e., smile), or body postures (i.e., tilting your head). *One caution, avoid doing things,that may embarrass students.*

Proximity Control

A fancy term, but you've probably used the technique quite frequently. You're aware how effective it is to stand near a student who is experiencing difficulty. Simply moving around the classroom can assist students in staying on task because of your "proximity" to them. This works well because the students know you're aware of what's going on, and allows the classroom teacher to continue without interrupting the lesson or the flow of the activity. As a caution, it's important not to reinforce the inappropriate behaviour or call attention to the student.

Ways to Increase Student Motivation

Motivation is a key to academic success for most students. There are a number of ways to increase the motivational level of students.

1. Relate the material to their life experience/s, in other words, make it relevant to them personally, thereby stimulating their interest.
2. Demonstrate an active interest in that child.
3. Demonstrate an active interest in the child's work or the activity.
4. Use lots of praise both verbally and nonverbally.

The Use of Humor

We're all aware of how a light, funny or amusing comment or statement can often decrease tension, or frustration and afterward allows everyone to feel a bit more comfortable.

Of all the techniques discussed here, humor can be the most prone to misuse and is not easy to master, especially if it's directed toward a particular child or group of children. We've all heard the expression "laugh with, not at children". Even the practice of laughing at one's own actions can sometimes be troublesome, particularly if it's negatively directed. Don't use sarcasm and don't belittle students. Be careful, because what you think is funny may not be funny to the student involved.

Humor can also be used constructively to decrease levels of anxiety and thereby increase students' academic perform.

Helping Students Through Tough Spots

All students eventually will come across a certain task, assignment or situation that causes them difficulty. Many will request assistance from teachers, staff or peers when appropriate. Others will simply stop working all together and not know what to do next. At these times, trouble can occur. We need to get them back on track. You can be most helpful in getting the student back on task by:

1. Doing (or solving) the problem with the student.
2. Reviewing the directions.
3. Providing another example or demonstrating.
4. Supplying them the correct answer as a model.

Appeal to Student Values

Often you can appeal to students' values when intervening in problem situations. Their desire is to be liked by others, to do the right thing, to be treated with respect, etc. You might:

1. Appeal to the relationship between yourself and the student.
2. Appeal to the natural consequences of a specific behaviour. (i.e., "I know you're frustrated, but if you break your pencil, then you'll have to replace it with your own money.")
3. Appeal to a student's need to be liked. (i.e., "Your friends may be disappointed with you if you continue to boss them around and interrupt them when they're speaking.")
4. Appeal to the student's self-respect. (i.e., "I know you'll be very upset with yourself by doing this.")

Removal of Nuisance Items

It is difficult for teachers to compete with certain objects, either found at school or brought from home (i.e., rubber bands, combs, etc.). Often times in order to gain students' undivided attention, you may be required to deal with these types of competing items. Often, however, the removal of such belongings will only lead to further conflict. One way to avoid such conflict is to simply state the choices:

1. You can either put it away immediately; or
2. I will put it away until the end of the day.

However, by taking a strong interest in the object and then politely asking to see and handle it. Once it is in your possession, you have the option of returning it, with a firm request that it disappear for the rest of the school day, week or year, or to keep it, with a promise to return it at the end of the day and/or week. This technique is most effective if you have established a relationship with the student.

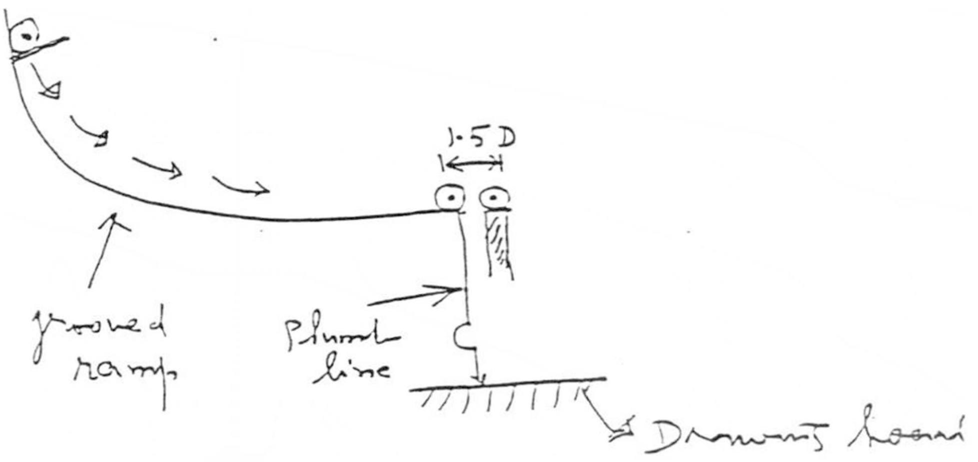
Materials adapted from: Baldwin J.D. and Baldwinn J.I. (1986). Behavior principals in everyday life (2nd Edition), Engle Wood Cliffs, New Jersey: Prentice Hall. and Martin, G. and Pear, J. (1992). Behavior Modification: What it is and how to do it. Engle Wood Cliffs, New Jersey: Prentice Hall.

EXPERIMENT NO. 1

COLLISION IN ONE-DIMENSION

Set up Draw the diagram and label the parts

Adjustments



i) Adjust the height of the set screw so that its tip is at the same level as the ramp and also along the same line as the groove on the ramp.

ii) Adjust the distance between the edge of the ramp and centre of the set screw to be 1.5D where D is the diameter of any hard sphere.

Observation

Keep the target sphere T on the set screw. Place the bullet sphere B at a suitable location (25cm) of the ramp with the help of a small scale held vertically. Release the bullet B.

Record your Observation

Answer the following question:

1. What is momentum (represent by $\bar{p}$) ?

2. What type of physical quantity it is ?

3. What physical quantities are to be measured to calculate momentum ?

4. What are transferred to sphere T by sphere B during the collision. What happens to the bullet sphere after collision?

5. What is the initial momentum of sphere B. sphere T and of the system before collision?

6. What is the final momentum of sphere B. sphere T of the system after collision?

How do you measure the momentum of sphere B. before collision and sphere T after collision ? Stop watch is not given.

What observed and measurable physical quantity can be considered as a measure of momentum? (Hint: Horizontally projected body).

What is range R basically a measure of ?

What other quantity it can represent assuming masses of spheres B and T as same. Similarly, what does R^2 represent?

