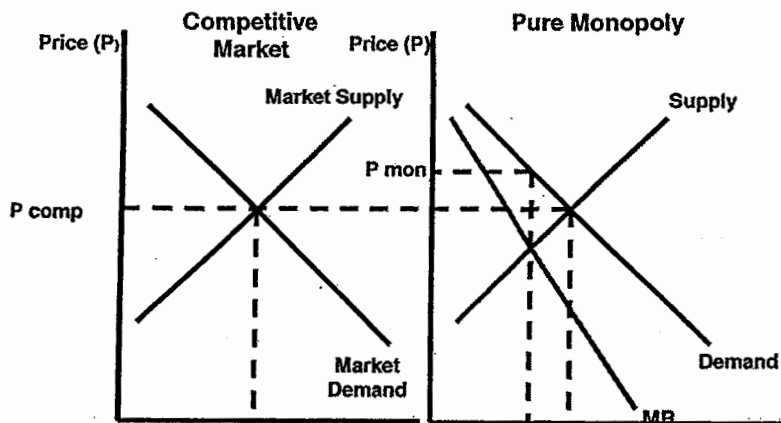


[Version 1 2007] [CRIBS]

SECTION A *Business Economics*
Answer not more than one question from this section

1 (a) Under what circumstances would you expect a monopoly to charge a higher price than if the industry was operating under perfect competition? [5]

Under normal circumstances it would be expected that a monopoly would charge a higher price than under perfect competition. A monopoly is a firm that is the sole supplier of a certain product (almost always by virtue of being protected by some or other barrier to entry certain barriers to entry). Thus a monopolist can take the market demand curve as its own demand curve and can enjoy some power over the setting of price or output as shown below. Conversely a perfectly competitive market has many suppliers each with an insignificant share of markers. Each firm is too small to affect price via a change in market supply – each individual firm is assumed to be a price taker (see below).



(b) Under what circumstances might monopolies provide economic and social benefits? [5]

The potential benefits of monopolies include the following. First, the ability to exploit economies of scale (this can be illustrated with a diagram). Second, the ability and stability to undertake significant research and development spending (although not pure monopolies there are examples in the aerospace and pharmaceuticals industries). Third, the ability to compete more effectively in international markets.

Some candidates may also discuss the potential benefits from natural monopolies.

(c) Using the Keynesian Consumption Function model, explain the potential impact of cutting income taxes on the level of aggregate demand. [5]

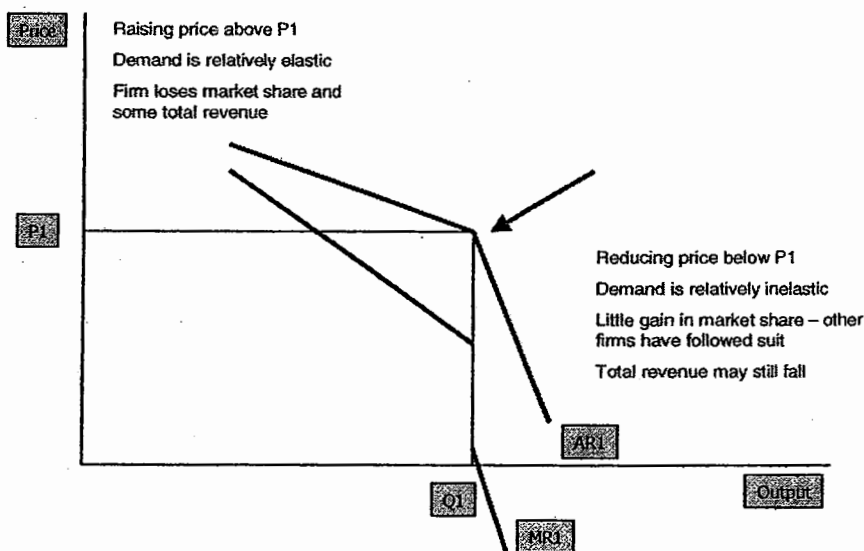
According to the Keynesian consumption function, current consumption (C) is determined by current personal disposable income (Y). This reduction in income taxes will lead to an increase in current personal disposable income leading to increase consumption and this arise in aggregate demand. This may lead to a rise in output or inflation (or possibly both) depending on availability of resources.

(d) Using either the Life Cycle model or the Permanent Income model, explain the potential impact of cutting income taxes on the level of aggregate demand. [5]

Both the Life Cycle model and the Permanent Income model use a notion of long-run or normal income. In both, the consumer's consumption is based on his/her long run average income. Thus, temporary income changes have less impact on spending than permanent income changes. If a tax cut is considered to be temporary, it will have little impact on consumption – and the extra disposable income may be saved in anticipation of compensating income tax rises in the future. If a tax cut is considered to be permanent it will lead to rise in consumption.

2 (a) Describe the kinked demand curve theory. [6]

Firms believe that their competitors would follow them if they were to reduce their price below the prevailing market price, and that it is therefore not possible to gain much in the way of extra sales by reducing price (i.e. demand is relatively inelastic below the market price). Firms believe that their competitors would not follow them if they were to raise their price above the prevailing market price (i.e. demand is relatively elastic above the market price). See below.



- (b) What is the principal-agent problem? Describe one example the problem. [5]

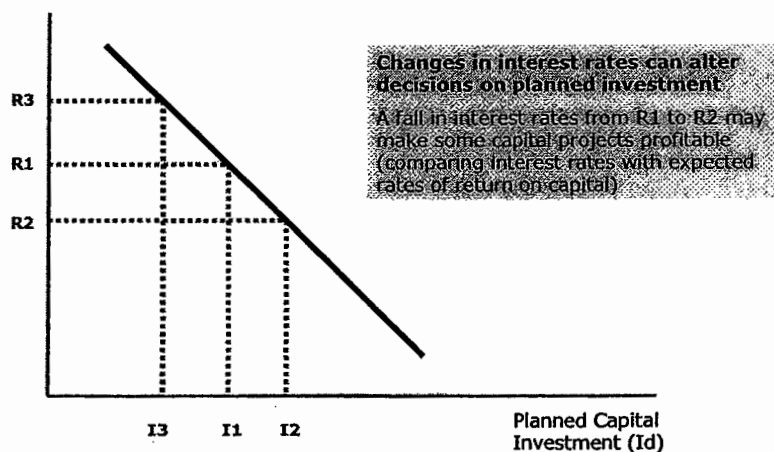
One person, the principal, hires an agent to perform tasks on their behalf but cannot ensure that the agent performs them in exactly the way the principal would like. The efforts of the agent are impossible or expensive to monitor and the incentives of the agent differ from those of the principal. The problem is created because of asymmetric information as the two parties share unequal information. It does not arise if an enforceable legal contract can be drawn up to specify all the duties of the agent.

Example of the problem: the management of companies on behalf of shareholders – managers wish to maximise their benefits (salaries and so on) whereas shareholders wish to maximise their returns (dividends and capital gains). In this example there are various ways to minimise the problem such as: share-ownership schemes; performance-related pay; incentive pay schemes; pay related to profitability; long-term employment contracts for senior management; and the take-over threat. Other examples can be given.

- (c) What impact would the following have on the level of investment in the macroeconomy:

- (i) a decrease in the rate of interest [3]

A decrease in the interest rate will increase the amount of autonomous investment in the macro economy. This is because the downward sloping marginal efficiency of capital curve indicates that there are more profitable investment opportunities at lower interest rates. See below.



(ii) an expected fall in future consumption [3]

Consumption is a major component of aggregate demand. If the level of expected demand falls, then the Accelerator model suggests that the desired size of capital stock will also fall. Net capital investment is the amount by which the required capital stock changes. It follows that the amount of investment depends on the size of the expected change in output. If the latter is decreasing, the desired size of the capital stock needs to be reduced – then investment will fall (and may become negative if existing capital stock is scrapped).

(iii) a reduction in the real exchange rate [3]

This may work through different mechanisms. First, if the lower exchange rate boosts trade it will boost demand and may stimulate income induced investment (see 2cii above). But this will mainly be in tradeable sectors – that is sectors that produce exports or import substitutes. Second, it may increase the price of imported capital goods – this may impact on investment depending on the availability of capital goods in the domestic economy. Third, if the change in the exchange rate has any subsequent impact on interest rates this may also influence investment through the impact on the marginal efficiency of capital.

SECTION B

3. (a) (i) Possible alternative types of wall that might be used are as follows:
- steel sheet piles, driven or jacked into the ground
 - reinforced concrete diaphragm walls, installed in slurry-supported trenches cut in the ground (first the trench is cut – typical plan dimensions 4m x 1m - then a reinforcement cage is lowered into the trench, then concrete is poured via a 'tremie' tube inserted to the bottom of the trench, displacing the slurry)
 - reinforced concrete secant pile wall in which alternate 'hard' and 'soft' bored piles are constructed – the 'soft' piles, comprising bentonite-cement are constructed first, and then the 'hard' reinforced concrete piles cutting into the 'soft' piles

(ii) Potential problems that can occur with water during excavation are as follows:

- leakage through a defect in the wall (caused, for example, by de-clutching during driving sheet piles or a problem with a joint between diaphragm wall panels)
- seepage around the bottom of the wall causing erosion of sands and possibly 'piping' failure
- uplift in cases where alternate clay and sand layers lead to high water pressure in the sands exceeding the weight of the clay layers above, thereby causing uplift

(b) (i) Stability ratio, N, is defined as

$$N = (\gamma z - \sigma_T) / S_u$$

where γ = unit weight of soil, z = depth of tunnel axis below ground level, σ_T = tunnel face pressure, and S_u = undrained shear strength of the clay

In the case of an open face tunnel $\sigma_T = 0$

In the case of the stiff clay for an open face tunnel

$$N = (25 \times 20) / 250 = 2.0$$

In the case of the soft clay for an open face tunnel

$$N = (25 \times 17) / 50 = 8.5$$

If N exceeds about 5 problems with face stability will occur. Hence tunnel construction with an open face will be satisfactory for the stiff clay (where $N = 2.0$) but will not be possible for the soft clay (where $N = 8.5$).

(ii) For N not to exceed 5 in the soft clay, a tunnel face pressure must be applied, using a closed face tunnelling machine:

$$N = (25 \times 17 - \sigma_T) / 50 = 5$$

4

$$\text{hence } \sigma_T = 175 \text{ kN/m}^2$$

(c) The Gaussian shape of the settlement trough results in a *sagging* mode directly above the tunnel centre-line and a *hogging* mode on either side of the centre-line. Masonry buildings are sensitive to differential settlement and curvature associated with these sagging and hogging modes. Sagging mode means that the bottom of the building (ie the foundations) experiences tension and the top compression, whereas the hogging mode means the reverse – the bottom of the building experiences compression and the top experiences tension. Masonry buildings are sensitive to tension (but are much stronger in compression) and are more likely to experience damage in hogging mode. Therefore a building directly above the tunnel centre-line may only experience the sagging mode and be less damaged than a building off-set from the centre-line experiencing the hogging mode.

4

Q.4 Solution

(a) Short term conditions

- Use undrained shear strength of clay to calculate horizontal total stress, σ_h

$$\text{active} \quad \sigma_h = \sigma_v - 2s_u$$

$$\begin{aligned} (\sigma_v &= \text{total vertical stress}) \\ s_u &= \text{undrained shear strength} \end{aligned}$$

$$\text{passive} \quad \sigma_h = \sigma_v + 2s_u$$

- For sand, calculate effective horizontal stress (σ_h') and add water pressure (u)
total horizontal stress σ_h

$$\sigma_h = \sigma_h' + u$$

$$\sigma_h' = K_a \sigma_v' \quad (\sigma_v' = \text{vertical effective stress})$$

$$K_a = \frac{1 - \sin \phi'}{1 + \sin \phi'} \quad (\text{for a smooth wall})$$

$$\phi' = 30^\circ \Rightarrow K_a = \frac{1 - \sin 30^\circ}{1 + \sin 30^\circ} = 0.33$$

Backfilled side (active pressures)

$$\text{At ground surface} \quad \sigma_v = \sigma_v' = 60 \text{ kN/m}^2$$

$$\sigma_h = \sigma_h' = K_a \sigma_v' = 0.33 \times 60 = \underline{20 \text{ kN/m}^2}$$

At 5m depth
(to water level)

$$\sigma_v = 60 + 5 \times 18 = 140 \text{ kN/m}^2$$

$$u = 0, \therefore \sigma_v' = \sigma_v; \quad \sigma_h = \sigma_h'$$

$$\sigma_h = \sigma_h' = 0.33 \times 140 = \underline{46.2 \text{ kN/m}^2}$$

At 7m depth
(sand/clay
interface)

At bottom of sand

$$\sigma_v = 60 + (5 \times 16) + (2 \times 19) = 178 \text{ kN/m}^2$$

$$u = 2 \times 10 = 20 \text{ kN/m}^2$$

$$\therefore \sigma_v' = 158 \text{ kN/m}^2$$

$$\sigma_h' = 0.33 \times 158 = 52.1 \text{ kN/m}^2$$

$$\therefore \sigma_h = \sigma_h' + u = 52.1 + 20 = \underline{72.1 \text{ kN/m}^2}$$

At top of clay

$$\sigma_v = 178 \text{ kN/m}^2 \text{ (see above)}$$

$$\sigma_h = \sigma_v - 2s_u = 178 - (2 \times 70) = \underline{38 \text{ kN/m}^2}$$

At 10m depth

$$\sigma_v = 60 + (5 \times 16) + (2 \times 19) + (3 \times 20) \\ = 238 \text{ kN/m}^2$$

$$\sigma_h = \sigma_v - 2s_u = 238 - (2 \times 70) = \underline{98 \text{ kN/m}^2}$$

