# Spatial conditional extremes 

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#### Abstract

Currently available models for spatial extremes suffer either from inflexibility in the dependence structures that they can capture, lack of scalability to high dimensions, or in most cases, both of these. We present an approach to spatial extreme value theory based on the conditional multivariate extreme value model, whereby the limit theory is formed through conditioning upon the value at a particular site being extreme. The ensuing methodology allows for a flexible class of dependence structures, as well as models that can be fitted in high dimensions. To overcome issues of conditioning on a single site, we suggest a joint inference scheme based on all observation locations, and implement an importance sampling algorithm to provide spatial realizations and estimates of quantities conditioning upon the process being extreme at any of one of an arbitrary set of locations. The modelling approach is applied to Australian summer temperature extremes, permitting assessment the spatial extent of high temperature events over the continent.


Keywords: conditional extremes; extremal dependence; importance sampling; Pareto process; spatial modelling

## 1 Introduction

### 1.1 Background

Conditional extreme value theory (Heffernan and Tawn, 2004; Heffernan and Resnick, 2007) focuses on the behaviour of a random vector $\boldsymbol{X}$, given that a component of that random vector, say $X_{j}$, is large. In contrast to classical multivariate extreme value theory, resulting limit distributions can offer non-trivial descriptions of vectors exhibiting asymptotic dependence or asymptotic independence (see Section 1.3 for definitions), whereas the classical framework provides neat characterizations for the asymptotic dependence case only.

Classical extreme value theory has been extended from the multivariate case (de Haan and Resnick, 1977) to the spatial case, resulting in max-stable processes for maxima (de Haan, 1984; Smith, 1990; Schlather, 2002), or more recently, Pareto processes (Ferreira and de Haan, 2014; Dombry and Ribatet, 2015) for suitable definitions of functional threshold-exceedances. In each case, the resulting theory is only applicable as a statistical model when asymptotic dependence is present in the process, and when sufficient convergence towards the limit has occurred. A seemingly more common situation in environmental data is asymptotic independence, whereby the dependence in the process becomes progressively weaker as the level of the event becomes more extreme. This is manifested by extreme events becoming more spatially localized at higher levels, and is a feature exhibited by all Gaussian processes that are not perfectly dependent.

In contrast to the asymptotic dependence case, little work has been done on developing asymptotically justifiable models for asymptotically independent extremes. Partly, this is because one has to reconsider the meaning of "asymptotically justifiable" when the limits from classical extreme value theory are trivial. Based on a subasymptotic argument, Wadsworth and Tawn (2012) suggested a class of models that might be broadly applicable to asymptotically independent extremes, whilst the Gaussian process forms another possibility (Bortot et al., 2000). The modelling approach that we propose in this paper is able to capture a variety of asymptotically independent and dependent structures, whilst also being motivated by limiting arguments.

Sometimes it is unclear whether data display asymptotic dependence or asymptotic independence, but models often only cover one dependence type. Huser et al. (2017) and Huser and Wadsworth (2018) present spatial models that can capture both possibilities, although the same dependence class must hold over the entire spatial domain of interest, and independence between sites at long range may not be possible. In some situations, it is plausible that asymptotic dependence may hold between sites that are reasonably close, but asymptotic independence prevails at longer distances. The class of models that we describe can capture this phenomenon.

Several of the models available in the literature on spatial extremes are challenging to fit in high dimensions, with $20-30$ sites being a common upper limit. For a certain class of Pareto processes, de Fondeville and Davison (2018) used gradient score methods to avoid computing costly components of the likelihood, and combined this with coding efficiencies to permit inference on several hundred sites. However, general application of spatial
extreme value modelling to large numbers of sites has yet to occur. The methods that we propose can be fitted to hundreds of sites.

The flexibility in the range of dependence structures that we can capture is achieved by conditioning on the process being extreme at an arbitrary reference location $s_{0}$. In practice, whilst this conditioning might suit certain applications, more generally it is desirable to condition on the process being extreme anywhere over the domain of interest, $\mathcal{S} \subset \mathbb{R}^{2}$. We will also demonstrate inference on quantities of interest conditioning upon the maximum of the process at any collection of locations being large. This is achieved via an importance sampling algorithm, which also leads to an approximate method to simulate directly from the distribution of the field conditionally upon the maximum being large, through the empirical distribution of the sampled processes and their importance weights.