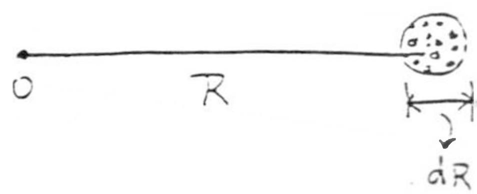
Measurement

Place a carbon paper with carbon side up and place over it a tracing paper. Fix them to the drawing board.

Step 1

Release the bullet sphere B from a suitable position (25cm) on the ramp. Note the position where target sphere exists the tracing paper. Release the sphere from the same point several times and note the distribution of points on the tracing paper.

To what degree is the velocity of the target sphere. after collision always same ?



Step 2

Bring down the set screw so that when sphere B is released from the same point as in step 1 moves down with out hitting the set screw.

Get the trace for several releases.

To what degree is the initial velocity always the same?

Mark the point "O" corresponding to the tip of the plumb line.

Draw momentum vectors OT and OB. Compare them.

Calculate R^2B and R^2T. Compare them.

Step 3

Repeat, by reloading the bullet sphere from two other positions.

Step 4

Tabulate your measurements

Conclusions:

i) Momentum

ii) Kinetic Energy

Step 5

Discussion:

1. Offer explanation for discrepancies, if any

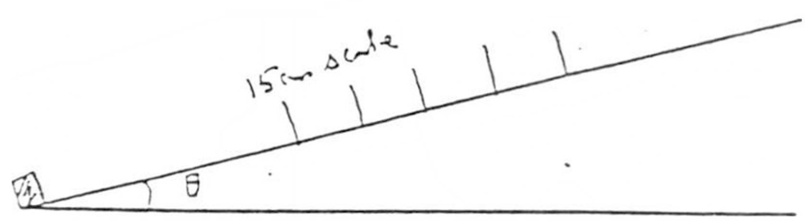
2. Sources of error and Methods to eliminate/minimize

Reference: PSSC Text, Lab guide and Teacher's guide.

EXPERIMENT NO. 2

GALILEO'S EXPERIMENT

Set up: Keep the grooved track (about 2.5m long) on a horizontal surface. Place the given smooth ball on it. Lift one end of the track till the ball slides off. Lift a little bit up and fix. (Give reason for this initial adjustment). Make sure that the track does not sag under its own weight. Provide suitable supports to ensure that the track is straight and inclined to the horizontal.



Step 1: Observation

Release the given smooth polished steel ball about 1cm in diameter from one end marked as starting point. Watch its motion (Judge the pitch of the sound)> what do you guess about the speed of the ball with respect to (a) time of travel and (b) distance travelled ? What type of motion is it ? Write down your guesses.

Step II: Guesses

Step III: How is acceleration defined ? Can we use this definition for our investigation? Give reason. A more convenient relation is needed. It could be worked out mathematically. Assuming the motion to be of uniform acceleration, we can derive the equation connecting distance travelled with time [your guess (a) above in step 1]. What is that expression? Suppose we set the initial velocity to be zero, then

$$S = \frac{1}{2}at^2 \quad \text{or} \quad \frac{S}{t^2} = \frac{a}{2} = \text{constant.}$$

Here, we have a definition for a uniformlly accelerated motion, more suitable for practical work. Distance-square of time graph is a straight line. Is this definition more convenient for practical investigation ? Why?

Step IV : Experimentation

a) Designing and Collection of Data

From the starting point, mark off distance S = 1.00mm, 1.25m, 1.50m, 1.75m, 2.00m and 2.50m (Mark your own tabulation). Measure the time (as average of best three readings, out of about five) taken to travel each distance. Tabulate.

b) Analysis of Data Collected and Drawing Conclusion:

Draw $S\text{-}t^2$ graph and conclude. What does the slope of the graph indicate? What is the acceleration for this inclination?

Step V: Further Questions and Investigation:

i) Draw S-t graph. What does this indicate?

ii) State the relation between speed and distance travelled in [your guess (b) in Step1] the case of a uniformly accelerated motion. Why is speed as a function of distance travelled not suitable for investigation?

iii) List further investigations that one can undertake with this set up?

iv) Considering the forces acting on the sphere, calculate the acceleration ?

Ref: Project Physics Harvard.

EXPERIMENT 3
Diffraction due to single slit

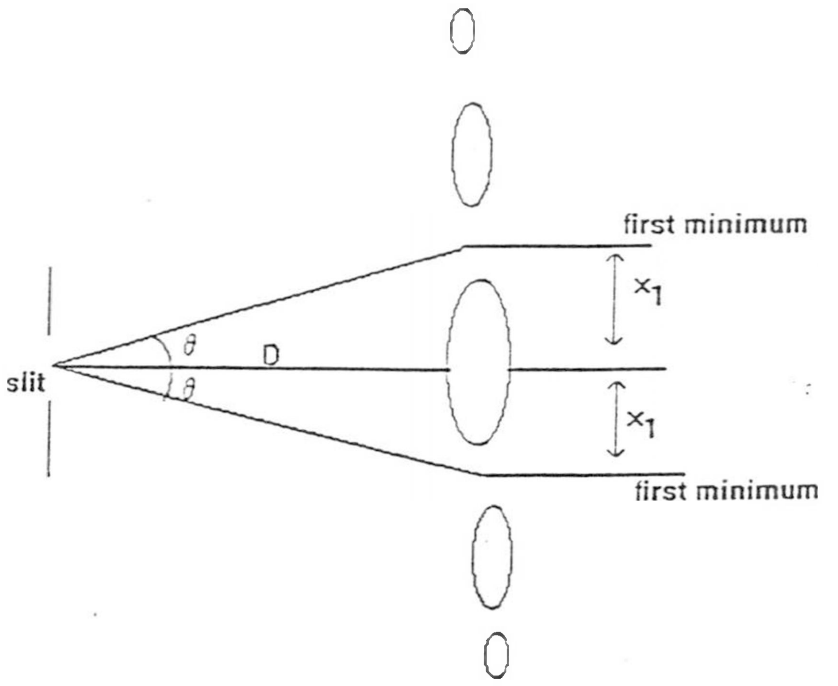
Aim : To study the profile of a diffraction pattern using single slit and laser beam.

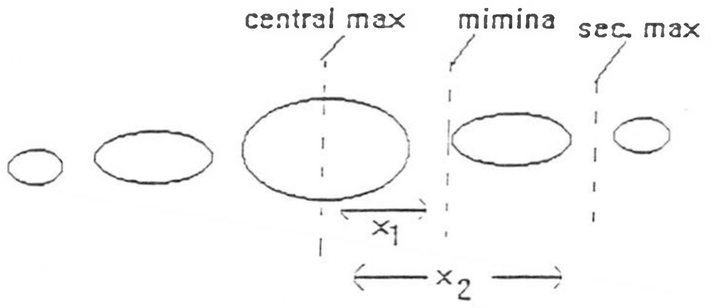
Materials Required: He-Ne laser source, single slit travelling microscope, measuring tape, drawing board, drawing paper, pins, graph sheet.