River side (passive pressures)

At 7m below top
of tunnel
(at river bed level)

$$\sigma_h = \sigma_v = u = 2 \times 10 = \underline{20 \text{ kN/m}^2}$$

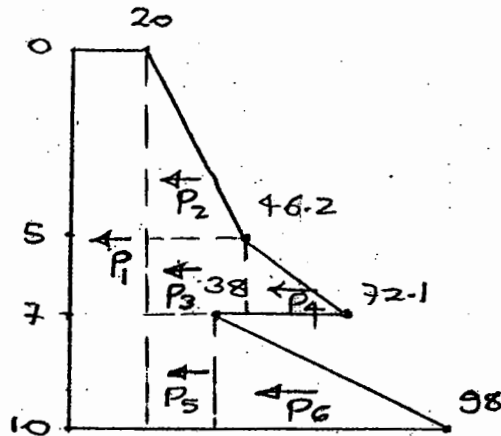
$$\sigma_h = \sigma_v + 2s_u = 20 + (2 \times 70) = \underline{160 \text{ kN/m}^2}$$

At 10m below top
of tunnel

$$\sigma_v = 20 + (3 \times 20) = 80 \text{ kN/m}^2$$

$$\sigma_h = 80 + (2 \times 70) = \underline{220 \text{ kN/m}^2}$$

Back filled side : total horizontal stresses
(active pressures)



$$P_1 = 10 \times 20 = 200 \text{ kN/m}$$

$$P_2 = \frac{1}{2} \times 5 (46.2 - 20) = 65.25 \text{ kN/m}$$

$$P_3 = 2 \times (46.2 - 20) = 52.4 \text{ kN/m}$$

$$P_4 = \frac{1}{2} \times 2 (72.1 - 46.2) = 25.9 \text{ kN/m}$$

$$P_5 = 3 \times (38 - 20) = 54 \text{ kN/m}$$

$$P_6 = \frac{1}{2} \times 3 (98 - 38) = 90 \text{ kN/m}$$

$$\Sigma (P_1 \dots P_6) = \underline{487.8 \text{ kN/m}}$$

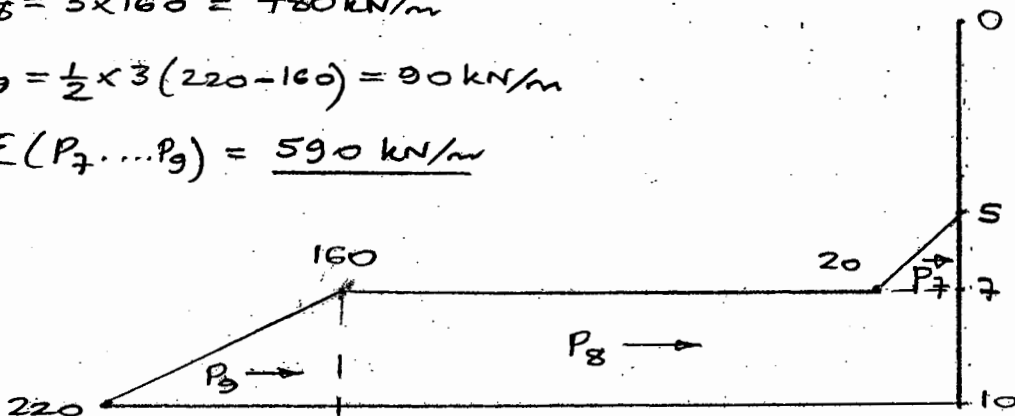
River side : total horizontal stresses
(passive pressures)

$$P_7 = \frac{1}{2} \times 2 \times 20 = 20 \text{ kN/m}$$

$$P_8 = 3 \times 160 = 480 \text{ kN/m}$$

$$P_9 = \frac{1}{2} \times 3 (220 - 160) = 90 \text{ kN/m}$$

$$\Sigma (P_7 \dots P_9) = \underline{590 \text{ kN/m}}$$



(b) Factor of safety = $\frac{\text{total passive force}}{\text{total active force}}$

$$= \frac{590}{487.8} = \underline{\underline{1.21}}$$

(c) Softening of clay on backfilled side to long term conditions

- horizontal stress only changes in the clay

At 7m depth
(sand/clay interface)

At top of clay

$$\sigma_v = 178 \text{ kN/m}^2 \text{ (see above)}$$

$$u = 2 \times 10 = 20 \text{ kN/m}^2$$

$$\therefore \sigma_v' = \sigma_v - u = 178 - 20 = 158 \text{ kN/m}^2$$

For clay, $\phi' = 20^\circ$ $K_a = \frac{1 - \sin 20}{1 + \sin 20} = 0.49$

$$\therefore \sigma_h' = K_a \sigma_v' = 0.49 \times 158 = 77.4 \text{ kN/m}^2$$

$$u = 20 \text{ kN/m}^2$$

$$\therefore \sigma_h = 77.4 + 20 = \underline{97.4 \text{ kN/m}^2}$$

At 10m depth

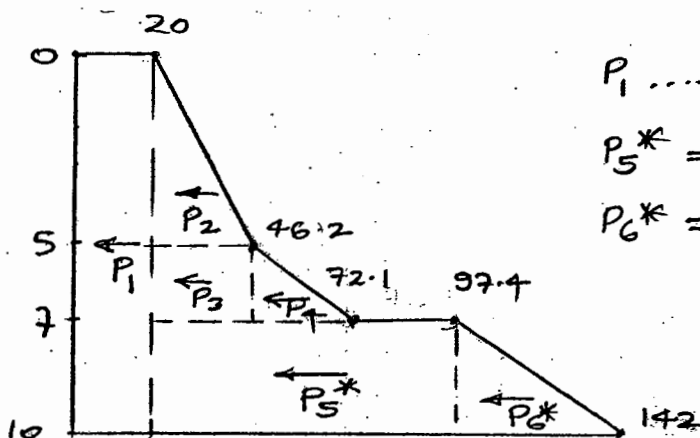
$$\sigma_v = 238 \text{ kN/m}^2 \text{ (see above)}$$

$$u = 5 \times 10 = 50 \text{ kN/m}^2$$

$$\therefore \sigma_v' = \sigma_v - u = 238 - 50 = 188 \text{ kN/m}^2$$

$$\sigma_h' = K_a \sigma_v' = 0.49 \times 188 = 92 \text{ kN/m}^2$$

$$\sigma_h = \sigma_h' + u = 92 + 50 = \underline{142 \text{ kN/m}^2}$$



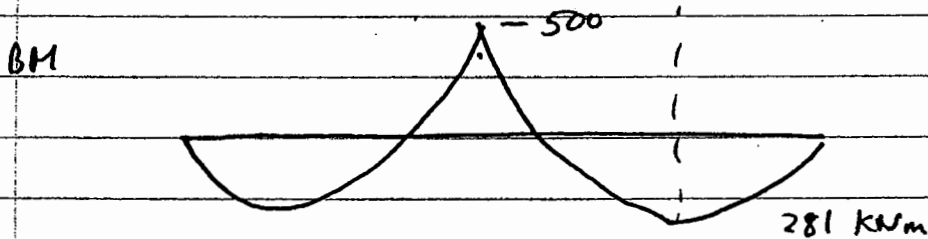
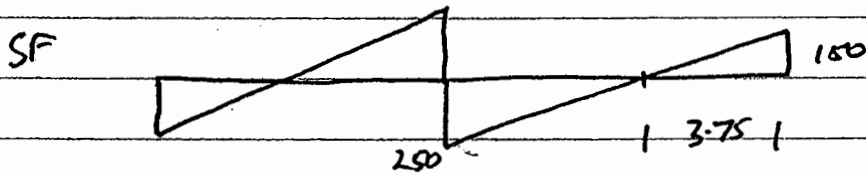
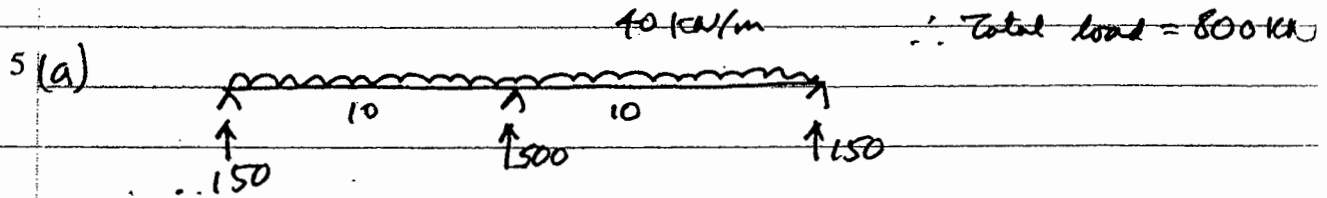
$P_1 \dots P_4$ same as before

$$P_5^* = 3 \times (97.4 - 20) = 232 \text{ kN/m}$$

$$P_6^* = \frac{1}{2} \times 3 (142 - 97.4) = 67 \text{ kN/m}$$

$$\sum P_1 + P_2 + P_3 + P_4 + P_5^* + P_6^* = \underline{643 \text{ kN}}$$

Factor of safety = $\frac{590}{643} = \underline{0.92}$



(b) Limit on singly reinforced moment capacity is

$$M = 0.15 f_{cu} b d^2 = 0.15 \cdot 40 \cdot 250 \cdot 500^2 \cdot 10^{-6}$$

$$= 375 \text{ kNm}$$

\therefore Can be singly reinforced in sagging but not in hogging.

(c) Sagging bending.

Assume neutral axis depth = $0.4d$ (x/d) (any reasonable value will do)

$$A_s \text{ required} = \frac{M}{0.87 \cdot f_y \cdot d (1 - x/2)} = \frac{281 \cdot 10^6}{0.87 \cdot 460 \cdot 500 (0.3)}$$

$$= 1755 \text{ mm}^2$$

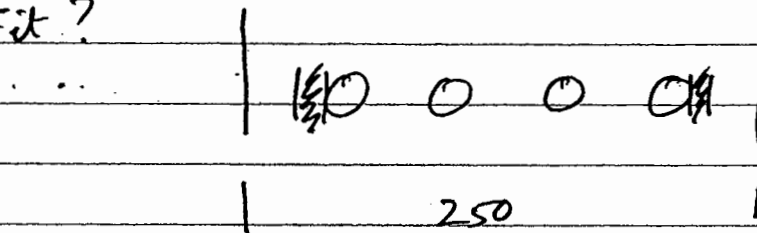
Check $x = \frac{2.175 f_y A_s}{f_{cu} b d} = 0.351$

Rebar $A_s = 1703 \text{ mm}^2$

$x = 0.340$
 (near enough)

$$4 \text{ } 25 \text{ mm bars} = \underline{\underline{1964 \text{ mm}^2}}$$

Fit?



$$2 \times \text{Cover (say } 30 \text{ mm)} + 2 \text{ shear links (say } 16) = 92$$

$$\therefore 92 \times 4 \times 25 = 92 \quad \therefore \text{Clears } \approx 20 \text{ mm}$$

(Just OK)

Hogging bending

Applied moment = 500 kNm

375 kNm can be carried by the concrete in compression.

$$\therefore \text{top steel needed to carry } 125 \text{ kNm}$$

Assume $d' = 60 \text{ mm}$ (any reasonable value 40 \Rightarrow 100 could be accepted)

$$0.75 f_y A_s' (d - d') = 125 \text{ kNm}$$

$$\therefore A_s' = \frac{125000000}{0.8760 \times 40} = 823 \text{ mm}^2$$

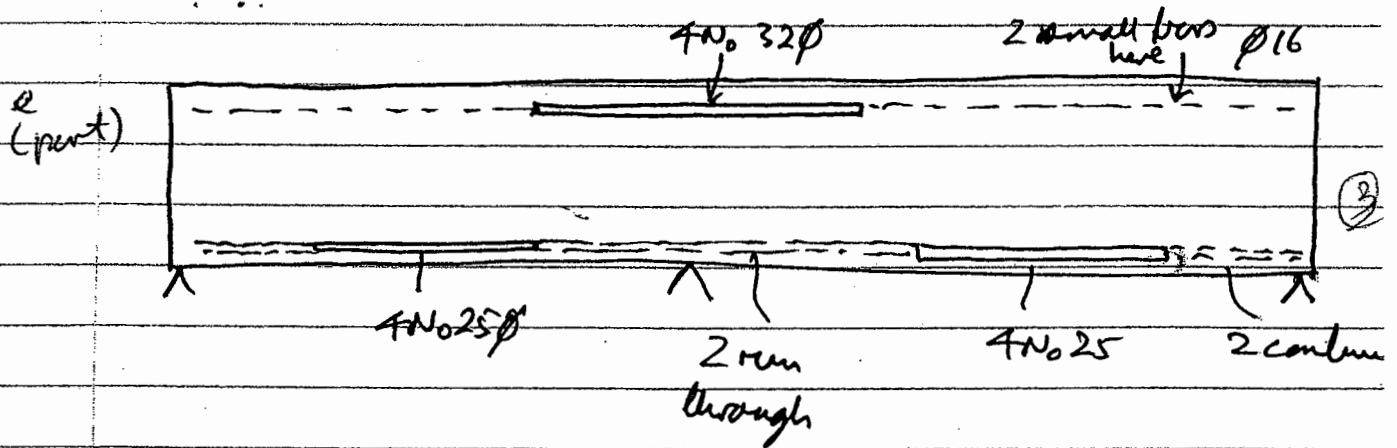
(2 No 25 mm bars would do $\approx 4 \times 20 \text{ mm}$)