### 1.2 Main assumption

Conditional extreme value theory requires standardization to exponential-tailed margins, achieved in practice via the probability integral transformation. Let $\left\{X(s): s \in \mathcal{S} \subset \mathbb{R}^{2}\right\}$ be a stationary stochastic process with continuous sample paths and margins satisfying $\mathrm{P}\left(X\left(s_{j}\right)>x\right) \sim e^{-x}, x \rightarrow \infty$. Where we wish to account for negative dependence, it is also assumed that $\mathrm{P}\left(X\left(s_{j}\right)<x\right) \sim e^{-|x|}$, as $x \rightarrow-\infty$. For such cases, Keef et al. (2013) proposed modelling with Laplace margins. We assume that one can find functions $\left\{a_{s-s_{0}}: \mathbb{R} \rightarrow \mathbb{R}, s \in \mathcal{S}\right\}$, with $a_{0}(x)=x$ and $\left\{b_{s-s_{0}}: \mathbb{R} \rightarrow(0, \infty), s \in \mathcal{S}\right\}$, such that for any $s_{0} \in \mathcal{S}$, any $d \in \mathbb{N}$ and any collection of sites $s_{1}, \ldots, s_{d} \in \mathcal{S}$,

$$
\begin{equation*}
\left.\left(\left\{\frac{X\left(s_{i}\right)-a_{s_{i}-s_{0}}\left(X\left(s_{0}\right)\right)}{b_{s_{i}-s_{0}}\left(X\left(s_{0}\right)\right)}\right\}_{i=1, \ldots, d}, X\left(s_{0}\right)-t\right) \right\rvert\, X\left(s_{0}\right)>t \xrightarrow{d}\left(\left\{Z^{0}\left(s_{i}\right)\right\}_{i=1, \ldots, d}, E\right), \quad t \rightarrow \infty ; \tag{1}
\end{equation*}
$$

that is, convergence in distribution of the normalized process to $\left\{Z^{0}(s): s \in \mathcal{S}\right\}$ in the sense of finite-dimensional distributions. In the limit, the variable $E \sim \operatorname{Exp}(1)$ is independent of the residual process $\left\{Z^{0}(s)\right\}$, which satisfies $Z^{0}\left(s_{0}\right)=0$ almost surely, but is non-degenerate for all $s \neq s_{0}$ and places no mass at $+\infty$. Note that in assumption (1) it is irrelevant whether the conditioning site $s_{0}$ represents one of the sites $\left\{s_{1}, \ldots, s_{d}\right\}$. When assumption (1) is employed for modelling (see Sections 3 and 4) one indeed needs to condition on observed values. However, when simulating new events for example (see Section 5), the conditioning site could be any site in the domain.

An equivalent limiting formulation under the existence of a joint density for the process $X$ is obtained by conditioning upon the precise value of $X\left(s_{0}\right)$. To see this, observe that L'Hôpital's rule provides

$$
\begin{aligned}
\lim _{t \rightarrow \infty} \frac{\mathrm{P}\left(\left\{\frac{X\left(s_{i}\right)-a_{s_{i}-s_{0}}\left(X\left(s_{0}\right)\right)}{b_{s_{i}-s_{0}}\left(X\left(s_{0}\right)\right)}\right\}_{i=1, \ldots, d} \leq \boldsymbol{z}, X\left(s_{0}\right)>t\right)}{\mathrm{P}\left(X\left(s_{0}\right)>t\right)} & =\lim _{t \rightarrow \infty} \frac{\frac{\partial}{\partial t} \mathrm{P}\left(\left\{\frac{X\left(s_{i}\right)-a_{s_{i}-s_{0}}\left(X\left(s_{j}\right)\right)}{b_{s_{i}-s_{0}}\left(X\left(s_{0}\right)\right)}\right\}_{i=1, \ldots, d} \leq \boldsymbol{z}, X\left(s_{0}\right)>t\right)}{\frac{\partial}{\partial t} \mathrm{P}\left(X\left(s_{0}\right)>t\right)} \\
& =\lim _{t \rightarrow \infty} \mathrm{P}\left(\left.\left\{\frac{X\left(s_{i}\right)-a_{s_{i}-s_{0}}(t)}{b_{s_{i}-s_{0}}(t)}\right\}_{i=1, \ldots, d} \leq \boldsymbol{z} \right\rvert\, X\left(s_{0}\right)=t\right) .
\end{aligned}
$$

The independence of the conditioning variable is also assured under this alternative formulation of the assumption, see e.g. Wadsworth et al. (2017, Proposition 5). Since all the processes that we will consider have densities, this alternative version may sometimes be useful (see Section 2).

The functions $a_{s-s_{0}}, b_{s-s_{0}}$ appearing in limit (1) can be characterized to some degree. Heffernan and Resnick (2007) detail various requirements in a bivariate setting under the assumption that the conditioning variable has a regularly varying tail, but with no marginal assumptions on the other variable. If the conditioning variable has an exponential tail, these requirements translate to the existence of functions $\psi_{s-s_{0}}^{1}, \psi_{s-s_{0}}^{2}$ such that for all $c \in \mathbb{R}$

$$
\begin{equation*}
\psi_{s-s_{0}}^{1}(c)=\lim _{t \rightarrow \infty} b_{s-s_{0}}(t+c) / b_{s-s_{0}}(t) \quad \psi_{s-s_{0}}^{2}(c)=\lim _{t \rightarrow \infty}\left\{a_{s-s_{0}}(t+c)-a_{s-s_{0}}(t)\right\} / b_{s-s_{0}}(t) \tag{2}
\end{equation*}
$$