Method of Approach

When light of wavelength λ undergoes diffraction at a single slit of width 'd', the angle of diffraction θ is given by $d\sin\theta = n\lambda$. In this experiment we determine the wavelength λ using this equation.

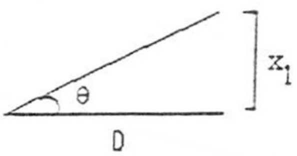
Keep the drawing board on which a graph sheet is fixed 5 m away from the He-Ne source. Place a slit which is wide open at a convenient distance from the source. Align all the three taking care to see that you do not directly see the laser beam with your naked eye. Adjust the slit width until a good diffraction pattern is obtained on the drawing board. Trace the profile (pattern) on the graph sheet.





Measure the distance between the drawing board and the slit D. Mark the centres of the central maxima and the other maxima on the graph sheet. Measure the distance x_1 between central maxima and first minima on either side of the pattern.

At the first minimum $d\sin\theta = \lambda$ since θ is small $d\sin\theta = d\tan\theta = \lambda$.
But



$$\tan\theta = \frac{x_1}{D} \qquad \therefore \frac{d\,x_1}{D} = \lambda$$

From this we can calculate λ. To measure 'd' the slit width use a travelling microscope. Repeat this for 2nd and 3rd orders if possible and calculate λ in each case.

Study what happens to the pattern if 'd' is varied. Tabulate your readings.

Sl No	d	n	x	λ
1	d_1	1	x_1	
		2	x_2	
2	d_2	1	x_1	
		2	x_2	

Try repeating the experiment for different values of D while keeping slit width 'd' a constant.

Questions What should be the order of the wavelength as compared to the slit width so that θ is measurable. In what way do interference and diffraction bands differ?

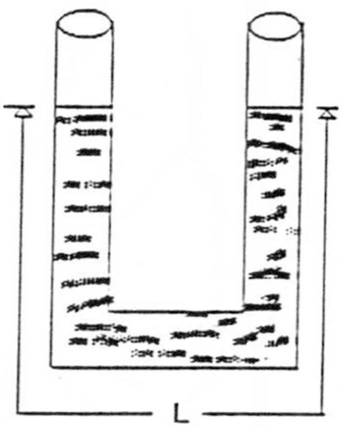
Further study Find out what happens to the diffraction pattern when the number is two or multiple in number.

EXPERIMENT NO : 4
OSCILLATIONS OF A LIQUID COLUMN

Aim (a) Determination of 'g' by study of the oscillations of a liquid column

Materials Required Long U tube diam. 2 cm, scale, stop watch, scale, adhesive tape

Method of Approach
The acceleration due to gravity 'g' is determined by considering the simple harmonic motion of a liquid column. Using the equation



$$T = 2\pi\sqrt{\frac{2L}{g}}$$

where L is the total length of the water column in the tube and g is the acceleration due to gravity.

A known length of the liquid column is taken in the tube as indicated in the figure above. Depress the water level in one side by blowing air so that the column oscillates. Measure the time for ten oscillations of the column a number of times. Repeat the experiment with different lengths of the liquid column by changing L in steps of 20 cms at least. Tabulate as shown below.

Sl.No	Length of water column L	√L	No. Of oscillations	Time 1	2	3	Time period

Plot a graph of T vs √L and calculate 'g'.

Aim (b) Study of damped oscillations

Materials Required Same as above

For one value of L of the liquid column, note the maximum displacement of the water level on either side of the equilibrium position for four or five successive oscillations.
Plot a graph of maximum displacement vs time period T.

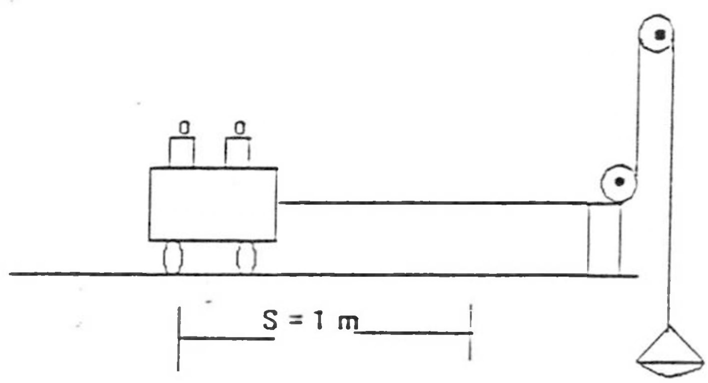
Questions

Does the maximum displacement remain the same?
What is the locus of the successive amplitudes on one side?
If the maximum displacement is different in the successive oscillation, why is it so?
Give some examples of other systems which exhibit this behaviour.

EXPERIMENT NO.5
NEWTON'S II LAW

Aim Study of the relationship between acceleration 'a' and unbalanced force 'F' for a constant mass.

Materials Required Two dynamic carts, pulleys, scale pan, thread, weight box, stop watch.



Method of Approach

For a body of mass m moving with a constant acceleration under the action of a constant force F

$S = \frac{1}{2} at^2$ if $u = 0$

s being the distance travelled in time t.

The apparatus is arranged as shown in the diagram. The cart is first loaded with weights (say about 100g). Give a gentle push to the cart and observe the motion. Transfer a few small weights from the cart to the scale pan such that the motion of the cart is uniform (as judged by your sight). Transfer 20 gm weight from the cart to the scale pan and note the time required by the cart to travel 1m. Repeat the experiment for the same force (weight of the pan + weight in the pan). Find the average time t and calculate a using it. Repeat the experiment for different values of **F**.

Mass of cart =

Mass on cart =

Mass of the scale pan + mass in the scale pan = m_1

Mass of the system =

Sl.No	$F = m_1g$	Time taken to cover 1 m			Average(t)	$a = 2s/t^2$
		1	2	3		

Questions

What does m_1 measure ?

Plot a graph of m_1 vs a. What do you infer?

When the cart was moving with uniform speed was there no force acting on the system? If your answer is yes, give reasons.

Draw the free body diagram of (1) cart, (2) scale pan (with weights in it) for the following situations (a) constant velocity, (b) constant acceleration

How will you find out the mass of an unknown object using this set up?

