## Two Hours

## THE UNIVERSITY OF MANCHESTER

## FRACTAL GEOMETRY

3rd June 2019
09:45-11:45

Answer ALL FOUR questions in Section A (48 marks in total) and
TWO of the THREE questions in Section B (32 marks in total).
If more than two questions from Section $B$ are attempted then credit will be given for the best two answers.

Electronic calculators may be used, provided that they cannot store text.

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## SECTION A

## Answer ALL FOUR questions

A1.
(a) Let $(X, d)$ be a metric space and $E \subset X$. Let $s \geq 0, \delta>0$.

Define what is meant by a $\delta$-cover of E. Define $\mathcal{H}_{\delta}^{s}(E)$.
(b) Define $\mathcal{H}^{s}(E)$, and justify why any limits that you used in your definition exist.
[4 marks]
(c) Give the definition of an outer measure on a metric space $(X, d)$. Prove that $\mathcal{H}^{s}$ satisfies the second condition of the definition of an outer measure (relating $\mathcal{H}^{s}(A)$ and $\mathcal{H}^{s}(B)$ where $A \subset B)$.
(d) Define $\operatorname{dim}_{H}(E)$ in terms of $\mathcal{H}^{s}(E)$. Prove that if $A \subset B$ then $\operatorname{dim}_{H}(A) \leq \operatorname{dim}_{H}(B)$.

A2.
(a) What does it mean for an outer measure $\mu$ on a space ( $X, d$ ) to be a mass distribution? (You do not need to define outer measure.)
(b) State and prove the mass distribution principle.
[6 marks]
(c) Use the mass distribution principle to prove that $\operatorname{dim}_{H}([0,1]) \geq 1$ (using the Euclidean metric on $\mathbb{R}$ ). You do not need to prove that the mass distribution you use is a mass distribution.
[4 marks]

A3.
(a) Let $(X, d)$ be a metric space. Define the space $\mathcal{S}(X)$ and the Hausdorff metric $d_{H}$ on $\mathcal{S}(X)$. For $X=\mathbb{R}$ with the Euclidean metric, $A=[1,2], B=\{0\}$, write down $d_{H}(A, B)$.
(b) What does it mean for an iterated function system to satisfy the open set condition? Write down an iterated function system $\Phi$ which satisfies the open set condition and for which the attractor is the Sierpinski triangle.
(c) Given an iterated function system $\Phi=\left\{\phi_{1}, \cdots, \phi_{N}\right\}$ where each $\phi_{i}$ is a similarity on $\left(\mathbb{R}^{n}, d\right)$ with contraction ratio $c_{i}<1$, state the definition of the similarity dimension $\operatorname{dim}_{\text {sim }}(\Phi)$. (Note: not all of the $c_{i}$ are the same.) In the special case that each contraction ratio $c_{i}=c$, prove that your above general definition of similarity dimension satisfies

$$
\operatorname{dim}_{s i m}(\Phi)=\frac{\log N}{-\log c}
$$

A4. Consider $\mathbb{R}$ equipped with the Euclidean metric and let $E \subset \mathbb{R}$. Consider the set

$$
E+1:=\{x+1: x \in E\}
$$

Prove that $\operatorname{dim}_{H}(E)=\operatorname{dim}_{H}(E+1)$.

## SECTION B

## Answer TWO of the THREE questions

## B5.

(a) Let maps $\phi_{1}, \phi_{2}, \phi_{3}: \mathbb{R}^{2} \rightarrow \mathbb{R}^{2}$ be given by

$$
\begin{aligned}
\phi_{1}(x, y) & =\left(\frac{x}{3}, \frac{y}{3}\right) \\
\phi_{2}(x, y) & =\left(\frac{x+1}{3}, \frac{y+2}{3}\right) \\
\phi_{2}(x, y) & =\left(\frac{x+2}{3}, \frac{y+1}{3}\right)
\end{aligned}
$$

Let

$$
E=\bigcup_{i=1}^{3} \phi_{i}(E) \subset \mathbb{R}^{2}
$$

Sketch the first two levels in the construction of $E$.
(b) Prove that $\overline{\operatorname{dim}}_{B}(E) \leq 1$.
(c) Without using the mass distribution principle, prove that $\operatorname{dim}_{H}(E) \geq 1$. Why can you now conclude that $\operatorname{dim}_{B}(E)$ exists?

B6.
(a) Let $\alpha \in(0,1]$. What does it mean for a map $f:(X, d) \rightarrow\left(X^{\prime}, d^{\prime}\right)$ to be $\alpha$-Hölder continuous with Hölder constant $K$ ? Let $f:(X, d) \rightarrow\left(X^{\prime}, d^{\prime}\right)$ be $\alpha$-Hölder continuous and $E \subset X$. Prove that

$$
\mathcal{H}^{\frac{s}{\alpha}}(f(E)) \leq K^{\frac{s}{\alpha}} \mathcal{H}^{s}(E)
$$

(b) State a corollary to the above statement relating $\operatorname{dim}_{H}(E)$ and $\operatorname{dim}_{H}(f(E))$.
(c) Let $F$ denote the middle- $\frac{1}{3}$ Cantor set. By defining a Hölder continuous map from $F$ to another space, give a lower bound for the dimension of $F$. You should prove that the map is Hölder continuous with the necessary constants.

B7.
(a) State and prove Hutchinson's Theorem. You may use the fact that the space $\mathcal{S}(X)$ is complete.
[8 marks]
(b) Let $\phi_{1}, \phi_{2}: \mathbb{R} \rightarrow \mathbb{R}$ be given by $\phi_{1}(x)=\frac{x}{3}, \phi_{2}(x)=\frac{x+1}{2}$. Let $E=\bigcup_{i=1}^{2} \phi_{i}(E)$. Prove that $\operatorname{dim}_{H}(E) \geq \operatorname{dim}_{\text {sim }}(\Phi)$ where $\Phi=\left\{\phi_{1}, \phi_{2}\right\}$.

## Two hours

## THE UNIVERSITY OF MANCHESTER

TOPOLOGY

21 May 2019
9:45-11:45

Answer ALL SIX questions

University approved calculators may be used.
(c) The University of Manchester, 2019
1.
(a) Define what is meant by a topology on a set $X$.
(b) Define what is meant by the usual topology on a subset $X \subset \mathbb{R}^{n}$.
(c) Define what is meant by saying that a function $f: X \rightarrow Y$ between topological spaces is continuous. Define what is meant by saying that $f$ is a homeomorphism.
(d) Consider the set $X=\{a, b, c\}$ and $\tau=\{\emptyset,\{a\},\{b, c\},\{a, b, c\}\}$. Show that $\tau$ defines a topology on $X$. Find all continuous maps from $X \rightarrow\{1,2\}$ with the usual topology on $\{1,2\}$.
[15 marks]
2.
(a) Define what is meant by saying that a topological space $X$ is path-connected.
(b) What is meant by saying that path-connectedness is a topological property?
(c) Define what is meant by the set $\pi_{0}(X)$ of path-components of a topological space $X$.
(d) Prove that a continuous map of topological spaces $f: X \rightarrow Y$ induces a map $f_{*}: \pi_{0}(X) \rightarrow \pi_{0}(Y)$ between the sets of path-components, taking care to prove that your function is well-defined. Prove that if $f$ is a homeomorphism then $f_{*}$ is a bijection.
(e) Use (d) to show that path-connectedness is a topological property.
(f) Consider the topological space $(X, \tau)$ from Question 1. (d), what are the path-components of this space? Justify your answer.
[15 marks]
3.
(a) Suppose that $q: X \rightarrow Y$ is a surjection from a topological space $X$ to a set $Y$. Define the quotient topology on $Y$ determined by $q$. State the universal property of the quotient topology.
(b) Suppose that $f: X \rightarrow Z$ is a continuous surjection from a compact topological space $X$ to a Hausdorff topological space $Z$. Define an equivalence relation $\sim$ on $X$ so that $f$ induces a bijection $F: X / \sim \rightarrow Z$ from the identification space $X / \sim$ of this equivalence relation to $Z$. Prove that $F$ is a homeomorphism. [State clearly any general results which you use.]
(c) Prove that the quotient space $\left(S^{1} \times[0,1]\right) /\left(S^{1} \times\{1\}\right)$ is homeomorphic to the closed unit disc $D^{2}$ (where $S^{1}$ and $D^{2}$ have the usual topology).
4.
(a) Define what is meant by saying that a topological space is Hausdorff.
(b) Suppose that $X$ and $Y$ are topological spaces. Define the product topology on the Cartesian product $X \times Y$. [It is not necessary to prove that this is a topology.]
(c) Consider the Sierpinski topology $\tau=\{\emptyset,\{a\},\{a, b\}\}$ on $X=\{a, b\}$. Write down the product topology on $X \times X$.
(d) Prove that if $X$ and $Y$ are Hausdorff spaces, then the product space $X \times Y$ is Hausdorff, as well.
(e) Prove that, if a topological space $X$ is Hausdorff, then all singleton subsets $\{x\}$ of $X$ (where $x \in X$ ) are closed. Give an example to show that the converse of this statement is false; that is, if a topological space has all singleton subsets closed, it is not necessarily Hausdorff.
[15 marks]
5.
(a) Define what is meant by a compact subset of a topological space.
(b) Prove by using only the definition of compactness, that $\left\{\left.\frac{1}{n} \right\rvert\, n \in \mathbb{N}\right\} \subset \mathbb{R}$ is not compact.
(c) Prove that every compact subset of a Hausdorff topological space is closed.
(d) Give an example to show that the Hausdorff condition in (c) cannot be omitted.
6.
(a) Suppose that $X_{1}$ is a subspace of a topological space $X$. Define what is meant by saying that $X_{1}$ is a retract of $X$.
(b) Show that $S^{1}=\left\{x \in \mathbb{R}^{2}| | x \mid=1\right\}$ is a retract of the punctured plane $\mathbb{R} \backslash\{0\}$.
(c) Explain how a continuous function of topological spaces $f: X \rightarrow Y$ induces a homomorphism of fundamental groups $f_{*}: \pi_{1}\left(X, x_{0}\right) \rightarrow \pi_{1}\left(Y, f\left(x_{0}\right)\right)$ for $x_{0} \in X$. You should indicate why $f_{*}$ is a well-defined homomorphism.
(d) Use the functorial properties of the fundamental group to prove that, if $X_{1}$ is a retract of $X$, then, for any $x_{0} \in X_{1}$, the homomorphism induced by the inclusion map

$$
i_{*}: \pi_{1}\left(X_{1}, x_{0}\right) \rightarrow \pi_{1}\left(X, x_{0}\right)
$$

is injective.
(e) Hence prove that $S^{1}$ is not a retract of the closed disc $D^{2}$.
[You may quote any calculation of fundamental groups that you need, without proof.]

## Two hours

## THE UNIVERSITY OF MANCHESTER

TOPOLOGY

21 May 2019
9:45-11:45

Answer ALL SIX questions

University approved calculators may be used.
(c) The University of Manchester, 2019
1.
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(c) Define what is meant by saying that a function $f: X \rightarrow Y$ between topological spaces is continuous. Define what is meant by saying that $f$ is a homeomorphism.
(d) Consider the set $X=\{a, b, c\}$ and $\tau=\{\emptyset,\{a\},\{b, c\},\{a, b, c\}\}$. Show that $\tau$ defines a topology on $X$. Find all continuous maps from $X \rightarrow\{1,2\}$ with the usual topology on $\{1,2\}$.
[15 marks]
2.
(a) Define what is meant by saying that a topological space $X$ is path-connected.
(b) What is meant by saying that path-connectedness is a topological property?
(c) Define what is meant by the set $\pi_{0}(X)$ of path-components of a topological space $X$.
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(e) Use (d) to show that path-connectedness is a topological property.
(f) Consider the topological space $(X, \tau)$ from Question 1. (d), what are the path-components of this space? Justify your answer.
[15 marks]
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(c) Prove that the quotient space $\left(S^{1} \times[0,1]\right) /\left(S^{1} \times\{1\}\right)$ is homeomorphic to the closed unit disc $D^{2}$ (where $S^{1}$ and $D^{2}$ have the usual topology).
4.
(a) Define what is meant by saying that a topological space is Hausdorff.
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(d) Prove that if $X$ and $Y$ are Hausdorff spaces, then the product space $X \times Y$ is Hausdorff, as well.
(e) Prove that, if a topological space $X$ is Hausdorff, then all singleton subsets $\{x\}$ of $X$ (where $x \in X$ ) are closed. Give an example to show that the converse of this statement is false; that is, if a topological space has all singleton subsets closed, it is not necessarily Hausdorff.
[15 marks]
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(b) Prove by using only the definition of compactness, that $\left\{\left.\frac{1}{n} \right\rvert\, n \in \mathbb{N}\right\} \subset \mathbb{R}$ is not compact.
(c) Prove that every compact subset of a Hausdorff topological space is closed.
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6.
(a) Suppose that $X_{1}$ is a subspace of a topological space $X$. Define what is meant by saying that $X_{1}$ is a retract of $X$.
(b) Show that $S^{1}=\left\{x \in \mathbb{R}^{2}| | x \mid=1\right\}$ is a retract of the punctured plane $\mathbb{R} \backslash\{0\}$.
(c) Explain how a continuous function of topological spaces $f: X \rightarrow Y$ induces a homomorphism of fundamental groups $f_{*}: \pi_{1}\left(X, x_{0}\right) \rightarrow \pi_{1}\left(Y, f\left(x_{0}\right)\right)$ for $x_{0} \in X$. You should indicate why $f_{*}$ is a well-defined homomorphism.
(d) Use the functorial properties of the fundamental group to prove that, if $X_{1}$ is a retract of $X$, then, for any $x_{0} \in X_{1}$, the homomorphism induced by the inclusion map

$$
i_{*}: \pi_{1}\left(X_{1}, x_{0}\right) \rightarrow \pi_{1}\left(X, x_{0}\right)
$$

is injective.
(e) Hence prove that $S^{1}$ is not a retract of the closed disc $D^{2}$.
[You may quote any calculation of fundamental groups that you need, without proof.]

## Two hours

## THE UNIVERSITY OF MANCHESTER

RIEMANNIAN GEOMETRY

04 June 2019
$09.45-11.45$

Answer ALL FOUR questions in Section A (50 marks in total)
Answer TWO of the THREE questions in Section B (40 marks in total)
If more than TWO questions in Section B are attempted,
the credit will be given for the best TWO answers.

The use of electronic calculators is not permitted.

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## SECTION A

## Answer ALL FOUR questions

A1.
(a) Explain what is meant by saying that $G$ is a Riemannian metric on a manifold $M$.
(b) Let $G$ be a Riemannian metric on a 2-dimensional manifold $M$ such that in local coordinates (u,v)

$$
G=\left\|g_{\alpha \beta}\right\|=\left(\begin{array}{ll}
A(u, v) & B(u, v) \\
C(u, v) & D(u, v)
\end{array}\right),
$$

where $A(u, v), B(u, v), C(u, v), D(u, v)$ are smooth functions.
Show that $C(u, v) \equiv B(u, v)$ and $A(u, v)>0, D(u, v)>0$.
(c) Show that $\operatorname{det}\left\|g_{\alpha \beta}\right\| \neq 0$.

A2.
(a) Explain what is meant by a Riemannian manifold being locally Euclidean.
(b) Consider the surface of the cylinder $x^{2}+y^{2}=9$ in $\mathbf{E}^{3}$

$$
\mathbf{r}(h, \varphi): \quad\left\{\begin{array}{l}
x=3 \cos \varphi \\
y=3 \sin \varphi \quad,-\infty<h<\infty, 0 \leq \varphi<2 \pi \\
z=h
\end{array}\right.
$$

Calculate the Riemannian metric $G$ on this surface induced by the Euclidean metric in $\mathbf{E}^{3}$ in coordinates $(h, \varphi)$.
Show that the surface of the cylinder is locally Euclidean by giving an example of local coordinates $(u, v)$, which are Euclidean coordinates.
(c) Give an example of other coordinates ( $\tilde{u}, \tilde{v}$ ) such that these coordinates are also Euclidean coordinates.

## A3.

(a) Explain what is meant by an affine connection on a manifold.
(b) Let $\nabla$ be an affine connection on a manifold $M$. Show that for arbitrary vector fields $\mathbf{X}$ and $\mathbf{Y}$ and for an arbitrary function $f$,

$$
e^{-f} \nabla_{\mathbf{x}}\left(e^{f} \mathbf{Y}\right)=\left(\partial_{\mathbf{x}} f\right) \mathbf{Y}+\nabla_{\mathbf{X}} \mathbf{Y}
$$

(c) Explain what is meant by a canonical flat connection on Euclidean space $\mathbf{E}^{n}$.

Calculate the Christoffel symbols $\Gamma_{r r}^{r}$ and $\Gamma_{\varphi \varphi}^{r}$ of the canonical flat connection in the Euclidean space $\mathbf{E}^{2}$, where $r, \varphi$ are polar coordinates $(x=r \cos \varphi, y=r \sin \varphi)$.

## A4.

(a) State the relation between the Riemannian curvature tensor of the Levi-Civita connection of a surface in $\mathbf{E}^{3}$ and its Gaussian curvature.
(b) Consider in $\mathbf{E}^{3}$ the surface of the saddle $z=k x y(k \neq 0)$ :

$$
\mathbf{r}(u, v):\left\{\begin{array}{l}
x=u \\
y=v \\
z=k u v
\end{array} \quad, \quad-\infty<u, v<\infty\right.
$$

Using the shape operator calculate the Gaussian curvature at the point $\mathbf{p}$ of the saddle with coordinates $x=y=z=0$.
Calculate the scalar curvature $R$ of the Riemannian curvature tensor of the induced Riemannian metric on the saddle at the point $\mathbf{p}$.
(c) Explain why the saddle is not a locally Euclidean surface.

## SECTION B

Answer TWO of the three questions

B5.
(a) Consider the stereographic projection of the sphere $x^{2}+y^{2}+z^{2}=1$ (without the north pole) on the plane $\mathbf{R}^{2}:(u, v)$ are standard coordinates on $\mathbf{R}^{2}$ and for arbitrary point $(x, y, z)$ on the sphere except the point $z=1$,

$$
\left\{\begin{array}{l}
u=\frac{x}{1-z} \\
v=\frac{y}{1-z}
\end{array} \quad, \quad\left\{\begin{array}{l}
x=\frac{2 u}{u^{2}+v^{2}+1} \\
y=\frac{2 v}{u^{2}+v^{2}+1} \\
z=\frac{u^{2}+v^{2}-1}{u^{2}+v^{2}+1}
\end{array}\right.\right.
$$

The metric

$$
\begin{equation*}
G=\frac{4\left(d u^{2}+d v^{2}\right)}{\left(u^{2}+v^{2}+1\right)^{2}} \tag{1}
\end{equation*}
$$

is induced on the plane $\mathbf{R}^{2}$ from the metric on the sphere by the stereographic projection.
Calculate the total area of $\mathbf{R}^{2}$ with respect to the Riemannian metric (1).
(b) Let $C$ be the intersection of the sphere with the plane $z=x$.

Let $C^{\prime}$ be the image of the curve $C$ under the stereographic projection.
Show that $C^{\prime}$ is a circle.
(c) The curve $C^{\prime}$ divides the plane $\mathbf{R}^{2}$ into two domains. Find the areas of these domains of $\mathbf{R}^{2}$ with respect to the Riemannian metric (1).
$B 6$.
(a) Give a detailed formulation of the Levi-Civita Theorem. In particular write down the expression for the Christoffel symbols $\Gamma_{k m}^{i}$ of the Levi-Civita connection in terms of the Riemannian metric $G=g_{i k}(x) d x^{i} d x^{k}$.
(b) Consider sphere of radius $R$ in stereographic coordinates. It is equipped with metric

$$
\begin{equation*}
G=\frac{4 R^{4}\left(d u^{2}+d v^{2}\right)}{\left(R^{2}+u^{2}+v^{2}\right)^{2}} . \tag{2}
\end{equation*}
$$

Show that Christoffel symbols of Levi-Civita connection $\nabla$ vanish at the point $u=v=0$ in coordinates $(u, v)$.
(c) Consider on the sphere a connection $\tilde{\nabla}$ such that the Riemannian metric (2) is invariant with respect to this connection, and the Christoffel symbols $\tilde{\Gamma}_{v v}^{u}$ and $\tilde{\Gamma}_{u u}^{v}$ of the connection $\tilde{\nabla}$ vanish at all points $(u, v)$.
Show that the connection $\tilde{\nabla}$ does not coincide with the Levi-Civita connection $\nabla$.
Does it contradict to Levi-Civita Theorem that the Riemannian metric (2) is invariant with respect to two different connections?
Justify the answer.
[20 marks]
B7.
(a) Let $(M, G)$ be a Riemannian manifold, let $C$ be a curve on $M$ starting at the point $\mathbf{p}_{1}$ and ending at the point $\mathbf{p}_{2}$.
Explain what is meant by parallel transport $P_{C_{\mathrm{P}_{1} \mathrm{P}_{2}}}$ along the curve $C$.
(b) Show that in the case that $\mathbf{p}_{1}=\mathbf{p}_{2}=\mathbf{p}$ ( $C$ is a closed curve), then $P_{C_{\mathbf{P}_{1} \mathbf{P}_{2}}}=P_{C_{\mathbf{P}}}$ is an orthogonal operator in the tangent space $T_{\mathrm{p}} M$.
(c) Let $M$ be the sphere $x^{2}+y^{2}+z^{2}=R^{2}$ in $\mathbf{E}^{3}$ with induced Riemannian metric, and let the curve $C$ on $M$ be an intersection of the sphere with a plane $x+z=R$. Take the point $\mathbf{p}=(R, 0,0)$ on the curve $C$ and consider vectors $\mathbf{Y}=\frac{\partial}{\partial y}$ and $\mathbf{Z}=\frac{\partial}{\partial z}$ which are tangent to the sphere at the point $\mathbf{p}$. Find the result of the parallel transport of these vectors along the closed curve $C$.

## Two hours

## THE UNIVERSITY OF MANCHESTER

## CODING THEORY

20 May 2019
14:00-16:00

Answer ALL THREE questions in Section A (40 marks in total). Answer TWO of the THREE questions in Section B (40 marks in total). If more than TWO questions from Section B are attempted, then credit will be given for the best TWO answers.