with local uniform convergence on compact subsets of $\mathbb{R}$; see also Papastathopoulos and Tawn (2016). Conditions (2) do not provide detailed information since the non-conditioning variable can have any marginal distribution, but any function $a_{s-s_{0}}$ or $b_{s-s_{0}}$ in (1), and later in Section 3, should satisfy these. Some modest additional structure is given in Proposition 3 of Appendix A under assumptions on the support of the limit distribution. Generally the literature on conditional extremes is split according to whether the non-conditioning variable(s) are assumed to have a standard marginal form, such as the exponential tails in assumption (1). Standardization occurs in most applied literature (e.g. Keef et al., 2013) with the prescription for the assumed normalization following from a variety of theoretical examples, in conjunction with checks on the modelling assumptions. We follow this general approach with the identical exponential-tailed margins in assumption (1). In more probabilistic literature (e.g. Das and Resnick, 2011; Drees and Janßen, 2017) standardization is typically not assumed, or involves heavy tails rather than exponential tails, such as in Heffernan and Resnick (2007).

### 1.3 Definitions and notation

Notions of asymptotic dependence and asymptotic independence are simply defined in the bivariate case, whilst the multivariate or spatial case requires more care. A bivariate random vector $\left(Y_{1}, Y_{2}\right)$, with $Y_{j} \sim F_{j}$, is termed asymptotically dependent if the limit $\chi=\lim _{q \rightarrow 1} \mathrm{P}\left(F_{1}\left(Y_{1}\right)>q, F_{2}\left(Y_{2}\right)>q\right) /(1-q)$ exists and is positive; $\chi=0$ defines asymptotic independence. In the spatial case, for $Y\left(s_{j}\right) \sim F_{s_{j}}$ we define

$$
\begin{equation*}
\chi\left(s_{1}, s_{2}\right)=\lim _{q \rightarrow 1} \mathrm{P}\left(F_{s_{1}}\left(Y\left(s_{1}\right)\right)>q, F_{s_{2}}\left(Y\left(s_{2}\right)\right)>q\right) /(1-q) \tag{3}
\end{equation*}
$$

We specialize to the case of stationary processes $Y$, for which $\chi\left(s_{1}, s_{2}\right) \equiv \chi(h), h=s_{1}-s_{2}$. A process is termed asymptotically dependent if $\chi(h)>0$ for all lags $h=\left(h_{1}, h_{2}\right)^{\top}$, and asymptotically independent if $\chi(h)=0$ for all $\|h\|>0$, with $\|\cdot\|$ the Euclidean distance. If $\chi(h)>0$ for $\|h\|<\Delta_{\theta}$, and $\chi(h)=0$ for $\|h\| \geq \Delta_{\theta}$, with $\theta=\arccos h_{1} /\|h\|$ the direction of $h$, we call the process directionally lag-asymptotically dependent, or simply lag-asymptotically dependent if the process is isotropic, so that $\Delta_{\theta} \equiv \Delta$.

Notationally, all vectors of length greater than one are expressed in boldface, with the exception of those denoting spatial location, e.g. $s_{j}=\left(s_{j, 1}, s_{j, 2}\right)^{\top}$, including spatial lags $h=s_{1}-s_{2}$. By convention, arithmetic operations on vectors are applied componentwise, with scalar values recycled as necessary. For example, if $f: \mathbb{R} \rightarrow \mathbb{R}, \boldsymbol{x} \in \mathbb{R}^{d}, c \in \mathbb{R}$, then $f(\boldsymbol{x}+c)=\left(f\left(x_{1}+c\right), \ldots, f\left(x_{d}+c\right)\right)^{\top}$.

## 2 Examples

We present examples of widely used spatial processes that satisfy limit (1), and use these to identify useful structures for model building. Table 1 summarizes the examples of Section 2.1-2.5.

Table 1: Normalization functions and limit processes for theoretical examples given in Sections 2.1-2.5. For the process of Huser and Wadsworth (2018), $V$ in (13) is taken as a marginally transformed Gaussian process. In the final row, $\ell_{s-s_{0}}(x)$ is a slowly-varying function of $x$, with $\lim _{x \rightarrow \infty} \ell_{s-s_{0}}(x)=0$, whose form is given by (10).

| Process | $a_{s-s_{0}}(x)$ | $b_{s-s_{0}}(x)$ | $Z^{0}(s)$ |
| :--- | :--- | :--- | :--- |
| Gaussian | $\rho\left(s-s_{0}\right)^{2} x$ | $1+a_{s-s_{0}}(x)^{1 / 2}$ | Gaussian |
| $t_{\nu}$ | $x$ | 1 | $t_{\nu+1}$ (transformed margins) |
| Brown-Resnick | $x$ | 1 | Gaussian |
| Huser and Wadsworth $(2018)(\lambda<1)$ | $\rho\left(s-s_{0}\right)^{2} x$ | $1+a_{s-s_{0}}(x)^{1 / 2}$ | Gaussian |
| Inverted Brown-Resnick | $\ell_{s-s_{0}}(x) x$ | $a_{s-s_{0}}(x) /(\log x)^{1 / 2}$ | Independence (reverse Gumbel margins) |