EXPERIMENT NO. 6

ANALYSIS OF MOTION BY TICKER TAPE TIMER

1. To analyse the motion of a body what information do we need ?
2. Tape-Timer is a device used to mark the positions of a body at equal intervals of time.
3. Fasten a tape to a moving body, your hand or an acceleration cart and pass the tape through the tape guide (under carbon paper - if white tape is used) so that the tape moves under the vibrator.
4. Examine the separation between the successive dots what can you say about the motion of your hand/the moving body ?
5. Estimate the number of dots and select a convenient time interval such that you have about ten observations (unit of time = say 10/15/20, etc. dots).
6. Use a stick-tape and fix the tape you have drawn on the table. Choose an origin and mark off units of time, as 1, 2, 3, etc.
7. Measure the distances of each mark from the chosen origin (i.e. position) and tabulate.

Position	S_0	S_1	S_2	S_3	etc.
Time	0	1	2	3	etc.

Time t (unit of time)	Position S cm	Average velocity $S_n - S_{n-1}$ cm/unit of time	Average acceleration cm/(unit of time)^2
0	$S_0 = 0$	$S_1 - S_0 = V_1 =$	$V_2 - V_1 = a_1 =$
1	$S_1 =$	$S_2 - S_1 = V_2 =$	$V_3 - V_2 = a_2 =$
2	$S_2 =$	$S_3 - S_2 = V_3 =$	
3	$S_3 =$		

Analysis 1

Draw the s-t graph and conclude.

Analysis 2

What does the distance between the successive marks $(S_1 - S_0)$, $(S_2 - S_1)$, etc. represent ? What other physical quantity does this represent ? What is its unit ?

Analysis 3

What is instantaneous velocity ? How is it defined on a s-t graph ? Determine its values for any five instants of time and draw the v-t graph. Infer.

Analysis 4

Using the v-t graph draw the a-t graph and infer. (Define a_{inst} - t graph. How is it defined on the V_{inst} - t graph).

Further Investigation

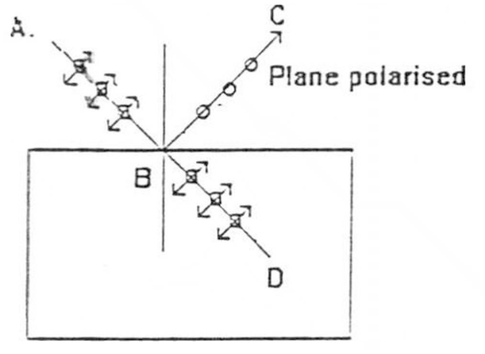
(i) Given v-t graph, how do you obtain (a-t) and (s-t) graphs ?

(ii) Given a-t graph, how do you obtain (v-t) and (s-t) graphs ?

EXPERIMENT NO. 7
BREWSTER'S LAW

Aim Determination of the refractive index of the material of the prism by measuring Brewster's angle.

Materials Required Polarisers, spectrometer, prism, reading lens, reading lamp.



Method of Approach

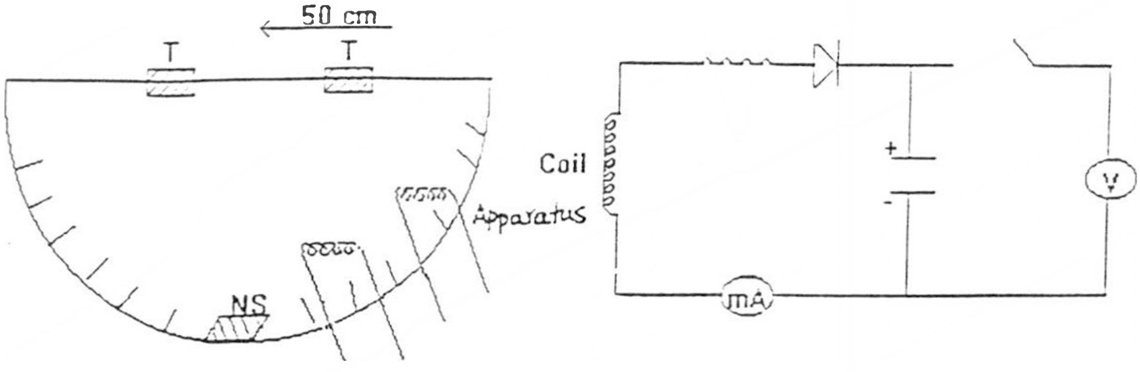
Make all the initial adjustments in the spectrometer. Mount the analyser on the telescope and rotate it to observe the variation in intensity. Remove the analyser. Adjust the telescope to receive reflected rays from the surface of the prism (mounted on the prism table) incident at an angle 50° as in the i–d curve experiment. Fix the telescope. Mount the prism and rotate it to get the reflected image at the cross wires.

Mount the analyser on the telescope and rotate it. Observe the variation in intensity. Mount the polariser on the collimator. Rotate the analyser and observe the position of minimum intensity. Note the angle of incidence θ_B for the minimum intensity positions. Calculate μ from

$\mu = \tan \theta_B$.

EXPERIMENT NO. 8
INDUCED EMF

Aim To study the emf induced as a function of the velocity of the magnet.
Materials Required E.M. kit (Rajasthan University), Multimeter.



Method of Approach

The experiment is based on Faraday's law of e.m. induction. The induced e.m.f. is given by

$$e = \frac{-d\theta}{dt}$$

The apparatus is set up as shown in the circuit diagram. The magnet is placed on the circular arc and allowed to oscillate through the coil by releasing it through different angles. The emf is measured for these various angles. Check if e α θ_o. This is done for a given setting of the masses on the arm. The time period can be varied by changing the positions of the masses. For constant θ measure the induced emf for different settings of the mass. Check if e α 1/T.

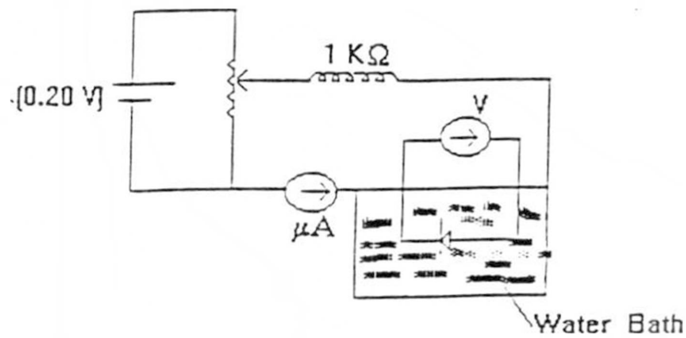
EXPERIMENT NO 9
ENERGY GAP OF A SEMICONDUCTOR

Aim Determination of the energy gap of a semiconductor.