As required to balance compressive force

$$0.87 A_s f_y = 0.75 f_y A_s' + \frac{b \cdot d}{2} \cdot 0.4 f_{cu}$$

$$\therefore 0.87 A_s f_y = 284000 + 1000000 = \text{1284} \cdot 10^3$$

$$\therefore A_s = 3208 \text{ mm}^2 \quad (4 \text{ No } 32 \text{ mm bar})$$



(d) Maximum shear force = 250 kN

$$\text{Nominal shear stress} = \frac{250 \cdot 10^3}{250 \cdot 500} = 2 \text{ MPa}$$

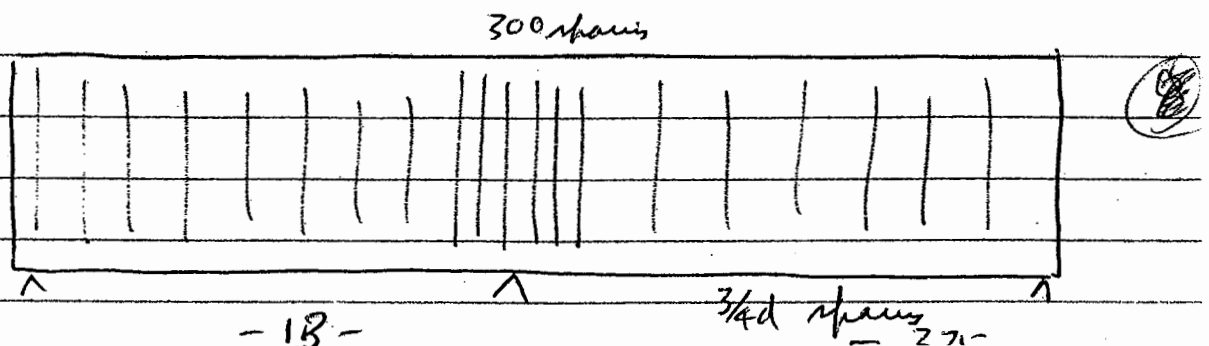
\therefore Shear links required

Shear formula gives $v_c = 0.83 \text{ N/mm}^2$

$$\therefore v_s = 1.17 \text{ N/mm}^2$$

$$s = \frac{f_y A_{sv}}{v} \quad \text{Assume } A_{sv} = 2 \text{ No } 12 \text{ bars} = 226 \text{ mm}^2$$

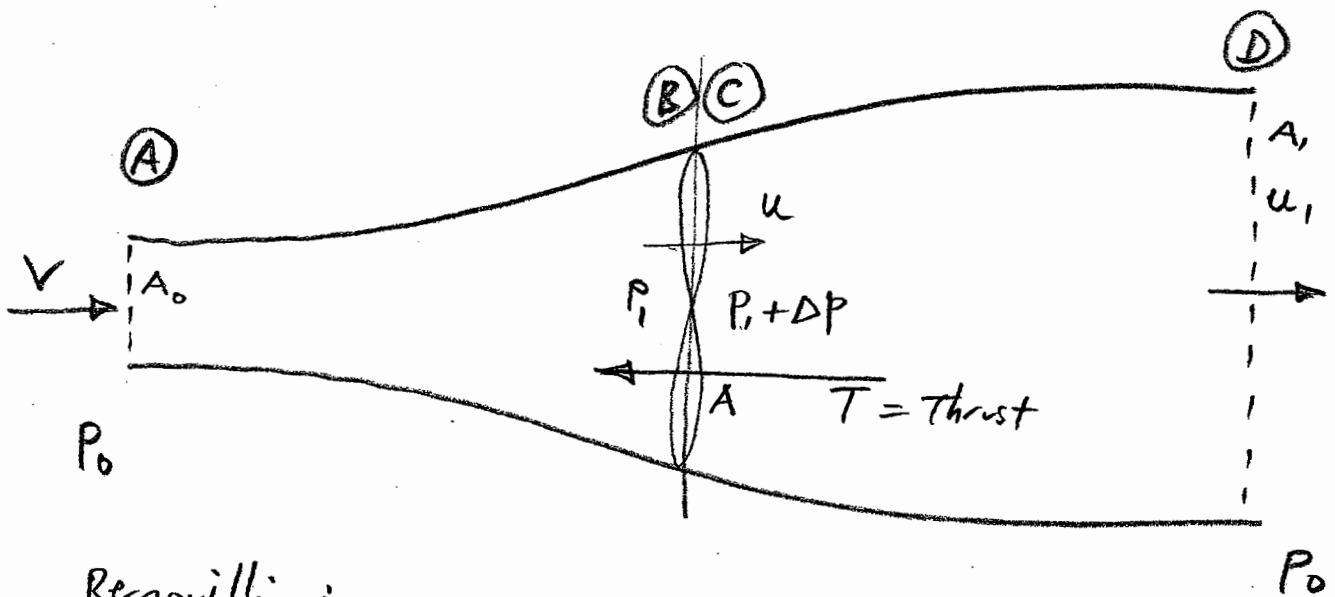
$$= \frac{460 \cdot 220}{250 \cdot 1.17} = 345 \text{ mm} \quad (\text{say } 300) < \frac{3}{4} d \therefore \text{OK}$$



Mechanics Materials Design

6 (a) Betz limit is the ratio of the power extracted from a "perfect" wind turbine to the power contained in free air moving through an area equal to the swept area of the turbine. It is based on certain assumptions eg lamina, irrotational, inviscid flow. If these assumptions are violated then the "limit" no longer holds and may be exceeded.

(b) Control volume for air passing through a turbine of swept area A



Bernoulli:

(A) → (B)

$$P_0 + \frac{1}{2} \rho V^2 = P_1 + \frac{1}{2} \rho u^2$$

(C) → (D)

$$P_0 + \frac{1}{2} \rho u_1^2 = P_1 + \Delta P + \frac{1}{2} \rho u^2$$

$$\text{Subtract } \therefore \Delta P = \frac{1}{2} \rho (u_1^2 - v^2) \quad (2)$$

$$\text{Thrust } T = -\Delta P A = \frac{1}{2} \rho A (v^2 - u_1^2)$$

Momentum flux through C.V.

$$\dot{m} v - \dot{m} u_1 = T$$

$$\therefore \dot{m} (v - u_1) = \frac{1}{2} \rho A (v - u_1)(v + u_1)$$

$$\therefore \dot{m} = \frac{1}{2} \rho A (v + u_1)$$

$$\text{but } \dot{m} = \rho A u \quad \therefore u = \frac{1}{2} (v + u_1)$$

Introduce induction factor a defined by

$$u = v(1-a)$$

$$\therefore v(1-a) = \frac{1}{2}(v + u_1)$$

$$\therefore u_1 = 2v - 2av - v = v(1-2a)$$

$$\text{Power} = T \cdot u \quad (\text{force} \times \text{velocity})$$

$$= \frac{1}{2} \rho A (v^2 - u_1^2) u$$

$$= \frac{1}{2} \rho A v^3 (1 - (1-2a)^2) (1-a)$$

$$= \frac{1}{2} \rho A v^3 (1 - 1 + 4a - 4a^2) (1-a)$$

$$= \frac{1}{2} \rho A v^3 4a(1-a)^2$$

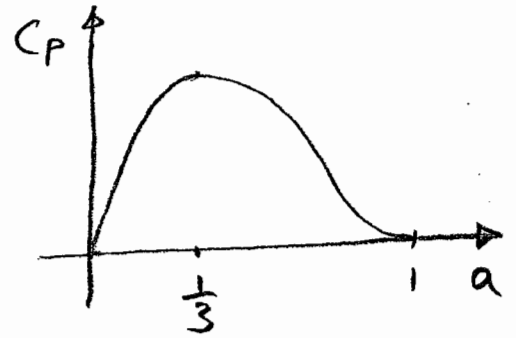
Power in the flow across an area $A = P_{\text{ref}}$

$$P_{\text{ref}} = \frac{1}{2} \dot{m} v^2 \quad \text{rate of flow of K.E.}$$

$$= \frac{1}{2} \rho A v^3 \quad \text{in free flow}$$

Power coefficient $C_p = \frac{P}{P_{ref}}$

$$C_p = 4a(1-a)^2$$



This is maximum when

$$\frac{1}{4} \frac{dC_p}{da} = 0$$

$$\therefore (1-a)^2 - 2a(1-a) = 0$$

$$\therefore (1-a)(1-a-2a) = 0$$

$$\therefore (1-a)(1-3a) = 0$$

$a=1$ is a minimum
 $a=\frac{1}{3}$ is a maximum

$$C_p(a=\frac{1}{3}) = 4 \cdot \frac{1}{3} \cdot \left(1 - \frac{1}{3}\right)^2 = \frac{16}{27} = 0.59$$

This is the Betz Limit.

For full marks an answer need to have:

- a good sketch showing the C.V.
- clear application of Bernoulli, momentum and power flow
- correct use of reduction factor
- correct differentiation to obtain $a = \frac{1}{3}$

Many candidates had rote-learned an answer and missed steps out

This question was attempted by nearly all candidates. Average mark 67%

(c) Thrust for $a = \frac{1}{3}$

$$T = \frac{P}{a} = \frac{\frac{1}{2} \rho A V^3 4a(1-a)^2}{V(1-a)}$$

$$= \frac{1}{2} \rho A V^2 4a(1-a)$$

$$= \frac{1}{2} \rho A V^2 4 \frac{1}{3} \left(1 - \frac{1}{3}\right)$$

$$= \frac{4}{9} \rho A V^2$$

Or, if (b) was omitted use momentum flow:

$$\text{Thrust} = \dot{m} \left(V - \frac{V}{3} \right) \quad \left(\frac{V}{3} \text{ given} \right)$$

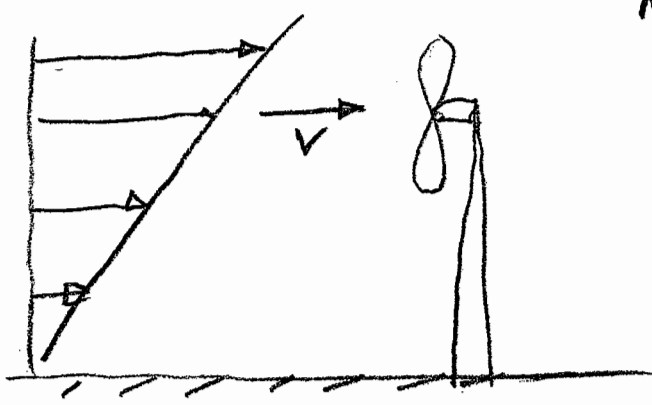
$$= \rho A u \frac{2}{3} V$$

$$u = V(1-a) = \frac{2}{3} V$$

$$\therefore \text{Thrust} = \frac{4}{9} \rho A V^2$$

(d) Betz limit can be exceeded if the assumptions (laminar, inviscid, irrotational etc) are violated.

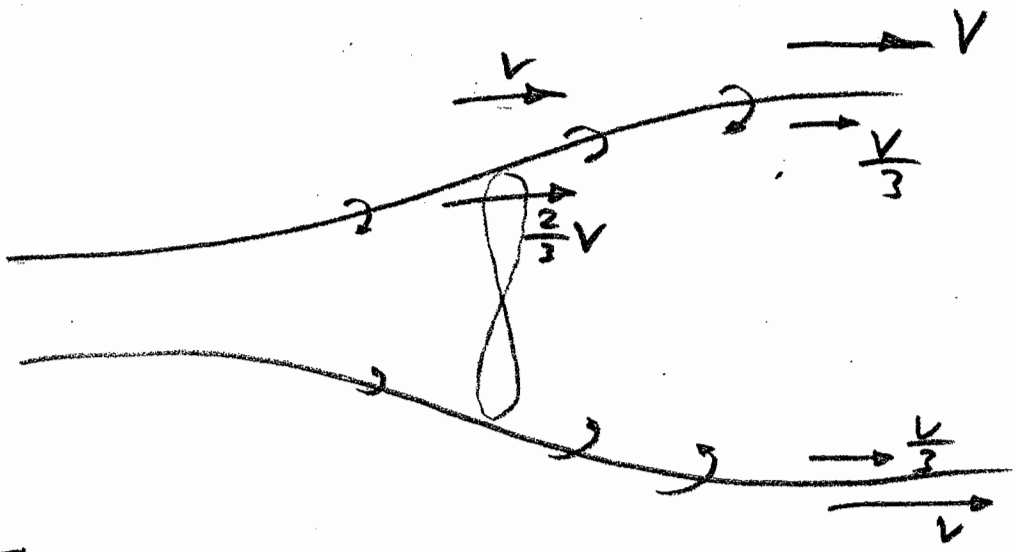
(i) The easiest violation is when the incoming velocity profile is not uniform. If V is measured at the



centre of the swept area then power can be higher than expected because

power $\propto V^3$ and the upper parts of the swept area see faster air.

(ii)



The step-change of velocity across the C.V. is unsustainable. Air outside the c.v. is damn in and some of its energy can be extracted - a bonus.

7 (a) (A significantly fuller answer than expected to this part.)

Material selection. The need for lightweight, high strength and stiffness in bending, tends to drive towards composites and wood. CFRP has good fatigue properties, while GFRP has a steeper S-N curve, so is more susceptible to fatigue. Nevertheless the cost of CFRP is prohibitive except in the spar. In other words, there is a trade-off between performance and cost. Difficulty in defining appropriate material properties against fatigue (including the effects of mean stress and stress concentration effects (see below), means that safety factors need to be increased on material properties.