### 2.1 Gaussian process

Let $\left\{Y(s): s \in \mathcal{S} \subset \mathbb{R}^{2}\right\}$ be a standard stationary Gaussian processes with correlation function $\rho(h) \geq 0$ and let $X(s)=-\log (1-\Phi(Y(s)))$ be the same process transformed to standard exponential margins. Then, taking $a_{s-s_{0}}(x)=\rho\left(s-s_{0}\right)^{2} x$ and $b_{s-s_{0}}(x)=x^{1 / 2}$, the limit process $Z^{0}(s)$ is Gaussian with zero mean and covariance structure defined by the matrix

$$
\Sigma_{0}=\left(2 \rho_{k, 0} \rho_{l, 0}\left(\rho_{k, l}-\rho_{k, 0} \rho_{l, 0}\right)\right)_{1 \leq k, l \leq d}
$$

with $\rho_{k, 0}=\rho\left(s_{k}-s_{0}\right)$ etc. This limit representation looks problematic as the process $Z^{0}$ becomes degenerate as $\rho\left(s_{k}-s_{0}\right) \rightarrow 0$, owing to the fact that the scale normalization, $X\left(s_{0}\right)^{1 / 2}$, is still present when $\rho\left(s_{k}-s_{0}\right)=0$ even though $X\left(s_{k}\right)$ and $X\left(s_{0}\right)$ are then independent. To avoid this, we can instead consider

$$
\left\{\frac{X\left(s_{i}\right)-\rho\left(s_{i}-s_{0}\right)^{2} X\left(s_{0}\right)}{\rho\left(s_{i}-s_{0}\right) X\left(s_{0}\right)^{1 / 2}+1}\right\}_{i=1, \ldots, d}
$$

for which the limit process $Z^{0}(s)$ is Gaussian with zero mean and covariance structure

$$
\Sigma_{0}=2\left(\rho_{k, l}-\rho_{k, 0} \rho_{l, 0}\right)_{1 \leq k, l \leq d} .
$$

Consequently, $Z^{0}(s)$ has the conditional distribution of a Gaussian process with correlation function $\rho$, conditional upon the event $Z^{0}\left(s_{0}\right)=0$. However, when $\rho\left(s_{k}-s_{0}\right)=0$, the limit process $Z^{0}$ does not have Gaussian margins, but identical margins to $X$. As such, there is still a discontinuity in the limit behaviour once independence is reached.

### 2.2 Brown-Resnick process

Let $\left\{X(s): s \in \mathcal{S} \subset \mathbb{R}^{2}\right\}$ be a Brown-Resnick process (Kabluchko et al., 2009) with Gumbel margins. That is, $X$ can be expressed as

$$
\begin{equation*}
X(s)=\bigvee_{i=1}^{\infty} E_{i}+W_{i}(s)-\sigma^{2}(s) / 2 \tag{4}
\end{equation*}
$$

where $E_{i}$ are points of a Poisson process on $\mathbb{R}$ with intensity $e^{-x} d x$, and $W_{i}$ are independent and identically distributed copies of a centred Gaussian process with stationary increments, and $\sigma^{2}(s)=\mathrm{E}\left\{W(s)^{2}\right\}$. The variogram of the process $W$ is given by

$$
\begin{equation*}
\gamma\left(s_{1}, s_{2}\right)=\mathrm{E}\left[\left\{W\left(s_{1}\right)-W\left(s_{2}\right)\right\}^{2}\right], \tag{5}
\end{equation*}
$$

and if $W(0)=0$ almost surely (a.s.), then $\sigma^{2}(s)=\gamma(s, 0)$. Note that the representation (4) is not unique, and e.g., Dieker and Mikosch (2015) provide alternative representations with the same distribution.

Engelke et al. (2015) show that, for such a process, the finite-dimensional distributions of extremal increments $\left\{X(s)-X\left(s_{0}\right) \mid X\left(s_{0}\right)>t\right\}$ converges as $t \rightarrow \infty$ to a multivariate Gaussian distribution with mean vector

$$
\begin{equation*}
\boldsymbol{m}_{0}=\left(-\gamma\left(s_{i}, s_{0}\right) / 2\right)_{i=1, \ldots, d} \tag{6}
\end{equation*}
$$

and covariance matrix

$$
\begin{equation*}
\Omega_{0}=\left(\gamma\left(s_{i}, s_{0}\right) / 2+\gamma\left(s_{k}, s_{0}\right) / 2-\gamma\left(s_{i}, s_{k}\right) / 2\right)_{1 \leq i, k \leq d} . \tag{7}
\end{equation*}
$$

As such, the diagonal elements of the covariance matrix are $\left(\gamma\left(s_{i}, s_{0}\right)\right)_{i=1, \ldots, d}$, i.e., $\boldsymbol{m}_{0}=-\operatorname{diag}\left(\Omega_{0}\right) / 2$. This is the same limiting formulation as (1), with $a_{s-s_{0}}(x)=x, b_{s-s_{0}}(x)=1$, and $Z^{0}$ a Gaussian process whose moment structure is determined by (6) and (7).