Materials Required Diode, thermometer, resistance 1kΩ (0A79), ammeter (0-100 uA), voltmeter (0-50V), oil bath, DC (0-20 V) or (0-50 V) source.

Method of Approach

Connect the circuit as shown below.



Find out the resistance of the diode at various temperatures.

Draw a suitable graph and determine the energy gap of the diode.

EXPERIMENT NO. 10
LCR-SERIES RESONANCE

Aim Study of the frequency response characteristics of a series resonant circuit.

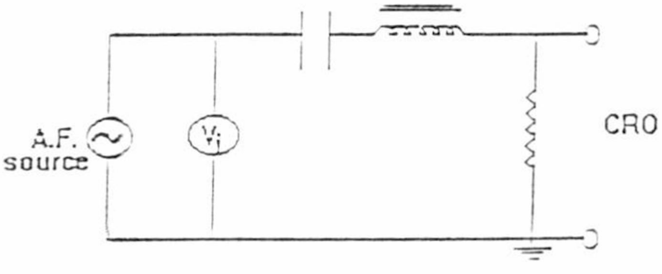
a) Resonant frequency (f_o) of the circuit.

b) Quality factor (Q) of the circuit.

c) Inductance (L) of the coil in the circuit.

d) Variations of phase difference with frequency.

Materials Required AC source, inductance 20mH, capacitance box, resistance box, 100Ω-2000Ω, CRO (double beam).



Method of Approach

Connect the circuit as shown. Choose an appropriate value of C such that the frequency response study can be made around a resonant frequency in a particular range of the AF source with a large spread of frequencies.

Questions

How will you study the phase difference using a double beam?

Experiment: Logic Gates

Objective: To verify the truth-table of AND, OR and NOT gates using Boolean logic.

Apparatus: Logic gate set up, connecting leads

Procedure : i) AND gate

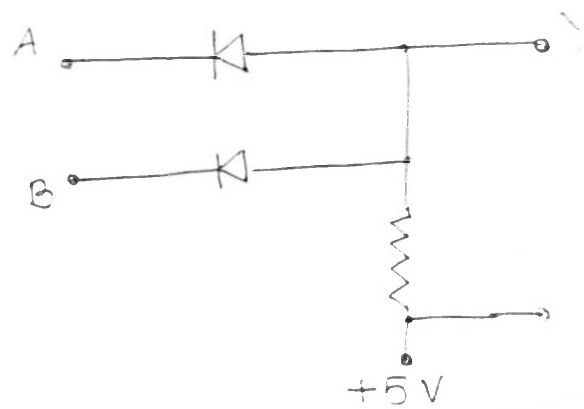
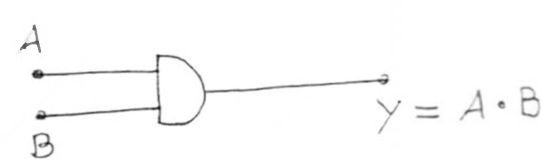


Fig 1

Connect the circuit as in Fig 1.

a) Connect the input terminals A and B to the "Set logic input switches". The terminal marked +5 is connected to the +5V (Hi) terminal of the regulated power supply.
b) The terminal Y is connected to the terminal marked '+' in the DVM(digital volt meter) and its negative is connected to the 0 V of the regulated power supply .
c) Give various possible inputs for A and B and find the corresponding outputs in the DVM and verify the following truth table.

A	B	Y=AB
0	0	0
0	1	0
1	0	0
1	1	1



ii) **OR gate:** Connect the circuit as in Fig. 2.

a) A and B are connected to the "set logic input terminals".
b) Connect the terminal marked 0 V to + 0 V of DVM and the Y terminal to '–' of DVM.

c) Verify the truth table.

A	B	Y=A+B
0	0	0
1	0	1
0	1	1
1	1	1

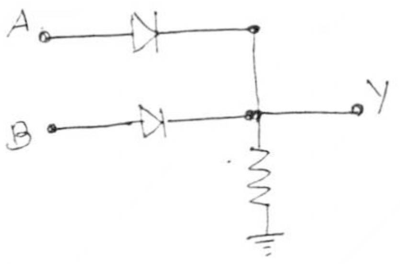
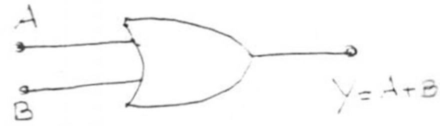


Fig 2



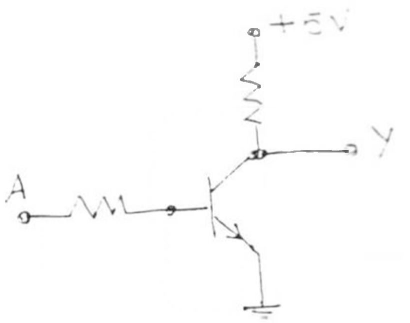
Logic Symbol.

iii) **NOT gate**

Connect the circuit as in Fig 3. Connect the input terminal to either A or B of the "set logic input switches" connect the '+5 V' marked terminal to +5 V of the regulated power supply '0' V terminal of power supply is connected to '-' terminal of DVM.

Verify the truth table.

Input	output
A	B
0	1
1	0



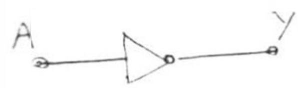


Fig 3

Further scope: Combination of gates such as NAND, NOR can be realized.

Note:

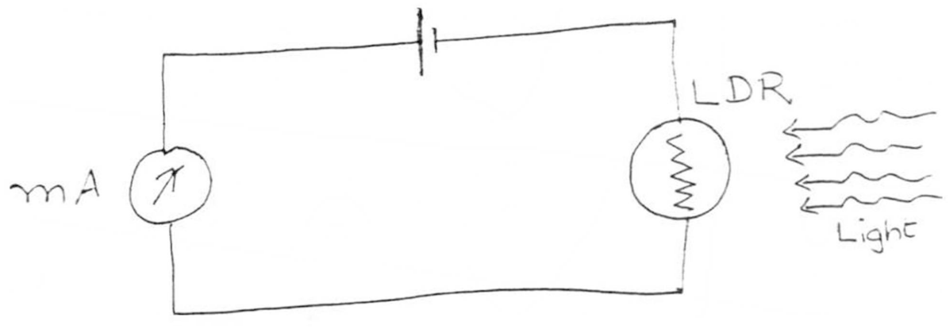
1. 1 in the table implies voltage greater than 3.2 V and 0 implies voltage less than 0.8 V at the output terminals as indicated by DVM.
2. Set logic input switches provide 5 V when in state 1 or high state and 0 V when in 0 or low state.
3. Circuit is completed only when the other terminal is connected to ground. In the case of set logic input switches the other terminal is grounded within the circuit board itself.