University approved calculators may be used
(C) The University of Manchester, 2019

## SECTION A

Answer ALL questions in this section (40 marks in total)

A1. (a) Define what is meant by:

- a word $\underline{x}$ of length $n$ in a finite alphabet $F$;
- a code $C$ of length $n$ in a finite alphabet $F$;
- the Hamming distance $d(\underline{\mathrm{x}}, \underline{\mathrm{y}})$ between two words $\underline{\mathrm{x}}, \underline{\mathrm{y}}$ of length $n$;
- a nearest neighbour of a word $\underline{x}$ in a code $C$.
(b) Let $C$ be the binary repetition code of length 6 . List the codewords of $C$. Write down words $\underline{\mathrm{x}}, \underline{\mathrm{y}} \in \mathbb{F}_{2}^{6}$ such that $\underline{\mathrm{x}}$ has exactly one nearest neighbour in $C$, and $\underline{\mathrm{y}}$ has more than one nearest neighbour in $C$.
[10 marks]
A2.
(a) Explain what is meant by a standard array for a linear $[n, k, d]_{q}$-code $C$. State the number of rows and the number of columns in a standard array for $C$. Explain how to decode a received vector using the standard array.

You are now given that $D=\{0000,0111, \underline{\mathrm{v}}, 1110\}$ is a binary linear code.
(b) Prove that the vector $\underline{v}$ is 1001. Assuming that a codevector of $D$ is sent via $\operatorname{BSC}(p)$, show that, with probability $p^{2}-p^{4}$, the received vector will contain an error which is not detected.
(c) Construct a standard array for $D$. Use the standard array you have constructed to give an example where one bit error occurs in a codevector $\underline{c}$ of $D$ and is not corrected by the standard array decoder.
(d) Assume that $\underline{\mathrm{y}}, \underline{\mathrm{z}} \in \mathbb{F}_{2}^{4}, \underline{\mathrm{y}} \neq \underline{\mathrm{z}}$, and that $H$ is a parity check matrix for the code $D$ such that $\underline{\mathrm{y}} H^{T}=\underline{\mathrm{z}} H^{T}$. Can $\underline{\mathrm{y}}$ and $\underline{\mathrm{z}}$ occur in the same column of the standard array that you constructed in part (c)? Justify your answer briefly.
[20 marks]
A3.
(a) Define what is meant by a line in the vector space $\mathbb{F}_{q}^{r}$. Explain how to construct a check matrix $H$ for a Hamming $\operatorname{Ham}(r, q)$-code. State without proof the number of rows and the number of columns of $H$.
(b) Write down a check matrix of a ternary $\operatorname{Ham}(2,3)$-code.
(c) Write down a check matrix of a $\operatorname{Ham}(2,5)$-code $C \subseteq \mathbb{F}_{5}^{6}$ such that $C$ contains the vector 111111 .

## SECTION B

Answer TWO of the three questions in this section (40 marks in total).
If more than TWO questions from this section are attempted, then credit will be given for the best TWO answers.

B4. In this question, $F$ is a finite alphabet of cardinality $q$, and $C \subseteq F^{n}$ is a code of cardinality $M>1$.
(a) Define what is meant by the minimum distance, $d(C)$, of the code $C$.
(b) Prove the Singleton bound which says that $M \leqslant q^{n-d(C)+1}$.
(c) Let $r$ be an integer between 0 and $n$. Define the Hamming sphere $S_{r}(\underline{\mathrm{u}})$ of radius $r$ centred at $\underline{\mathrm{u}} \in F^{n}$. State and prove the formula for the cardinality of $S_{r}(\underline{\mathrm{u}})$ in terms of $r, n$ and $q$.
(d) State without proof the Hamming bound on $M$ in terms of $n, q$ and $d(C)$. Define what is meant by saying that $C$ is a perfect code.
(e) Show that if $n=q=60$ and $C$ is perfect, then $d(C) \neq 7$. Hint: you may use the Singleton bound.
[20 marks]
B5. In this question, $C \subseteq \mathbb{F}_{q}^{n}$ is a linear code of dimension $k$.
(a) Define what is meant by the inner product $\underline{x} \cdot \underline{y}$ of two vectors $\underline{x}, \underline{y} \in \mathbb{F}_{q}^{n}$, and what is meant by the dual code $C^{\perp}$. State without proof the length ${ }^{-1}$ and the dimension of $C^{\perp}$.
(b) Recall that $C$ is self-orthogonal if $C \subseteq C^{\perp}$. Prove that if $C$ has a generator matrix $G$ such that $G G^{T}$ is a zero matrix, then $C$ is self-orthogonal.
(c) A self-orthogonal code $D \subseteq \mathbb{F}_{q}^{4}$ has generator matrix $\left[\begin{array}{rrrr}1 & 2 & -1 & 2 \\ 2 & 1 & 2 & -1\end{array}\right]$. Find a prime number $q>2$ for which this is possible. Then write down a check matrix for $D$, explaining how it is obtained.
(d) Assume that $q=2$ and $C \subseteq C^{\perp}$. Let $W_{C}(x, y)=\sum_{i=0}^{n} A_{i} x^{n-i} y^{i}$ be the weight enumerator of $C$. Prove that $A_{2} \leqslant k$ and that $A_{2}\left(A_{2}-1\right) / 2 \leqslant A_{4}$.
[20 marks]
B6. In this question, we assume that vectors in $\mathbb{F}_{q}^{n}$ are identified with polynomials, in $\mathbb{F}_{q}[x]$, of degree less than $n$ in the usual way.
(a) Let $C \subseteq \mathbb{F}_{q}^{n}$ be a cyclic code. Explain what is meant by a generator polynomial of $C$. Show that there exists at most one generator polynomial of $C$.
(b) Let $Z \subseteq \mathbb{F}_{7}^{8}$ be the cyclic code with generator polynomial $x-1$. Write down a generator matrix for $Z$. Prove that $Z=\left\{\left(v_{1}, v_{2}, v_{3}, v_{4}, v_{5}, v_{6}, v_{7}, v_{8}\right) \in \mathbb{F}_{7}^{8}: \sum_{i=1}^{8} v_{i}=0\right.$ in $\left.\mathbb{F}_{7}\right\}$.
You may use the factorisation $(x-1)(x+1)\left(x^{2}+1\right)\left(x^{2}-3 x+1\right)\left(x^{2}+3 x+1\right)$ of $x^{8}-1$ into irreducible monic polynomials in $\mathbb{F}_{7}[x]$ to answer the following questions.
(c) Determine: (i) the total number of cyclic codes $C \subseteq \mathbb{F}_{7}^{8}$; (ii) the number of cyclic codes $C \subseteq \mathbb{F}_{7}^{8}$ such that $C \subseteq Z$ where $Z$ is the code from part (b). Justify your answers briefly.
(d) Prove that none of the cyclic codes in $\mathbb{F}_{7}^{8}$ is a Hamming code. Any facts from the course can be freely used.

## END OF EXAMINATION PAPER

## 2 hours

## THE UNIVERSITY OF MANCHESTER

## HYPERBOLIC GEOMETRY

30 May 2019
14:00-16:00

Answer THREE of the FOUR questions.
If all four questions are attempted then credit will be given for the three best answers.

Electronic calculators may be used, provided that they cannot store text.

Notation: Throughout, $\mathbb{H}$ denotes the upper half-plane, $\partial \mathbb{H}$ denotes the boundary of $\mathbb{H}, \mathbb{D}$ denotes the Poincaré disc, and $\partial \mathbb{D}$ denotes the boundary of $\mathbb{D}$.
1.
(i) What does it mean to say that $\gamma$ is a Möbius transformation of $\mathbb{H}$ ?
(ii) Let $\gamma \in \operatorname{Möb}(\mathbb{H})$. How is $\tau(\gamma)$ defined? Briefly state (without proof) how to use $\tau(\gamma)$ to determine if $\gamma$ is hyperbolic, parabolic or elliptic.

Let

$$
\gamma_{1}(z)=\frac{7 z-8}{2 z-1}, \quad \gamma_{2}(z)=\frac{2 z-1}{-7 z+8} .
$$

Determine whether each of $\gamma_{1}$ and $\gamma_{2}$ is hyperbolic, parabolic or elliptic.
(iii) Suppose that $\gamma_{1}, \gamma_{2} \in \operatorname{Möb}(\mathbb{H})$. Prove that the composition $\gamma_{1} \gamma_{2} \in \operatorname{Möb}(\mathbb{H})$.
(iv) It was proved in the course that, in $\mathbb{C}$, vertical straight lines and circles with real centres have equations of the form

$$
\alpha z \bar{z}+\beta z+\beta \bar{z}+\gamma=0
$$

where $\alpha, \beta, \gamma \in \mathbb{R}$.
Suppose that $A$ is either a vertical straight line or a circle with a real centre and that $\gamma \in$ Möb $(\mathbb{H})$. Show that $\gamma(A)$ is also either a vertical straight line or a circle with a real centre in $\mathbb{C}$.
(v) One can also prove the following fact (you do not need to prove this yourself): Let $A$ a circle in $\mathbb{C}$ (not necessarily with a real centre). Let $\gamma \in \operatorname{Möb}(\mathbb{H})$. Then $\gamma(A)$ is either a straight line in $\mathbb{C}$ (not necessarily vertical) or a circle in $\mathbb{C}$ (not necessarily with a real centre).
Now suppose that $A$ is a circle in $\mathbb{H}$ that touches the real axis at a single point, as illustrated in Figure 1. Using the above fact, describe (in words) precisely all the possibilities for $\gamma(A)$, justifying your answer.


Figure 1: See Question 1(v).
2. Throughout this question you may use, without proof, the formula

$$
\cosh d_{\mathbb{H}}(z, w)=1+\frac{|z-w|^{2}}{2 \operatorname{Im}(z) \operatorname{Im}(w)}
$$

You may also assume that in a hyperbolic right-angled triangle with angles $\alpha, \beta, \pi / 2$ and sides of hyperbolic length $a, b, c$ (as illustrated below) we have

$$
\begin{aligned}
& \sin \alpha=\frac{\sinh a}{\sinh c} \\
& \cos \alpha=\frac{\tanh b}{\tanh c} .
\end{aligned}
$$



You may also use the trig and hyperbolic trig identities: $\cosh ^{2} x-\sinh ^{2} x=1, \sinh 2 x=2 \sinh x \cosh x$, $\sin (x+y)=\sin x \cos y+\cos x \sin y, \sinh (x+y)=\sinh x \cosh y+\cosh x \sinh y$.
(i) Let $\Delta$ be an arbitrary hyperbolic right-angled triangle with sides of hyperbolic lengths $a, b, c$. Suppose that the side of hyperbolic length $c$ is opposite the right-angle. Prove the Hyperbolic Pythagoras' Theorem:

$$
\cosh c=\cosh a \cosh b .
$$

(If you reduce $\Delta$ to a special case by applying a Möbius transformation then you should briefly explain why you can do this.)
(ii) Let $\Delta$ be an arbitrary acute hyperbolic triangle with internal angles $\alpha, \beta, \gamma$ and opposite sides of hyperbolic length $a, b, c$ (see Figure 2 on the following page). Prove the hyperbolic sine rule:

$$
\frac{\sin \alpha}{\sinh a}=\frac{\sin \beta}{\sinh b}=\frac{\sin \gamma}{\sinh c} .
$$

Hint: drop a perpendicular from a vertex to the opposite side to create two right-angled triangles.
[12 marks]
(iii) Let $\Delta$ be a hyperbolic triangle with internal angles $\alpha, \beta$ and $\gamma$. State, without proof, the Gauss-Bonnet theorem for a hyperbolic triangle.
(iv) Consider a tiling of the hyperbolic plane by regular hyperbolic triangles with $k$ triangles meeting at each vertex. Use the Gauss-Bonnet Theorem to show that $k \geq 7$.
(v) Sketch a tiling of the Poincaré disc by regular hyperbolic triangles when $k=\infty$.


Figure 2: An acute hyperbolic triangle; see Question 2(ii).

## 3.

(a) (i) Let $\Gamma$ be a Fuchsian group acting on $\mathbb{H}$. What does it mean to say that an open set $F \subset \mathbb{H}$ is a fundamental domain for $\Gamma$ ?
(ii) Briefly explain how to construct a Dirichlet region for a Fuchsian group $\Gamma$. (Your answer should include a definition of the terms $[p, \gamma(p)], L_{p}(\gamma), H_{p}(\gamma), D(p)$ that were defined in the lectures.)
(iii) It was proved in the lectures that the perpendicular bisector of the arc of geodesic $\left[z_{1}, z_{2}\right]$ between two points $z_{1}, z_{2} \in \mathbb{H}$ is given by

$$
\begin{equation*}
\left\{z \in \mathbb{H} \mid d_{\mathbb{H}}\left(z, z_{1}\right)=d_{\mathbb{H}}\left(z, z_{2}\right)\right\} . \tag{1}
\end{equation*}
$$

Write $z_{1}=x_{1}+i y_{1}, z_{2}=x_{2}+i y_{2}$. Show that (1) can be written in the form

$$
\begin{equation*}
\left\{z \in \mathbb{H}\left|y_{2}\right| z-\left.z_{1}\right|^{2}=y_{1}\left|z-z_{2}\right|^{2}\right\} . \tag{2}
\end{equation*}
$$

(You may use the fact that

$$
\left.\cosh d_{\mathbb{H}}(z, w)=1+\frac{|z-w|^{2}}{2 \operatorname{Im}(z) \operatorname{Im}(w)} .\right)
$$

(b) (i) Let $\gamma(z)=z+b$ where $b \in \mathbb{R}$. Let $p=i$. Use the formula (2) to calculate the perpendicular bisector of $[p, \gamma(p)]$.
(ii) Let $\Gamma=\left\{\gamma_{n} \mid \gamma_{n}(z)=z+4 n, n \in \mathbb{Z}\right\}$. Use the procedure you outlined in (a)(ii) and the result of (b)(i) above to show that $D(i)=\{z \in \mathbb{H} \mid-2<\operatorname{Re}(z)<2\}$.
(iii) Draw the resulting tessellation in $\mathbb{H}$. What does this tessellation look like in $\mathbb{D}$ ?
(iv) Describe (either in words or by drawing a picture) what $D(p)$ will look like for an arbitrary choice of $p \in \mathbb{H}$.
Find a fundamental domain for $\Gamma$ which is not of the form $D(p)$ for some $p \in \mathbb{H}$.
4.
(i) Let $\mathcal{E}=v_{1} \rightarrow v_{2} \rightarrow \cdots \rightarrow v_{n}$ be an elliptic cycle. Let the internal angle at $v_{j}$ be $\angle v_{j}$.

Define $\operatorname{sum}(\mathcal{E})$.
What does it mean to say that $\mathcal{E}$ satisfies the elliptic cycle condition?
What does it mean to say that $\mathcal{E}$ is an accidental cycle?
(ii) Consider the hyperbolic hexagon illustrated in Figure 3 with the side-pairings indicated. Here $s, t, u \geq 2$ are real numbers and $\theta_{1}+\theta_{2}+\theta_{3} \leq \pi$. (You may assume that such a hyperbolic hexagon exists.)


Figure 3: A hyperbolic hexagon; see Question 4(ii).

Show that there are 4 elliptic cycles. In each case, calculate the elliptic cycle transformations and the angle sum.
(iii) Determine conditions on $\theta_{1}, \theta_{2}, \theta_{3}, s, t, u$ so that the side-pairing transformations generate a Fuchsian group $\Gamma$. Give a presentation for $\Gamma$ in terms of generators and relations.
(iv) How is the signature of a cocompact Fuchsian group defined?

Assume that $\theta_{1}, \theta_{2}, \theta_{3}, s, t, u$ are such that the side-pairing transformations in Figure 3 generate a Fuchsian group $\Gamma$. Use Euler's formula to calculate the genus $g$ of $\mathbb{H} / \Gamma$. Hence write down the possibilities for the signature of $\Gamma$. Briefly describe $\mathbb{H} / \Gamma$ in each case.

## Two hours

## THE UNIVERSITY OF MANCHESTER

## ALGEBRAIC GEOMETRY

17 May, 2019
14:00-16:00

Answer THREE of the FOUR questions. If more than THREE questions are attempted, then credit will be given for the best THREE answers.

Electronic calculators may be used in accordance with the University regulations

1. In this question $K$ denotes an arbitrary field and unless specified otherwise, varieties are defined over $K$.
(a) (i) Give the definition of the affine algebraic variety $\mathcal{V}(J) \subseteq K^{n}$ determined by an ideal $J \triangleleft K\left[x_{1}, x_{2}, \ldots, x_{n}\right]$.
(ii) Define the ideal $\mathcal{I}(V)$ and the co-ordinate ring $K[V]$ of an affine algebraic variety $V \subseteq K^{n}$.
(iii) Let $J=\left\langle(x-y+z)^{2}, x^{2}-y^{2}+z^{2}\right\rangle \triangleleft K[x, y, z]$. Show that $\mathcal{I}(\mathcal{V}(J)) \neq J$.
(b) (i) Explain what it means to say that an affine algebraic variety $W$ is irreducible.
(ii) Prove that if $\mathcal{I}(W)$ is not prime, then $W$ is reducible. (You may use any other results from the course that you need.)
(iii) This part of the question is over $\mathbb{C}$. Find the irreducible components of the variety

$$
W=\mathcal{V}\left(\left\langle x y z-x z, x^{2}+2 y^{2}-z^{2}-2\right\rangle\right) \subset \mathbb{A}^{3}(\mathbb{C})
$$

(The irreducibility of certain varieties was proved in the course, you can use this if you identify correctly what kind of a variety each component is.)
(c) (i) Let $V \subseteq \mathbb{A}^{m}(K)$ and $W \subseteq \mathbb{A}^{n}(K)$ be affine algebraic varieties. Define the notion of a morphism $\varphi: V \rightarrow W$ and state how the ring homomorphism $\varphi^{*}: K[W] \rightarrow$ $K[V]$ is defined. (You are not asked to prove that $\varphi^{*}$ is a homomorphism.)
(ii) This part of the question is over $\mathbb{R}$. Let $C=\mathcal{V}\left(\left\langle x^{2}+y^{2}-1\right\rangle\right) \subset \mathbb{R}^{2}$ be the unit circle. Show that if $\varphi: \mathbb{A}^{1}(\mathbb{R}) \rightarrow C$ is a morphism, then it is a constant function, i. e., it maps the whole of $\mathbb{A}^{1}(\mathbb{R})$ to a single point. (Hint: Assume that there exists a non-constant morphism $\varphi: \mathbb{A}^{1}(\mathbb{R}) \rightarrow C$ and derive a contradiction.)
2.
(a) Let $K$ be a field. Let $V \subseteq \mathbb{A}^{n}(K)$ be an affine algebraic variety and let $P \in V$.
(i) Let $\mathbf{v} \in K^{n}$ be a non-zero vector. Let $\ell=\{P+t \mathbf{v} \mid t \in K\}$ be the line through $P$ with direction vector $\mathbf{v}$. Explain what is meant by saying that $\ell$ is tangent to $V$ at $P$ and give the definition of the tangent space $T_{P} V$ to $V$ at $P$.
(ii) Let $f_{1}, f_{2}, \ldots, f_{r}$ be a generating set for $\mathcal{I}(V)$ and let

$$
J_{P}=\left(\begin{array}{cccc}
\left.\frac{\partial f_{1}}{\partial x_{1}}\right|_{P} & \left.\frac{\partial f_{1}}{\partial x_{1}}\right|_{P} & \cdots & \left.\frac{\partial f_{1}}{\partial x_{n}}\right|_{P} \\
\left.\frac{\partial f_{2}}{\partial x_{1}}\right|_{P} & \left.\frac{\partial f_{2}}{\partial x_{2}}\right|_{P} & \cdots & \left.\frac{\partial f_{2}}{\partial x_{n}}\right|_{P} \\
\vdots & \vdots & \ddots & \vdots \\
\left.\frac{\partial f_{r}}{\partial x_{1}}\right|_{P} & \left.\frac{\partial f_{r}}{\partial x_{2}}\right|_{P} & \cdots & \left.\frac{\partial f_{r}}{\partial x_{n}}\right|_{P}
\end{array}\right)
$$

be the Jacobian matrix evaluated at $P \in V$.
State and prove the criterion in terms of $J_{P}$ which describes when the line $\ell=$ $\{P+t \mathbf{v} \mid t \in K\}\left(\mathbf{v} \in K^{n} \backslash\{\mathbf{0}\}\right)$ is tangent to $V$ at $P$.
(b) This part of the question is over $\mathbb{C}$. Let $x, y$ be co-ordinates on $\mathbb{A}^{2}(\mathbb{C})$ and let $t$ be the co-ordinate on $\mathbb{A}^{1}(\mathbb{C})$. Let

$$
f=x^{3}-6 x y^{2}-y^{4}-8 y^{2}
$$

and let $C=\mathcal{V}(\langle f\rangle) \subset \mathbb{A}^{2}(\mathbb{C})$. (You may assume without proof that $f$ is an irreducible polynomial, therefore $C$ is irreducible and $\mathcal{I}(C)=\langle f\rangle$.)
(i) Show that $\varphi(t)=\left(t^{4}+2 t^{2}, t^{3}\right)$ is a morphism $\varphi: \mathbb{A}^{1}(\mathbb{C}) \rightarrow C$.
(ii) Let $\psi(x, y)=\frac{x y+4 y}{2 x+y^{2}}$. Show that $\psi$ is a rational map $C \rightarrow \mathbb{A}^{1}(\mathbb{C})$ and that $\psi \circ \varphi: \mathbb{A}^{1}(\mathbb{C}) \rightarrow \mathbb{A}^{1}(\mathbb{C})$ is the identity map.
(iii) Find the three singular points of $C$.
3. In this question $K$ denotes an arbitrary field.
(a) (i) Let $n$ be a positive integer and let $I \triangleleft K\left[X_{0}, X_{1}, \ldots, X_{n}\right]$ be a homogeneous ideal. State the definition of $\mathcal{V}(I)$, the projective algebraic variety defined by $I$.
(ii) Let $W \subseteq \mathbb{P}^{n}$ be a projective algebraic variety for some $n \geq 1$. Define $\mathcal{I}(W)$ and $K[W]$.
(b) Let $m, n$ be positive integers and let $A$ be a $(n+1) \times(m+1)$ matrix with entries from $K$ other than the 0 matrix. Define $\Phi$ by $\Phi\left(X_{0}: X_{1}: \ldots: X_{m}\right)=\left(Y_{0}: Y_{1}: \ldots: Y_{n}\right)$, where $\left(\begin{array}{c}Y_{0} \\ Y_{1} \\ \vdots \\ Y_{n}\end{array}\right)=A\left(\begin{array}{c}X_{0} \\ X_{1} \\ \vdots \\ X_{m}\end{array}\right)$.
(i) Show that $\Phi$ is a rational map $\mathbb{P}^{m} \longrightarrow \mathbb{P}^{n}$.
(ii) Prove that if $m=n$ and $A$ is invertible, then $\Phi$ is a morphism.
(iii) Prove that if $m>n$ and the rows of $A$ are not all scalar multiples of one of the rows, then $\Phi$ is not a morphism.
(c) We identify $\mathbb{P}^{1}(K)$ with $K \cup\{\infty\}$ by identifying $\left(X_{0}: X_{1}\right)$ with $X_{0} / X_{1} \in K$ if $X_{1} \neq 0$ and $\left(X_{0}: 0\right)$ with $\infty$ as usual. Recall that if $z_{1}, z_{2}, z_{3}, z_{4} \in K \cup\{\infty\}$ with at least 3 of them pairwise distinct, their cross ratio is defined to be