## $2.3 t$ process

The $t$ process, with $\nu>0$ degrees of freedom, arises as a particular Gaussian scale mixture. Specifically, taking

$$
\begin{equation*}
Y(s)=R W(s) \tag{8}
\end{equation*}
$$

with $W$ a standard Gaussian process with correlation function $\rho(h)$, and $R^{-2} \sim \operatorname{Gamma}(\nu / 2, \nu / 2)$, the finitedimensional distributions $\boldsymbol{Y}=\left(Y\left(s_{1}\right), \ldots, Y\left(s_{d}\right)\right)$ have density

$$
f_{\nu}^{d}(\boldsymbol{y} ; \boldsymbol{\mu}, \Sigma)=C_{\nu}^{d}\left[1+(\boldsymbol{y}-\boldsymbol{\mu})^{\top} \Sigma^{-1}(\boldsymbol{y}-\boldsymbol{\mu}) / \nu\right]^{-(\nu+d) / 2}
$$

with normalization constant

$$
C_{\nu}^{d}=\frac{\Gamma((\nu+d) / 2)}{\Gamma(\nu / 2)|\Sigma|^{1 / 2}(\nu \pi)^{d / 2}}
$$

we write $\boldsymbol{Y} \sim \operatorname{St}^{d}(\boldsymbol{\mu}, \Sigma)$. Calculations for the bivariate $t_{\nu}$ distribution were given in Keef (2006); here we extend these to arbitrary dimension. Suppose $\boldsymbol{Y} \sim \operatorname{St}_{\nu}^{d}(\mathbf{0}, \Sigma)$, with dispersion matrix $\Sigma=\left(\rho_{k, l}\right)$ a correlation matrix, and let $\boldsymbol{Y}_{-j}$ represent $\boldsymbol{Y}$ without component $j$. Then

$$
\begin{equation*}
\frac{\boldsymbol{Y}_{-j}-\boldsymbol{\rho}_{j} Y_{j}}{\left(\nu+Y_{j}^{2}\right)^{1 / 2}}(\nu+1)^{1 / 2} \sim \operatorname{St}_{\nu+1}^{d-1}\left(\mathbf{0},\left(\rho_{k, l}-\rho_{j, k} \rho_{j, l}\right)_{k, l \neq j}\right) \tag{9}
\end{equation*}
$$

with $\boldsymbol{\rho}_{j} \in(-1,1)^{d-1}$ the $j$ th column of $\Sigma$, without the $j$ th component. Let $T: \mathbb{R} \rightarrow \mathbb{R}$ be a monotonic increasing transformation and consider $X=T(Y)$, such that $\mathrm{P}\left(X\left(s_{j}\right)>x\right) \sim e^{-x}$, with $\boldsymbol{X}=T(\boldsymbol{Y})$ the transformed finite-dimensional random vector. A suitable choice of transformation $T$ satisfies $T^{-1}(x) \sim K^{1 / \nu} e^{x / \nu}$, where $K=C_{\nu}^{1} \nu^{(\nu-1) / 2}$. We then have

$$
\begin{aligned}
\mathrm{P}\left(\boldsymbol{X}_{-j}-t \leq \boldsymbol{z} \mid X_{j}=t\right) & =\mathrm{P}\left(T\left(\boldsymbol{Y}_{-j}\right) \leq \boldsymbol{z}+t \mid T\left(Y_{j}\right)=t\right) \\
& =\mathrm{P}\left(\left.\frac{\boldsymbol{Y}_{-j}-\boldsymbol{\rho}_{j} T^{-1}(t)}{\left(\nu+T^{-1}(t)^{2}\right)^{1 / 2}} \leq \frac{T^{-1}(\boldsymbol{z}+t)-\boldsymbol{\rho}_{j} T^{-1}(t)}{\left(\nu+T^{-1}(t)^{2}\right)^{1 / 2}} \right\rvert\, Y_{j}=T^{-1}(t)\right)
\end{aligned}
$$

which, by (9) is the cdf of the $\operatorname{St}_{\nu+1}^{d-1}\left(\mathbf{0},\left(\rho_{k, l}-\rho_{j, k} \rho_{j, l}\right)_{k, l \neq j} /(\nu+1)\right)$ distribution. Using the asymptotic form of $T^{-1}(x)$, we have that the argument of this distribution function,

$$
\frac{T^{-1}(\boldsymbol{z}+t)-\boldsymbol{\rho}_{j} T^{-1}(t)}{\left(\nu+T^{-1}(t)^{2}\right)^{1 / 2}} \rightarrow e^{\boldsymbol{z} / \nu}-\boldsymbol{\rho}_{j}, \quad t \rightarrow \infty
$$