Experiment : Variation of light intensity with distance

Objective: To investigate the variation of intensity of light with distance using an LDR (Light Dependent Resistor).

Apparatus: LDR, Source of light (15 W electric bulb) milliammeter, battery eliminator, meter scale, convex lens etc.

Circuit:



Part (i) : To study the variation of intensity of light with distance.

Procedure: Connect the circuit as in Fig.(1). Keeping the light source at a distance of about 1.50 m measure the current in the milliammeter. Move the source toward LDR and measure the current for different distances say 1.40 m , 1.30 m etc. Represent the observation in a graph by plotting the current versus distance. Interpret the graph. What modified graph is to be plotted to get a straight line ? Plot and infer the result.

Further investigations: Use light of different intensity or use filters to see the effect of wavelength. Investigate whether the intensity depends on the angle of incidence of light on LDR.

Part (ii): Place a convex lens in between the LDR and the source such that the LDR is at the focal length of the lens. Vary the position of the lens and see the variation in current. Interpret your result ? Can you use this to determine the focal length of a double convex lens ?

Solar Cells

Crystalline silicon with deliberately added impurities is an essential ingredient of a silicon PV Cell.

In a p-n junction the free electrons in N side see free holes on the P side and hence rush to fill them in but only near the junction in the process the charge neutrality is disrupted. This forms a barrier to other electrons on the N side to cross to the P side. In equilibrium we have an electric field separating the two sides (Fig.1 a). Thus a PV Cell has p and n type silicon in contact, between which an electric field is set.

The electric field makes the junction to act as a diode, in which electrons can move only in one direction.

When light 'HITS' the solar cell, each photon with sufficient energy frees one electron (and results in a free hole as well). If the freed electron or the hole happens to wander into the range of electric field of the diode, the field will send the electron to N side and the hole to P side. This causes a disruption of electrical neutrality. If we provide an external current path electrons would flow through this path to P side to unite with hole there which the electric field had created (Fig.1 b).

The flow of electrons provide current and the junction electric field causes a voltage. With both current and voltage we get power.

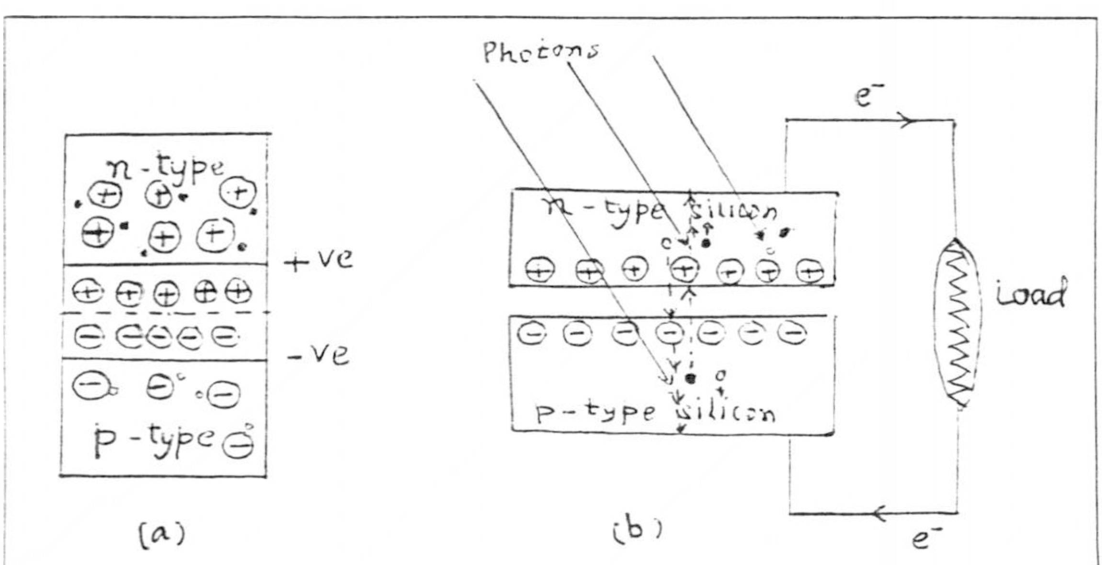


Fig.1. Operation of a P V Cell, ∘ hole, • electron, ⊕ ⊖ bound ions

It may be useful to note that only about 15% of sunlight's energy is useful for Solar Cell. This is because light photons have a wide range of energy, some of them do not have enough energy to form an electron hole pair. Still other photons may have too much energy than that required, then also the extra energy is lost (unless photon has twice the required energy to create one more electron hole pair). This speaks of the quantisation of energy in nature, eg. if the energy of photon is 1.5 times that is needed for the formation of electron-hole pair, 0.5 part of the energy goes waste as heat. These two effects alone cause loss of about 70% of radiation energy incident on the cell.

Optimal band gap for Solar Cell

If we choose a material with a low band gap we can make use of more incident photons. But what we get in the form of extra current, we loose by having a small Voltage. Balancing these two effects a band gap of 1.4 eV has been found suitable for a cell made from a single material.

Other requirements:

1. The incident photons have to reach the junction hence one side of the junction should be left open as window. The other side is covered with a metal (acting as anode) for good conduction. Sometimes a transparent window of conducting material is provided over the upper n type silicon which acts as the cathode of the cell.

2. Silicon being very shiny material the photons that are reflected away by it cannot be used by the cell. Hence an anti-reflective coating is applied to the window of the cell.

Finally, the cell is protected by a glass cover plate. In one unit 36 such Cells are connected in series and parallel combinations and mounted over sturdy frame to achieve satisfactory levels of voltage and current.