$$
\left(z_{1}, z_{2} ; z_{3}, z_{4}\right)=\frac{\left(z_{1}-z_{3}\right)\left(z_{2}-z_{4}\right)}{\left(z_{1}-z_{4}\right)\left(z_{2}-z_{3}\right)} \in K \cup\{\infty\}
$$

with appropriate rules for arithmetic involving $\infty$ and for division by 0 .
Let $z_{1}, z_{2}, z_{3}, z_{4} \in K \cup\{\infty\}$ with at least 3 of them pairwise distinct. Show that $\left(z_{1}, z_{2} ; z_{3}, z_{4}\right)=\left(\varphi\left(z_{1}\right), \varphi\left(z_{2}\right) ; \varphi\left(z_{3}\right), \varphi\left(z_{4}\right)\right)$ for any projective transformation $\varphi: \mathbb{P}^{1} \rightarrow$ $\mathbb{P}^{1}$.
(d) Find $a, b, c, d \in \mathbb{C}$ such that $\varphi: \mathbb{C} \cup\{\infty\} \rightarrow \mathbb{C} \cup\{\infty\}, \varphi(z)=\frac{a z+b}{c z+d}$ satisfies $\varphi(-1)=-5, \varphi(3)=3$ and $\varphi(5)=4$.
4.
(a) Let $K$ be a field. Let $E \subset \mathbb{P}^{2}(K)$ be a non-singular cubic curve and let $O$ be a point of $E$. (You should not assume that the equation of $E$ is in a special form or that $O$ is an inflection point of $E$.)
(i) Describe without proof how one can define the addition operation on the points of $E$ to obtain a commutative group in which $O$ is the zero element.
(ii) Let $A \in E$. Describe how $-A$ is calculated and prove that this point $-A$ indeed satisfies $A+(-A)=O$.
(b) Let $F$ be the elliptic curve given by the affine equation

$$
y^{2}=x^{3}-4 x^{2}+8 x+4
$$

over $\mathbb{C}$ together with the point $O=(0: 1: 0)$ in homogeneous co-ordinates, which is taken to be the zero element of the group law as usual.
Let $A=(1,3), B=(4,-6)$. Calculate $A+B$ and $2 A$. (You do not have to verify that $A$ and $B$ lie on $F$.)
(c) The elliptic curve $G$ is defined by the affine equation

$$
9 y^{2}-18 x y-12 y=27 x^{3}-9 x^{2}
$$

over $\mathbb{C}$. Find $p, q \in \mathbb{C}$ such that the curve defined by the equation $y^{2}=x^{3}+p x+q$ is isomorphic to $G$.

## Two hours

## THE UNIVERSITY OF MANCHESTER

## GREEN'S FUNCTIONS, INTEGRAL EQUATIONS AND APPLICATIONS

## 29 May 2019

14:00-16:00

Answer ALL 6 questions.

The use of electronic calculators is not permitted.

Given equations:

- For functions $f$ and $g$ defined in either two or three dimensions, and a domain $D$ with smooth boundary in either two or three dimensions

$$
\begin{equation*}
\int_{D}\left(f(\mathbf{x}) \nabla^{2} g(\mathbf{x})-g(\mathbf{x}) \nabla^{2} f(\mathbf{x})\right) d \mathbf{x}=\int_{\partial D}(f(\mathbf{x}) \nabla g(\mathbf{x})-g(\mathbf{x}) \nabla f(\mathbf{x})) \cdot \mathbf{n}(\mathbf{x}) d s(\mathbf{x}) \tag{1}
\end{equation*}
$$

where the boundary $\partial D$ is oriented by the outward pointing unit normal $\mathbf{n}(\mathbf{x})$ at each point $\mathbf{x}$ in $\partial D$.

- The surface area of the unit sphere $\mathbb{S}^{2}$ is $4 \pi$ :

$$
\int_{\mathbb{S}^{2}} d s=4 \pi
$$

(C) The University of Manchester, 2019

Answer ALL questions (there are a total of 100 marks possible). To receive full marks you must show your work.
1.
(a) Is the following a regular Sturm-Liouville eigenvalue problem? Justify your answer.

$$
\left\{\begin{array}{c}
\cos (x) \phi^{\prime \prime}(x)-\sin (x) \phi^{\prime}(x)+\lambda \phi(x)=0, \quad x \in\left(-\frac{\pi}{2}, \frac{\pi}{2}\right) \\
\phi(-\pi / 2)=0, \quad \phi(\pi / 2)=0 .
\end{array}\right.
$$

(b) Suppose that $G$ is the Green's function for a boundary value problem

$$
\left\{\begin{aligned}
\mathcal{L} u(x)=f(x), & x \in(a, b) \\
\mathcal{B}, &
\end{aligned}\right.
$$

where $\mathcal{L}$ is a second order linear ordinary differential operator, and $\mathcal{B}$ is a set of two boundary conditions. Also suppose that $G^{*}$ is the adjoint Green's function.
(i) What boundary value problem does $G^{*}$ satisfy?
(ii) What is the relationship between $G$ and $G^{*}$ ?
(c) Give the free space Green's function for the Laplacian in three dimensions. You do not need to justify the formula for the Green's function, just write it down.
(d) Consider the integral equation

$$
u(x)=\int_{1}^{2} e^{-(x-y)^{2}} u(y) \mathrm{d} y+f(x)
$$

(i) What type of integral equation is this?
(ii) What is the kernel?
2. Consider the following boundary value problem:

$$
\left\{\begin{array}{c}
u^{\prime \prime}(x)+4 x u^{\prime}(x)+2\left(1+2 x^{2}\right) u(x)=f(x), \quad x \in(0,1) \\
u(0)=0, u(1)=0
\end{array}\right.
$$

A complementary solution for the ODE in this boundary value problem is

$$
u_{c}(x)=c_{1} e^{-x^{2}}+c_{2} x e^{-x^{2}}
$$

Find the Green's function.
3. Suppose that $p$ and $p^{\prime}$ are continuous, and $p(x)>0$ for all $x$. Consider the following boundary value problem:

$$
(\mathrm{BVP}) \quad\left\{\begin{array}{c}
p(x) u^{\prime \prime}(x)+p^{\prime}(x) u^{\prime}(x)=f(x), \quad x \in(0,1) \\
u(0)=0, u(1)=0
\end{array}\right.
$$

(a) Show that (BVP) is fully self-adjoint.
(b) Use the Fredholm Alternative to show that there exists a unique solution for (BVP).
4. Consider the following integral equation:

$$
\text { (INTEQ) } \quad u(x)=\lambda \int_{0}^{1} x^{3} y^{2} u(y) \mathrm{d} y+f(x)
$$

(a) For what values of $\lambda \in \mathbb{C}$ is there a unique solution $u(x)$ to (INTEQ) for any continuous function $f$ ?
[5 marks]
(b) For the values of $\lambda$ found in part (a), find a formula for the solution $u(x)$ of (INTEQ).
[5 marks]
(c) When $\lambda$ is not one of the values found in part (a) find a condition on $f$ so that there is a solution for (INTEQ). Also write a general solution for (INTEQ) in this case.
5. For time harmonic waves on an infinite string where the physical displacement from equilibrium is given by the real part of $e^{-i \omega t} u(x)$, the function $u$ will satisfy the Helmholtz equation

$$
\text { (HELMHOLTZ) } \quad u^{\prime \prime}(x)+k(x)^{2} u(x)=f(x),
$$

provided the oscillations are small. Assume that both $k(x)=k_{0}$ (a constant) and $f(x)=0$ when $|x|$ is sufficiently large.
(a) Describe the radiation conditions on $u$ which model the situation in which all waves are created by the source $f$. You should give the formulas for the radiation conditions, and explain how they are derived.
[8 marks]
(b) Assume that $u(x)=e^{i k_{0} x}+u_{s c}(x)$, where $u_{s c}$ satisfies the radiation conditions from part (a), is a solution of (HELMHOLTZ) with $f(x)=0$ for all $x$. Show that $u(x)$ satisfies a Fredholm integral equation of the second kind. You may use without justification the fact that the Green's function for (HELMHOLTZ) with $k(x)=k_{0}$ everywhere and the radiation conditions is

$$
G\left(x, x_{0}\right)=\frac{e^{i k_{0}\left|x-x_{0}\right|}}{2 i k_{0}} .
$$

(c) What is the Born approximation for the solution of the integral equation from part (b)? You do not need to evaluate the integral which appears in the approximation.
[6 marks]
[Total 22 marks]
6. In three dimensions we can also consider the Helmholtz equation

$$
\nabla^{2} u(\mathbf{x})+k^{2} u(\mathbf{x})=Q(\mathbf{x})
$$

where $\mathbf{x} \in \mathbb{R}^{3}$. In this question we will assume that $k$ is constant and use the notation

$$
G\left(\mathbf{x}, \mathbf{x}_{0}\right)=-\frac{1}{4 \pi} \frac{e^{i k\left|\mathbf{x}-\mathbf{x}_{0}\right|}}{\left|\mathbf{x}-\mathbf{x}_{0}\right|}
$$

where x and $\mathrm{x}_{0}$ are in $\mathbb{R}^{3}$.
(a) Show that $G$ satisfies

$$
\nabla_{\mathbf{x}}^{2} G\left(\mathbf{x}, \mathbf{x}_{0}\right)+k^{2} G\left(\mathbf{x}, \mathbf{x}_{0}\right)=0
$$

when $\mathrm{x} \neq \mathrm{x}_{0}$.
[9 marks]
(b) Let $f$ be a function on $\mathbb{R}^{3}$ which is twice continuously differentiable, and such that $f(\mathbf{x})=0$ when $|\mathbf{x}|$ is sufficiently large. Also, suppose that for $\epsilon>0, B_{\epsilon}\left(\mathbf{x}_{0}\right)$ is the ball of radius $\epsilon$ around $\mathbf{x}_{0}$. Show that

$$
\int_{\mathbb{R}^{3} \backslash B_{\epsilon}\left(\mathbf{x}_{0}\right)} G\left(\mathbf{x}, \mathbf{x}_{0}\right)\left(\nabla^{2} f(\mathbf{x})+k^{2} f(\mathbf{x})\right) \mathrm{d} \mathbf{x}=\frac{e^{i k \epsilon}}{4 \pi \epsilon^{2}} \int_{\partial B_{\epsilon}\left(\mathbf{x}_{0}\right)}(1-i k \epsilon) f(\mathbf{x})+\nabla f(\mathbf{x}) \cdot\left(\mathbf{x}-\mathbf{x}_{0}\right) d s(\mathbf{x})
$$

(c) Take the limit $\epsilon \rightarrow 0^{+}$in the formula from part (b) to show that

$$
\int_{\mathbb{R}^{3}} G\left(\mathbf{x}, \mathbf{x}_{0}\right)\left(\nabla^{2} f(\mathbf{x})+k^{2} f(\mathbf{x})\right) \mathrm{d} \mathbf{x}=f\left(\mathbf{x}_{0}\right)
$$

[HINT: You should change variables $r=\left(x-x_{0}\right) / \epsilon$.]

## Two hours

## THE UNIVERSITY OF MANCHESTER

WAVE MOTION

5th June 2019
09:45-11:45

Answer ALL SIX questions.

Electronic calculators are permitted, provided they cannot store text.

1. You are given that a harmonic signal (in the usual notation) $y(x, t)=A \exp [i(k x-\omega t)]$, satisfies the equation

$$
\frac{\partial y}{\partial t}+a \frac{\partial y}{\partial x}-b \frac{\partial^{3} y}{\partial x^{3}}=0
$$

Here $a, b>0, x$ is a spatial coordinate and $t$ is time.
(i) Obtain the dispersion relation for this problem.
(ii) Define both the phase speed $c$ and group velocity $c_{g}$, and hence determine each in terms of $a, b$ and $k$.
(iii) Sketch the behaviour of $\omega, c$ and $c_{g}$ vs $k>0$. Is this wave system dispersive?
(iv) Which propagates faster, phase or energy?
2. Consider a physical signal $f(x, y, t)=A \exp [i(k x+l y-\omega t)]$, in the usual notation, which satisfies the governing equation

$$
\frac{\partial f}{\partial t}-\beta R^{2} \frac{\partial f}{\partial x}=R^{2} \frac{\partial}{\partial t}\left(\nabla^{2} f\right)
$$

where $R$ and $\beta$ are positive constants and $\nabla^{2}$ is the two-dimensional Laplacian.
(i) Obtain the dispersion relation, and show that it can be written in the form

$$
\left(k-k_{0}\right)^{2}+l^{2}=A,
$$

where $k_{0}$ and $A$ are to be determined in terms of $\beta, R, \omega$.
(ii) For fixed values of $0<\omega<\omega_{c}$ (where $\omega_{c}$ is to be determined in terms of $\beta$ and $R$ ) sketch solutions of the dispersion relation in the $k-l$ wavenumber plane. Indicate a wavenumber vector and a group velocity vector on your sketch.
(iii) Using the sketch, explain how shorter wavelength waves behave differently to those with a longer wavelength.
3. Small-amplitude water waves exist on an inviscid, incompressible fluid layer of depth $h$. The waves (of amplitude $A$ ) have a velocity potential

$$
\phi(x, z, t)=\frac{g A}{\omega} \frac{\cosh (\hat{K}[z-h])}{\cosh \hat{K} h} \cos (\hat{K} x-\omega t),
$$

where $\hat{K} \tanh (\hat{K} h)=\omega^{2} / g$, with $g$ being the acceleration due to gravity. The coordinates $(x, z)$ have $z$ measured downwards from the mean location of the free surface.

In the motion resulting from the potential $\phi$ above, write down the equations of motion for a particle with position $x(t), z(t)$. Find approximations for $x(t)$ and $z(t)$ near a point $\left(x_{0}, z_{0}\right)$ and hence give an equation for the particle paths at leading order. Sketch the shape of the paths for varying $z_{0}$.
[Total: 10 marks]
4. Consider an impermeable sphere centred at the origin, surrounded by an inviscid, compressible fluid of sound speed $c$. The radius $a(t)$ of the sphere changes, and this generates sound waves in the fluid; these waves are to be described in terms of a potential $\phi(r, t)$, where $r$ is the distance from the origin.
(i) State the boundary conditions to be satisfied by $\phi(r, t)$.
(ii) Show that the wave equation (given below) can be satisfied by a velocity potential of

$$
\phi(r, t)=-\frac{1}{4 \pi r} m(t-r / c),
$$

where $m$ is a (twice differentiable) function.
(iii) For small deflections of the sphere radius about a mean value $a_{0}$, obtain an ordinary-differential equation for $m(t)$ in terms of the volume of the sphere.

You may use: The wave equation for a spherically symmetric velocity potential is

$$
\frac{\partial^{2} \phi}{\partial t^{2}}=c^{2} \frac{1}{r^{2}} \frac{\partial}{\partial r}\left(r^{2} \frac{\partial \phi}{\partial r}\right)
$$

5. A wave motion exists in an incompressible, inviscid, fluid (occupying $z>0$ ), covered by an elastic sheet, such that the usual linearised surface conditions are altered to be

$$
\frac{\partial \eta}{\partial t}=\frac{\partial \phi}{\partial z} \quad \text { on } \quad z=0
$$

and

$$
-\rho \frac{\partial \phi}{\partial t}=-\rho g \eta+T \frac{\partial^{2} \eta}{\partial x^{2}}-R \frac{\partial^{4} \eta}{\partial x^{4}} \quad \text { on } \quad z=0
$$

Here $\rho, g, T$ and $R$ are positive constants associated with the density of the fluid, the local gravitational acceleration, tension in the sheet and rigidity of the sheet respectively and $x$ is measured along the interface. In addition, $\phi$ is the velocity potential of the wave and $\eta$ is the free surface deflection caused by the wave.
(i) State the equation to be satisfied by $\phi$ in $z>0$ and the condition on $\phi$ as $z \rightarrow \infty$. Show that the surface conditions above can be reduced to a single boundary condition for $\phi$.
(ii) Use separation of variables to obtain a solution $\phi(x, z, t)$ that is periodic in $x$ and $t$. Hence obtain the dispersion relation for these waves in terms of a frequency $\omega$ and a wavenumber $K$ in the $x$ direction.
[6 marks]
(iii) Show that (for $K>0$ ) the phase speed has a minimum value, and find the value $K=K_{\text {min }}$ associated with this minimum phase speed.
[5 marks]
(iv) Obtain an expression for the ratio $c_{g} / c$ in terms of $T, R, g, \rho, K$. Evaluate this expression for $K=K_{\text {min }}, K \rightarrow 0, K \rightarrow \infty$, and interpret the results.
6. A compressible fluid of sound speed $c$ and mean density $\rho_{0}$ fills a pipe with rigid impermeable walls. A Cartesian coordinate system is assumed, such that $x$ is measured along the axis of the pipe.
(i) For linear acoustic waves in the pipe, state the equation to be satisfied in the fluid and the boundary condition to be imposed at the walls for a velocity potential $\phi$. Hence verify that, for a pipe with a constant cross section, a solution for the velocity potential exists in the form

$$
\phi(x, t)=B \exp \{i \omega(t-x / c)\}
$$

for any constant $B$, where $\omega$ is a frequency, $t$ is time.
(ii) If the pipe has cross-sectional area $A$ then give the mass flux $Q$ down the pipe in terms of $\rho_{0}, A$ and $\phi$.
(iii) You are given that the pressure perturbation of the acoustic field is

$$
p=-\rho_{0} \frac{\partial \phi}{\partial t} .
$$

Now consider a pipe that has a piecewise constant cross sectional area, being $A_{1}$ for $x \leq 0, A_{2}$ for $x>0$, with a velocity potential of the form

$$
\phi(x, t)=\phi_{I}(x, t)+\phi_{R}(x, t),
$$

for $x \leq 0$ and

$$
\phi(x, t)=\phi_{T}(x, t),
$$

for $x>0$.
For an incoming wave $\phi_{I}(x, t)=B_{I} \exp \{i \omega(t-x / c)\}$, give appropriate expressions for $\phi_{T}$ and $\phi_{R}$. Hence determine the amplitudes of $\phi_{T}$ and $\phi_{R}$ relative to $\phi_{I}$, given that $p$ and $Q$ must be continuous at $x=0$.
Evaluate $\left|\phi_{T}\right| /\left|\phi_{I}\right|$ for the particular case of $A_{1}=2 A_{2}$.
[10 marks]
(iv) Interpret your results in the special cases $A_{2} \ll A_{1}$ and $A_{2} \approx A_{1}$.

## Two hours

## THE UNIVERSITY OF MANCHESTER

## MATHEMATICAL BIOLOGY

21 May 2019
14.00-16.00

Answer ALL three questions in Section A (30 marks in total).
Answer TWO of the THREE questions in Section B (50 marks in total).
If more than TWO questions from Section B are attempted,
then credit will be given for the best TWO answers.

Electronic calculators may be used in accordance with the University regulations
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## SECTION A

## Answer ALL three questions

A1. A fish population of size $N(t)$ that is subject to harvesting at rate $c>0$ is described by the ordinary differential equation

$$
\begin{equation*}
\frac{\mathrm{d} N}{\mathrm{~d} t}=r N\left(1-\frac{N}{K}\right)-c . \tag{1}
\end{equation*}
$$

(a) Give a physical interpretation of the parameters $r$ and $K$.
(b) Writing $N=K x(\tau)$ and $t=\tau / r$, re-write (1) in the form $\mathrm{d} x / \mathrm{d} \tau=f(x)$, where $f$ includes the parameter $\alpha=c / r K$.
(c) By considering the form of the function $f(x)$, find the largest value of $\alpha$ for which a stable fish population can be maintained.

A2. A gene with expression level $X(t)$ that represses its own transcription is described by the delaydifferential equation

$$
\frac{\mathrm{d} X}{\mathrm{~d} t}=\frac{1}{1+X(t-T)}-\frac{1}{1+X_{0}}
$$

where $X_{0}$ and $T$ are positive constants.
(a) By setting $X(t)=X_{0}+z(t)$, where $|z| \ll X_{0}$, derive the linear delay-differential equation

$$
\frac{\mathrm{d} z}{\mathrm{~d} t}=-\frac{z(t-T)}{\left(1+X_{0}\right)^{2}}
$$

(b) Assuming $z(t)=a e^{\lambda t}$ for some constants $a$ and $\lambda$ (which may be complex), derive a condition that must be satisfied by $\lambda$.
(c) Find the smallest value of $T$ for which $z$ exhibits neutrally stable oscillations (for which $\lambda$ is purely imaginary).
(d) Show that the resulting oscillations have period $4 T$.

A3. A model for an animal population migrating steadily across a spatial domain is given by

$$
\begin{equation*}
-c u_{z}=D u_{z z}+r u(1-u) \tag{2}
\end{equation*}
$$

Here $c>0$ is a migration speed, $D$ is a diffusion coefficient, and $u(z) \geq 0$ measures population size (relative to a reference value) as a function of distance $z$.
(a) To describe the solution where $u \ll 1$, we may replace $(1-u)$ in (2) with 1 . Seek a solution of the resulting equation of the form $u=A e^{\lambda z}$ for some constant $A$, subject to the condition that $u \rightarrow 0$ as $z \rightarrow \infty$. You may assume that $\lambda$ is real. Show that $c$ cannot be less than $2 \sqrt{D r}$.
[4 marks]
(b) Suppose now that $c=2 \sqrt{D r}$ and find $\lambda$. Give a physical interpretation of the quantity $\sqrt{D / r}$.
[3 marks]
(c) For this value of $c$, and writing $u=1+\hat{u}$, linearise (2) assuming $|\hat{u}| \ll 1$. Setting $\hat{u}=B e^{\mu z}$ for some constant $B$, and assuming $\hat{u} \rightarrow 0$ as $z \rightarrow-\infty$, show that $\mu=\alpha \sqrt{r / D}$ where $\alpha$ is a constant to be determined.