Consequently, $\boldsymbol{X}-X_{j} \mid X_{j}>t \xrightarrow{d} \boldsymbol{Z}^{j}$, where $Z_{j}^{j}=0$ and

$$
\mathrm{P}\left(\boldsymbol{Z}_{-j}^{j} \leq \boldsymbol{z}\right)=F_{\nu+1}^{d-1}\left(e^{\boldsymbol{z} / \nu} ; \boldsymbol{\rho}_{j},\left(\rho_{k, l}-\rho_{j, k} \rho_{j, l}\right)_{k, l \neq j} /(\nu+1)\right)
$$

where $F_{\nu+1}^{d-1}(\cdot, \boldsymbol{\mu}, \boldsymbol{\Sigma})$ is the cdf of $\operatorname{St}_{\nu+1}^{d-1}(\boldsymbol{\mu}, \boldsymbol{\Sigma})$. As such, we conclude for the spatial process that $a_{s-s_{0}}(x)=x$, $b_{s-s_{0}}(x)=1$. Note that this distribution has mass on lines through $-\infty$, which arise due to the fact that in representation (8), large values of $R$ cause both large and small values of $Y$, depending upon the sign of $W$.

The process $Z^{0}(s)$, defined through its finite-dimensional distributions $\boldsymbol{Z}^{0}=\left\{Z^{0}\left(s_{1}\right), \ldots, Z^{0}\left(s_{d}\right)\right\}$, is thus a transformed version of the $t_{\nu+1}$ process with $Z^{0}\left(s_{0}\right)=0$ a.s., with positive probability of being equal to $-\infty$ at any other site. That probability is smaller the stronger the dependence with the conditioning site; i.e., locally around the conditioning site, there is a high probability of $Z^{0}>-\infty$.

### 2.4 Inverted Brown-Resnick process

Wadsworth and Tawn (2012) introduced the class of inverted max-stable processes as those processes whose upper joint tail has the same dependence structure as the lower joint tail of a max-stable process. That is, if $Y$ is a max-stable process and $T$ represents a monotonically-decreasing marginal transformation, then $T(Y)$ is an inverted max-stable process. Applying the transformation $T(x)=e^{-x}$ to (4) yields the inverted Brown-Resnick process, with standard exponential margins.

Papastathopoulos and Tawn (2016) consider conditional limits of the bivariate margins of the inverted Brown-Resnick process. They show that the normalizations required to obtain a non-degenerate limit are

$$
\begin{align*}
& a_{s_{i}-s_{0}}(x)=x \exp \left\{\gamma\left(s_{i}, s_{0}\right) / 4-\left(\gamma\left(s_{i}, s_{0}\right) / 2\right)^{1 / 2}(2 \log x)^{1 / 2}+\left(\gamma\left(s_{i}, s_{0}\right) / 2\right)^{1 / 2} \frac{\log \log x}{(2 \log x)^{1 / 2}}\right\}  \tag{10}\\
& b_{s_{i}-s_{0}}(x)=a_{s_{i}-s_{0}}(x) /(\log x)^{1 / 2} \tag{11}
\end{align*}
$$

with limiting marginal distribution for $Z^{0}\left(s_{i}\right)$ given by

$$
\begin{equation*}
\left.\mathrm{P}\left(Z^{0}\left(s_{i}\right) \leq z\right)=1-\exp \left\{-\left(\gamma\left(s_{i}, s_{0}\right) / 2\right)^{1 / 2} \exp \left\{2 z / \gamma\left(s_{i}, s_{0}\right)^{1 / 2}\right)\right\} /(8 \pi)^{1 / 2}\right\} \tag{12}
\end{equation*}
$$

The multiplier of $x$ in equation (10) is a slowly varying function $\ell_{s-s_{0}}(x)$, meaning for $c>0, \lim _{x \rightarrow \infty} \ell_{s-s_{0}}(c x) / \ell_{s-s_{0}}(x)=$ 1. In the case of (10), $\lim _{x \rightarrow \infty} \ell_{s-s_{0}}(x)=0$, i.e., $a_{s-s_{0}}(x)=o(x), x \rightarrow \infty$. For the inverted Brown-Resnick process, the normalization and limiting distribution appear rather unintuitive when considering how $\gamma$ affects the dependence. Indeed, as $\gamma\left(s_{i}, s_{0}\right) \rightarrow \infty$, the max-stable process, and thus the inverted max-stable process, approaches independence. Yet, for finite $x$, the normalization $a_{s_{i}-s_{0}}(x)$ becomes large, and the mass of the limiting distribution of $Z^{0}\left(s_{i}\right)$ is placed at smaller values. Note that replacing $z$ by $\left\{\gamma\left(s_{i}-s_{0}\right) / 2\right\}^{1 / 2} z-\gamma\left(s_{i}-\right.$ $\left.s_{0}\right)^{1 / 2} \log \left\{\gamma\left(s_{i}-s_{0}\right) / 2\right\} / 4$ removes dependence of the limit distribution on $\gamma$; this is equivalent to modifying the normalization functions to

$$
\begin{aligned}
\tilde{a}_{s_{i}-s_{0}}(x) & =a_{s_{i}-s_{0}}(x)-b_{s_{i}-s_{0}}(x) \gamma\left(s_{i}, s_{0}\right)^{1 / 2} \log \left\{\gamma\left(s_{i}, s_{0}\right) / 2\right\} / 4 \\
\tilde{b}_{s_{i}-s_{0}}(x) & =\left\{\gamma\left(s_{i}-s_{0}\right) / 2\right\}^{1 / 2} b_{s_{i}-s_{0}}(x)
\end{aligned}
$$