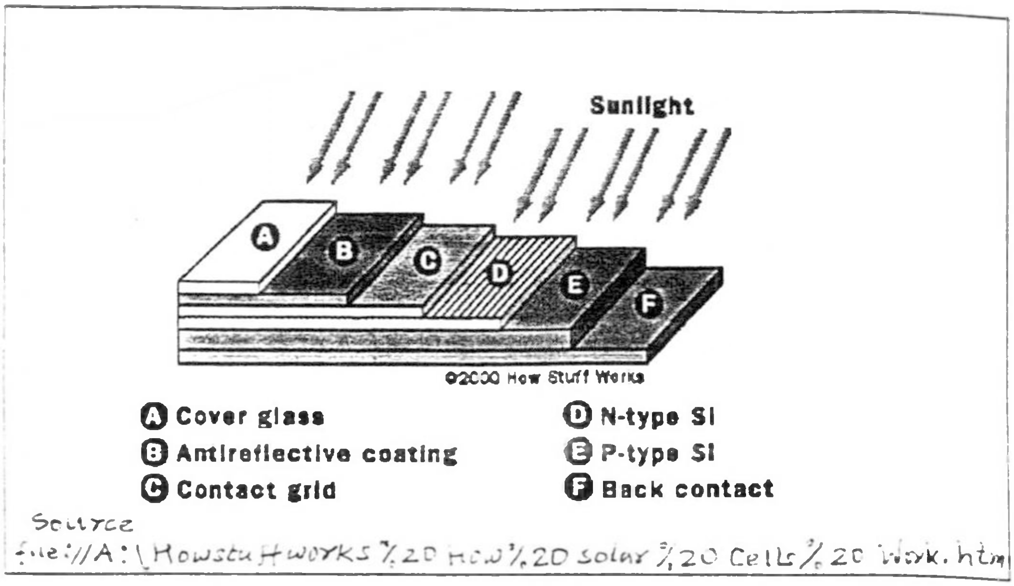


Fig.2. Basic structure of Silicon PV Cell

These days other materials such as GaAs, $CuInSe_2$, CdTe and amorphous Silicon are being used in PV Cells. Now even two different kinds of materials (high band gap on the window side and low band gap on the base side) are being tried. Such cells are more efficient and have been identified as multijunction Cells.

The Solar Cells that you see on Calculators and Satellites are photovoltaic (PV) cells also known as modules.

On a bright sunny day we receive about 1000 W of energy per square meter. We venture into collecting most of it to power our homes. If it could be done we will have "Solar revolution". However, it is limited today to power electrical systems on satellites and frequently for emergency road signs, remote tracks, on buoys and in calculators etc.

Identification of Transistor terminals

Transistor is a device made from two p-n junctions connected so that we have either p-n + n-p ⇒ p n p or n-p + p-n ⇒ npn configuration. Still it cannot be got from two p-n diode connected in the above manner. This is essentially because the intermediate doped semiconductor (also called base) is very thin and lightly doped. Also, base should not draw any current when in circuit (ideally).

When forward biased, however, both p-n junctions allow flow of current very easily i.e., the forward bias resistance between emitter and base and between collector and base is usually very small.

The base is common to both these junctions. Hence if transistor terminals are to be identified with a multimeter (in Ohm meter mode) base terminal is that which shows low resistance when tested with other two terminals separately. But in opposite polarity the p-n junctions get reverse biased hence resistance should be very large between base and any of the two remaining terminals. Fig.1 illustrates this point.

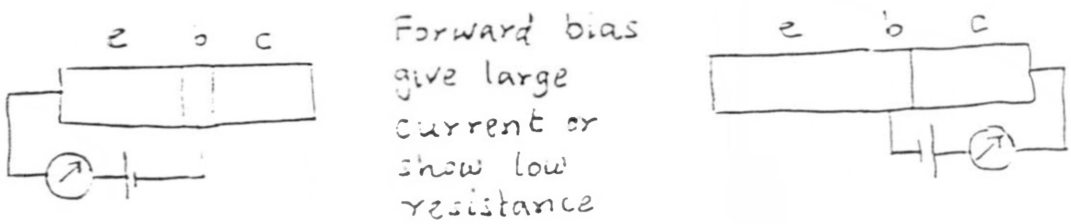
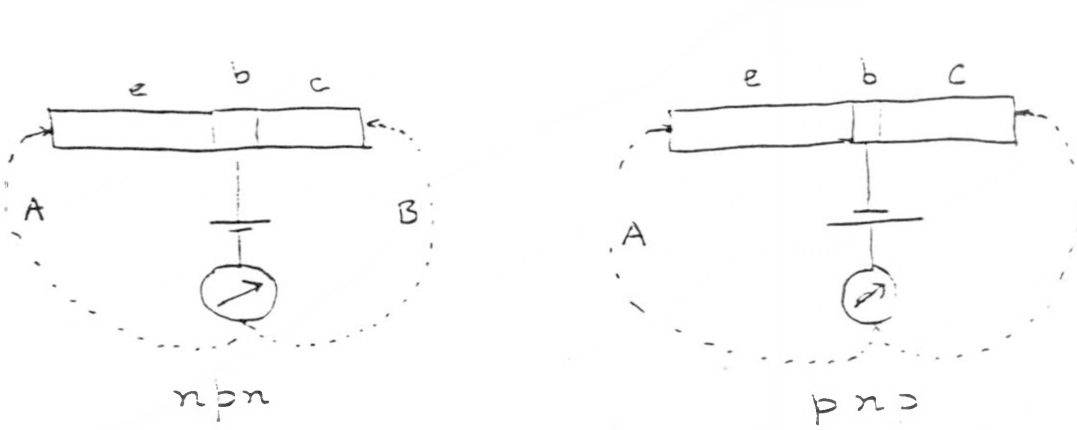


Fig.1

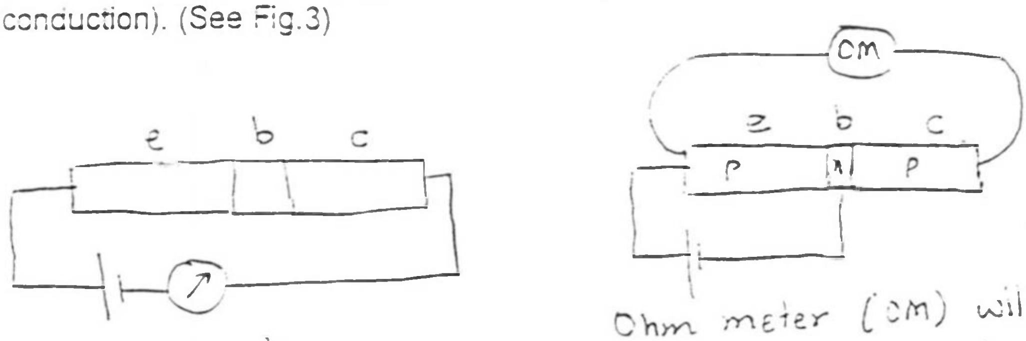
When junction is in forward bias mode and fixed +ve terminal of multimeter (MM in Ohmmeter mode) is connected to the base then the transistor is n p n. But if in forward bias mode fixed terminal is –ve connected to base of multimeter, then the transistor is p n p. (See Fig.2). The common terminal is identified as base.