Loading. For larger blades gravity gives substantial component which reverses each cycle, while wind loading gives bending (compression on downwind side, tension on upwind side). Very large numbers of cycles are applied, e.g. 10^8 . Wind loading is random in nature, so that the exact loading will depend on the location and safety factors/risk factors need to be incorporated. Some locations will have higher mean wind speeds and loads, others will be more gusty. However the control system will prevent overload in the event of storms e.g. feathering the blades, so that this aspect of the operational control strategy needs to be integrated. Variation of load over the blade depends on the shape of the blade, via aerodynamic loading factors.

Overall design. The overall shape of the blade gives rise to an increasing bending moment towards the hub. Hence, to keep the loads within the fatigue life, more material needs to be used towards the hub. However aerodynamic design factors makes it acceptable to have a deeper beam towards the hub where the torque and power contribution is lower, countering this effect. Stiffness considerations (which again dictate a variation of material along the length of the blade), may also affect this conclusion. But the overall design strategy, within these constraints, is to ensure that the amount of material is just below the fatigue life-time, with an appropriate safety margin.

Detailed design. Fatigue is sensitive to stress concentrations, so much attention needs to be paid to features and attachment points. Moreover construction may give rise to defects (e.g. knots if wood is used or waviness or ply drops if fibre composites) which can seriously degrade the fatigue life.

Testing. The very large number of cycles presents a problem for testing, as does the scale of larger blades. Nevertheless some full scale tests need to be done to validate the structure. However it is also important (and cheaper) to test coupons and substructures to evaluate the fatigue life-time. [7]

[Examiner's comments: relatively unpopular question, with an average of 62%. Part (a) was not answered so well, with a tendency to go on at length about loading details but to omit other important factors. Parts (b) and (c) were reasonably well answered, though by no means everyone had mastered the rainflow algorithm and a significant minority treated half-cycles of loading as complete cycles for use in the fatigue law. People were not penalised for sensible estimates of σ_{15} in (c), as it wasn't so obvious from the question what value to choose.]

(b). See the marked up diagram below which uses the rainflow method to identify four half-cycles from the troughs and corresponding half-cycles from the peaks (which are the same as the record starts and ends at the same point. The table below then quotes the peak to peak stress range $\Delta\sigma$ and mean stress σ_m [5]

(c) The table below then calculates the effective stress range $\Delta\sigma_0$ for zero mean stress, assuming that the tensile strength σ_{ts} for use in Goodman's rule is 150 MPa (i.e. half the value of stress with $N = 1$, noting that the S-N law quoted is for a peak-to-peak range). [This value for σ_{ts} wasn't obvious from the question and any sensible assumption was OK.]

$$\Delta\sigma = \Delta\sigma_0 \left(1 - \frac{\sigma_m}{\sigma_{ts}} \right)$$

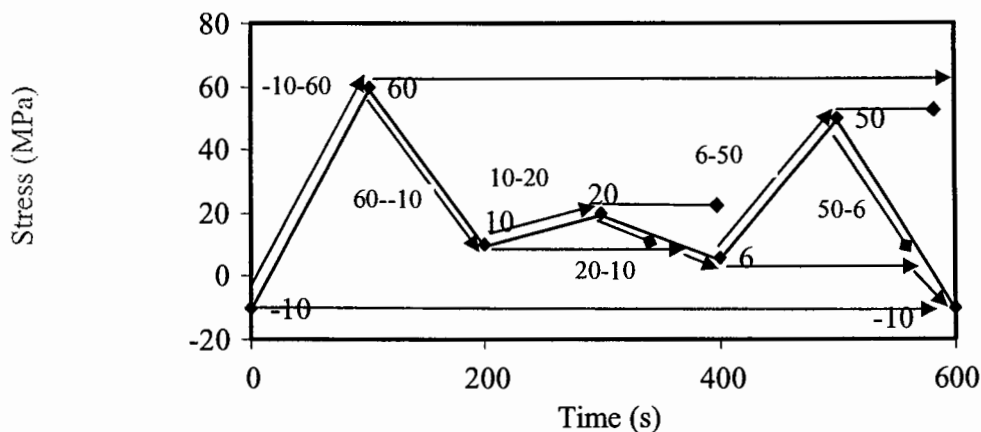
Next find the number of cycles to failure for each stress cycle N_{fi} from the given S-N curve putting in the corresponding value of $\Delta\sigma_0$.

$$N_{fi} = \left(\frac{300}{\Delta\sigma_0} \right)^8$$

Finally Miner's rule is applied, summing up the effect of the given cycle of loading and hence finding the number of repeats N needed for the total life time to be used up via

$$N \times \sum_i \frac{1}{N_{fi}} = 1 \quad \text{and} \quad \text{Life} = N \times 600 \text{ s} = 0.49 \text{ years (clearly too short - need to redesign)} \quad [5]$$

(d) In practice we need a much longer load cycle to get good statistics on stress mean and range. In theory it would be possible to do a calculation summing up the effect of each load cycle, but simpler would be to sort the cycles into bands of range and mean stress (constructing a probability matrix) and evaluate the contribution from each band of mean stress and range. Then calculate the number of repeats of the total cycle as above. Alternatively analytical forms of the load cycle distribution could be used and integrated. [3]



σ_{min}	σ_{max}	$\Delta\sigma$	σ_m	$\Delta\sigma_0$	N_{fi}	$1/N_{fi}$
-10	60	70	25	84	26500	3.77×10^{-5}
10	20	10	15	11.1	2.82×10^{11}	3.54×10^{-12}
6	50	44	28	54.1	894000	1.12×10^{-6}

All stresses in MPa

8. (a) Alternatives discussed in lecture notes in comparison with wind: tidal stream, tidal barrage, wave.

Advantages of wind:

- reliable, established technology with low capital and maintenance costs
- fast construction, easy to decommission and potentially recycle
- adaptable scale from single home multi-kW machines to multi-MW machines for grid
- significant % of UK need possible: planned installation of over 10 GW

Disadvantages of wind:

- intermittency and unpredictability of supply, requiring backup
- current manufacturing capacity for suitable steel and GFRP, CFRP is stretched
- variable public acceptability (noise, visual impact, proximity to wild land, impact on birdlife)

Tidal, in comparison:

- predictable but most systems cyclic with tides
- barrages proven technology, but only in limited locations, and single large installations
- tidal stream technologies unproven, and relatively small capacity (150-300kW)
- underwater technology: hostile environment, transmission and maintenance more difficult
- negative impact on marine environment, fishing, estuary silting etc

Wave, in comparison:

- intermittency and unpredictability of supply, requiring backup
- range of technologies from 75kW to 2MW
- straightforward technology, but largely prototype – efficiency of scale-up unknown
- surface/shoreline technology: easier for maintenance (similar to offshore wind)
- impact on marine environment similar to offshore wind (better for birdlife)

(b) Life cycle analysis (LCA) considers the environmental impact of materials and processes throughout the lifetime of a product, typically divided into:

material production

manufacturing

use (including transportation, installation and maintenance)

disposal (including recycling processing and transport).

Streamlined LCA restricts the analysis to one dominant measure of impact (usually energy, either consumed directly, or embodied in the material). CO₂ emission is an alternative, but is mostly found to correlate strongly with energy. The analysis focuses on identifying the principal contributions simply and quickly. Conventional LCA tends to itemise every contribution in great detail, breaking environmental impact down to many different factors (energy, different greenhouse gases, different pollutants and toxic emissions etc). Aggregating these diverse measures introduces a greater data burden with greater subjectivity and uncertainty in the data, adding little value to the conclusions for the effort expended.

When applied to an energy-producing product such as a wind turbine, the total energy consumption over the product lifetime can be estimated. Given the average actual annual energy output of the machine in service, the time taken to generate the total lifetime energy consumption of the machine can be calculated – this is the “energy payback period” (typically under one year).

For a typical (large) onshore machine, material production energy dominates the input side of the energy payback (manufacturing around 10%, and maybe 10% recovery by recycling). (This could be illustrated simply with a sketch of the relative energy consumptions/returns in material production/manufacturing/disposal stages of the wind turbine life cycle).

The expected changes in this LCA are:

(i) 50% larger power rating primarily means larger blade diameter and tower height. For the example shown in lectures (2MW and 3MW onshore machines), the energies associated with blades, power generating equipment, tower and foundation broadly scale up in proportion. Since annual energy production also scales up by 50%, the payback period is essentially unchanged.

(ii) moving offshore shifts the balance of environmental impact: increased material energy in the transmission system, and higher transport energy for maintenance (e.g. use of helicopters), but potentially lower impact foundation and tower (all steel, instead of extensive use of concrete). The capacity factor (proportion of possible energy produced in a year) is greater offshore. This reduces the energy payback period – the higher energy production offshore more than offsets any additional impact of the machine itself.

Examiner's comments:

Most answers were of a high standard. A handful of students considered nuclear as the alternative energy source, on the basis of the very long timescale for which nuclear fuel will be available. This does not conform with the usual definition of "renewable" as discussed in the lecture notes. In (b), few students mentioned that conventional life cycle analyses often try to aggregate many different impacts on the environment (pollution, gases, waste etc), so failed to make much distinction with streamlined LCA. In discussing the changes in LCA of the wind turbine, many neglected to discuss the changes in energy output and payback period, only mentioning the changes in energy expended in the machine installation.

IB PAPER 8

SECTION D - AEROTHERMAL ENGINEERING

$$9. a) \frac{d\left(\frac{L}{D}\right)}{dc_L} = -624 c_L + 178 = 0 \text{ @ } \left(\frac{L}{D}\right)_{\max}$$

$$\therefore c_{L, \text{optimum}} = \frac{178}{624} = 0.2853, \left(\frac{L}{D}\right)_{\max} = 25.39$$

$$c_L = \frac{L}{\frac{1}{2} \rho A V^2} = \frac{L}{\frac{1}{2} \rho A M^2 \gamma R T} = \frac{L}{\frac{1}{2} A M^2 \gamma \rho} \quad [L = W \text{ @ cruise}]$$

$$\Rightarrow p_{\text{atm}} = \frac{166000 \times 9.81}{\frac{1}{2} \cdot 750 \cdot 0.8^2 \cdot 1.4 \cdot 0.2853} = 16990 \text{ Pa}$$

$$\text{Interpolating, initial cruise altitude} = 41000 + \frac{17.9 - 16.99}{17.9 - 16.2} \times 2000$$

$$= \underline{\underline{42070 \text{ ft}}}$$

NOTE - This is high for current civil aircraft, but typical for "flying wings" which have to fly higher for peak performance.

$$b) s = \frac{-V \frac{L}{D}}{g \cdot sfc} \ln \left(\frac{W_{\text{end}}}{W_{\text{start}}} \right) = H \ln \left(\frac{W_{\text{start}}}{W_{\text{end}}} \right)$$

$$e^{\frac{s}{H}} = \frac{W_c + W_p + W_f}{W_c + W_p} = 1 + \frac{W_f}{W_c + W_p}$$

$$\therefore \left(e^{\frac{s}{H}} - 1 \right) (W_c + W_p) = W_f \Rightarrow \underline{\underline{\frac{W_f}{s W_p} = \frac{1}{s} \left(1 + \frac{W_c}{W_p} \right) \left(e^{\frac{s}{H}} - 1 \right)}}$$

$$9c) H = \frac{V \frac{L}{D}}{g \cdot sfc} = \frac{0.8 \sqrt{8 \cdot R \cdot 216.65} \times 25.39}{9.81 \cdot 0.016 \times 10^{-3}} = 38190 \text{ km}$$

$$\text{From Breguet, } L^{\frac{3}{4}} = \frac{W_{start}}{W_e + W_p}$$

$$\therefore W_p = \frac{W_{start}}{L^{\frac{3}{4}}} - W_e = \frac{166000}{\exp\left(\frac{8000}{38190}\right)} - 110000$$

$$W_p = 24625 \text{ kg}$$

$$\text{No. of passengers} = \frac{24625}{100} = \underline{\underline{246}} \quad \left(\begin{array}{l} \text{maximum since} \\ (\frac{L}{D})_{max} \text{ was used} \end{array} \right)$$

$$W_f = 166000 - 110000 - 24625 = 31375 \text{ kg}$$

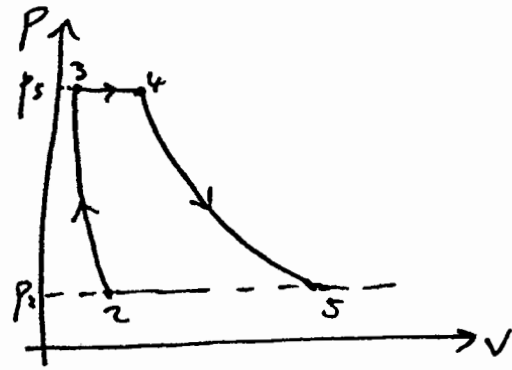
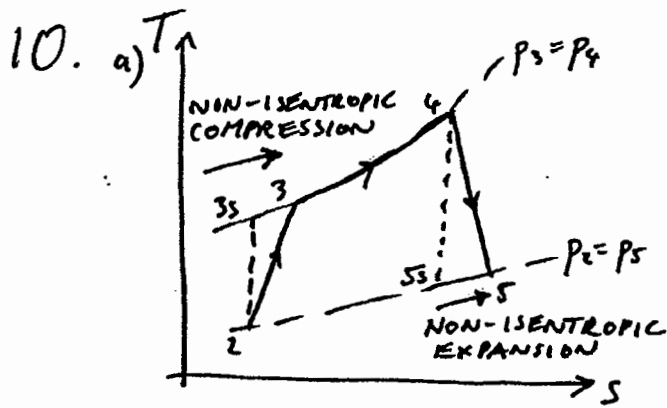
$$\frac{W_f}{SN_p} = \frac{31375}{8000 \cdot 246} = \underline{\underline{0.0159 \text{ kg fuel per passenger-km}}}$$

d) i) Reducing W_e reduces fuel burn per pass-km, because $\left(1 + \frac{W_e}{W_p}\right)$ reduces.

ii) Reducing sfc by 5% increases H by 5%. $L^{\frac{3}{4}}$ thus decreases and the fuel burn per pass-km decreases.

iii) If altitude is reduced (keeping M constant) ρ_{atm} increases leading to lower C_L and lower $\left(\frac{L}{D}\right)$. Thus H is reduced and the fuel burn per pass-km increases.