## SECTION B

## Answer TWO of the three questions

B4. An infectious disease spreads through a population of individuals. The population may be divided into those who have not yet caught the disease (i.e. are susceptible, $S(t)$ ), those who are infected $(I(T))$ and those who have recovered $(R(t))$. These three sub-populations evolve according to

$$
\frac{\mathrm{d} S}{\mathrm{~d} t}=-\beta I S, \quad \frac{\mathrm{~d} I}{\mathrm{~d} t}=\beta I S-\gamma I, \quad \frac{\mathrm{~d} R}{\mathrm{~d} t}=\gamma I
$$

(a) Let $S+I+R=N$. Show that $N$ is constant during the course of the disease outbreak. [2 marks]
(b) Writing $S(t)=N u(\tau), I(t)=N v(\tau)$, where $t=\tau / \gamma$, derive the simplified system

$$
\frac{\mathrm{d} u}{\mathrm{~d} \tau}=-R_{0} u v, \quad \frac{\mathrm{~d} v}{\mathrm{~d} \tau}=v\left(R_{0} u-1\right)
$$

where $R_{0}$ is a dimensionless constant (to be determined) that represents the virulence of the disease.
(c) You may assume that $u(0)=1, v(0)=0$ and $R_{0}>1$. By constructing $\mathrm{d} v / \mathrm{d} u$, or otherwise, show that the disease evolves under the constraint

$$
\begin{equation*}
v=1-u+\frac{\log u}{R_{0}} \tag{3}
\end{equation*}
$$

Plot graphs of $u-1$ and $(\log u) / R_{0}$ to demonstrate that (3) has a solution $u=u_{\star}, v=0$ satisfying $0<u_{\star}<1 / R_{0}$.
(d) Use (3) to describe the evolution of the outbreak, identifying in particular the largest proportion $v_{\max }$ of the population that is infected at any time. You may find it helpful to sketch a trajectory in the $(u, v)$-plane. What does the quantity $N\left(1-u_{\star}\right)$ represent?
(e) Show that $v_{\text {max }}$ is an increasing function of $R_{0}$.
(f) Returning to the original formulation of the problem, imagine that recovered individuals slowly lose their immunity to the disease, and so return to the susceptible population at rate $\nu$ (where $\nu$ has dimensions of $1 /$ time ). Write down the modified model for the evolution of $S(t), I(t)$ and $R(t)$.

B5. An enzyme $E$ catalyses the transformation of a substrate $S$ to a product $P$ through the reaction

$$
S+E \underset{k_{-1}}{\stackrel{k_{1}}{\rightleftharpoons}} C_{1} \xrightarrow{k_{2}} P+E
$$

A drug $D$ is added to the system which binds competitively to the enzyme according to

$$
D+E \stackrel{k_{3}}{\underset{k_{-3}}{\rightleftharpoons}} C_{2}
$$

Here $k_{1}, k_{-1}, k_{2}, k_{3}$ and $k_{-3}$ are reaction rates and $C_{1}$ and $C_{2}$ are different complexes. Initially, concentrations of $S, E, C_{1}, C_{2}, D, P$ are $S_{0}, E_{0}, 0,0, D_{0}, 0$ respectively.
(a) Using the principles of mass action kinetics, write down the six ordinary differential equations (ODEs) describing the evolution of the concentrations of six chemical species $\left(S, E, C_{1}, C_{2}, D\right.$ and $P$ ) in terms of five different reactions.
(b) Write down the stoichiometric matrix for this system. Hence, or otherwise, identify three conserved quantities, and reduce the system to

$$
\begin{aligned}
\dot{C}_{1} & =k_{1}\left(S_{0}-C_{1}-P\right)\left(E_{0}-C_{1}-C_{2}\right)-k_{2} C_{1}-k_{-1} C_{1} \\
\dot{C}_{2} & =k_{3}\left(D_{0}-C_{2}\right)\left(E_{0}-C_{1}-C_{2}\right)-k_{-3} C_{2} \\
\dot{P} & =k_{2} C_{1} .
\end{aligned}
$$

(c) Assuming that $C_{1}$ and $C_{2}$ are close to equilibrium, and that $S$ and $D$ are sufficiently abundant for their concentrations to be assumed constant (i.e. $S_{0} \gg C_{1}+P$ and $D_{0} \gg C_{2}$ ), show that

$$
\begin{gathered}
S_{0}\left(E_{0}-C_{1}-C_{2}\right)=L C_{1}, \\
D_{0}\left(E_{0}-C_{1}-C_{2}\right)=K C_{2},
\end{gathered}
$$

where $K=k_{-3} / k_{3}$ and $L=\left(k_{-1}+k_{2}\right) / k_{1}$. Hence derive the equation

$$
\frac{\mathrm{d} P}{\mathrm{~d} t}=\frac{k_{2} S_{0} E_{0}}{S_{0}+L+\left(D_{0} L / K\right)}
$$

(d) Briefly discuss the impact of the drug on the yield of product, making comparison with the case $D_{0}=0$.

B6. In a simple formulation of Turing's reaction-diffusion mechanism for morphogenesis, $X(\theta, t)$ and $Y(\theta, t)$ are the concentrations of two morphogens that interact via

$$
\begin{equation*}
\frac{\partial X}{\partial t}=f(X, Y)+\mu \frac{\partial^{2} X}{\partial \theta^{2}}, \quad \frac{\partial Y}{\partial t}=g(X, Y)+\nu \frac{\partial^{2} Y}{\partial \theta^{2}} \tag{4}
\end{equation*}
$$

where $\mu$ and $\nu$ are positive real numbers. $X$ and $Y$ are defined over the spatial domain $0 \leq \theta \leq 2 \pi$ and satisfy periodic boundary conditions, so that

$$
X(0, t)=X(2 \pi, t), \quad Y(0, t)=Y(2 \pi, t)
$$

You may assume there are numbers $X^{\star}$ and $Y^{\star}$ such that $f\left(X^{\star}, Y^{\star}\right)=g\left(X^{\star}, Y^{\star}\right)=0$.
(a) Show that the spatially uniform solution $X(\theta, t)=X^{\star}, Y(\theta, t)=Y^{\star}$ satisfies (4).
(b) Consider solutions of the form $X(\theta, t)=X_{\star}+x(\theta, t)$ and $Y(\theta, t)=Y_{\star}+y(\theta, t)$, where $x$ and $y$ are assumed to be small in magnitude. Linearise (4a,b) to obtain equations of the form

$$
\begin{equation*}
\frac{\partial x}{\partial t}=a x+b y+\mu \frac{\partial^{2} x}{\partial \theta^{2}}, \quad \frac{\partial y}{\partial t}=c x+d y+\nu \frac{\partial^{2} y}{\partial \theta^{2}} \tag{5}
\end{equation*}
$$

Explain how the parameters $a, b, c$ and $d$ appearing in (5) relate to the functions $f$ and $g$. [8 marks]
(c) Consider solutions to (5) of the form $x(\theta, t)=u(t) \cos (n \theta)$ and $y(\theta, t)=v(t) \cos (n \theta)$ with $n$ a positive integer. Show that the functions $v(t)$ and $u(t)$ evolve according to

$$
\frac{\mathrm{d} u}{\mathrm{~d} t}=\left(a-\mu n^{2}\right) u+b v, \quad \frac{\mathrm{~d} v}{\mathrm{~d} t}=c u+\left(d-\nu n^{2}\right) v
$$

(d) Give two conditions on $a, b, c$ and $d$ that guarantee that the spatially uniform solution is stable when $n=0$.
(e) Assuming the conditions in (d) hold, show that stability is guaranteed for all $n>0$ provided

$$
a d-b c>\frac{\mu \nu}{4}\left(\frac{a}{\mu}+\frac{d}{\nu}\right)^{2}
$$

## END OF EXAMINATION PAPER

## Two hours

# THE UNIVERSITY OF MANCHESTER 

SYMMETRY IN GEOMETRY AND NATURE

16 May 2019
14.00-16.00

## Answer ALL FIVE questions

[^0]1.
(a) Name the symmetry groups of the following three planar shapes, and state in each case the order of the group and whether the group is Abelian.

[6 marks]
(b) Given a function $f: \mathbb{R} \rightarrow \mathbb{R}$, its extended symmetry group consists of all pairs $(T, \sigma)$ for which $T$ is a transformation of $\mathbb{R}$ and $\sigma \in \mathbb{Z}_{2}=\{ \pm 1\}$ satisfying $f(T x)=\sigma f(x)$ (in particular, $f(T x)= \pm f(x)$ ).
Describe the extended symmetry group of each of the functions $f$ and $g$ of one variable whose graphs are shown below. (It is sufficient to give generators of each of the groups.)

$y=f(x)$

$y=g(x)$
2. Suppose a finite group $G$ acts on a set $X$.
(a) Define the orbit $G \cdot x$ and stabilizer $G_{x}$ of an element $x \in X$. Show that the stabilizer is always a subgroup of $G$.
(b) State the orbit-stabilizer theorem.
(c) Let $h \in G$ and $y=h \cdot x$. Show that the subset
$$
G_{x, y}=\{g \in G \mid g \cdot x=y\}
$$
is a left coset of $G_{x}$, and explain briefly how this can be used to prove the orbit-stabilizer theorem (details of the proof are not required).
(d) The 'octahedral group' $\mathbb{O}$ of order 24 acts by rotations on the cube. On each face of the cube the 2 diagonals are drawn. Use the orbit-stabilizer theorem to find the order of the stabilizer of one of these diagonals.
3.
(a) Define the action on $\mathbb{R}^{2}$ of an element $(A \mid \mathbf{v}) \in \mathrm{E}(2)$, where $A \in \mathrm{O}(2)$ and $\mathbf{v} \in \mathbb{R}^{2}$. Show that multiplication in the group is given by
$$
(A \mid \mathbf{v})(B \mid \mathbf{u})=(A B \mid \mathbf{v}+A \mathbf{u})
$$
and deduce the Seitz symbol of the inverse transformation $(A \mid \mathbf{v})^{-1}$.
(b) Consider the following lattice in the plane:
$$
L=\left\{(x, y) \in \mathbb{R}^{2} \mid x \in \mathbb{Z}, y \in \mathbb{Z}, x+y \in 3 \mathbb{Z}\right\}
$$
(i) Show that the two vectors $\mathbf{u}_{1}=(3,0)$ and $\mathbf{u}_{2}=(0,3)$ both belong to $L$. Sketch a diagram showing all the points of $L$ that lie in the rectangle $-3 \leq x \leq 3$ and $-3 \leq y \leq 3$ and find a point in $L$ which is not contained in $\mathbb{Z}\left\{\mathbf{u}_{1}, \mathbf{u}_{2}\right\}$.
(ii) Find two vectors $\mathbf{a}$ and $\mathbf{b}$, as required for the classification of symmetry types of lattices; that is, such that $L=\mathbb{Z}\{\mathbf{a}, \mathbf{b}\}$, and $|\mathbf{a}| \leq|\mathbf{b}| \leq|\mathbf{b}-\mathbf{a}| \leq|\mathbf{b}+\mathbf{a}|$, proving carefully that your choice satisfies this.
(iii) Identify which of the 5 types of lattice $L$ is (Oblique, Rectangular, Centred rectangular, Square or Hexagonal).
[16 marks]
4. Consider a system of 3 coupled cells governed by the system of differential equations
\[

\left\{$$
\begin{array}{l}
\dot{x}=x-y z  \tag{*}\\
\dot{y}=y-x z \\
\dot{z}=z-x y
\end{array}
$$\right.
\]

(a) Show that the system $(*)$ has $S_{3}$ symmetry, and draw a cell diagram illustrating the system.
(b) Deduce that the diagonal, where $x=y=z$, is invariant under the evolution of the system, stating carefully any results used.
(c) Find all the equilibrium points of the system. Divide them into group orbits and find the Burnside type of the set of equilibrium points.
(d) Let $\mathbf{x}(t)$ be the solution to this system with initial value $(x, y, z)=\left(\frac{1}{2}, \frac{1}{2}, \frac{1}{2}\right)$. Find which of the equilibria is equal to $\lim _{t \rightarrow \infty} \mathbf{x}(t)$. [You are not expected to solve the differential equation.]
[4 marks]
5.
(a) Define the symmetry group of a periodic solution $\gamma(t)$ with fundamental period $T$ of a differential equation with symmetry group $G$.
(b) Suppose a mechanical system in the plane has potential energy, with $\mathrm{D}_{2}$ symmetry; that is $V(x, y)=$ $V(-x, y)=V(x,-y)$. Let $\gamma$ be a periodic orbit with period $T$ and symmetry group $\widetilde{C_{2}}$, and let $\delta(t)$ be a periodic solution of the same period and with symmetry $\widetilde{D_{1}}$. Write down in each case what the symmetry group tells us about $x(t)$ and $y(t)$ for each periodic orbit. Below is shown numerical data taken from two experiments with this symmetry showing both $x(t)$ and $y(t)$. Determine which solution is $\gamma(t)$ and which is $\delta(t)$, giving a brief justification of your answer.



## TWO Hours

## THE UNIVERSITY OF MANCHESTER

## NUMERICAL ANALYSIS 2

3rd June 2019
09:45-11:45

Answer ALL questions in SECTION A (40 marks in total)
and
Answer TWO of the three questions in SECTION B (20 marks each).
If more than TWO questions from Section B are attempted, then credit will be given for the best TWO answers.

Electronic calculators may be used in accordance with the University regulations
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## SECTION A

## Answer ALL five questions

## A1.

(a) Let $w(x)$ be a given weight function. Define the $L_{2, w}$ norm of a function $f \in C[a, b]$, denoted by $\|f\|_{2, w}$. State any properties that $w(x)$ must satisfy. In addition, explain what is meant by the phrase $g$ is a "best weighted $L_{2}$ approximation" to $f$ from a subspace $G \subset C[a, b]$.
(b) Suppose we want to find an approximation to $f \in C[a, b]$ of the form

$$
g(x)=\sum_{i=0}^{n} c_{i} \phi_{i}(x),
$$

where $\left\{\phi_{i}(x)\right\}_{i=0}^{n}$ are linearly independent. Define the normal equations that can be solved to find the best weighted $L_{2}$ approximation. Explain why it is advantageous to use orthogonal polynomials $\phi_{i}$ (where $\phi_{i}$ has degree $i$ ) and in that case, state the solution $c_{i}, i=0,1, \ldots, n$.

A2. Recall that the Chebyshev polynomial of degree $n$ is defined by $T_{n}(x)=\cos \left(n \cos ^{-1}(x)\right)$ for $x \in[-1,1]$ and $n=0,1,2, \ldots$.

Define the set of $n+1$ Chebyshev points $x_{i}, i=0,1, \ldots, n$ associated with the Chebyshev polynomial $T_{n+1}(x)$ and briefly explain why it is beneficial to use Chebyshev points as interpolation points when constructing polynomial approximations $p_{n}$ of degree $\leq n$ to a function $f \in C[-1,1]$.

A3. Divide $[0,1]$ into $n$ equal intervals $\left[x_{i}, x_{i+1}\right]$ of width $h$, where $x_{i}=i h, i=0,1, \ldots, n$ and consider the repeated trapezium rule

$$
I(f)=\int_{0}^{1} f(x) d x \approx \sum_{i=0}^{n-1} \frac{h}{2}\left(f\left(x_{i}\right)+f\left(x_{i+1}\right)\right)=T(h) .
$$

Assuming that $f$ is sufficiently differentiable, the Euler-Maclaurin summation formula gives an asymptotic expansion for the error of the form

$$
I(f)-T(h)=a_{2} h^{2}+a_{4} h^{4}+a_{6} h^{6}+\ldots
$$

for some real constants $a_{2}, a_{4}, a_{6}, \ldots$.
(a) Use the Euler-Maclaurin formula to determine what happens to the error when we double the number of intervals $\left[x_{i}, x_{i+1}\right]$.
(b) By taking an appropriate linear combination of the rules $T(h)$ and $T\left(\frac{h}{2}\right)$, derive a new quadrature scheme that gives an error that is $\mathcal{O}\left(h^{4}\right)$.

A4. Suppose that we approximate $I(f)=\int_{0}^{1} f(x) d x$ using the following three-point quadrature rules

$$
\frac{1}{6}(f(0)+4 f(1 / 2)+f(1)), \quad \frac{1}{3}(2 f(1 / 4)-4 f(1 / 2)+2 f(3 / 4)),
$$

and suppose that small errors are made in evaluating $f$ at the chosen quadrature points $x_{i}$. Let $\epsilon_{i}$ be the error in $f\left(x_{i}\right)$ and suppose $\left|\epsilon_{i}\right| \leq 10^{-4}$. Briefly analyse the effect of this data error for both quadrature schemes. Why are negative weights not desirable?

A5. The $\ell$-step Adams linear multistep methods for the initial value problem $y^{\prime}=f(x, y), x \geq x_{0}$, $y\left(x_{0}\right)=y_{0}$ have the form

$$
y_{n+1}=\sum_{i=1}^{\ell} a_{i} y_{n+1-i}+h \sum_{i=0}^{\ell} b_{i} f_{n+1-i},
$$

where $a_{i}, b_{i}$ are given constants, $h$ is the step size, and $f_{n}=f\left(x_{n}, y_{n}\right)$.
(a) Explain how we can derive the $\ell$-step Adams-Bashforth method and the $\ell$-step AdamsMoulton method by replacing the integrand in

$$
y\left(x_{n+1}\right)=y\left(x_{n}\right)+\int_{x_{n}}^{x_{n+1}} f(x, y(x)) d x
$$

with a suitable polynomial. Briefly comment on the accuracy of the two $\ell$-step methods and their computational costs.
(b) By introducing the appropriate polynomial, derive the 1-step Adams-Bashforth method. By what name is this scheme better known?

## SECTION B

Answer TWO of the three questions

B6.
(a) State two reasons why Padé approximants offer more flexibility than single polynomials for approximating functions. In addition, explain briefly why it is beneficial to consider continued fraction representations of Padé approximants.
(b) Consider

$$
r_{n}(x)=\frac{c_{1} x}{1+\frac{c_{2} x}{1+\frac{c_{3} x}{1+\cdots+\frac{c_{n-2} x}{1+\frac{c_{n-1} x}{1+c_{n} x}}}}}
$$

with coefficients

$$
c_{1}=1, \quad c_{2 j}=\frac{j}{2(2 j-1)}, \quad c_{2 j+1}=\frac{j}{2(2 j+1)}, \quad j=1,2, \ldots
$$

Show that

$$
r_{3}(x)=\frac{x+x^{2} / 6}{1+2 x / 3}
$$

and hence that $r_{3}(x)$ is the [2/1] Padé approximant of $f(x)=\log (1+x)$.
(c) The Top-Down algorithm for evaluating the continued fraction

$$
\begin{equation*}
r_{n}(x)=b_{0}+\frac{a_{1} x}{b_{1}+\frac{a_{2} x}{b_{2}+\frac{a_{3} x}{b_{3}+\cdots+\frac{a_{n-1} x}{b_{n-1}+\frac{a_{n} x}{b_{n}}}}}} \tag{1}
\end{equation*}
$$

is given by

$$
\begin{array}{ll}
1 & p_{-1}=1, q_{-1}=0, p_{0}=b_{0}, q_{0}=1 \\
2 & \text { for } j=1: n \\
3 & p_{j}=b_{j} p_{j-1}+a_{j} x p_{j-2} \\
4 & q_{j}=b_{j} q_{j-1}+a_{j} x q_{j-2} \\
5 & \text { end } \\
6 & r_{n}=p_{n} q_{n}^{-1}
\end{array}
$$

Prove by induction that $p_{n} / q_{n}$, where $p_{n}, q_{n}$ are computed via the Top-Down algorithm, is equal to $r_{n}(x)$ defined in (1).

B7. Consider the integral

$$
I(f)=\int_{a}^{b} w(x) f(x) d x
$$

where $w(x)$ is a valid weight function and let $\left\{\phi_{i}(x)\right\}$ be the system of polynomials that are orthogonal with respect to $w(x)$ on $[a, b]$.
(a) Using the idea of interpolatory quadrature, derive an expression for the weights $w_{i}$ in the $n$-point rule

$$
I(f) \approx G_{n}(f)=\sum_{i=1}^{n} w_{i} f\left(x_{i}\right)
$$

that ensures the rule is exact for all polynomials of degree $\leq n-1$.
(b) If the weights $w_{i}$ are chosen as in part (a) and the nodes $x_{1}, \ldots, x_{n}$ are chosen as the zeros of the polynomial $\phi_{n}(x)$ of degree $n$, prove that the resulting Gauss rule $G_{n}(f)$ is in fact exact for polynomials of degree $\leq 2 n-1$.
(c) Derive the two point Gauss-Hermite rule

$$
\int_{-\infty}^{\infty} e^{-x^{2}} f(x) d x \approx w_{1} f\left(x_{1}\right)+w_{2} f\left(x_{2}\right)
$$

using the method of undetermined coefficients. You may use the results

$$
\int_{-\infty}^{\infty} e^{-x^{2}} d x=\sqrt{\pi}, \quad \int_{-\infty}^{\infty} x^{2} e^{-x^{2}} d x=\frac{1}{2} \sqrt{\pi}
$$

without proof.
HINT: You may find it useful to define a quadratic polynomial $\Phi(x)=x^{2}+a x+b$ whose roots coincide with $x_{1}$ and $x_{2}$, and find $a$ and $b$.