However, whilst stabilizing the limit in $\gamma$, this still does not lead to an easily interpretable normalization in the sense of $a$ and/or $b$ decreasing monotonically with $\gamma$. To understand this, it is helpful to consider how such a process is formed. From equation (4), with $\sigma^{2}(s)=\gamma(s)$, the Brown-Resnick process is the pointwise maximum of location-adjusted Gaussian processes with negative drift. The more negative the drift (i.e., the larger $\gamma$ ) the more likely it is that the pointwise maxima from two locations will stem from different underlying Gaussian processes, which is why independence is achieved in the limit as $\gamma \rightarrow \infty$. Now, large values of the inverted Brown-Resnick process correspond to small values of the uninverted process, which are likely to be at the intersection points whereby different Gaussian processes contribute to the suprema. This rather complex construction thus leads to the seemingly unintuitive behaviour. Concerning the limiting process $Z^{j}$, Papastathopoulos and Tawn (2016) state that this corresponds to pointwise independence, and as such all structure lies in the functions $a_{s-s_{0}}, b_{s-s_{0}}$, and reverse Gumbel type margins (12).

### 2.5 Process of Huser and Wadsworth (2018)

Huser and Wadsworth (2018) present a model for spatial extremes obtained by scale mixtures on Pareto margins, or location mixtures on exponential margins. Suppose that $V(s)$ is an asymptotically independent process with unit exponential margins, and $Q$ is an independent unit exponential variate. Then

$$
\begin{equation*}
X^{*}(s)=\delta Q+(1-\delta) V(s) \tag{13}
\end{equation*}
$$

exhibits asymptotic independence for $\delta \leq 1 / 2$ and asymptotic dependence for $\delta>1 / 2$. Here, for simplicity of presentation, we reparameterize to $X^{* *}(s)=X^{*}(s) /(1-\delta)$, and set $\lambda=\delta /(1-\delta) \in(0, \infty)$, with $\lambda \in(0,1]$ corresponding to asymptotic independence. The marginal distribution of $X^{* *}$ is

$$
\mathrm{P}\left(X^{* *}\left(s_{j}\right)>x\right)=\frac{1}{1-\lambda} e^{-x}-\frac{\lambda}{1-\lambda} e^{-x / \lambda}
$$

so that the leading order term is $e^{-x} /(1-\lambda)$ for $\lambda<1$ and $\lambda e^{-x / \lambda} /(\lambda-1)$ for $\lambda>1$; the case $\lambda=1$ is obtained as $(1+x) e^{-x}$ upon taking the limit. Therefore, for $\lambda<1$, the transformation $X(s)=X^{* *}(s)+\log (1-\lambda)$ satisfies $\mathrm{P}\left(X\left(s_{j}\right)>x\right) \sim e^{-x}, x \rightarrow \infty$.

The following proposition establishes the behaviour of interest in this case; in particular we find that the conditional limit distribution of the modified process $X(s)$ requires the same normalization and has the same limit distribution as the process $V(s)$, if the scale normalization required for $V$ is increasing in $V\left(s_{0}\right)$. For notational purposes in the following we set $V_{0}=V\left(s_{0}\right), X_{0}=X\left(s_{0}\right), \boldsymbol{V}=\left(V\left(s_{1}\right), \ldots, V\left(s_{d}\right)\right), \boldsymbol{X}=\left(X\left(s_{1}\right), \ldots, X\left(s_{d}\right)\right)$, $\boldsymbol{a}(x)=\left(a_{s_{1}-s_{0}}(x), \ldots, a_{s_{d}-s_{0}}(x)\right)$, and $\boldsymbol{b}(x)=\left(b_{s_{1}-s_{0}}(x), \ldots, b_{s_{d}-s_{0}}(x)\right)$.
Proposition 1. Suppose that $\boldsymbol{V}$ has identical exponential-tailed margins, $P\left(V_{l}>v\right) \sim e^{-v}, v \rightarrow \infty$, and for $\boldsymbol{a}(v)$ and $\boldsymbol{b}(v)$ with twice-differentiable components $a_{l}, b_{l}$ satisfying $a_{l}^{\prime}(v) \sim \alpha_{l}, a_{l}^{\prime \prime}(v)=o(1), b_{l}^{\prime}(v) / b_{l}(v)=o(1)$, as $v \rightarrow \infty, l=1, \ldots, d$,

$$
P\left(\left.\frac{\boldsymbol{V}-\boldsymbol{a}\left(V_{0}\right)}{\boldsymbol{b}\left(V_{0}\right)} \leq \boldsymbol{z} \right\rvert\, V_{0}=v\right) \rightarrow G(\boldsymbol{z}), \quad v \rightarrow \infty
$$