- Alternatively, if the cruise Mach number is reduced as well as the altitude to try to maintain $\left(\frac{L}{D}\right)_{max}$ then $H = \frac{V \frac{L}{D}}{g \cdot sfc}$ is still reduced. Therefore the fuel burn per passenger-km still increases.



$$b) \eta_{\text{cycle}} = \frac{W_{\text{net}}}{Q} = \frac{m c_p (T_4 - T_5) - m c_p (T_3 - T_2)}{m c_p (T_4 - T_3)}$$

$$\eta_{\text{cycle}} = \frac{(T_4 - T_3) - (T_5 - T_2)}{T_4 - T_3} = 1 - \frac{T_5 - T_2}{T_4 - T_3}$$

BOOKWORK

i) Increasing T_4 increases η_{cycle} because $\frac{(T_5/T_2 - 1)}{(T_4/T_2 - T_3/T_2)}$ decreases.

T_4 limited by: ① Material capability (creep at high temperature)
② Cooling technology (complex passages/coatings)

ii) There is an optimum overall pressure ratio that depends on T_4/T_2 , η_{turb} and η_{comp} (see pg. 36 in the notes)

Trade-off between (T_5/T_2) and (T_3/T_2) changing with pressure ratio.

P_3/P_2 limited by compressor materials and size.

iii) Increasing η_{comp} increases η_{cycle} as T_3/T_2 is reduced.
Limited by aerodynamic performance, $\eta_{\text{comp}} < 90\%$.

iv) Increasing η_{turb} increases η_{cycle} as T_5/T_2 is reduced.
Limited by aerodynamic performance.

See T-s diagram. Typically, $\eta_{\text{turb}} < 92\%$.

c) Static test $\Rightarrow p_u = p_{02}$, $F_G = \text{Net Thrust}$, F_N

w = winter test

s = summer test

10. c) contd.

$$i) \left(\frac{F_a + p_a A_N}{p_{02} A_N} \right)_s = \left(\frac{F_a + p_a A_N}{p_{02} A_N} \right)_w \quad A_N = \text{constant}$$

$$\frac{F_{a,s}}{p_{02,s}} = \frac{F_{a,w}}{p_{02,w}} \Rightarrow F_{a,w} = F_{a,s} \times \frac{p_{02,w}}{p_{02,s}} = 50 \times \frac{102.44}{98.5}$$

$$= \underline{\underline{52.0 \text{ kN}}}$$

$$\left(\frac{w \sqrt{c_p T_{02}}}{A_N p_{02}} \right)_s = \left(\frac{w \sqrt{c_p T_{02}}}{A_N p_{02}} \right)_w$$

$$\therefore \text{Ratio of mass flows, } \frac{w_s}{w_w} = \frac{p_{02,s}}{p_{02,w}} \frac{\sqrt{T_{02,w}}}{\sqrt{T_{02,s}}}$$

$$= \frac{98.5}{102.44} \cdot \sqrt{\frac{267}{300}} = \underline{\underline{0.907}}$$

ii) Non-dimensional group with T_{04} is: $\frac{T_{04}}{T_{02}}$

Non-dimensional group with N is: $\frac{ND}{\sqrt{\gamma R T_{02}}}$ ($D = \text{diameter}$)

$$\frac{T_{04,s}}{T_{04,w}} = \frac{T_{02,s}}{T_{02,w}} = \underline{\underline{1.12}} \quad \frac{N_s}{N_w} = \sqrt{\frac{T_{02,s}}{T_{02,w}}} = \underline{\underline{1.06}}$$

Engine wear increases with turbine entry temperature, T_{04} , and engine component stress (stress $\propto N^2$).

\therefore Greater wear in the summer test.

11. a) High bypass ratio engines have a high total mass flow of air and a low jet velocity, V_j . This leads to a high propulsive efficiency since $\eta_p = \frac{2V}{V_j + V}$

Low V_j also leads to lower noise.

The maximum ISPR is limited by:

- i) Installation drag
- ii) Mechanical stress (in fan root)
- iii) Engine weight increase
- iv) Airframe factors (the engine must fit under the wing)
- v) Transportation problems.

$$b) X_N = \dot{m}(V_j - V) \Rightarrow \dot{m}_{TOTAL} = \frac{X_N}{V_j - V} = \frac{25000}{360 - 250} = 227.3 \text{ kg/s}$$

$$\dot{m}_{core} = \frac{\dot{m}_{TOTAL}}{BPR + 1} = \frac{227.3}{13} = \underline{\underline{17.5 \text{ kg/s}}}$$

$$\eta_p = \frac{2V}{V + V_j} = \frac{2 \cdot 250}{250 + 360} = \underline{\underline{82.0\%}}$$

c) Mass flow is 90% of choking value.

$$\therefore \frac{\dot{m}_{TOTAL}}{A_f} = 0.9 \times 1.281 \times \frac{P_{02}}{\sqrt{\gamma T_{02}}}, \quad A_f = \frac{227.3 \sqrt{1005.250}}{0.9 \times 1.281 \times 38 \times 10^3} = 2.60 \text{ m}^2$$

$$A_f = \pi r_t^2 (1 - HTR^2) \Rightarrow r_t = \sqrt{\frac{A_f}{\pi (1 - HTR^2)}} = \sqrt{\frac{2.6}{\pi (1 - 0.3^2)}}$$

$$r_t = 0.954 \text{ m} \quad \therefore D_{fan} = 2 \times 0.954 = \underline{\underline{1.907 \text{ m}}}$$

11. d) LP Turbine. Work = Work input to fan

$$\dot{w}_{\text{core}} W_x = \dot{w}_{\text{TOTAL}} c_p \Delta T_0^{\text{FAN}}$$

$$\therefore W_x = \frac{\dot{w}_{\text{TOTAL}}}{\dot{w}_{\text{core}}} \frac{(V_j^2 - V^2)}{2\eta_{\text{fan}}} = (12.1) \frac{(360^2 - 250^2)}{2 \cdot 0.9}$$

$$\underline{\underline{W_x = 484.6 \text{ kJ/kg}}}$$

$$\Delta h_0^{\text{LPT}} = 484.6 \text{ kJ/kg}, \quad U^{\text{LPT}} = r \Omega = 0.55 \times 3000 \times \frac{2\pi}{60} = 172.8 \text{ m/s}$$

$$N_{\text{stage}} > \frac{\Delta h_0^{\text{LPT}} / U^2}{3} = \frac{484600}{172.8^2 \cdot 3} = 5.41$$

\therefore Number of stages required = 6

$$\text{e) } \dot{w}_{\text{TOTAL, new}} = \dot{w}_{\text{TOTAL, old}} \times \left(\frac{r_{t, \text{new}}}{r_{t, \text{old}}} \right)^2 = 227.3 \times (1.1)^2 = 275.0 \text{ kg/s}$$

$$\text{BPR}_{\text{new}} = \frac{\dot{w}_{\text{TOTAL, new}}}{\dot{w}_{\text{core}}} - 1 = \frac{275}{17.5} - 1 = \underline{\underline{14.7}}$$

$$V_{j, \text{new}} = \frac{X_N}{\dot{w}_{\text{TOTAL, new}}} + V = \frac{25000}{275} + 250 = 340.9 \text{ m/s}$$

$$\therefore \eta_{p, \text{new}} = \frac{2V}{V + V_{j, \text{new}}} = \frac{2 \cdot 250}{250 + 340.9} = \underline{\underline{84.6\%}}$$

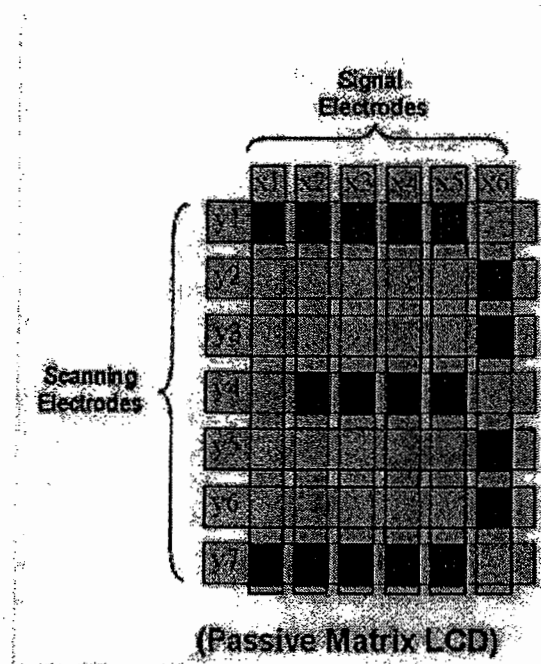
i.e. An increase of 2.6% for a significantly larger (and heavier) fan system.

Solutions
Section E Electrical Engineering

Q.12(a)

PASSIVE Matrix Addressed LCD

The simplest type of liquid crystal display is the passive matrix addressed LCD, where the front and back cell wall are covered with stripes of conducting material (typically Indium Tin Oxide -ITO). In order to bring a selected pixel into the on-state a voltage is applied to one of the row electrodes and a voltage of opposite polarity onto the column. If the voltages are appropriately chosen only the pixel at the crossing of the electrodes will be selected- see below. However the current standard for LCDs is the active matrix (AMLCD) based on thin film transistor (TFT) technology.



3:1 Dynamic Addressing Scheme

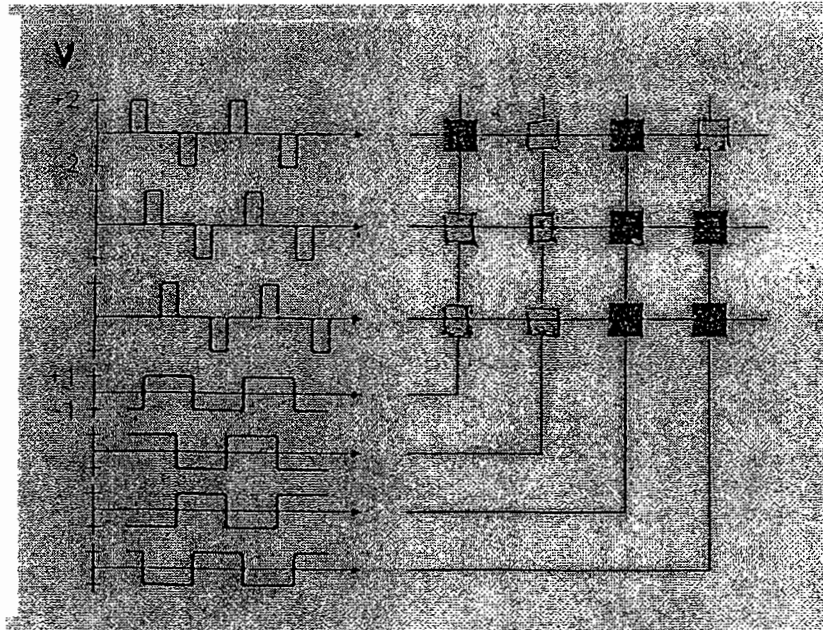
There is a threshold voltage below which the liquid crystal will stay in its “off-state”.

In order to switch it “on” a rms voltage above this level will have to be applied. Fig. below shows a commonly used address scheme for a 3 row-4 column display.

This is the 3:1 scheme where it can be seen that during the first 3 time intervals the 3 rows are addressed one after the other by a voltage of + 2V. When a row is selected the desired “on-state” pixels get a voltage of -V applied to the column electrode. Those which are to remain “off” get +V applied to the column, i.e. for the pixels selected to be

on, 3V is applied across the cell and for those to remain off, V volts (i.e. below the threshold voltage) are applied.

3.1 addressing scheme. The top three graphs on the left side show the time dependence of the row electrode voltages which are the same for any selected pattern. The bottom four graphs contain the column voltages which are necessary to generate the select pattern displayed on the right.



Since a pixel is selected only during one time interval in three, the rms voltage on the selected pixel is

$$(V^2 + V^2 + (3V)^2)^{1/2} = (11/3)^{1/2} = 1.91 V$$

For all pixels not selected the V remains as V.

After the first 3 time intervals the sequence is repeated (but inverted to ensure a zero DC component).

Thus by using a 3:1 scheme a rms voltage which is 91% greater than the “off” voltage is applied to each which we want to be in the “on” state.

It is easy to see if we increase the number of rows significantly (i.e. in any useful display) then the relative voltage difference between selected and unselected pixels will get smaller and smaller – dynamic addressing this way thus becomes impracticable as the contrast differences will become lower and lower.

This led to the use of active matrix addressing.

Q.12 (b)

AMLCD

Here each display pixel is equipped with its own TFT switch. Thus the glass substrate contains a matrix of TFTs whose gates are connected to the row lines and drain electrodes are connected to the column lines. The pixels are not constructed as the intersection of row and column lines (as in the case of passive matrix addressing), but rather by individual electrode regions connected to the sources of the TFTs. The counter electrode extends over the whole pixel area of the opposite glass substrate. See below. The addressing scheme is very straight forward – see Fig. The TFTs simply act as switches. The rows are opened sequentially and when open, the voltage on the column is transferred to the liquid crystal cell. The charge needed to maintain the electric field can be stored in a capacitor located on every pixel.