B8. Consider the numerical method

$$
y_{n+1}=y_{n}+\frac{h}{2}\left(f\left(x_{n}, y_{n}\right)+f\left(x_{n}+h, y_{n}+h f\left(x_{n}, y_{n}\right)\right)\right),
$$

for the initial value problem $y^{\prime}(x)=f(x, y(x)), x \geq x_{0}$ with $y\left(x_{0}\right)=y_{0}$, where $h$ is the chosen stepsize and $y_{n}$ is an approximation to $y\left(x_{n}\right)$ at the grid point $x_{n}$ for $n=0,1, \ldots$.
(a) Explain how the method in ( $\star$ ) can be interpreted as:
(1) an 'improved' Euler scheme,
(2) a 2-stage Runge-Kutta scheme.
(b) Define the terms local truncation error (denoted $\tau(h)$ ) and order of a numerical method for initial value problems.
(c) Show that for the numerical method ( $\star$ ),

$$
\tau(h)=y\left(x_{n+1}\right)-y\left(x_{n}\right)-\frac{h}{2} y^{\prime}\left(x_{n}\right)-\frac{h}{2} f\left(x_{n}+h, y\left(x_{n}\right)+h y^{\prime}\left(x_{n}\right)\right) .
$$

(d) Using part (c), show that the numerical method in ( $\star$ ) has order two.

HINT: you may find the following identity useful

$$
\frac{d^{2} y}{d x^{2}}=\frac{\partial f}{\partial x}+\frac{\partial f}{\partial y} f .
$$

## Two hours

## THE UNIVERSITY OF MANCHESTER

MARKOV PROCESSES

03 June 2019
9.45-11.45

Answer ALL four questions.

The use of electronic calculators is prohibited.

1 of 5
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1. a)
(i) Define the terms "recurrent state" and "transient state" in the context of discrete time Markov Chains.
(ii) Define the stationary distribution of a discrete time Markov Chain.
(iii) Using a transition diagram, give an example of a Markov Chain with two transient states, six recurrent states and $i, j \in S$ such that

$$
P\left(X_{n}=j \text { for some } n \geq 1 \mid X_{0}=i\right)=\frac{2}{3} .
$$

(iv) Give an example of a four state Markov Chain such that, $\forall i, j \in S$,

$$
\lim _{n \rightarrow \infty} p_{i j}(n) \text { does not exist. }
$$

(v) Give an example of a Markov Chain which has more than one stationary distribution.
[2 marks]
b) Consider the discrete time Markov Chain $\left\{X_{n}: n \geq 0\right\}$ with state space $\{1,2,3,4,5\}$ and transition probability matrix

$$
\left[\begin{array}{lllll}
0.4 & 0 & 0 & 0.6 & 0 \\
0.3 & 0 & 0 & 0 & 0.7 \\
0.5 & 0.1 & 0.4 & 0 & 0 \\
0.4 & 0 & 0 & 0.6 & 0 \\
0 & 0.7 & 0 & 0 & 0.3
\end{array}\right]
$$

(i) Show that state 5 is transient.
(ii) Determine a stationary distribution for the Markov chain.
(iii) Carefully stating any results to which you appeal, determine $\lim _{n \rightarrow \infty} p_{11}(n)$ and $\lim _{n \rightarrow \infty} p_{52}(n)$.
(iv) Write down the mean recurrence time of state 1 .
2. Let $\left\{X_{n}, n \geq 0\right\}$ be a Markov chain, with state space the set of non-negative integers and transition probabilities

$$
\begin{aligned}
p_{0,0} & =p \\
p_{i, i+1} & =q, \quad 0 \leq i \\
p_{i, i-1} & =p, \quad 1 \leq i .
\end{aligned}
$$

where $0<p<1$ and $q=1-p$.
(i) Determine a stationary distribution for this Markov chain, stating the range of $p$ for which a stationary distribution exists.
(ii) Giving a reason, state the range of $p$ for which the Markov chain is positive recurrent. For $p$ in this range, write down the mean recurrence time of state 0 .
(iii) Prove that, if a Markov chain with transition matrix $P$ has stationary distribution $\underline{\pi}$, then so does a Markov chain with transition matrix $\frac{1}{2}(I+P)$.
(iv) Let $\left\{Y_{n}, n \geq 0\right\}$ be a Markov chain, with state space the set of non-negative integers and transition probabilities

$$
\begin{aligned}
p_{0,0} & =\frac{5}{6} \\
p_{i, i+1} & =\frac{1}{6}, \quad 0 \leq i \\
p_{i, i-1} & =\frac{1}{3}, \quad 1 \leq i \\
p_{i, i} & =\frac{1}{2}, \quad 1 \leq i .
\end{aligned}
$$

Using the result of part (iii) or otherwise, determine a stationary distribution for this Markov chain.
(v) For the Markov chain of part (iv), determine the mean recurrence time of state zero and hence, or otherwise, the expected time of first visit to state zero starting at state 1.
3. a) At time 0 a population comprises 1 individual. Lifetimes of individuals are independent and each is exponentially distributed with parameter $\lambda$. At death each individual splits into 2 new individuals. Let $X(t)$ denote the number of individuals alive in the population at time $t$.
(i) With careful reference to the lack of memory property of the exponential distribution, and to the distribution of the minimum of a number of independent exponential random variables, derive the generator matrix $Q$ for this process.
(ii) For $X(t): t \geq 0$, write down the forward equations in matrix form and show they include

$$
p_{1, k}^{\prime}(t)=(k-1) \lambda p_{1, k-1}(t)-k \lambda p_{1, k}(t) \quad 1 \leq k
$$

b) A telephone network connects 4 subscribers. When not engaged in a conversation, each makes calls in a Poisson process with rate $\lambda$ independently of the others. The recipient of a call is equally likely to be any of the other 3 subscribers. If that recipient is not engaged, a conversation begins; otherwise the call is lost. Call durations are independently exponentially distributed with parameter $\mu$.

Let the Markov chain $X(t): t \geq 0$, with state space $S=\{0,1,2\}$, denote the number of conversations in progress at time $t$.
(i) Stating any results to which you appeal, derive the generator matrix $Q$ for $X(t)$ and calculate its stationary distribution.
(ii) Determine the expected time of first visit to state 2 , starting from state 0 .
4. a) Consider a Markov queue where customers arrive in a Poisson process, rate $\lambda$, each customer is served immediately on arrival (effectively, there is an unlimited number of servers), and service times are independent, exponentially distributed with parameter $\mu$, and independent of the arrival process.
(i) Carefully stating any results to which you appeal, determine the generator matrix $Q$.
(ii) Determine the stationary distribution of the number of customers in the system and write down its expected value.
b) Let the Markov chain $X(t): t \geq 0$, with state space $S=\{0,1,2 \ldots m\}$, denote the number of items in a storage facility. When the facility contains at least one item, the intervals between removals are exponentially distributed with parameter $\lambda$. Following the removal of the final item, the facility is replenished with $m$ items after an exponential, parameter $\mu$, distributed interval. All the exponential random variables are independent.
(i) Determine the stationary distribution of the Markov chain.
(ii) Determine the expected time taken for the facility to first empty following replenishment.
(iii) The Markov chain is observed when stationary. Determine the expected time elapsed before the facility is first empty.

## Statistical Tables to be provided

Two hours

THE UNIVERSITY OF MANCHESTER

Time Series Analysis

30 May 2019
09.45-11.45

Electronic calculators are permitted, provided they cannot store text.

Answer ALL Four questions in Section A (32 marks in all)
and

ALL Two questions in Section B (24 marks each).

Where verbal answers are required, they should be concise and clear-typically, a single phrase or sentence for each worthy thought.
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## SECTION A <br> Answer ALL four questions

A1. Let $\left\{X_{t}\right\}$ be a stationary time series, such that

$$
X_{t}=\phi_{1} X_{t-1}+\varepsilon_{t},
$$

where $\left|\phi_{1}\right|<1,\left\{\varepsilon_{t}\right\} \sim \mathrm{WN}\left(0, \sigma^{2}\right)$, and $\mathrm{E} \varepsilon_{t} X_{t-k}=0$ for $k>0$. Let $\left\{u_{t}\right\}$ be white noise, $\mathrm{WN}\left(0, \sigma_{u}^{2}\right)$, independent of $\left\{X_{t}\right\}$ and $\left\{\varepsilon_{t}\right\}$. Consider the time series $\left\{Y_{t}\right\}$, such that

$$
Y_{t}=X_{t}+u_{t} .
$$

Find the autocovariance function of $\left\{Y_{t}\right\}$ for all lags (give the expressions in terms of $\phi_{1}$ and the variances of the white noise sequences).

A2. The graphs below show the sample autocorrelations and sample partial autocorrelations of an observed time series, $\left\{x_{t}\right\}$. Based on these graphs, what models would you consider for $\left\{x_{t}\right\}$ ? Which model would be your first choice? (Justify your answers concisely.)

## Series $x_{t}$



Series $x_{t}$

[8 marks]

## A3.

a) (2 marks) The sample autocorrelations, $\hat{\rho}_{k}$, computed from a sequence of independent identically distributed random variables of length $n$, are approximately independent and have normal (Gaussian) distributions with mean zero and variance $1 / n$. Consider the statistic

$$
\mathrm{BP}=n \sum_{k=1}^{h} \hat{\rho}_{k}^{2} .
$$

Give an informal argument that the distribution of the statistic BP is approximately $\chi_{h}^{2}$.
b) ( 6 marks) An ARMA(2,1) model was fitted to an observed time series, $\left\{x_{t}\right\}, t=1, \ldots, 200$. As part of the diagnostic checking of the fitted model, the values of the Ljung-Box statistic,

$$
\begin{equation*}
Q_{L B}(h)=n(n+2) \sum_{k=1}^{h} \frac{\hat{\rho}_{k}^{2}}{n-k}, \tag{1}
\end{equation*}
$$

were computed for $h=1, \ldots, 8$. The table below gives the values of the statistics $Q_{L B}(h)$ and the corresponding p -values.

|  | 1 | 2 | 3 | 4 | 5 | 6 | 7 | 8 |
| ---: | ---: | ---: | ---: | ---: | ---: | ---: | ---: | ---: |
| X-squared | 0.0131 | 0.0736 | 0.0736 | 0.1334 | 1.0152 | 1.1412 | 4.1835 | 4.2584 |
| p.value | 0.9090 | 0.9639 | 0.9948 | 0.7150 |  | 0.7671 | 0.3817 | 0.5128 |

Complete the table by computing the omitted p-value for $h=5$. What do $\hat{\rho}_{k}, k=1, \ldots, h$ stand for in Equation (1)? What can you infer about the quality of the fitted model?

A4. A seasonal ARIMA model was fitted to a monthly time series. A summary of the fitted model and some residual diagnostics are given below.
Write down the mathematical equation of the fitted model in operator form. Is there evidence that the model is not adequate? If so, suggest a modification of the model that may lead to a better fit.

Call:
$\operatorname{arima}(\mathrm{x}=\mathrm{y}$, order $=c(2,1,1)$, seasonal $=$ list (order $=c(0,1,0)$, period $=12)$ )
Coefficients:

|  | $\operatorname{ar1}$ | $\operatorname{ar2}$ | ma1 |
| :--- | ---: | ---: | ---: |
|  | 0.8498 | -0.4475 | 0.5438 |
| s.e. | 0.0508 | 0.0482 | 0.0466 |

sigma^2 estimated as 1.582: log likelihood $=-803.8$, aic $=1615.61$

## Standardized Residuals



## ACF of Residuals


p values for Ljung-Box statistic


## SECTION B <br> Answer ALL two questions

B5.
a) (6 marks) Consider the time series $\left\{X_{t}\right\}, t \geq 0$, defined by the equations

$$
\begin{aligned}
X_{0} & =0 \\
X_{t} & =0.5 X_{t-1}+\varepsilon_{t}, \quad t \geq 1,
\end{aligned}
$$

where $\left\{\varepsilon_{t}\right\} \sim \mathrm{WN}(0,1)$.
Check if the time series $\left\{X_{t}\right\}, t \geq 0$, is stationary and report your conclusion. (No marks will be awarded for unsubstantiated yes/no guesses.)
b) (18 marks) Let $Y$ be a random variable with mean $\mathrm{E} Y=a$ and variance $\operatorname{Var}(Y)=\sigma_{y}^{2}$. Consider the time series $\left\{X_{t}\right\}, t \geq 0$, defined by the equations

$$
\begin{aligned}
& X_{0}=Y \\
& X_{t}=c+\phi X_{t-1}+\varepsilon_{t}, \quad t \geq 1,
\end{aligned}
$$

where $c$ is a constant, $|\phi|<1,\left\{\varepsilon_{t}\right\} \sim \mathrm{WN}\left(0, \sigma^{2}\right)$, and $\mathrm{E} \varepsilon_{t} Y=0$ for all $t \geq 1$.
i) Obtain a condition on the mean, $a$, of $Y$, that ensures that $\mathrm{E} X_{t}$ is constant.
ii) Assuming that $\mathrm{E} X_{t}$ is constant (i.e., $a$ is chosen according to part (i) above), show that $\operatorname{Cov}\left(X_{t}, X_{t-1}\right)=\phi \operatorname{Var}\left(X_{t-1}\right)$.
iii) Find necessary and sufficient conditions for stationarity of $\left\{X_{t}\right\}, t \geq 0$.

B6. Consider the following seasonal ARIMA model for a time series $\left\{X_{t}\right\}$ :

$$
(1-\mathbf{B})\left(1-\mathbf{B}^{7}\right)(1-0.5 B) X_{t}=\varepsilon_{t},
$$

where $\left\{\varepsilon_{t}\right\} \sim \mathrm{WN}\left(0, \sigma^{2}\right)$.
a) (2 marks) What other property of $\left\{\varepsilon_{t}\right\}$ is assumed implicitly in the specification of ARIMA models, such as the above? (It may be referred to as orthogonality or causality condition.)
b) (2 marks) Identify the model as an $\operatorname{ARIMA}(p, d, q)\left(p_{s}, d_{s}, q_{s}\right)_{s}$ model, i.e., give the values of $s$, $p, d, q, p_{s}, d_{s}$, and $q_{s}$.
c) (4 marks) Let $f(h)=\hat{X}_{60+h \mid 60,59, \ldots}$. be the $h$-step ahead best linear predictor at time $t=60$. Give recursive equations for computation of $f(h)$ for $h=1,2, \ldots$.
d) (4 marks) Give the variance of the prediction errors for horizons $h=1$ and $h=2$.
e) (6 marks) Give a state-space form of the model for $\left\{X_{t}\right\}$.
f) (6 marks) Consider processes $\left\{Y_{t}\right\}$ and $\left\{Z_{t}\right\}$ defined as follows

$$
\begin{aligned}
& Y_{t}=Y_{t-1}+\varepsilon_{t} \\
& Z_{t}=\phi Y_{t-1}+u_{t}
\end{aligned}
$$

where $\phi \neq 0,\left\{\varepsilon_{t}\right\}$ and $\left\{u_{t}\right\}$ are white noise sequences, and $\left\{u_{t}\right\}$ is independent of $\left\{\varepsilon_{t}\right\}$ (and therefore of $\left.\left\{X_{t}\right\}\right)$. Show that $\left\{Y_{t}\right\}$ and $\left\{Z_{t}\right\}$ are cointegrated and give a cointegrated relation and the corresponding cointegration vector.
[24 marks]

## Two Hours

Statistical tables are provided at end of paper

THE UNIVERSITY OF MANCHESTER

## GENERALISED LINEAR MODELS

24 May 2019
9:45-11:45

Answer ALL THREE questions in Section A.
Answer TWO of the THREE questions in Section B.
If more than TWO questions from section $B$ are attempted then credit will be given to the best TWO answers.

Electronic calculators may be used in accordance with the University regulations
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Page 1 of 8

## SECTION A

Answer ALL of the three questions

A1.
(a) Give the general form of a distribution from the exponential family and indicate which is the natural parameter and which is the dispersion parameter.
(b) Show that the distribution with probability mass function

$$
P(Y=y)=e^{-\lambda y}\left(1-e^{-\lambda}\right), y=0,1,2, \ldots
$$

belongs to the exponential family, with $\lambda>0$ being the natural parameter.
(c) Use Property 1 of distributions from the exponential family to find the mean and variance of the distribution in (b) and give your answers in terms of $\lambda$.
(d) Find the canonical link for a generalised linear model (GLM) with response following the distribution in (b).
(e) Write the mean response $\mu$ for the GLM in (d) with canonical link in terms of the linear predictor $\eta$ and state what the formula can be used for after fitting the model.
(f) The Newton-Raphson algorithm is identical to which two other algorithms for the fitting of the model in (d) with canonical link? Give the names of the algorithms in full.
[2 marks]
[20 marks in total for this question.]

A2.
(a) Write down a generalised linear model for $y_{i} \sim \operatorname{Pois}\left(\mu_{i}\right)$ on a single explanatory variable $x$ taking values $x_{i}, i=1, \ldots, n$. Use the square root link $g(\mu)=\sqrt{\mu}$ and linear predictor $\eta=\beta_{0}+\beta_{1} x$ in this model.
(b) Derive the log-likelihood function $\ell(\underset{\sim}{\beta})$ for the model in (a), where $\underset{\sim}{\beta}=\left(\beta_{0}, \beta_{1}\right)^{\top}$. Make sure you use the link function specified above.
(c) Following on from (a), let

$$
\xi_{i}=\frac{1}{2}\left(\beta_{0}+\beta_{1} x_{i}+\frac{y_{i}}{\beta_{0}+\beta_{1} x_{i}}\right), i=1, \ldots, n
$$

Find the mean vector and variance-covariance matrix of $\underset{\sim}{\xi}=\left(\xi_{1}, \ldots, \xi_{n}\right)^{\top}$.
(d) By differentiating the log-likelihood function in (b), show that the gradient vector or score is

$$
\frac{\partial \ell}{\partial \underset{\sim}{\beta}}=4 \sum_{i=1}^{n}\left(\xi_{i}-\beta_{0}-\beta_{1} x_{i}\right)\binom{1}{x_{i}}
$$

where $\xi_{i}$ is as defined in (c).
(e) Use the result in (d) to show that the MLE of $\underset{\sim}{\beta}$ satisfy

$$
X^{\top} X \underset{\sim}{\beta}=X^{\top} \underset{\sim}{\xi},
$$

where $X$ is the data matrix and $\underset{\sim}{\xi}$ is as defined in (c).
(f) Explain in detail how the equation in (e) is solved iteratively and interpret what happens at each iteration.
[20 marks in total for this question.]

## SECTION B

## Answer TWO of the three questions

B1. The Gamma distribution with probability density function (pdf) written as

$$
f(y ; \alpha, \lambda)=\frac{1}{\Gamma(\alpha)} \lambda^{\alpha} y^{\alpha-1} e^{-\lambda y}, y>0
$$

is known to belong to the exponential family with dispersion parameter $\phi=\alpha$ and $a(\phi)=1 / \phi$. Both parameters $\alpha$ and $\lambda$ are positive, and $\Gamma(\alpha)=\int_{0}^{\infty} x^{\alpha-1} e^{-x} \mathrm{~d} x$ is the Gamma function.
(a) Find the natural parameter $\theta$ and the function $b(\theta)$.
(b) Use Property 1 of distributions from the exponential family or otherwise to show that the mean is $\mu=\alpha / \lambda$ and find the variance function $V(\mu)$.
(c) Show that the scaled deviance of a generalised linear model fitted to Gamma responses $y_{1}, \ldots, y_{n}$ is

$$
D^{*}=2 \alpha \sum_{i=1}^{n}\left(\frac{y_{i}}{\hat{y}_{i}}-\log \frac{y_{i}}{\hat{y}_{i}}-1\right)
$$

where $\hat{y}_{i}, i=1, \ldots, n$ are fitted values.
(d) Suppose the model in (c) with linear predictor $\beta_{0}+\beta_{1} x$ is fitted to $n=22$ observations with

$$
\sum_{i=1}^{n} \frac{y_{i}}{\hat{y}_{i}}=31.84 \text { and } \sum_{i=1}^{n} \log \frac{y_{i}}{\hat{y}_{i}}=5.37 .
$$

Use these results to estimate $\alpha$.
(e) If the true value of $\alpha$ is 3 , does the model above appear to provide an adequate fit to the data? Use the scaled deviance to answer this question.
[20 marks in total for this question.]

B2. The Insurance dataset from the MASS library has information on the following variables.

- Claims $=$ number of claims on a policy during a set period of time.
- Holders $=$ number of policy holders on a policy (who can drive the car insured and make a claim).
- District $=$ geographic location (4 categories, $4=$ large cities $)$.
- Group $=$ Car group by engine size (from $1=$ small to $4=$ large)
- Age $=$ Age group of policy holder (from $1=$ young to $4=$ old $)$

A Poisson model with an offset was fitted to the data with the following results.

```
> fit <- glm(Claims ~ District + Group + Age + offset(log(Holders)),
    + data = Insurance, family = poisson)
    > summary(fit)
    Coefficients:
                            Estimate Std. Error z value Pr(>|z|)
    (Intercept) -1.82174 0.07679 -23.724 < 2e-16 ***
    District2 0.02587 0.04302 0.601 0.547597
    District3 0.03852 0.05051 0.763 0.445657
    District4 0.23421 0.06167 3.798 0.000146 ***
    Group2 0.16134 0.05053 3.193 0.001409 **
    Group3 0.39281 0.05500 7.142 9.18e-13 ***
    Group4 0.56341 0.07232 7.791 6.65e-15 ***
    Age2 -0.19101 0.08286 -2.305 0.021149 *
    Age3 -0.34495 0.08137 -4.239 2.24e-05 ***
    Age4 -0.53667 0.06996 -7.672 1.70e-14 ***
    (Dispersion parameter for poisson family taken to be 1)
    Null deviance: 236.26 on 63 degrees of freedom
    Residual deviance: 51.42 on 54 degrees of freedom
```

(a) Write the model that was fitted in mathematical/statistical terms, stating clearly any assumptions associated with it. You may use claims $i_{i}$ and holders ${ }_{i}$ to denote the claims and policy holder numbers respectively on policy $i$ with factor levels $\operatorname{district}_{i}$, group $_{i}$ and age $i, i=1, \ldots, 64$, and $\alpha_{j}, \beta_{j}, \gamma_{j}(j=1,2,3,4)$ for the main effects of the factors.
[4 marks]
(b) Explain the use of the offset term.
(c) Does the model provide an adequate fit to the data? Give a reason for your answer. [4 marks]
(d) What can be said of the different levels of District?
(e) Find the combination of factor levels that gives the highest probability of a claim per policy holder and estimate this probability.
[4 marks]
[20 marks in total for this question.]