When transistor is forward biased whether A or B mode of connection is used.

Fig.2

If transistor is OK the resistance between emitter and collector is ideally infinite The resistance between collector and emitter ishowever, ideally zero when emitter base junction is forward biased (transistor is said to be in the state of conduction). (See Fig.3)

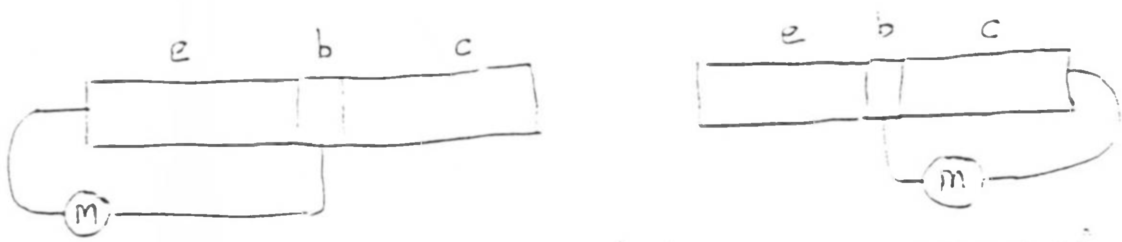


no current or infinite resistance

Ohm meter (OM) will show zero resistance between e & c when e & b are forward biased

Fig.3

Theoretically emitter is heavily doped than the collector yet construction wise emitter-base forward resistance is a few Ohms larger than the base-collector forward resistance. By measuring the resistance of the two p-n junctions of transistor separately, we can identify which should be collector and which should be emitter. (Fig.4)

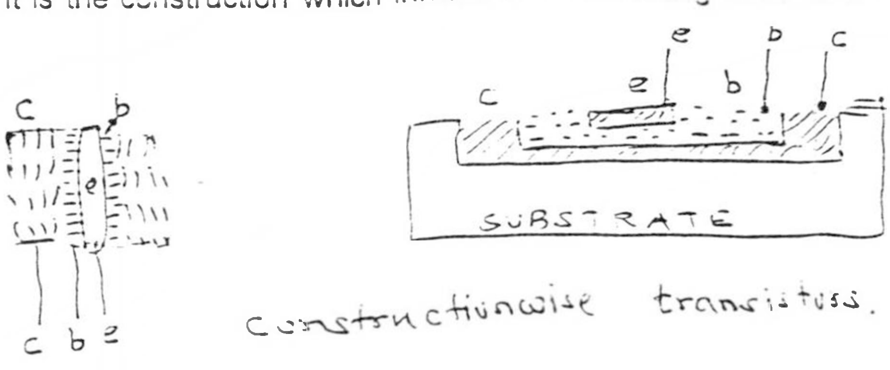


Forward bias resistance as measured by multimeter for bc terminals is slightly less than for eb terminals.

Fig.4

Usually manufacturers provide a dot or some mark near the collector terminal. for ease in identification.

Note: It is the construction which inhibits collector being used as emitter and vice-versa.



Constructionwise transistors.

21 day Training Programme in Physics for the PGTs of Navodaya Vidyalayas
2.6.2003 – 22.6.2003

Provisional Time Table

Date/Day	9.00 am – 11.00 am	11.30 am - 1.00 pm	2.00 pm – 4.00 pm	4.15 pm – 5.30pm
2.6.03 Monday	Registration & Inauguration	Identification of difficult areas	Lab work-planning RN+MNB	Discussion on Lab work NRN
3.6.03 Tuesday	Pretest	Problem solving approach-PRR	Lab work PRL+RN	Discussion/seminar SSR
4.6-03 Wednesday	MNB	Errors & significant figures PRR	Lab work PRL+RN	Discussion/seminar NRN
5.6.03 Thursday	PRR	NRN	Lab work PRL+RN	Discussion/seminar SSR
6.6.03 Friday	MNB	KV	Lab work NRN+RN	Discussion/seminar SSR
7.6.03 Saturday	MNB	PRR	Lab work SSR+PRL	Discussion/seminar RN
8.6.03 Sunday	---------------------- Project Work----------------------			
9.6.03 Monday	NRN	MNB	Lab work MNB+PRL	Discussion/seminar RN
10.6.03 Tuesday	SSR	NRN	Lab work RN+MNB	Discussion/seminar PRL
11.6.03 Wednesday	MAC	SSR	Lab work SSR+RN	Discussion/seminar MNB
12.6.03 Thursday	---------COMPUTER LAB---- SSR+PRL+MNB		Lab work RN+PRL	Discussion/seminar SSR
13.6.03 Friday	CG	MAC	Lab work MNB+MMS	Discussion/seminar MMS+PRL
14.6.03 Saturday	CRN	CG	Lab work MNB+RN	Discussion/seminar Library work
15.6.03 Sunday	---------------Project Work ------------			
16.6.03 Monday	CRN	VDB	Lab work MMS+SSR	Discussion/seminar MNB
17.6.03 Tuesday	NNP	MAC	Lab work NRN+MMS	Discussion/seminar SSR
18.6.03 Wednesday	NNP	VDB	Lab work NRN+MNB	Discussion/seminar MMS
19.6.03 Thursday	Popular Talk	LIBRARY	Lab work PRL+MMS	Discussion/seminar RN
20.6.03 Friday	CRN	KV	Popular Talk	
21.6.03 Saturday	CRN	CRN	---------Library Work---------	
22.6.03 Sunday	Post Test	Discussion	Feedback	Valedictory session

PRR	-	P.Ramachandra Rao	SSR	-	S.S.Raghavan
MAC	-	M.A.Chandrasekhar	PRL	-	P.R.Lalitha
C.G.	-	C.Gurumurthy	MNB	-	M.N.Bapat
CRN	-	C.R.Natraraj	NRN	-	N.R.Nagaraja Rao
KV	-	Kalpana Venugopal	VDB	-	V.D.Bhat
NNP	-	N.N. Prahallada	MMS	-	M.M. Sahajwani

Popular Talk (1) Evolution of Stars - Prof. G.T. Narayana Rao

(2) Management Skills - Prof. K. Shamanna

R.Narayanan
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