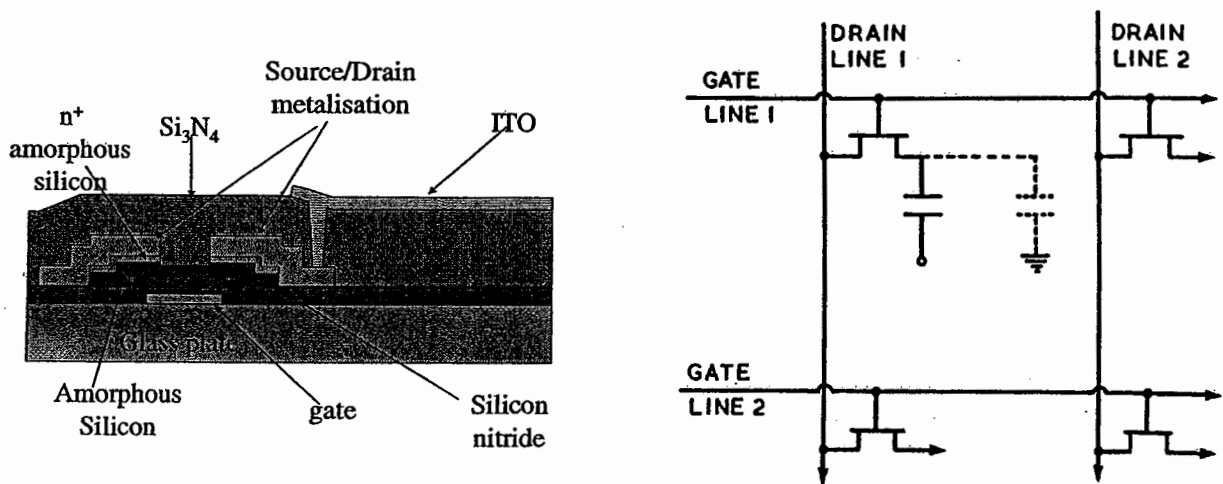


Fig. Cross section of the (a-Si:H) TFT in the active matrix and schematic of TFT at each pixel

Q. 12 (c)

Backlights:

Most LCDs require a backlight, and design of the light source is a very important part of the LCD module. In these displays, about 90% of the electrical power is consumed by the backlight – this severely impacts battery operation time.

Source	Filament Lamp (Halogen)	Electro luminescent Foil	Light Emitting Diode	Cold Cathode Fluorescent Lamp
Power Consumption	1.5 W	0.5-1.5 W	1-3 W	0.5 W
Intrinsic Colour	White	orange-blue (white possible)	red-green (blue and white expensive)	Different phosphors (white possible)
Life Span	2000 h	3000-5000 h	100000 h	10000-20000 h
Remarks	many shapes, simple power supply	very flat, high AC supply voltage (inverter)	Many shapes, mounting on circuit board possible	High AC supply voltage (inverter)

Comparison of backlighting sources. The power consumption is stated for 100 cd/m² from a surface of 100 cm². Note that the cold cathode fluorescent lamp has the lowest power consumption, which is why it is used in laptops.

Because of uniformity issues, LEDs may soon be the backlight choice although many groups are now working on backlights based on field emission from carbon nanotubes.

The only sensible alternative to AMLCDs for TVs >40" diagonal at present is the Plasma Display (some may say OLED – not a sensible answer for these large areas at present or FED – not yet on the market and may never be).

Q.13

(a) $V = 2V$, $\sigma = 32 \text{ /ohm.cm}$, $v_s = 10^5 \text{ m/s}$,
 $\mu = 0.1 \text{ m}^2/\text{V.s}$

$$\sigma = Ne\mu, \text{ therefore}$$

$$N = 32 / (1.6 \cdot 10^{-19} \cdot 0.1) = 2 \cdot 10^{21} \text{ m}^{-3}.$$

(b) $t = L/V$, $v = \mu E$, $E = V/L$,

$$\text{so } t = L^2 / \mu V \text{ and } L = (t\mu V)^{1/2} \text{ or } L = 2 \cdot 10^{-6} \text{ m.}$$

$$E = V/L = 2 / 2 \cdot 10^{-6} = 10^6 \text{ V/m.}$$

(c) $W/L = 20$, so $W = 4 \cdot 10^{-5} \text{ m}$

$$I = (V\sigma A) / L = (2 \cdot 32 \cdot 10^{-6} \cdot 4 \cdot 10^{-5}) / 2 \cdot 10^{-6} = 1.28 \text{ mA}$$

(d) $V = Ned^2 / \epsilon$ $d = \text{channel thickness, } 10^{-6} \text{ m.}$

$$V = 2 \cdot 10^{21} \cdot 1.6 \cdot 10^{-19} \cdot 10^{-12} / (12 \cdot 8.85 \cdot 10^{-12}) = 1.51 \text{ V}$$

Q.14

(a) $N = 4.1 / (3.61 \cdot 10^{-10})^3 = 8.5 \cdot 10^{28} / \text{m}^3$

$$\sigma = Ne\mu$$

$$\mu = 6 \cdot 10^7 / (8.5 \cdot 10^{28} \times 1.6 \cdot 10^{-19}) = 0.0044 \text{ m}^2 / \text{V.s}$$

$$\sigma (\text{semiconductor}) = Ne\mu = 4 \cdot 10^{21} \cdot 1.6 \cdot 10^{-19} \cdot 0.0044 = 2.82 \text{ ohm}^{-1} \text{ m}^{-1}.$$

(b) bookwork.

$$\text{energy} = \frac{1}{2} mv^2 / e = 7.1 \times 10^{-3} \text{ eV}$$

(c) $v = \mu E$

$$\text{thus, } E = 5 \cdot 10^4 / 0.0044 = 1.14 \cdot 10^7 \text{ V/m.}$$

$$V = 2 \text{ V, thus } L = 2 / 1.14 \cdot 10^7 = 1.76 \cdot 10^{-7} \text{ m.}$$

$$t = L/v = 1.76 \cdot 10^{-7} / 5 \cdot 10^4 \sim 3.5 \cdot 10^{-12} \text{ s}$$

Paper 8 Information Engineering
SECTION F

Q15(a) - low pass filter to reduce effect of noise in differential for edge detection

- low pass filter to blur image (remove high spatial frequencies)

- low pass filter to band-limit before subsampling to avoid aliasing.

(4 marks)

$$(b) \quad S(x, y) = \sum_{i=-n}^{-n} \sum_{j=-n}^{-n} I(x-i, y-j) g_{\sigma}(i, j)$$

where $g_{\sigma}(i, j) = \frac{1}{2\pi\sigma^2} e^{-\frac{(i^2+j^2)}{2\sigma^2}}$

(4 marks)

$$(c) \quad S(x, y) = I(x, y) * g_{\sigma}(x, y)$$

$$= [I(x, y) * g_{\sigma}(x)] * g_{\sigma}(y)$$

where $g(x) = \frac{1}{\sigma\sqrt{2\pi}}$

$$g(x) = \frac{e^{-\frac{x^2}{2\sigma^2}}}{\sigma\sqrt{2\pi}}$$

due to separability of $g(x, y)$.

Computational saving, if kernel is $(N=2n+1 \text{ pixels}) N \times N$ per smoothed pixel.

(4 marks)

$$= \frac{N^2}{2N} = \frac{(2n+1)^2}{(2n+1) \times 2} = \frac{2n+1}{2} = n + \frac{1}{2}$$

per pixel.

(d) Filter size for $\sigma=1$.

Discard sample if $e^{-\frac{(n+1)^2}{2\sigma^2}} < \frac{1}{1000}$ (a. tails are smaller
low peak)

$$\therefore n > 3.7\sigma - 1$$

$\therefore \sigma=1$, kernel is 7 pixels big with samples of gaussian $g(x)$:

$g(-3)$	$g(-2)$	$g(-1)$	$g(0)$	$g(1)$	$g(2)$	$g(3)$
0.004	0.054	0.242	0.399	0.242	0.054	0.004

Check that $\sum_{i=-3}^3 g(i) = \underline{1}$ ✓ (4 marks)

(e) High-pass filter $H(x,y) = \alpha I(x,y) - \beta g_{\sigma}(x,y) * I(x,y)$
(α * original image minus β * low-pass filtered version).

Used in edge enhancement or edge detection. (See example Q1)

(4 marks)

Paper 8 - Information Engineering

Q16

(a) Band-pass images with Laplacian of Gaussian, $L(x, y) = \nabla^2 G(x, y)$ to find blobs in images at finite range of scales (2)

(i) Produce scale-space representation at $\sigma_i = 2^{\frac{i}{s}} \sigma_0$, $s = \# \text{ images}$

$$G(x, y, \sigma_i) = G(x, y, \sigma_0) * I(x, y)$$
 (2)

(ii) For each smoothed, low-pass filtered image compute $\nabla^2 S(x, y)$ (2)

(iii) Search for local maxima in image and across scales. Mark as interest point. Record scale of image. (2)

(iv) Efficient implementation with difference of Gaussians (2)

(b) (i) Intensity patch: Take $N = 16 \times 16$ pixels and normalize by σ and form a vector of pixel values. (size $16 \times 16 = 256$)

$$Z(x, y) = \frac{I(x, y) - \mu}{\sigma}$$

so has zero mean and unit norm; $\therefore |Z| = 1$

Patch compared by comparing vector of pixels Z by $\text{scalar} = \text{vector}$ (i.e. normalized cross correlation) (3)

(ii) Convolve pixel patch with filter bank 8 Laplacian + Gaussian
 4 Gaussians
 36 oriented filters at 1

Response of 48 filters characterizes patch of pixels. (3)

(b) (iii) Orientation histograms

Bin all edge gradients in 16×16 pixels to give a histogram of edge orientations. Edges are robust to brightness and contrast changes & encode shape information for matching.

(2 marks)

(iv) SIFT

8 orientations.

16 bins/histograms of orientation concatenated to give a vector of size $N = 128$ which is normalised to unit norm to give invariance to lighting.

(2 marks)

SOLUTIONS (Q3)

- 17 (a) Different metrics can be used, for example Euclidean distance

$$d_n^2 = \sum_m (x_{nm} - x_{N+1,m})^2$$

or Absolute or Manhattan distance

$$d_n = \sum_m |x_{nm} - x_{N+1,m}|$$

A simple algorithm is:

initialize $d_{min} = \infty$, $n_{min} = 0$

for $n = 1$ to N :

if $d_n < d_{min}$ then $d_{min} \leftarrow d_n$ and $n_{min} \leftarrow n$

endfor

return (d_{min}, n_{min})

A full answer might discuss that for large N but moderate or small M it would be a good idea to use some sensible data structure, such as a KD-tree, so that not all N points need to be tested.

- (b) The probability of an image under this model is:

$$r_n = \prod_m p_m^{x_{nm}} (1 - p_m)^{(1-x_{nm})}$$

For large M these probabilities can get *very* small so it's better to use log probabilities. We can use exactly the above algorithm except using r_n instead of d_n to find the least probable image. The least probable images in a collection of images are likely to be outliers, as long as the model is a good model of the collection of images. So the least probable images could be removed.

- (c) For a feature m to discriminate between mountain and sunset images, the value of p_m should be different from q_m . A simple measure of difference is $|p_m - q_m|$, but this measure is not very good (since for probabilities, 0 is much more different to 0.1 than 0.4 is to 0.5). [Not needed for credit: A better measure might be $\log \left| \frac{p_m (1-q_m)}{q_m (1-p_m)} \right|$]. Using these differences, one can run the same kind of algorithm as in (a) to find the top ten largest differences. The algorithm stores the list of currently top ten items at each step, and tests to see if this list needs to be updated at each step. Sunset images tend to have reddish colors and some low-frequency texture elements, whereas mountain images tend to have blue and grey colors and some higher frequency texture elements. These are likely to be discriminative.

Paper 8 Section G Engineering for the Life Sciences 2006/7 Crib

Question 18

a)

i) Reward is a scalar value which represents the degree to which a state or action is desirable. It represents the immediate desirability of a particular state ignoring future or past states. A value function is a mapping from states to a scalar where the value of a state represents the long-term total reward achieved starting from that state, and executing a particular policy. Therefore the value function is the expected sum of (possibly discounted) rewards that are associated with being in that state and following a particular policy

ii) A policy is the decision-making function (control strategy) of the agent, which represents a mapping from situations to actions. It defines an action (or more generally the probability of choosing each possible action) given a particular state

iii) A discount factor is a scalar value between 0 and 1 which determines the present value of future rewards. If the discount factor is 0, the agent is concerned with maximizing immediate rewards. As the discount factor approaches 1, the agent takes more future rewards into account. In general if $\gamma \in [0,1]$ is the discount factor and r_t the

reward at step t then the overall value to be optimized is $= r_0 + \sum_{t=1}^{\infty} \gamma^t r_t$

(b)

(i) A. Subject A has the lower loss. Subject A total loss = $2^2 + 2^2 = 8$. Subject B total loss = $1^2 + 3^2 = 10$

B. Subject A and B have the same loss. Subject A total loss = $2 + 2 = 4$. Subject B total loss = $1 + 3 = 4$

C. Subject B has the lower loss. Subject A total loss = $2^{0.5} + 2^{0.5} = 2.83$. Subject B total loss = $1^{0.5} + 3^{0.5} = 2.73$

(ii) The two main benefits to moving slower are that there is that as motor noise is signal-dependent (standard deviation scales with mean motor command) less noise is injected into the system. In addition longer movements allows more time for feedback and error correction.