## B3.

(a) A random sample of $n$ subjects are classified according to characteristic A with $I$ levels and characteristic B with $J$ levels. Let $y_{i j}$ be the number of subjects in the combined category $(i, j), i=1, \ldots, I, j=1, \ldots, J$. What is the name of the joint distribution of the random variables behind these observations? Write down the joint probability.
[3 marks]
(b) Show that the joint distribution in (a) can be obtained from independent Poisson random variables $\left\{Y_{i j}, i=1, \ldots, I, j=1, \ldots, J\right\}$ by conditioning on their sum.
[5 marks]
(c) A random sample of 433 men were observed for obesity and snoring and the data are as follows.

| Obese | Snoring | Total (y) |
| :---: | :---: | :---: |
| no | no | 77 |
| yes | no | 10 |
| no | yes | 272 |
| yes | yes | 74 |

(i) Use the following results to test independence between obesity and snoring at the $5 \%$ level.
[4 marks]

```
> fit <- glm(y~obese*snoring, family=poisson)
> anova(fit, test="Chisq")
Analysis of Deviance Table
Terms added sequentially (first to last)
    Df Deviance Resid. Df Resid. Dev Pr(>Chi)
\begin{tabular}{lrrrrll} 
NULL & & & 3 & 344.83 & & \\
obese & 1 & 174.224 & 2 & 170.60 & \(<2 \mathrm{e}-16 * * *\) \\
snoring & 1 & 165.810 & 1 & 4.79 & \(<2 \mathrm{e}-16 * *\) \\
obese:snoring & 1 & 4.793 & 0 & 0.00 & \(0.02857 *\)
\end{tabular}
```

(ii) Interpret the obeseyes:snoringyes term below as a log odds ratio and estimate the odds ratio.

```
> summary(fit)
Coefficients:
```

| Estimate | Std. Error $z$ value | $\operatorname{Pr}(>\|z\|)$ |  |  |
| ---: | ---: | ---: | ---: | ---: |
| 4.3438 | 0.1140 | 38.117 | $<2 \mathrm{e}-16$ | $* * *$ |
| -2.0412 | 0.3361 | -6.073 | $1.26 \mathrm{e}-09$ | $* * *$ |
| 1.2620 | 0.1291 | 9.776 | $<2 \mathrm{e}-16$ | $* * *$ |
| 0.7395 | 0.3608 | 2.050 | 0.0404 | $*$ |

(iii) What are the fitted probabilities under this model?
[4 marks]
[20 marks in total for this question.]

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| $d f$ | $q$-value |  |  |  |  |  |  |  |  |  |  |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| $\nu$ | 0.005 | 0.01 | 0.025 | 0.05 | 0.10 | 0.50 | 0.90 | 0.95 | 0.975 | 0.99 | 0.995 |
| 1 | . $0^{4} 39$ | . $0^{3} 16$ | . $0^{3} 98$ | . 0039 | . 0158 | 455 | 2.706 | 3.841 | 5.024 | 6.635 | 7.879 |
| 2 | . 0100 | . 0201 | . 0506 | . 103 | . 211 | 1.386 | 4.605 | 5.991 | 7.378 | 9.210 | 10.597 |
| 3 | . 0717 | . 115 | 216 | . 352 | . 584 | 2.366 | 6.251 | 7.815 | 9.348 | 11.345 | 12.838 |
| 4 | . 207 | . 297 | . 484 | . 711 | 1.064 | 3.357 | 7.779 | 9.488 | 11.143 | 13.277 | 14.860 |
| 5 | . 412 | . 554 | . 831 | 1.145 | 1.610 | 4.351 | 9.236 | 11.070 | 12.832 | 15.086 | 16.750 |
| 6 | . 676 | . 872 | 1.237 | 1.635 | 2.204 | 5.348 | 10.645 | 12.592 | 14.449 | 16.812 | 18.548 |
| 7 | . 989 | 1.239 | 1.690 | 2.167 | 2.833 | 6.346 | 12.017 | 14.067 | 16.013 | 18.475 | 20.278 |
| 8 | 1.344 | 1.646 | 2.180 | 2.733 | 3.490 | 7.344 | 13.362 | 15.507 | 17.535 | 20.090 | 21.955 |
| 9 | 1.735 | 2.088 | 2.700 | 3.325 | 4.168 | 8.343 | 14.684 | 16.919 | 19.023 | 21.666 | 23.589 |
| 10 | 2.156 | 2.558 | 3.247 | 3.940 | 4.865 | 9.342 | 15.987 | 18.307 | 20.483 | 23.209 | 25.188 |
| 11 | 2.603 | 3.053 | 3.816 | 4.575 | 5.578 | 10.341 | 17.275 | 19.675 | 21.920 | 24.725 | 26.757 |
| 12 | 3.074 | 3.571 | 4.404 | 5.226 | 6.304 | 11.340 | 18.549 | 21.026 | 23.337 | 26.217 | 28.300 |
| 13 | 3.565 | 4.107 | 5.009 | 5.892 | 7.042 | 12.340 | 19.812 | 22.362 | 24.736 | 27.688 | 29.819 |
| 14 | 4.075 | 4.660 | 5.629 | 6.571 | 7.790 | 13.339 | 21.064 | 23.685 | 26.119 | 29.141 | 31.319 |
| 15 | 4.601 | 5.229 | 6.262 | 7.261 | 8.547 | 14.339 | 22.307 | 24.996 | 27.488 | 30.578 | 32.801 |
| 16 | 5.142 | 5.812 | 6.908 | 7.962 | 9.312 | 15.338 | 23.542 | 26.296 | 28.845 | 32.000 | 34.267 |
| 17 | 5.697 | 6.408 | 7.564 | 8.672 | 10.085 | 16.338 | 24.769 | 27.587 | 30.191 | 33.409 | 35.718 |
| 18 | 6.265 | 7.015 | 8.231 | 9.390 | 10.865 | 17.338 | 25.989 | 28.869 | 31.526 | 34.805 | 37.156 |
| 19 | 6.844 | 7.638 | 8.907 | 10.117 | 11.651 | 18.338 | 27.204 | 30.144 | 32.852 | 36.191 | 38.582 |
| 20 | 7.434 | 8.260 | 9.591 | 10.851 | 12.443 | 19.337 | 28.412 | 31.410 | 34.170 | 37.566 | 39.997 |
| 21 | 8.034 | 8.897 | 10.283 | 11.591 | 13.240 | 20.337 | 29.615 | 32.671 | 35.479 | 38.932 | 41.401 |
| 22 | 8.643 | 9.542 | 10.982 | 12.338 | 14.041 | 21.337 | 30.813 | 33.924 | 36.781 | 40.289 | 42.796 |
| 23 | 9.260 | 10.196 | 11.688 | 13.091 | 14.848 | 22.337 | 32.007 | 35.172 | 38.076 | 41.638 | 44.181 |
| 24 | 9.886 | 10.856 | 12.401 | 13.848 | 15.659 | 23.337 | 33.196 | 36.415 | 39.364 | 42.980 | 45.558 |
| 25 | 10.520 | 11.524 | 13.120 | 14.611 | 16.473 | 24.337 | 34.382 | 37.652 | 40.646 | 44.314 | 46.928 |
| 26 | 11.160 | 12.198 | 13.844 | 15.379 | 17.292 | 25.336 | 35.563 | 38.885 | 41.943 | 45.642 | 48.290 |
| 27 | 11.808 | 12.879 | 14.573 | 16.151 | 18.114 | 26.336 | 36.741 | 40.113 | 43.194 | 46.963 | 49.645 |
| 28 | 12.461 | 13.565 | 15.308 | 16.928 | 18.939 | 27.336 | 37.916 | 41.337 | 44.461 | 48.278 | 50.993 |
| 29 | 13.121 | 14.256 | 16.047 | 17.708 | 19.768 | 28.336 | 39.087 | 42.557 | 45.722 | 49.588 | 52.336 |
| 30 | 13.787 | 14.953 | 16.791 | 18.493 | 20.599 | 29.336 | 40.256 | 43.733 | 46.979 | 50.892 | 53.672 |
| 31 | 14.458 | 15.656 | 17.539 | 19.281 | 21.434 | 30.436 | 41.422 | 44.985 | 48.232 | 52.191 | 55.003 |
| 32 | 15.134 | 16.362 | 18.291 | 20.072 | 22.271 | 31.336 | 42.585 | 46.194 | 49.480 | 53.486 | 56.328 |
| 33 | 15.815 | 17.074 | 19.047 | 20.867 | 23.110 | 32.336 | 43.745 | 47.400 | 50.725 | 54.776 | 57.649 |
| 34 | 16.501 | 17.879 | 19.806 | 21.664 | 23.952 | 33.336 | 44.903 | 48.602 | 51.966 | 58.061 | 58.964 |
| 35 | 17.192 | 18.509 | 20.569 | 22.465 | 24.797 | 34.336 | 46.059 | 49.802 | 53.203 | 57.342 | 60.275 |
| 36 | 17.887 | 19.233 | 21.336 | 23.269 | 25.643 | 35.336 | 47.212 | 50.999 | 54.437 | 58.619 | 61.581 |
| 37 | 18.586 | 19.960 | 22.106 | 24.075 | 26.492 | 36.336 | 48.363 | 52.192 | 55.668 | 59.893 | 62.883 |
| 38 | 19.289 | 20.691 | 22.879 | 24.884 | 27.343 | 37.335 | 49.513 | 53.384 | 56.896 | 61.162 | 64.181 |
| 39 | 19.996 | 21.462 | 23.654 | 25.695 | 28.196 | 38.335 | 50.660 | 54.572 | 58.120 | 62.428 | 65.476 |
| 40 | 20.707 | 22.164 | 24.433 | 26.509 | 29.051 | 39.335 | 51.805 | 55.759 | 59.342 | 63.691 | 66.766 |
| 42 | 22.139 | 23.650 | 25.999 | 28.144 | 30.765 | 41.335 | 54.090 | 58.124 | 61.777 | 66.206 | 69.336 |
| 44 | 23.584 | 25.148 | 27.575 | 29.788 | 32.487 | 43.335 | 56.369 | 60.641 | 64.201 | 68.710 | 71.893 |
| 46 | 25.041 | 26.657 | 29.160 | 31.439 | 34.215 | 45.335 | 58.641 | 62.830 | 66.617 | 71.201 | 74.437 |
| 48 | 26.511 | 28.177 | 30.755 | 33.098 | 35.949 | 47.335 | 60.907 | 65.171 | 69.023 | 73.683 | 76.969 |
| 50 | 27.991 | 29.707 | 32.357 | 34.764 | 37.689 | 49.335 | 63.167 | 67.505 | 71.420 | 76.154 | 79.490 |
| 55 | 31.735 | 33.571 | 36.398 | 38.958 | 42.060 | 54.335 | 68.796 | 72.312 | 77.381 | 82.292 | 85.749 |
| 60 | 35.535 | 37.485 | 40.482 | 43.188 | 46.459 | 59.335 | 74.397 | 79.082 | 82.298 | 88.379 | 91.952 |
| 65 | 39.383 | 41.444 | 44.603 | 47.450 | 50.883 | 64.334 | 79.973 | 84.821 | 89.177 | 94.422 | 98.105 |
| 70 | 43.275 | 45.442 | 48.758 | 51.739 | 55.329 | 69.334 | 85.527 | 90.531 | 95.023 | 100.43 | 104.22 |

Table 1: Percentage points $\chi_{\nu, q}^{2}$ of the chi-squared distribution. The percentage point $\chi_{\nu, q}^{2}$ of the chi-squared distributed random variable $\chi_{\nu}^{2}$ is defined as $\operatorname{Pr}\left(\chi_{\nu}^{2} \leq \chi_{\nu, q}^{2}\right)=q$
Percentage points of the F-distribution $\quad q=0.95$


# Two Hours <br> Statistical tables are attached UNIVERSITY OF MANCHESTER 

## MEDICAL STATISTICS

$31^{\text {st }}$ May 2019
09:45-11:45

Please use SEPARATE booklets for your answers to each Section.

Answer ALL five questions in SECTION A (40 marks in total).
Answer TWO of the three questions in SECTION B (40 marks in total). If more than two questions from Section B are attempted, then credit will be given for the two best answers.

The total number of marks for the paper is 80 .

University-approved calculators may be used.
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A1.
A randomised controlled trial of an electronic glucose monitor (E) against standard finger-prick glucose monitoring $(\mathbf{F})$ in patients with Type 1 diabetes has been designed to use deterministic minimisation based on two baseline factors. The minimisation factors are trial centre (3 levels: Manchester; Birmingham, Cambridge) and HbA1c level (2 levels: $7.5 \%-9.0 \%$; >9.0\% to 11.0).

Note: HbA1c is glycated haemoglobin, a measure of average glucose level over the past 2-3 months.

The number of participants with each baseline factor level after 20 participants have been allocated to the intervention groups are given in the table below.

| Baseline | Trial Centre |  |  |  |  |  | HbA1c level |  |  |  |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
|  | Manchester |  | Birmingham |  | Cambridge |  | 7.5\%-9.0\% |  | >9.0\%-11.0\% |  |
| Treatment | E | F | E | F | E | F | E | F | E | F |
| Number of participants | 4 | 2 | 3 | 3 | 3 | 5 | 5 | 6 | 5 | 4 |

(i) How many participants have been allocated to each treatment group?
(ii) The characteristics of the next participant (participant 21) entered into the trial is:

Trial Centre: Manchester
HbA1c level: 7.5\%-9.0\%.
Determine the intervention group to which this participant has to be allocated.
[2 marks]
(iii) What type of bias has an increased risk when deterministic minimisation is used rather than stratified randomisation? How would the use of stochastic minimisation reduce this risk of bias?
[2 marks]
(iv) If stochastic randomisation is used with allocation probabilities 0.7 and 0.3 , use the random number 0.60131 (from a uniform distribution on [0,1]) and a suitably-defined rule to allocate participant 21 to a treatment group.

## A2.

Van de Port (2012) reported testing for differences between treatment groups for 35 baseline variables, 4 of which had $p$-values less than 0.05 . In the Discussion section of their paper they report some limitations of their trial, which included the text:
"Firstly, although the trial was well powered, including 250 patients with few drop-outs, we found significant baseline differences in favour of the circuit training group for a few secondary outcomes. All analyses, however, were adjusted for these covariates at baseline."
(i) Explain why performing significance tests for baseline differences between treatment groups is not recommended in randomised controlled trials.
(ii) When should the between-groups comparison of continuous outcome measures in randomized controlled trials be adjusted for baseline variables? Is it appropriate to use significance testing to select the baseline covariates to include in an adjusted analysis with the aim of reducing potential bias? (Justify your answer.)

A3.
A randomised controlled trial is carried out to compare two doses of a new vaccine for the prevention of H 1 N 1 influenza. The effectiveness of the vaccine is tested by measuring the immune response in the blood measured 9 days following vaccination (as this measures successful vaccination and protection against influenza). The results are summarised in the table below: the numbers in the table are the respective counts for the different combinations of Dose ( $15 \mathrm{mg} ; 30 \mathrm{mg}$ ) and Immune response (Yes; No), with the totals randomised to each of the dose levels.

| Dose |  | 15 mg | 30 mg |
| :--- | :---: | :---: | :---: |
| Immune <br> response | Yes | 200 | 220 |
|  | No | 50 | 30 |
|  | Total | 250 | 250 |

(i) Estimate the odds ratio for the effectiveness of a 30 mg dose as compared to a 15 mg dose.
(ii) Calculate the $95 \%$ confidence interval of this odds ratio.
(iii) Interpret the results of this trial.

## A4.

In a meta-analysis of $k$ trials, suppose that $\hat{\theta}_{i}$ is an estimate of the treatment effect for the $i^{\text {th }}$ trial and let $\operatorname{Var}\left[\hat{\theta}_{i}\right]$ be its sampling variance. The minimum variance estimate is defined by:

$$
\hat{\theta}_{M V}=\frac{\sum_{i=1}^{k} w_{i} \hat{\theta}_{i}}{\sum_{i=1}^{k} w_{i}}
$$

where $w_{i}=1 / \operatorname{Var}\left[\hat{\theta}_{i}\right]$, and $\operatorname{Var}\left[\hat{\theta}_{M V}\right]=\frac{1}{\sum_{i=1}^{k} \frac{1}{\operatorname{Var}\left[\hat{\theta}_{i}\right]}}$.

The table below summarises the outcome of three trials of Neuromuscular Electrical Stimulation (NMES) versus no treatment for the condition of patellofemoral pain (PFP). The outcome measures was a Visual Analogue Scale (VAS) Pain score (range 0-10, with 0 representing no pain and 10 representing worst imaginable pain) during normal activities, assessed at the end of treatment. The treatment effect for each study $\left(\theta_{i}, i=1,2,3\right)$ is the difference in mean VAS pain for the two treatments (NMES group No treatment group).

| Study <br> (Date of publication) | Difference <br> $\hat{\theta}_{i}$ | $\operatorname{Var}\left[\hat{\theta}_{i}\right]$ |
| :---: | :---: | :---: |
| Akarcali(2002) | 1.12 | 0.3969 |
| Bily (2008) | 0.55 | 1.2769 |
| Tunay (2003) | 1.90 | 0.1296 |

(i) Compute the minimum variance estimate of the overall treatment effect, $\hat{\theta}_{M V}$, and determine its $95 \%$ confidence interval, stating any assumptions that you make.
[6 marks]
(ii) What can you conclude from the meta-analysis regarding the performance of NMES?
(iii) It is known that the minimal clinically important difference for VAS pain is in the range 1.5 to 2.0. By comparing these values with the $95 \%$ confidence interval that you computed in (i), provide interpretation of the confidence interval in the context of clinical importance.

## A5.

Consider a randomized controlled trial. Suppose the patient population can be divided into three latent sub-groups as follows:

- Compliers: patients who will comply with the allocated treatment;
- Always control treatment: patients who will receive control treatment regardless of allocation;
- Always new treatment: patients who will receive the new treatment regardless of allocation.

This assumes that there are no defiers, namely patients who will always receive the opposite of the treatment to which they are randomized. Assuming that the proportion and characteristics of compliers, always control treatment, always new treatment is the same in both arms and that randomization can only affect the outcome through the receipt of treatment, show that:
(i) An intention-to-treat estimate of the treatment effect is biased towards the null hypothesis of no treatment effect.
(ii) A per-protocol estimate of the treatment effect may be biased either towards or away from the null hypothesis of no treatment effect.

## B1.

For an $A B / B A$ crossover trial, a standard model for a continuous outcome $y_{i j}$ for the $i^{\text {th }}$ participant in the $j^{\text {th }}$ period is:

$$
\begin{array}{ll}
y_{i 1}=\mu+\xi_{i}+\varepsilon_{i 1} & \text { for a participant in sequence } A B \text { in period } 1 \\
y_{i 2}=\mu+\tau+\phi+\xi_{i}+\varepsilon_{i 2} & \text { for a participant in sequence } A B \text { in period } 2 \\
y_{i 1}=\mu+\tau+\xi_{i}+\varepsilon_{i 1} & \text { for a participant in sequence } B A \text { in period } 1 \\
y_{i 2}=\mu+\phi+\xi_{i}+\varepsilon_{i 2} & \text { for a participant in sequence } B A \text { in period } 2
\end{array}
$$

where $\mu$ is the mean for the sequence $B A$ in period $1, \tau$ is the treatment effect of $A$ compared to B , and $\phi$ is the period effect, with $\xi_{i}$ and $\varepsilon_{i j}$ being independent random variables with $\xi_{i} \sim N\left[0, \sigma_{B}^{2}\right]$ and $\varepsilon_{i j} \sim N\left[0, \sigma_{\varepsilon}^{2}\right]$. Defining $d_{i}=y_{i 2}-y_{i 1}$, let $\bar{d}_{A B}, \bar{d}_{B A}, \mu_{A B}^{d}$ and $\mu_{B A}^{d}$ be the sample and population means of $d_{i}$ for sequences $A B$ and $B A$ respectively.
(i) Show that $\hat{\tau}=\frac{\bar{d}_{A B}-\bar{d}_{B A}}{2}$ is an unbiased estimator of $\tau$, that is

$$
E[\hat{\tau}]=E\left[\frac{\bar{d}_{A B}-\bar{d}_{B A}}{2}\right]=\tau
$$

Hence also show that a test of the null hypothesis $H_{0}: \mu_{B A}^{d}=\mu_{A B}^{d}$ is the same as a test of the null hypothesis of no treatment effect $H_{0}: \tau=0$.
(ii) Show that the variance of $\hat{\tau}$ is given by

$$
\operatorname{Var}[\hat{\tau}]=\operatorname{Var}\left[\frac{\bar{d}_{A B}-\bar{d}_{B A}}{2}\right]=\frac{\sigma_{\varepsilon}^{2}}{2}\left(\frac{1}{n_{A B}}+\frac{1}{n_{B A}}\right)
$$

(You may assume the general result $\operatorname{Var}(\bar{x})=\frac{\sigma^{2}}{n}$ where $\operatorname{Var}(x)=\sigma^{2}$, and $n$ is the sample size.)
(iii) Given that the variance of the estimator based on the first period observations from the two groups, which is equivalent to a parallel-design trial treatment effect estimator, is given by:

$$
\left(\sigma_{B}^{2}+\sigma_{\varepsilon}^{2}\right)\left(\frac{1}{n_{A B}}+\frac{1}{n_{B A}}\right)
$$

what does the result given in (iii) suggest is an advantage of using an $A B / B A$ crossover design rather than a parallel-design trial? State the condition when this advantage is at its greatest.
(iv) An $A B / B A$ crossover trial was designed to compare two treatments for hypertension (high blood pressure): Drug A and Drug B. Thirty-eight participants, all suffering from hypertension, were randomised: 18 to the sequence Drug $A$ then Drug B, and 20 to the sequence Drug B then Drug A. A measurement of the primary outcome measure, diastolic blood pressure (in mm Hg ), was taken at the end of each treatment period and are summarised in the table below. Test the hypothesis $H_{0}: \tau=0$ versus $H_{1}: \tau \neq 0$ using a $5 \%$ significance level. Also compute and briefly interpret the $95 \%$ confidence interval for $\tau$.

| Sequence | Period 1 | Period 2 | Period 2-Period 1 |  |  |  |  |
| :---: | :---: | :---: | :---: | :---: | :--- | :---: | :---: |
|  | Mean | S.D. | Mean | S.D. | Mean | S.D. | $N$ |
| AB | 84.50 | 7.01 | 83.39 | 6.57 | -1.11 | 2.47 | 18 |
| BA | 84.10 | 8.94 | 86.65 | 8.12 | 2.55 | 2.26 | 20 |