(ii) Redundancy allows multiple ways to achieve a task. Because of signal-dependent noise each way will generate a different distribution of final position, velocity, force, joint angles etc. Therefore by choosing different solutions it is possible to control the statistics of action. Redundancy therefore allows the movement to be chosen that best matches the statistical requirements of a task.

(c)

(i) A *forward* dynamic model is an internal model that converts motor outputs to estimates of future states (or sensory inputs). These are called forward as they model the

causal (forward) relationship between actions and their consequences. A forward model anticipates how the motor system's configuration will change as a result of a motor command

(ii) Internal models that convert desired sensory inputs to motor outputs are known as *inverse* models. Such a model could transform a desired sensory consequence into the motor commands that would achieve it. Because an inverse model can determine the motor command required to achieve some desired result, such a model is useful in controlling movement.

(iii) As forward models are predictive the training signal, the actual outcome is available for learning. In inverse models the desired output, the optimal motor command, is not known. In addition forward dynamic models of the human body are many-to-one mapping they are easy to represent unlike inverse models which can be one to many (a desired state can be reached by more than one motor command)

(iv) The concept of motor prediction was first considered by Helmholtz when trying to understand how we localize visual objects. To calculate the location of an object relative to the head, the central nervous system must take account of both the retinal location of the object and also the gaze position of the eye within the orbit. Helmholtz's ingenious suggestion was that the brain, rather than sensing the gaze position of the eye, predicted the gaze position based on a copy of the motor command acting on the eye muscles, termed efference copy, running through a forward mode. He used a simple experiment on himself to demonstrate this. When the eye is moved without using the eye muscles (e.g. by covering one eye and gently press with your finger on your open eye through the eyelid), the retinal locations of visual objects change, but the predicted eye position is not updated, leading to the false perception that the world is moving. [Alternative evidence would be predictive grip force, sensory cancellation of self-produced tactile stimuli].

19(a) Describe briefly the "modified random walk" strategy by which bacteria such as *Salmonella typhimurium* and *Escherichia coli* seek nutrient molecules in their aqueous environment. [4]

Bacteria such as *S. typhimurium* and *E. coli* are observed (via the microscope) to swim with a 3D "random walk". They swim in essentially a straight line for a few seconds; then stop for a fraction of a second; before setting off again in a straight line in a direction which makes a random angle with the previous one. Such a "random walk" would result in no overall motion if the "legs" or "runs" of straight-line motion were all of equal length.

But the bacterium is seeking "food", or nutrient molecules, to enable it to grow and divide. It achieves this aim by extending the lengths of "runs" if it finds that it is moving into a region of higher concentration of the sought nutrient. The bacterium has detectors of nutrient molecules ("receptors") which measure the rate of encounter with these molecules over time; and if the rate of detection increases with time, the length of that "run" or "leg" is increased. Thus the bacterium moves, overall, towards a supply of nutrient.

In the smooth-swimming "runs" the motors turning the 6-or-so corkscrew-like flagella all rotate in the same direction, and the flagella form a single propulsive bundle. At the end of the "run", some or all of the motors go into reverse; in consequence the flagella change their wave-form from a LH to a RH helix; the bundle "flies apart" and the bacterium "tumbles" on-the-spot. When the motors resume their "forward" sense of rotation, the bacterium moves off in a random direction.

The internal machinery for converting data on the rate of encounter with nutrient molecules into instructions to the flagella motors is at present imperfectly understood.

(b) Discuss the various circumstances in which a bacterial flagellar filament can switch between members of a family of twelve discrete helical waveforms. What is the structural basis of these different forms? [4]

Experiments on isolated flagella filaments demonstrate the ability of these slender corkscrew-like structures to adopt different helical forms in the following circumstances.

1. Change in the PH of the surrounding water.
2. Change in the temperature of the surrounding water..
3. Co-polymerisation of flagella molecules ("building-blocks") from different mutant strains of bacteria.
4. Sequential polymerisation of flagella from different strains.
5. Mechanical torque, arranged by streaming water past filaments tethered to the microscope slide (which is analogous to reversal of the direction of rotation of the bacterium's motors).

The flagellum building blocks are something like stone "voussoirs" for building an arch; but they are "twisted", so as to build a tubular "tower". There are strict geometrical constraints on the geometry of the blocks if they are to assemble in a particular building pattern.

The flagella molecules have a distinctive feature whereby they can adopt two slightly different 3D conformations: they can switch between two "bi-stable" forms, say A and B. It turns out that if all blocks are in state A the geometry is not right for building a complete tube. Neither is it

right if all blocks are in state B. But if there is a mix of blocks in the two states, building can proceed, with the help of a little elastic distortion of the blocks. Then, if the blocks in the two states have slightly different lengths -as they do- they can build a curved tube (A on one side, B on the other); and hence a vehicle or corkscrew-like filament.

The geometry of the protein-molecule blocks is slightly different in different mutant strains; and at different PH and Temperature; and under mechanical stress. This explains the 5 items above.

(c) Describe briefly how the protein devices known as bacterial receptors work; and how they may be "re-engineered" to recognise different molecules, such as TNT (trinitrotoluene).

Receptors are protein devices which sit in the cell-wall of bacteria - and indeed of other cells. As described in (a) above, they can recognise individual molecules of a specific chemical; and when they do so a signal is sent into the interior of the cell, which contributes eventually to instructions for controlling the flagella motors.

A receptor has a sort of "open mouth". Many kinds of molecule drift in and out of the mouth; but only one kind is "recognised", by having exactly the right shape, and also the right disposition of hydrogen-bond donors and acceptors to match the acceptors and donors which are present on the amino-acids which line the "palate" of the mouth.

When a molecule of the right species is recognised in this way, the hydrogen bonds all form, and the mouth "snaps shut".

This change in geometry of the mouth in turn activates further conformational changes by which a signal is transmitted along the proteins embedded in the cell membrane, to the interior of the cell.

The specificity of the recognition-process depends on the layout of about a dozen amino acids in the mouth-region of the protein. The techniques of genetic engineering can be used to change some of these amino acids so that the modified receptor will recognise a different kind of molecule from its natural one. But to decide what changes are needed in order to make the receptor recognise, say, a TNT molecule, requires formidably complicated "molecular dynamics" computations.

Bacteria or plants which have been engineered to recognise TNT can be used to detect buried land-mines, etc..

(d) Biological organisms have a common scheme of growth by cell-division, in which each "daughter" cell contains the full DNA instructions for constructing the entire organism. Explain briefly how cells can "specialise" in order to build the different parts of multi-cellular organisms.

[4]

Each multi-cellular organism - whether a fly, worm, plant or animal begins as a single cell. [N.B. Bacteria are organisms that only consist of a single cell.] This cell then divides into two daughter cells. That is a complicated process, because division does not occur until a duplicate copy of the cell's DNA has been made - equivalent to xeroxing the entire contents of a large college library. The process of cell-division proceeds, until all of the very many cells of the fully-grown organism have been set up.

Each cell contains the full DNA “blueprint” for the entire organism. There are many different kinds of cell required for proper operation of any multi-cellular organism. E.g. an animal requires special cells for muscle, bones, eyes, brain, liver, blood etc., etc.. Each of these specialised cells transcribes only a very small part of the DNA which it contains, in order to manufacture the various proteins which are required to enable that cell to work properly, in association with its neighbours.

The process of specialisation begins early in the life of any such organism. E.g. studies of developing embryos show how the overall layout of the organism’s body is set up.

(e) Describe the principle of Darwin’s “tree of life”. What is the range of organisms included in this tree? How may the tree be mapped in detail? [4]

In Darwin’s time (mid C19) there was wide acknowledgement that species evolved: there were many obvious similarities between, e.g., dogs and wolves, and between different kinds of fish. Furthermore, farmers and pigeon-fanciers were able, by controlling watering, to produce different varieties of animals, plants (e.g. flowers) and pigeons, etc..

Darwin tried to think of natural processes which could make new varieties occur over time. After reading Malthus’s book on the way in which potentially exponential growth of a population is limited by food supply, he thought that naturally occurring random mutations might provide certain specimens with a particular advantage – e.g. a more hairy individual animal would be better able to survive an ice-age; which might then kill off the less hairy animals, leaving the hairy specimen to flourish and produce a new variety of that animal.

Many small changes, over the very long timescale which geologists were finding evidence for, could explain, said Darwin, all known life-forms. He drew in his note-book a “tree of life” in which the branching-points represent the emergence of new varieties/species. This had the implication that all life started with some sort of “primordial ancestor” at the base of tree; and in particular that mammals occupied a small region on the periphery.

Darwin did not understand the discrete nature of inheritance – he was unaware of Mendel’s mathematical conclusions from his experiments on the inheritance of certain characteristics of peas - but his work has nevertheless stood the test of time. For example, the fact that all living things - bacteria, worms, flies, fish birds, plants etc. – use exactly the same DNA and its code for making proteins provides strong support for the idea of a single “tree of all life”.

The tree of life may be mapped in several different ways.

One method is by careful examination of anatomical features - often in the skeletons - of different organisms. A modern method has been made possible by the recent development of DNA-sequencing technology. For example, the protein “cytochrome C” is found in all organisms which use oxygen. This is a similar molecule in all species, yet its small difference in DNA sequence enable a unique “polygenetic tree” to be mapped.

The “tree of life” is found to be essentially the same by whatever method was used to set it up.

Question 20

(a)

(i) σ is stress, ϵ is strain, F is force, and the subscripts 1 and 2 are for each component and with no subscript is for the whole system.

Recognizing that the components are in parallel and the strains are equal:

$$\epsilon = \epsilon_1 = \epsilon_2 = \frac{\sigma_1}{E_1} = \frac{\sigma_2}{E_2}$$

$$\frac{F_1}{A_1 E_1} = \frac{F_2}{A_2 E_2} = \frac{F - F_1}{A_2 E_2}$$

$$F_1 \left(\frac{1}{A_1 E_1} + \frac{1}{A_2 E_2} \right) = \frac{F}{A_2 E_2}$$

$$\frac{F_1}{F} = \frac{A_1 E_1}{A_1 E_1 + A_2 E_2}$$

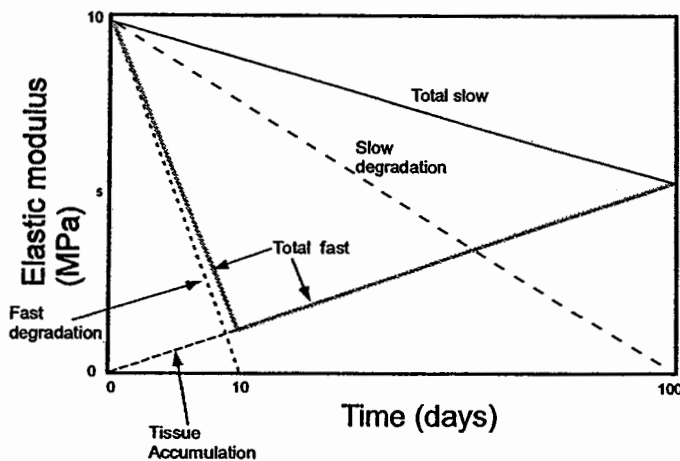
1 = titanium, 2 = bone

$$\frac{F_1}{F} = \frac{7 A_2 E_2}{7 A_2 E_2 + A_2 E_2} = \frac{7}{8}$$

(ii) The bone is shielded from positive bone-making force levels such that the bone around the implant degrades, potentially contributing to failure of the implant.

(b)

(i) The slow-rate polymer degrades linearly over 100 days (dashed line) while the fast-rate polymer degrades over 10 days (dotted line). The tissue fills in linearly over 100 days (dashed rising line). The system response curves are the sum of the polymer and the tissue responses. Straight line is for polymer with slow degradation; curve with minimum is for fast degradation material.



(ii) Choose the slower resorption rate material as it will allow for a sufficiently stiff composite material in the intermediate phase during tissue ingrowth. The faster degradation rate material with its composite stiffness minimum would expose the new tissue to potentially damaging mechanical loading before it was fully formed.

(c)

(i) The adhesive pad area of an animal is expected to scale with $m^{2/3}$ (surface-to-volume ratio). To support the animal's body weight, the adhesive force per pad area needs to scale with

$$\frac{F}{A} \propto \frac{m}{m^{2/3}} = m^{1/3}$$

Since the adhesive setae occupy a constant proportion of the pad area,

$$N_A \propto r^{-2}$$

The adhesive strength (force per area) is

$$\frac{F}{A} = N_A \cdot f \propto r^{-2} r^{3/2} = r^{-1/2} \propto (N_A)^{1/4}$$

To sustain the animal's body weight, the density of adhesive hairs has to be

$$(N_A)^{1/4} \propto m^{1/3}$$

$$N_A \propto m^{4/3}$$

i.e. hair density is expected to increase with body mass to the power of 4/3.

(ii) Geckos could achieve a greater adhesive strength than other animals, because finer hairs are expected to conform better to small scale surface roughness. Moreover, reducing the size of contacts below a critical length given by the Griffith criterion can prevent crack propagation and thus enhance adhesion.

The greater seta density in geckos could also simply be a consequence of the fact that adhesion in geckos and spiders is 'dry', there are no adhesive fluids involved. Fluid-based adhesive systems can cope with small length scales of surface roughness even with relatively blunt seta tips, whereas dry systems need very fine tips to come close enough to the substrate.