B2.
A randomised controlled trial is planned to compare a treatment $(A)$ with the current standard therapy (B). For an outcome measure $Y$, let $\overline{y_{A}}, \overline{y_{B}}, \mu_{A}, \mu_{B}$ be the sample and population means of $Y$ for each treatment. Let s and $\sigma$ be the common within-group sample and population standard deviations of $Y$. Assume that the null hypothesis of no treatment effect will be tested by the statistic $T=\frac{\overline{y_{A}}-\overline{y_{B}}}{s \lambda}$, with $\lambda=\sqrt{1 / n_{A}+1 / n_{B}}$, where $n_{A}$ and $n_{B}$ are the number of participants allocated to each treatment.
(i) Write down an expression for the power (1- $\beta$ ) of the test for a two-tailed alternative hypothesis to detect a treatment effect $\tau$, stating the assumptions you make.
[4 marks]
(ii) Assuming that participants are allocated in the ratio of $1: \mathrm{k}$ such that $n_{B}=k . n_{A}$, show that the total sample size required to give a power of $(1-\beta)$ for a two-tailed $\alpha$ sized test is:

$$
n=\frac{(k+1)^{2}}{k} \frac{\sigma^{2}}{\tau^{2}}\left(z_{\alpha / 2}+z_{\beta}\right)
$$

(iii) Show that the minimum sample size is obtained by using an equal allocation ratio.
(iv) In a randomised controlled trial it is planned to allocate 240 participants to two treatments. The pooled within-group standard deviation is thought to be 12 units. Estimate the power of a study to detect a treatment effect of 5 units for two-tailed 0.05 -sized test, if an allocation ratio of 1:2 (treatment: control) is used.
[4 marks]
(v) In a randomised controlled trial, it is decided to allocate twice as many participants to the control group as to the treatment group, using blocked randomisation with a block size of 3 . Briefly describe how to create a randomisation list to implement this randomisation scheme by using random numbers with a simple allocation rule.
[2 marks]
[Total 20 marks]

## B3.

In a parallel-group non-inferiority trial, a new treatment $T$ is being compared with a control treatment $C$ using a normally distributed outcome measure $Y$. Assume that large values of $Y$ represent a worse outcome for the participant. Let $\overline{y_{T}}, \mu_{T}$ and $n_{T}$ and $\overline{y_{C}}, \mu_{C}$ and $n_{C}$ be the sample mean, population mean and the sample size for the new treatment and for the control treatment, respectively. Suppose $\sigma$ is the population standard deviation of both treatments. The treatment effect is defined as $\tau=\mu_{T}-\mu_{C}$.
(i) Provide an example, with justification, of a situation where a non-inferiority trial rather than a superiority trial should be used.
(ii) Explain why a significance test of the hypothesis $H_{0}: \tau=0$ versus $H_{1}: \tau>0$ would not be appropriate in a non-inferiority trial.
(iii) For a non-inferiority trial, the null hypothesis $H_{0}: \tau \geq \tau_{N}$ is rejected if the upper limit for a $100(1-\alpha) \%$ one-sided confidence interval for $\tau$ is less than the limit of noninferiority $\tau_{N}$, for which the probability, under $H_{0}$, is given by:

$$
\operatorname{Pr}\left[\operatorname{Reject} H_{0} \mid \tau\right]=\Phi\left(\left(\frac{\tau_{N}}{\sigma \lambda}-z_{\alpha}\right)-\frac{\tau}{\sigma \lambda}\right),
$$

where $\Phi$ is the cumulative distribution function of the standard normal distribution and $\lambda=\sqrt{1 / n_{T}+1 / n_{C}}$.

Given this, show that $\operatorname{Pr}\left[\operatorname{Reject} H_{0} \mid \tau\right]$ has a maximum under $H_{0}$ when $\tau=\tau_{N}$. Hence, show that this procedure has a Type I error probability less than or equal to $\alpha$.
(iv) A randomised controlled non-inferiority trial was carried out to test whether Nurse-led Telephone follow-up (Telephone Group) was non-inferior to the current standard care of Consultant-led Hospital follow-up (Hospital Group) for women following treatment for cancer of the endometrium (the inner lining of the uterus). The primary outcome was measured on a continuous scale using a questionnaire-based measure of anxiety, with higher scores representing greater anxiety. Follow-up data were collected on 111 participants from the Telephone Group and 106 from the Hospital Group. The computer output from the statistical analysis is given below.

```
Two-sample t test with equal variances
```

| Group | Obs | Mean | Std. Err. | Std. Dev. | [90\% Conf | Interval] |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| telephon | 111 | 33.00901 | 1.045759 | 11.01775 | 31.27428 | 34.74374 |
| hospital | 106 | 35.5283 | 1.265849 | 13.03271 | 33.42763 | 37.62897 |
| combined | 217 | 34.23963 | . 8201926 | 12.08219 | 32.88472 | 35.59454 |
| diff |  | -2.519293 | 1.635633 |  | -5.221312 | . 1827265 |
| diff $=$ mean(telephon) - mean(hospital) |  |  |  |  | $t=-1.5403$ |  |
| Ho: diff |  |  |  | degrees of freedom |  | 215 |

Ha: diff < 0
$\operatorname{Pr}(T<t)=0.0625$

Ha: diff != 0
$\operatorname{Pr}(|T|>|t|)=0.1250$

Ha: diff > 0
$\operatorname{Pr}(T>t)=0.9375$

A difference of less than 3.5 points on the anxiety scale was considered by researchers to be an unimportant difference between treatments. Using the computer output given above, test whether the Nurse-Led Telephone follow-up was non-inferior to the Consultant-Led Clinic Hospital-based follow-up for the primary outcome measure. Interpret your findings in the context of the trial. Can you also conclude superiority of the Nurse-led Telephone follow-up on the primary outcome measure?

## END OF EXAMINATION PAPER

## Two hours

Tables of the cumulative distribution function are provided.

## THE UNIVERSITY OF MANCHESTER

## MATHEMATICAL MODELLING IN FINANCE

4 June 2019
14:00-16:00

Answer ALL SIX questions.

Electronic calculators may be used in accordance with the University regulations
(C) The University of Manchester, 2019

Page 1 of 4

1. By using no arbitrage argument, derive put-call parity for the European put option, $P$, and the European call option, $C$, with the same exercise price $X$ and expiry date $T$. Assume the constant risk-free interest rate $r$.
2. Draw the expiry payoff diagrams for each of the following portfolios:
(i) Short two shares, long three calls with an exercise price $X$.
(ii) Short two shares, long three puts with exercise price $X_{1}$, long four calls with exercise price $X_{2}$. Consider only the case $X_{1}<X_{2}$.
[12 marks]
3. The change in price of an asset $S$ satisfies the stochastic differential equation

$$
d S=0.5 S d t+(1+t) \sqrt{S} d W
$$

where $W(t)$ is a Wiener process. Using Ito's Lemma along with the delta-hedge argument, derive the partial differential equation satisfied by the price $V(S, t)$ of an option on the underlying asset.
[12 marks]
4. Consider the Black-Scholes equation for the price for an option, $V$ of an option, on a stock $S$ that pays no dividends, namely

$$
\frac{\partial V}{\partial t}+\frac{1}{2} \sigma^{2} S^{2} \frac{\partial^{2} V}{\partial S^{2}}+r S \frac{\partial V}{\partial S}-r V=0
$$

where $r$ and $\sigma$ are both constants.
(i) Show by substitution that

$$
V(S, t)=A e^{r t+\beta \ln S}
$$

where $A$ and $\beta$ are constants, is a solution of the Black-Scholes equation.
(ii) Determine the values of the constant $\beta$.
(iii) Find the partial differential equation satisfied by $V(Z, t)$, where $Z=-\ln S$.
5. You are given that the solution for a vanilla (non-dividend paying) European call option is

$$
C(S, t)=S N\left(d_{1}\right)-X e^{-r(T-t)} N\left(d_{2}\right)
$$

where

$$
\begin{aligned}
& d_{1}=\frac{\ln (S / X)+\left(r+\frac{1}{2} \sigma^{2}\right)(T-t)}{\sigma \sqrt{T-t}}, \\
& d_{2}=\frac{\ln (S / X)+\left(r-\frac{1}{2} \sigma^{2}\right)(T-t)}{\sigma \sqrt{T-t}},
\end{aligned}
$$

and $X$ is the exercise price and $N(x)$ is the cumulative distribution function, namely

$$
N(x)=\frac{1}{\sqrt{2 \pi}} \int_{-\infty}^{x} e^{-\frac{1}{2} y^{2}} d y
$$

(i) Find the present value of a three-month European call option on a stock with the exercise price $\$ 100$, when the current stock price is $\$ 100$ and $\sigma=1$. The risk-free interest rate is $0 \%$ p.a.
(ii) Find the limits

$$
\lim _{X \rightarrow 0} C(S, t), \quad \lim _{\sigma \rightarrow \infty} C(S, t), \quad \lim _{\sigma \rightarrow 0} C(S, t)
$$

(iii) By using the formula

$$
\Delta=\frac{\partial C}{\partial S}=N\left(d_{1}\right)
$$

find the explicit expression for

$$
\Gamma=\frac{\partial^{2} C}{\partial S^{2}}
$$

and sketch the graph of $\Delta$ as a function of $S$ in the limit $\sigma \rightarrow 0$.
(iv) Show that

$$
\frac{\partial C}{\partial X}=f(t) N\left(d_{2}\right)
$$

and find the function $f(t)$.
6. Suppose that an asset of value $S$ pays out a dividend $D S d t$ over the time period $d t$ (i.e. a constant dividend yield $D$ ). A put option $P(S, t)$ on this asset satisfies the modified Black-Scholes equation

$$
\frac{\partial P}{\partial t}+\frac{1}{2} \sigma^{2} S^{2} \frac{\partial^{2} P}{\partial S^{2}}+(r-D) S \frac{\partial P}{\partial S}-r P=0
$$

where $r$ and $\sigma$ are constants.
(i) Let $P(S, t)$ be a solution of the modified Black-Scholes equation for $D=r$. We set $S=e^{z}$, $\tau=T-t$ ( $T$ is the expiry date), $P=v(z, \tau)$. Find the partial differential equation satisfied by $v(z, \tau)$.
(ii) Find the solution of the modified Black-Scholes equation for $D=r$ in the form

$$
P(S)=A S^{\beta}
$$

where $A$ and $\beta$ are constants. Find the two possible values of $\beta$.
(iii) Consider a perpetual American put option on an asset paying a continuous constant dividend yield, which is given by $D=r$. If the exercise boundary is denoted by $S_{f}$, write down the two conditions to be imposed on $S_{f}$. State why only one of the solutions in (ii) is retained, and which one it is.
(iv) Use the conditions at $S=S_{f}$ to determine both the value of the constant $A$ and $S_{f}$ itself. Sketch the graph of $S_{f}$ as a function of $\beta$. What happens to $S_{f}$ as $D \rightarrow 0$ ?

## END OF EXAMINATION PAPER

## Two hours

Statistical Tables provided

## THE UNIVERSITY OF MANCHESTER

## ACTUARIAL MODELS 2

23 May 2019
9:45-11:45

Answer ALL FIVE questions.
The total number of marks in the paper is 80 .

University approved calculators may be used.
(c) The University of Manchester, 2019

1. Consider the following observed values of the possibly left-truncated and right-censored survival times of 12 independent homogeneous individuals, where + denotes a censored value:

$$
\begin{equation*}
(1,3] \quad(5,9+] \quad(5,10] \quad(5,10+] \quad(4,11] \quad(3,12] \quad(5,12] \quad(1,16] \quad(4,17+])(4,20+] \quad(0,23+] \tag{4,25}
\end{equation*}
$$

(a) Use the above survival data to calculate the Kaplan-Meier estimate of the survival function of the survival time distribution of the individuals and plot your estimate (you should show the details of your calculations).
[8 marks]
(b) Estimate, by using the Kaplan-Meier estimate, the probability that an individual fails before time 20.
(c) Suppose the survival data corresponding to the first individual changes from $(1,3]$ to $(0,3]$. Explain if with this change in the data the estimate of the probability in part (b) goes either (i) up, (ii) down or (iii) remains the same. (Note that you can answer this question without needing to recompute the estimate.)
[3 marks]
(d) For a positive random variable $X$ and a scalar $b \geq 0$, we have

$$
\mathbb{E}\left[X \mathbf{1}_{\{X \leq b\}}\right]=\int_{0}^{b} \mathbb{P}(X>x) \mathrm{d} x-b \mathbb{P}(X>b)
$$

Estimate, by using the Kaplan-Meier estimate and the above identity, the conditional expectation of the failure time of an individual given that the individual has failed before time 15 .
2. A clinical trial took place testing a particular treatment for Alzheimer's disease. For this trial a number of patients were randomly divided into a treatment group and a control group. The original plan was to observe patients for a duration of 24 months but due to a suspiciously high number of patients dying while getting this particular treatment, the study was halted after only 8 months. The researchers involved still want to use the data obtained during these 8 months for estimating and testing the effect of the treatment on the survival time distribution of Alzheimer's patients. To this end the survival times of patients who were still alive when the study was abruptly stopped are considered to be censored.

Explain briefly what independent censoring is and explain whether or not the particular type of censoring described above is likely to be independent or not.
[Total 5 marks]
3. Consider the following observed values of the survival times of five independent homogeneous individuals, where + denotes a censored value:

$$
2 \quad 5 \quad 5+\quad 6 \quad 7+
$$

Assume the following parametric form of the common hazard function of the individuals:

$$
\mu(t)=\mu(t ; \alpha, \lambda)=\alpha \lambda^{\alpha} t^{\alpha-1}, \quad t \geq 0
$$

with parameters $\alpha, \lambda>0$. Assume further that the times at which individuals are censored are all deterministic quantities and do not depend on the parameters $\alpha$ and $\lambda$.
(a) Derive, from first principles, an explicit expression (in term of $\alpha$ and $\lambda$ ) for the log-likelihood function of the parameters $\alpha$ and $\lambda$ given the observations. (Note that directly applying the (log)likelihood formula from the notes is not sufficient for getting full marks.) Here any constants that do not effect the maximum likelihood estimation can be ignored.
(b) We want to test whether $\alpha=1$ and $\lambda=0.20$, i.e. whether the hazard rate is a constant given by $\lambda=0.20$. Therefore we want to test the null hypothesis

$$
H_{0}: \mu(t)=\lambda \text { for all } t \geq 0 \quad \text { versus } \quad H_{A}: \mu(t) \neq \lambda \text { for some } t \geq 0
$$

where $\lambda=0.20$. The following 1 -sample test says to reject $H_{0}$ at significance level $\alpha$ if

$$
\left|\frac{N_{\infty}-\lambda \int_{0}^{\infty} R_{s} \mathrm{~d} s}{\sqrt{\lambda \int_{0}^{\infty} R_{s} \mathrm{~d} s}}\right|>z_{\alpha / 2},
$$

where $N_{t}$ denotes the number of individuals that are observed to have failed before or at time $t, R_{t}$ denotes the number of individuals at risk just before time $t$ and $z_{\alpha}>0$ is such that $\Phi\left(z_{\alpha}\right)=1-\alpha$, where $\Phi(\cdot)$ is the cumulative distribution function of the standard normal distribution.
(i) Given the survival data at the beginning of the question, carry out this 1 -sample test (with $\lambda=0.20$ ) at significance level 0.05 and report your conclusions.
[5 marks]
(ii) Show that under $H_{0}$ the process $Z=\left\{Z_{t}: t \geq 0\right\}$ given by $Z_{t}=N_{t}-\lambda \int_{0}^{t} R_{s} \mathrm{~d} s$ is a martingale. You can use here without proof that the process $\left\{M_{t}: t \geq 0\right\}$ given by $M_{t}=N_{t}-\int_{0}^{t} \mu(s) R_{s} \mathrm{~d} s$ is a martingale.
(iii) Show that under $H_{0}, \mathbb{E}\left[Z_{t}\right]=0$ and $\operatorname{Var}\left(Z_{t}\right)=\lambda \mathbb{E}\left[\int_{0}^{t} R_{s} \mathrm{~d} s\right]$ for any $t \geq 0$. (Hint: use part (ii).)
4. For a study into recidivism a number of convicted persons who spent time in jail were observed for some period of time. The data below displays for each individual how long it took to commit another offence carrying a prison sentence after he/she had been released from jail. Here a + next to the survival time indicates that the corresponding person did not carry out another major offence while being observed. Also recorded and displayed in the table are the amount of time each person spent in prison and whether he/she has a low or high income.

| time to reoffence (in months) | 1 | $4+$ | 6 | $12+$ | 15 |
| :--- | :--- | :--- | :--- | :--- | :--- |
| amount of time jailed (in months) | 5 | 2 | 9 | 10 | 10 |
| income (h=high, l=low) | l | l | h | h | h |

For modelling the rate of recidivism (i.e. the hazard rate of the time to recommit an offence after being released from prison) consider a Cox proportional hazards model with the amount of time jailed and income acting as covariates.
(a) Give the form of the hazard rate of an (arbitrary) individual in this Cox proportional hazards model. Here you should clearly define each piece of notation.
(b) Derive an explicit expression in terms of the two regression coefficients for the partial likelihood in this Cox proportional hazards model given the above data.
(c) Although none of the survival times in the above table are left-truncated, explain within the context of this study, what it means if a survival time of an individual would be left-truncated.
(d) Suppose, within this Cox proportional hazards model, that you want to use the likelihood ratio test to check whether or not the amount of income has a significant effect (for a given significance level) on the rate of recidivism. To this end, (i) give the null and alternative hypotheses, (ii) state the form of the test statistic and describe when one rejects the null hypothesis and (iii) explain for both possible outcomes of this test what you can conclude from it within the context of this particular study.

Note that you do not have to carry out this test, i.e. you do not have to compute a numerical value for the test statistic.
5. Assume a proportional frailty model for the mortality of a certain population where (i) the frailty random variable $Z$ has an atom at 1 with mass $p \in[0,1]$ (i.e. $\mathbb{P}(Z=1)=p$ ) and has a density on $(0, \infty)$ given by $f_{Z}(z)=(1-p) \mathrm{e}^{-z}$ and (ii) the individual hazard function is given by

$$
\mu(t \mid Z)=0.01 \cdot t \mathrm{e}^{-b t} Z, \quad t \geq 0
$$

where $b \geq 0$.
(a) Find an expression for the population hazard function in terms of $p$ and $b$. (Hint: treat the cases $b>0$ and $b=0$ separately.)

The remaining questions can be answered without knowing the answer to part (a).
(b) A research team is investigating the age at which a particular disease amongst humans occurs. When looking at the population hazard function of the age at which the disease occurs they observe that this hazard rate has a peak at around age 20 and decays afterwards. The research team thinks the explanation for this is that the physical changes the body undergoes during early adulthood makes humans more susceptible to develop the disease, whereas at later ages when the human body has reached full maturity, the disease is less likely to strike.
Consider the proportional frailty model stated in the beginning of the question and consider the following four sets of parameter choices for $p$ and $b$.

- $p=1$ and $b=0$;
- $p=1$ and $b=0.05$;
- $p=0.75$ and $b=0$;
- $p=0.75$ and $b=0.05$.

Which of the four sets of parameter choices would make the proportional frailty model align most closely with the explanation of the research team about the observed shape of the population hazard function corresponding to the age at which the disease occurs? Explain your answer.
(c) (i) Explain how the frailty effect can provide a completely different explanation for the observed shape of the population hazard function corresponding to the age at which the disease occurs.
(ii) Which of the four sets of parameter choices in (b) would make the resulting proportional frailty model be in contradiction with the explanation of the research team but instead match with the frailty effect being the explanation for the observed shape of the aforementioned population hazard function? Explain your answer.

## 2 Hours

## THE UNIVERSITY OF MANCHESTER

## CONTINGENCIES 2

17 May 2019
14:00-16:00

Answer ALL EIGHT questions
The total number of marks in the paper is 100 .

Electronic calculators may be used, provided that they cannot store text.

1 of 8
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Please note the following before you get started.

- You can find a formula sheet on the last page of this exam paper.
- The marking is based on your answer booklet only, so make sure to write down your working there.

1. Using the following extracts from two simplified mortality tables, one for life $x$ and one for life $y$, calculate the probabilities below (in all cases assume the first life is life $x$ and the second is life $y$ ):

| age $x$ | $l_{x}$ | age $y$ | $l_{y}$ |
| :---: | :---: | :---: | :---: |
| 30 | 10,000 | 20 | 10,000 |
| 31 | 9,900 | 21 | 9,950 |
| 32 | 9,800 | 22 | 9,900 |
| 33 | 9,700 | 23 | 9,850 |

a) $p_{32: 22}$
b) ${ }_{2} p_{30: 20}$
c) ${ }_{2 \mid} q_{30: 20}$
d) ${ }_{2 \mid} q_{30: 20}$
2. For two lives both aged 40 exact and assuming that both lives are statistically independent, calculate the expected present value of the following assurances, assuming a rate of interest of $3 \%$ p.a. and mortality of PFA90C20 for both lives:
a) $A_{40: 40}$
b) $A_{40: 40}^{1}$
c) $A_{\overline{40: 40}}$
3. An insurance company issues a policy to two lives $x$ and $y$ which pays a sum assured of $£ 10,000$ immediately on the first death. Premiums are payable continuously until the first death occurs.

Assume a mortality model for life $x$ which has $\mu_{x}=0.01$ for all $x \geq 0$ and for life $y$ has $\mu_{y}=0.012$ for all $y \geq 0$.
a) Show that the premium for this policy is independent of the force of interest $\delta$.
b) Calculate the premium.
c) If the policy was a whole life last survivor policy paying the same sum assured immediately on death, with premiums also payable continuously until the second death, would the amount of premium be dependent on the value of $\delta$ ? Give an explanation for your answer.
4. Consider a 3 -state sickness model with states $H$ (healthy), $S$ (sick) and $D$ (dead) and the following forces of transition for a life aged $x$

- $\mu_{x}^{H S}$ on transition from healthy to sick,
- $\mu_{x}^{H D}$ on transition from healthy to dead,
- $\mu_{x}^{S H}$ on transition from sick to healthy,
- $\mu_{x}^{S D}$ on transition from sick to dead.
a) Write down an integral expression for the expected present value (EPV) of a lump sum sickness benefit of $£ 20,000$ payable immediately on becoming ill for the first time for a healthy life aged 40 exact.
b) Write down an integral expression for the expected present value (EPV) of a lump sum sickness benefit of $£ 10,000$ payable immediately on becoming ill for the second or any subsequent time for a healthy life aged 40 exact.
c) Assume the above transition intensities take the following constant values, for all $x \geq 0$ :
- $\mu_{x}^{H S}=0.10$
- $\mu_{x}^{H D}=0.01$
- $\mu_{x}^{S H}=0.08$
- $\mu_{x}^{S D}=0.02$

Calculate the EPV for a sickness benefit of $£ 2,000$ p.a. for a life aged 30 exact who is currently in the sick state where the benefit is paid until the life either recovers or dies. Assume a force of interest of 0.03 .
5. Consider a 3 -state multiple decrement model with states $a$ (active), $r$ (retired) and $d$ (dead).
a) Given an independent probability of retiring when aged 60 of 0.1 and a constant force of mortality of 0.01 for all ages between 60 exact and 65 exact, calculate $(a q)_{60}^{d}$ and $(a q)_{60}^{r}$
b) If retirement can only occur at the end of a year what is the probability of a life aged 60 exact being active at age 61 exact?
6. A pension scheme provides a pension on normal health retirement, calculated as follows:

Annual amount of pension payable for life from normal health retirement at age $x=\frac{N}{80} \times F P S_{x}$ where

- $F P S_{x}=$ the average salary earned over the three years ending at age $x$ and
- $N$ is the total period of time the life has been a member of the pension scheme.

Now, consider a person aged 50 exact who joined the pension scheme 20 years ago. You are told that the person has earned a salary of $£ 25,000$ in the previous year.
a) Using the data and functions provided in R , assuming an interest rate of $4 \%$ p.a. and assuming the member retires between the age of 63 exact and age 65 exact, calculate the expected present value (EPV) of the pension benefits the life will receive.
[14 marks]
b) The pension scheme is now changed so that there is a maximum value placed on $F P S_{x}$ of $£ 30,000$. Again, assuming the member retires between the age of 63 exact and age 65 exact, calculate the EPV of the pension benefits the life will receive as a result of this change.
c) Instead of $F P S_{x}$ having a maximum value of $£ 30,000, S_{x}$, the salary earned between ages $x$ and $x+1$, has this same maximum amount applied. Does this change:
i) increase the EPV calculated in part b), or
ii) decrease the EPV calculated in part b), or
iii) leave the EPV calculated in part b) unchanged?
7. An insurer writes a 5 year unit linked endowment policy for a life aged 35. Death benefits are payable at the end of the year of death and the maturity value on completion of the 5 year term. Level premiums are paid yearly in advance. Benefits are paid at the end of the year.

You are given the following tables of data (note numbers shown in brackets are negative numbers):
Unit Fund

| year | Premium <br> paid | Premium <br> allocated | bid <br> offer <br> spread | net premium <br> allocated <br> to policyholder | fund before <br> management <br> charge | management <br> charge | Unit <br> Fund at <br> year end |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| 1 | 10,000 | 8,000 | 320 | 7,680 | 8,064 | 81 | 7,983 |
| 2 | 10,000 | 10,500 | 420 | 10,080 | 18,966 | 190 | 18,776 |
| 3 | 10,000 | 10,500 | 420 | 10,080 | 30,299 | 303 | 29,996 |
| 4 | 10,000 | 10,500 | 420 | 10,080 | 42,080 | 421 | 41,659 |
| 5 | 10,000 | 10,500 | 420 | 10,080 | 54,326 | 543 | 53,783 |

## Non-Unit Fund

| year | Non allocated <br> premium | bid/offer <br> spread | Expenses | interest on <br> fund at 4\% | management <br> charge | Profit |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| 1 | 2,000 | 320 | 3070 | $(30)$ | 81 | $(699)$ |
| 2 | $(500)$ | 420 | 120 | $(8)$ | 190 | $(18)$ |
| 3 | $(500)$ | 420 | 120 | $(8)$ | 303 | 95 |
| 4 | $(500)$ | 420 | 120 | $(8)$ | 421 | 213 |
| 5 | $(500)$ | 420 | 120 | $(8)$ | 543 | 335 |

In addition assume :

- $(a p)_{x}=0.99$ for all $x$
- Interest on reserves $=3 \%$ p.a.
a) State the allocation percentage and the bid offer spread (as a percentage) for this policy for each year.
b) What is the assumed investment return on the Unit Fund each year implied by this data?
c) What is the amount payable on death implied by this data?
d) Zeroise the profit vector for this policy.
e) If the insurance company decides to introduce a penalty to the policyholder on surrender equal to $10 \%$ of the value of the Unit Fund, what is the new Profit Vector. Assume the dependent probability of surrender, $(a q)_{x}^{s}=0.05$ for all ages $x$.

8. 

a) List four factors that may affect the levels of mortality within a population.
b) Mortality for a population has, for 20 years, consistently shown improvements. Recently these improvements have stopped. Referring to the factors affecting mortality, give four reasons why this might have happened.

## Formula sheet Contingencies 2

## 1. Compound interest

$i$ is the interest rate per annum, $v=(1+i)^{-1}$ and $\delta=\log (1+i)$
$a_{\bar{n}}=v+v^{2}+\ldots+v^{n}$ is the present value of an annuity certain in arrears with term $n$ years

## 2. Mortality tables and laws

$l_{x}$ is the number of lives at age $x$
$d_{x}=l_{x}-l_{x+1}$ is the number of lives who die between ages $x$ and $x+1$
${ }_{t} p_{x}$ is the probability that a life aged $x$ will survive at least $t$ years and ${ }_{t} q_{x}$ is the probability that a life aged $x$ will die within $t$ years
$\left.{ }_{m}\right|_{n} q_{x}$ is the probability that a life aged $x$ will die between ages $x+m$ and $x+m+n$
$K_{x}$ is the curtate future lifetime for a life aged $x$; $e_{x}=\mathbb{E}\left[K_{x}\right]$
$T_{x}$ is the complete (exact) future lifetime for a life $\operatorname{aged} x ; \stackrel{\circ}{e}_{x}=\mathbb{E}\left[T_{x}\right]$
$\mu_{x}$ is the force of mortality at age $x$ :

$$
{ }_{t} p_{x}=\exp \left(-\int_{0}^{t} \mu_{x+s} \cdot d s\right)
$$

## 3. Assurances

$A_{x}$ is the EPV of a whole life assurance that pays at the end of the year of death of a life aged $x$
$\bar{A}_{x}$ is the EPV of a whole life assurance that pays at the moment of death of a life aged $x$

$$
\begin{aligned}
\mathbb{E}\left[v^{K_{x}+1}\right] & =A_{x} ; \operatorname{Var}\left(v^{K_{x}+1}\right)=A_{x}^{i^{*}}-\left(A_{x}\right)^{2} \\
\mathbb{E}\left[v^{T_{x}}\right] & =\bar{A}_{x} ; \operatorname{Var}\left(v^{T_{x}}\right)=\bar{A}_{x}^{i^{*}}-\left(\bar{A}_{x}\right)^{2}
\end{aligned}
$$

(where $i^{*}=(1+i)^{2}-1$. Similar relationships hold for other assurances.
$A_{x y}$ is the EPV of a whole life assurance that pays at the end of the year in which the first death of two lives aged $x$ and $y$ occurs.
$A_{\overline{x y}}$ is the EPV of a whole life assurance that pays at the end of the year in which the second death of
two lives aged $x$ and $y$ occurs.
$A_{x y}^{1}$ is the EPV of a whole life assurance that pays at the end of the year in which a life aged $x$ dies assuming that a life aged $y$ remains alive at that time.
$A_{x y}^{2}$ is the EPV of a whole life assurance that pays at the end of the year in which a life aged $x$ dies assuming that a life aged $y$ is already dead at that time.

## 4. Annuities

$a_{x}$ is the EPV of an annuity for life in arrears for a life aged $x$
$\ddot{a}_{x}=1+a_{x}$ is the EPV of an annuity for life in advance for a life aged $x$
$\bar{a}_{x}$ is the EPV of a continuously paying annuity for life for a life aged $x$

$$
\begin{gathered}
\mathbb{E}\left[\ddot{a}_{\overline{K_{x}+1}}\right]=\ddot{a}_{x} ; \operatorname{Var}\left(\ddot{a}_{\overline{K_{x}+1}}\right)=\frac{A_{x}^{i^{*}}-\left(A_{x}\right)^{2}}{(1-v)^{2}} \\
\mathbb{E}\left[\bar{a}_{\bar{T}_{x}}\right]=\bar{a}_{x} ; \operatorname{Var}\left(\bar{a}_{\overline{T_{x}}}\right)=\frac{\bar{A}_{x}^{i^{*}}-\left(\bar{A}_{x}\right)^{2}}{\log (v)^{2}}
\end{gathered}
$$

(where $i^{*}=(1+i)^{2}-1$. Similar relationships hold for temporary annuities with status ${ }_{x: \bar{n}}$ )
$a_{x y}$ is the EPV of an annuity for the joint lifetime of two lives aged $x$ and $y$, paid in arrears.
$a_{\overline{x y}}$ is the EPV of an annuity payable in arrears so long as at least one of two lives aged $x$ and $y$ is alive.
$a_{x \mid y}$ is the EPV of a reversionary annuity payable to a life aged $y$ after the death of a life aged $x$, paid in arrears.

$$
\begin{aligned}
& \ddot{a}_{x}^{(m)} \approx \ddot{a}_{x}-\frac{m-1}{2 m} ; \\
& \quad \ddot{a}_{x: \bar{n})}^{(m)} \approx \ddot{a}_{x: \bar{n} \mid}-\frac{m-1}{2 m}\left(1-v^{n}{ }_{n} p_{x}\right) . \\
& A_{x}=1-(1-v) \ddot{a}_{x} ; \quad \bar{A}_{x}=1-\delta \bar{a}_{x}
\end{aligned}
$$

## Two hours

## THE UNIVERSITY OF MANCHESTER

## RISK THEORY

31 May 2019
14:00-16:00

Answer ALL FIVE questions
The total number of marks in the paper is 90 .

University approved calculators may be used
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Page 1 of 6

Note: attached to this exam paper you find the list of distributions that you have previously used during the course.

1. For a portfolio of products, an insurance company models the evolution of capital levels by means of a Cramér-Lundberg capital process denoted $\left(U_{t}\right)_{t \geq 0}$. That is, the amount of capital at time $t$ (measured in years) is given by

$$
U_{t}=u+c t-S_{t}=u+c t-\sum_{k=1}^{N_{t}} X_{k} \quad \text { for } t \geq 0
$$

where $u \geq 0$ is the initial capital, $c>0$ the premium rate, $\left(N_{t}\right)_{t \geq 0}$ a Poisson process with intensity $\lambda>0$, and finally $\left\{X_{k}\right\}_{k=1,2, \ldots}$ is a sequence of independent and identically distributed (iid) positive random variables representing the claim sizes. Assume that the claim sizes have finite mean.
(a) Suppose, in this part (a) only, that the company has an initial capital of 3 and that premiums are received at a rate of 2 per year. Further, assume that during the first year two claims are submitted, one of size 2 after 3 months and one of size 3 after 9 months (for simplicity, a month is $1 / 12$-th of a year). Draw the corresponding trajectories of $\left(N_{t}\right)_{t \geq 0},\left(S_{t}\right)_{t \geq 0}$ and $\left(U_{t}\right)_{t \geq 0}$ for this year. Also indicate whether or not ruin happens during this first year, and if so at which point in time.
(b) In the Cramér-Lundberg model, premiums are assumed to be paid continuously by the policyholders. Consider (in this part (b) only) changing the model so that all premiums to be paid during the first year are paid at the beginning of the first year instead. Would the probability of ruin during the first year increase or decrease due to this change? Briefly motivate your answer.
(c) Suppose that the Net Profit Condition holds. Show that the expected amount of capital at any time $t>0$ is larger than the initial capital $u$.
(d) Prove that the probability that the first incoming claim leads to ruin can be expressed as

$$
\int_{0}^{\infty}\left(1-F_{X_{1}}(u+c s)\right) \lambda e^{-\lambda s} \mathrm{~d} s
$$

where $F_{X_{1}}$ is the common cdf of the claim sizes.
Hint: consider conditioning on the arrival time of the first claim.
2. The insurance company from question 1 owns a second portfolio of products, for which the evolution of capital levels is also modelled as a Cramér-Lundberg capital process $\left(U_{t}\right)_{t \geq 0}$ as introduced in question 1. For this portfolio, the premium rate is $c=1 / 2$, the claim intensity is $\lambda=1$ and the claim sizes have a common pdf (probability density function) $f$ given by

$$
f(x)= \begin{cases}\frac{1}{8} e^{-2 x}+\frac{15}{4} e^{-4 x} & \text { if } x>0 \\ 0 & \text { if } x \leq 0\end{cases}
$$

The initial capital is $u \geq 0$.
(a) Verify whether or not the Net Profit Condition holds.
(b) Determine the Laplace exponent $\xi$ of $\left(U_{t}\right)_{t \geq 0}$ and state its domain.
(c) Find the Lundberg coefficient $R$ of $\left(U_{t}\right)_{t \geq 0}$.

Note: if you could not do part (c), then use $R=1$ in the rest of this question where you need the value of $R$ (this is not the correct value).
(d) Consider the following four functions, each defined for all $u \geq 0$ :

$$
\begin{aligned}
& \psi_{1}(u)=\frac{17}{32} e^{-\frac{1}{2} u}, \quad \psi_{2}(u)=\frac{15}{64} e^{-\frac{3}{2} u}+\frac{9}{64} e^{-2 u}, \\
& \psi_{3}(u)=\frac{18}{8} e^{-\frac{3}{2} u}-\frac{55}{32} e^{-2 u}, \quad \psi_{4}(u)=\frac{25}{64} e^{-\frac{3}{2} u}+\frac{9}{64} e^{-\frac{5}{2} u}
\end{aligned}
$$

One of these four is the correct expression for the ruin probability of $\left(U_{t}\right)_{t \geq 0}$ (as a function of the initial capital $u \geq 0$ ). Determine which one that is. Motivate your answer.
Hint: consider using elimination i.e. check which functions cannot be the ruin probability.
[5 marks]
(e) Suppose now that $u=0$. The company is contemplating whether a change in its current premium level of $c=1 / 2$ could result in a lower ruin probability. However it knows from experience that the more premium it charges, the more its policyholders are keen to get value for their money, resulting in more claims being submitted and larger claim sizes.
More precisely, it is assumed that these effects can be approximated as follows for $c \in(0,3)$. If the current value $c=1 / 2$ is changed by an amount $h$ (positive or negative), then

- the common mean of the claim sizes changes by $0.2 h \mathbb{E}\left[X_{1}\right]$ (where $\mathbb{E}\left[X_{1}\right]$ is the current common mean of the claim sizes),
- and the common mean of the claim interarrival times changes by $-0.3 h$.

Which premium level in the interval $(0,3)$ yields the smallest ruin probability?
3. This question concerns premium principles.
(a) State the Esscher premium principle, and state the Monotonicity property for premium principles.
[6 marks]
(b) Show that the standard deviation premium principle with loading factor $\beta=1.1$ satisfies the Scale Invariance property, but not the No Rip-off property.
Hint: for the latter, consider constructing a counterexample.
[8 marks]
[Total 14 marks]
4. Every week you buy a bag containing 10 treats for your pet dog Kevin. You know that Kevin's attitude towards any given type of treats is pretty straightforward (yet unobservable), namely one of the following two states:

$$
\theta_{1}: \text { he really likes them, } \theta_{2}: \text { he likes them less. }
$$

Your model uses the 'uninformative' prior that assigns probability $1 / 2$ to both possible states.
Further, you model the amount of treats that Kevin eats per week by a $\operatorname{Binomial}(10,0.7)$ distribution for treats that he really likes and a $\operatorname{Binomial}(10,0.3)$ distribution for treats that he likes less.

Now, your supermarket has started selling a new type, Squirrel Treats, and you decide to buy them. By the end of the week Kevin has eaten 7 of the Squirrel Treats.
(a) Using the 0-1 loss function, show that Bayes estimate for Kevin's state is $\theta_{1}$.
[6 marks]
(b) Suppose now that you would like to estimate Kevin's state using the loss function $l$ given by

$$
l(\theta, a)=\left\{\begin{array}{ll}
c & \text { if } a \neq \theta \\
1 & \text { if } a=\theta
\end{array} \quad \text { for some constant } c>1\right.
$$

Show that Bayes estimate for Kevin's state under this loss function is (also) $\theta_{1}$.
(c) The observation that we find the same result in parts (a) and (b) is not a coincidence. Indeed, consider a more general model in which Kevin's state can be one of $\theta_{1}, \theta_{2}, \ldots, \theta_{n}$ with any prior distribution and any likelihood distribution. Suppose that Kevin eats $x \in\{0,1, \ldots, 10\}$ treats. Show that Bayes estimate for Kevin's state under the 0-1 loss function is the same as Bayes estimate for Kevin's state under the loss function given in part (b).
5. Suppose that the total amount a policyholder claims per year can be modelled by an $\operatorname{Exp}(\theta)$ distribution, where $\theta>0$ is a parameter. It is assumed that these claim amounts are independent over the years. Further it is initially assumed that the parameters are distributed over all policyholders according to a $\operatorname{Gamma}(\alpha, \beta)$ distribution, for some $\alpha, \beta>0$.

Now, consider a policyholder who generated total claim amounts $x_{1}, x_{2}, \ldots, x_{n}$ during the past $n$ years.
(a) Show that the Bayes credibility estimate for the total claim amount for next year for this policyholder is given by

$$
\begin{equation*}
\frac{\beta+\sum_{i=1}^{n} x_{i}}{\alpha+n-1} \tag{1}
\end{equation*}
$$

Hint: you may find the following identity for the Gamma function helpful: $\Gamma(z+1)=z \Gamma(z)$ for any $z>0$.
[10 marks]
(b) Without doing any computations, briefly explain why the Bühlmann credibility estimate for the total claim amount for next year for this policyholder is also given by (1).
(c) Give an advantage that the Bühlmann credibility estimate in general (so not restricted to the problem in this question) has compared to the Bayes credibility estimate.

## List of distributions

## Discrete distributions

| Name | Notation | Parameters | Pmf | Mean | Variance |
| :--- | :--- | :--- | :--- | :--- | :--- |
| Bernoulli | Bernoulli $(p)$ | $p \in(0,1)$ | $f(0)=1-p, f(1)=p$ | $p$ | $p(1-p)$ |
| Binomial | Binomial $(n, p)$ | $n \in\{1,2, \ldots\}$ <br> $p \in(0,1)$ | $f(k)=\binom{n}{k} p^{k}(1-p)^{n-k}$ <br> for $k=0,1, \ldots, n$ | $n p$ | $n p(1-p)$ |
| Geometric | Geometric $(p)$ | $p \in(0,1)$ | $f(k)=(1-p)^{k-1} p$ <br> for $k=1,2, \ldots$ | $\frac{1}{p}$ | $\frac{1-p}{p^{2}}$ |
| Poisson | Poisson $(\lambda)$ | $\lambda>0$ | $f(k)=\frac{\lambda^{k}}{k!} e^{-\lambda}$ <br> for $k=0,1, \ldots$ | $\lambda$ | $\lambda$ |

## Continuous distributions

| Name | Notation | Parameters | Pdf |  | Mean | Variance |
| :--- | :--- | :--- | :--- | :--- | :--- | :--- |
| Beta | $\operatorname{Beta}(\alpha, \beta)$ | $\alpha, \beta>0$ | $f(x)=\frac{x^{\alpha-1}(1-x)^{\beta-1}}{B(\alpha, \beta)}$ | for $x \in(0,1)$ | $\frac{\alpha}{\alpha+\beta}$ | $\frac{\alpha \beta}{(\alpha+\beta)^{2}(\alpha+\beta+1)}$ |
| Exponential | $\operatorname{Exp}(\lambda)$ | $\lambda>0$ | $f(x)=\lambda e^{-\lambda x}$ for $x>0$ | $\frac{1}{\lambda}$ | $\frac{1}{\lambda^{2}}$ |  |
| Gamma | $\operatorname{Gamma}(\alpha, \beta)$ | $\alpha, \beta>0$ | $f(x)=\frac{\beta^{\alpha}}{\Gamma(\alpha)} x^{\alpha-1} e^{-\beta x} \quad$ for $x>0$ | $\frac{\alpha}{\beta}$ | $\frac{\alpha}{\beta^{2}}$ |  |
| Normal | $\mathcal{N}\left(\mu, \sigma^{2}\right)$ | $\mu \in \mathbb{R}$ | $f(x)=\frac{1}{\sqrt{\sqrt{2 \pi \sigma^{2}}} \exp \left(-(x-\mu)^{2} /\left(2 \sigma^{2}\right)\right)}$ | $\mu$ | $\sigma^{2}$ |  |
|  |  | $\sigma^{2}>0$ | for $x \in \mathbb{R}$ |  |  |  |
| Uniform | Unif $(a, b)$ | $a, b \in \mathbb{R}$ <br> $a<b$ | $f(x)=\frac{1}{b-a}$ for $x \in(a, b)$ | $\frac{a+b}{2}$ | $\frac{(b-a)^{2}}{12}$ |  |

## Notes

- For the pdf's above we understand $f(x)=0$ for those $x \in \mathbb{R}$ for which no value is specified. For example, in the case of a $\operatorname{Beta}(\alpha, \beta)$ distribution we have that $f(x)=0$ for $x \leq 0$ and for $x \geq 1$.
- $\Gamma$ denotes the Gamma function which can be defined as

$$
\Gamma(x)=\int_{0}^{\infty} t^{x-1} e^{-t} \mathrm{~d} t \quad \text { for } x>0
$$

Note that for integers $x \in\{1,2, \ldots\}$ we have that $\Gamma(x)=(x-1)$ !.

- $B$ denotes the Beta function which can be defined as

$$
B(x, y)=\int_{0}^{1} t^{x-1}(1-t)^{y-1} \mathrm{~d} t \quad \text { for } x, y>0 .
$$

Note that for integers $x, y \in\{1,2, \ldots\}$ we have that

$$
B(x, y)=\frac{(x-1)!(y-1)!}{(x+y-1)!}
$$

## END OF EXAMINATION PAPER


[^0]:    Electronic calculators may be used, provided that they cannot store text.