Engineering Tripos Part IB 2007

Paper 8 Section H: Manufacturing, Management and Design

Answers

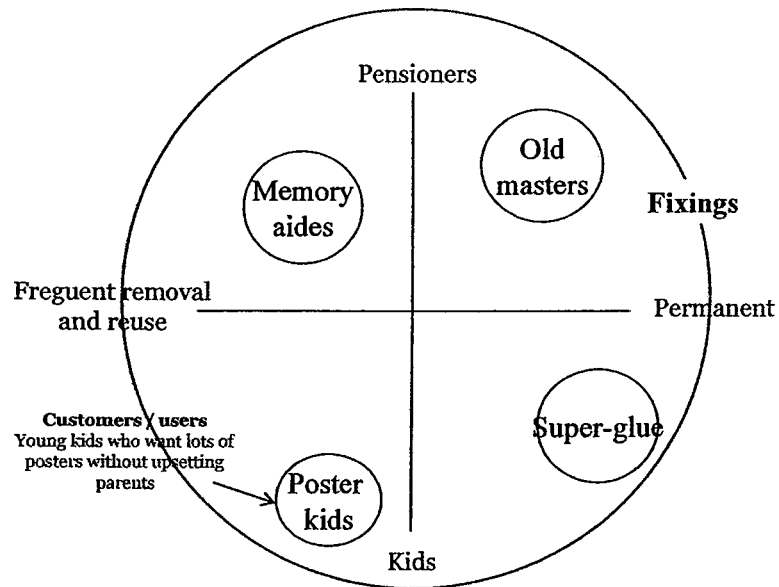
Question 21

- a) Students should discuss the challenge of adequately defining the market for this technology. The market could be defined broadly – such as ‘fixings’, encompassing competitive solutions such as nails, screws, pins, tacks, blu-tack, glue, tape etc. Or, the market could be defined very narrowly to focus specifically on ‘poster fixings’.

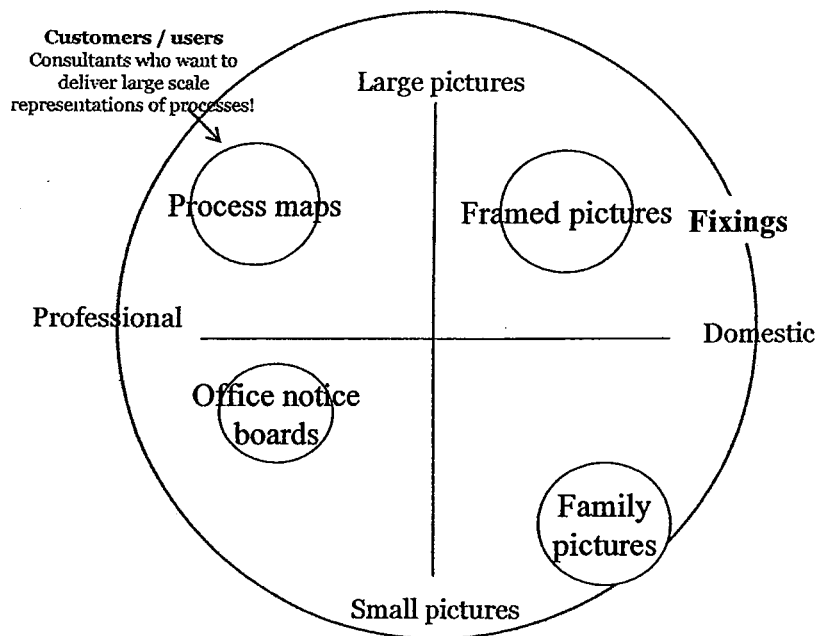
Having defined the market, there are a number of ways in which it can be segmented. Students should be able to describe that a market can be segmented by:

- **Benefits delivered:** in this case that could be long lasting stickiness, no marks, easy to clean, no damage to posters etc.
- **By user or consumer:** including things such as consumer age, lifestyle, social group or possibly attitudes. In this case, examples could be students, pensioners, middle class families etc.
- **By product attributes:** specific properties of the product – in this case, that could be stickiness, residue, weight, density, cost, price etc
- **By usage:** frequency, duration etc. In this case, it could be ‘mass poster stickers, occasional users, office workers, nurses etc

To segment the market, students must pick two ‘sensible’ dimensions to divide the market space and identify *homogenous* and *heterogeneous* segments – people with the segment have the same needs, and these are different to those of people in other segments. This is typically represented on a perceptual map.



This chart would point towards potentially targeting young kids as the audience for this product.



This chart would point towards maybe professional consultants being a potential market for this product.

- b) Students should note the difference between research methods that are aimed at Validating and Qualifying assumptions and research that is aiming to generate new Insights and Inspiration.

The plan should include data on which stakeholders you would seek to gain information from ...

- purchase stakeholders: initiator, influencer, decider, buyer or user. It might be most important for example to understand the needs of the store at point of sale. It might be necessary to understand the needs of the parents or the users themselves.
- Lead users: people at the cutting edge of 'poster sticking' – perhaps fly-poster stickers!

[approx 4 marks]

Students should consider how many users they would ask – more for validation, less for inspiration.

To gather inspiration, would use 'ethnographic methods' – such as user observation – to get under the skin of potential users.

To gather validation – would use interviews and surveys to test viewpoints and opinions.

Students could expand here with more specific information.

- c) A persona is an 'archetype' of a user and aims to capture their motivations and provide the design team with a 'live' character in the design process. Good personas should include insight into the goals of the user.

Strengths: help make design decisions more objective, ensure that the team is focused on the user. Can help in idea generation and also provides a focal point for concept evaluation "would Brenda like it"?

Weaknesses: No-one is truly 'average' and there is no such thing in practice as an archetype. If they are not based on sensible research, then they can be misleading.

Question 22

(a)

Debt = borrowing money from a bank or specialist finance organisation. You can only borrow money if you can convince the bank that you can repay the money, plus interest, exactly when they want it. This is usually very hard for a new company to do.

Banks hardly ever lend money to early stage ventures as they regard them as too risky. Banks need to be able to apply measures to a company based upon their past performance so that they can assess the likelihood of them being an acceptable risk for lending. Start-ups have no track record, so banks cannot assess them.

Equity = selling part of the ownership (shares) of your business in return for cash. The assumption is that whoever buys part of your business will wish to sell this part of the business to someone else in the future at a higher price. The only reason for doing this is if the investor can be convinced that:

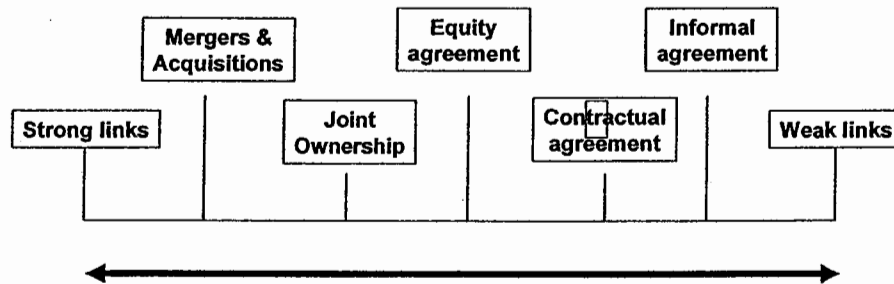
- (a) the business is really going to grow (and quite fast) and that

(b) there will be someone else willing and able to buy their share in the company at a later date

The sources of equity funding are:

- Business Angels are wealthy individuals who choose to invest some of their own money in new business ventures. The term 'smart' is used to refer to those business angels who are able to bring not only money but also expertise of particular technologies, markets and industries, based upon their own experience.
- Venture Capital Funds (known as 'VCs') are a form of high risk, high return investment. VCs raise money from large institutions such as pension funds and then re-invest portions of this money in high growth potential businesses. They know that a number of their investments will fail, so they seek to ensure that those that do succeed do so in such a manner that not only covers the costs of the failed investments, but also generates the very high levels of returns required to please their investors and cover the VC fund's own management costs.
- IPO ('Initial Public Offering') refers to the selling of shares in a business on a public market such as the London Stock Exchange or NASDAQ. This is usually only done when the company has a track record, but can happen at the early stages of development of a business. During the 'dot com' investment boom of the late 1990s, this was quite a common route to getting money into a new business.

(b)(i)



Informal agreement – This may be a loose collaboration around sharing some ideas.

Contractual agreement – There is a legal agreement between the two organisations to work together on some particular activity.

Equity agreement – One of the organisations buys a part of the ownership of the other organisation.

Joint ownership – Both organisations agree to form a new organisation that they will both own.

Mergers and acquisitions – One organisation takes over ownership of the partner organisation.

Each of these different types of partnership are appropriate in different situations. The choice of partnership type depends on the aims of both partners in seeking to work together

(ii)

Needs analysis – What does your business intend to do 'in-house' and what do you intend to 'outsource'?

Identification – How do you find the best potential partner organisations?

Selection – How do you choose between the various possible partner organisations?

Negotiation – What type of partnership will be best (in the short, medium and long term)?

How much effort are you willing to put in to get the best agreement?

Formation – When is this going to happen, and who is going to be responsible for making it work?

Management – How are you going to make sure that the partnership is really working? How will you deal with disagreements?

Evolution – What if you, or the partner organisation, changes strategy? What if the partner goes out of business? How do you get out of the agreement? Do you have 'Plan B'?

(iii)

Examples given in class that they could use include: CDT, CSR, Microsoft, Hypertag, Palm, but they can use their own examples. Typical issues that will have been raised that they could present include:

Size difference and proximity – There are some basic practical issues around bringing together a company with a few employees with one having tens of thousands. Who should you talk to in the larger company? How to deal with the fact that people frequently change roles within larger firms? Also, if the two companies are located at opposite sides of the world, this can lead to some real management problems. One quote from a start-up that had many problems working with a larger company reveals how this situation can feel: "We felt like a small speedboat trying to dock with a supertanker".

Strategy and business models – Both sides of any partnership will have their own strategy and business models. Partnerships are formed when there is mutual strategic need. But strategies and business models tend to be dynamic. What happens when one partner's strategy changes?

Sector and organisational 'clockspeed' – Companies and industries have particular 'clockspeeds'. For example, in business to business selling, the time between making first contact with a customer and receiving payment may be either a few days to several months (for aerospace, it may be years). If one company is used to operating at the short deal cycle end and the other at the long deal cycle end, this can cause practical cash flow problems.

Resources and funding – Partnerships take a lot of time to make them work. For a small company with very few people, the proportion of time that they are devoting to making the partnership work can be very significant, and may lead to insufficient time being available for other management tasks. For the bigger firm, they may be able to devote substantial resource to making the partnership work. For the smaller firm, the partnership may become so time consuming and distracting that it becomes harmful to the overall success of the business.

Partnering capability – The ability to work with other firms has become a key skill for many high tech firms. The level of skill is often linked to amount of experience the companies have at managing partnerships, and whether lessons learned are converted into management practices. For a start-up, a good question to ask of a big firm is 'Have you ever worked with a company small as ours?'

Question 23

(a) The usual structure of a UK patent is: abstract, description, claims, and often drawings. The description usually includes a statement of the prior art and of the problem which the invention solves. It then describes in general terms the way in which the invention is carried out, and in very specific detail one or more examples of the invention, sufficient for someone with appropriate skills to carry out (or manufacture) the invention. The claims are the most important part of the patent as they define in precise terms the extent of the monopoly granted. They are usually constructed as a nested sequence of descriptions, proceeding from more general to less general. The purpose of this structure is to achieve as wide protection as possible, yet to prevent the whole patent being invalidated by one claim being rendered invalid (e.g. through challenge after the patent has been granted).

To be patentable the invention must satisfy four strict tests:

- it must be novel – i.e. must never have been disclosed in public anywhere in the world before the application date;
- involve an inventive step: i.e. not be obvious to someone skilled but completely unimaginative in the light of what has been done before (the ‘prior art’);
- have a practical application: be capable of being made or used in some kind of ‘industry’ – but not only in a purely intellectual/aesthetic activity; and
- not be excluded (e.g. scientific theory or mathematical method, method of doing business, perpetual motion machine....)

A patent provides a monopoly to prevent anyone else from making, using, selling or importing the invention. The protection lasts for up to 20 years, but an annual fee must be paid to keep the patent in force and protection only exists in the country in which the patent is held. The patent is published and is available to the public worldwide; it must provide full details of the invention. This means that the invention could be used quite legitimately by a competitor in a country where there is no patent protection. To obtain patents in all major world markets would involve major costs. Furthermore, the availability of full details of the invention may stimulate competitors to further inventiveness, and allow them to develop competing processes more readily than they would have done without that knowledge. Benefits of patenting include the protection outlined above, and the deterrent effects of the patent on competitors. The question refers to a manufacturing process rather than a product, and it may not be easy to prove that a competitor is indeed infringing a patent simply from examination of the product of the process, so that defending the patent might be difficult. Whatever the invention, the patent must be defended by the patent holder, and unless they are prepared to incur substantial legal costs in doing this, the deterrent effect of the patent may be small.

(b)

The product is a mixture of iron oxide particles, other chemicals and water. It seems that the special properties of the mixture result from the nature of the particles (produced by a special process), the particular blend of chemicals added, and the way in which the mixture is made (by heating and agitation). All these features are needed in order for the product to be effective. It would probably be possible for a competitor to ‘reverse-engineer’ the mixture and determine the nature of the oxide particles (composition, particle size and shape etc) as well as the chemical composition of the solution. So these could be protected by patenting (perhaps separate patents for the oxide particles and the overall composition of the slurry). The process by which the mixture is made (heating and agitation) would not be evident from analysis of the slurry, so it might be better to treat this as confidential know-how and not

patent that aspect, since then publication of the details would be required. In this case a combination of patenting the composition and keeping the manufacturing process as a trade secret might be appropriate. Another valid approach might be just to keep the process secret, without patenting the other features, which would avoid the costs of obtaining, maintaining and possibly defending patents. This could be a sound approach if it was indeed very difficult for anyone to achieve the right properties in the product without knowing how the mixing process was carried out.