Suppose further that all first and second order partial derivatives with respect to components of $\boldsymbol{z}$ converge, as specified in Lemma 1 of Appendix A. Then for $\boldsymbol{X}=\boldsymbol{V}+\lambda Q+\log (1-\lambda)$, with $Q \sim \operatorname{Exp}(1)$ independent of $\boldsymbol{V}$, and $\lambda \in(0,1)$,
$P\left(\left.\frac{\boldsymbol{X}-\boldsymbol{a}\left(X_{0}\right)}{\boldsymbol{b}\left(X_{0}\right)} \leq \boldsymbol{z} \right\rvert\, X_{0}=x\right) \rightarrow \begin{cases}G(\boldsymbol{z}), & \min _{1 \leq l \leq d} b_{l}(x) \rightarrow \infty \\ \left.\int_{0}^{q^{\star}} G\left(\boldsymbol{z}+\frac{(\boldsymbol{\alpha}-1)(\lambda q+\log (1-\lambda))}{\lim _{x \rightarrow \infty} \boldsymbol{b}(x)}\right)\right)(1-\lambda) e^{-(1-\lambda) q} \mathrm{~d} q, & \text { otherwise }\end{cases}$
as $x \rightarrow \infty$, with $\boldsymbol{\alpha}=\left(\alpha_{1}, \ldots, \alpha_{d}\right)^{\top}$, and $q^{\star}=\lim _{x \rightarrow \infty}\left(\left[-\log (1-\lambda)-v_{*}+\min \left\{\min _{l}\left(a_{l}(x)+b_{l}(x) z_{l}\right), x\right\}\right] / \lambda\right)_{+}$ for $v_{*}$ the lower endpoint of the support of $V$.
Corollary 1. The normalization and limit distribution for $X$ are the same as those for $V$ when either:
(i) All $b_{s-s_{0}}(x) \rightarrow \infty, x \rightarrow \infty$, for all $s, s_{0}$.
(ii) $a_{s-s_{0}}(x) \sim x$ and $b_{s-s_{0}}(x) \sim 1, x \rightarrow \infty$, for all $s, s_{0}$, as arises under asymptotic dependence for $V$.

The proof of Proposition 1 is in Appendix A. For the application of Proposition 1, convergence of the partial derivatives needs to be established. Supposing concretely that $V$ is a Gaussian process with margins transformed to be exponential, Lemma 2 and Remark 1 of Appendix A provides this result.

For $\lambda>1$, asymptotic dependence arises, and with a rescaling, one can express $X(s)=Q+V(s) / \lambda+\log (1-$ $1 / \lambda)$ such that $\mathrm{P}(X(s)>x) \sim e^{-x}$. In this case, $a_{s-s_{0}}(x)=x, b_{s-s_{0}}(x)=1$ and the limiting dependence structure is determined by the distribution of $V$.

## 3 Statistical modelling

Motivated by the limit assumption (1) and the examples in Section 2, we suppose that for a high threshold $u$,

$$
\begin{equation*}
\left\{X(s) \mid X\left(s_{0}\right)>u: s \in \mathcal{S}\right\} \stackrel{d}{\approx}\left\{a_{s-s_{0}}\left(X\left(s_{0}\right)\right)+b_{s-s_{0}}\left(X\left(s_{j}\right)\right) Z^{0}(s): s \in \mathcal{S}\right\} \tag{14}
\end{equation*}
$$

for some choice of functions $a_{s-s_{0}}, b_{s-s_{0}}$, residual process $Z^{0}$, and that $X\left(s_{0}\right)-u \mid X\left(s_{0}\right)>u \sim \operatorname{Exp}(1)$ is independent of $Z^{0}$. This specifies a process model conditioning on a particular site being extreme. With $d$ points at which the process is observed, for inference we are interested in $d$ such specifications, taking each observation site as the conditioning site $s_{0}$. Inference is described in Section 4, whilst simulation conditioning both on $\left\{X\left(s_{0}\right)>u\right\}$ and $\left\{\max _{i_{1} \leq i \leq i_{m}} X\left(s_{i}\right)>u\right\}$, for any collection of sites $\left\{s_{i_{1}}, \ldots, s_{i_{m}}\right\}$ with each $s_{i_{j}} \in \mathcal{S}$, is dealt with in Section 5. For now, we address the specification of $a_{s-s_{0}}, b_{s-s_{0}}$ and $Z^{0}$.

### 3.1 Functions $a_{s-s_{0}}$ and $b_{s-s_{0}}$

As noted in Section 1, the function $a_{s-s_{0}}: \mathbb{R} \rightarrow \mathbb{R}$ must satisfy $a_{0}(x)=x$. Further, if the process $X$ is asymptotically dependent, then $a_{s-s_{0}}(x)=x$ for all $s$, whilst if it is (directionally) lag-asymptotically dependent up to lag $\Delta_{\theta}$, then $a_{s-s_{0}}(x)=x$ for $\left\|s-s_{0}\right\| \leq \Delta_{\theta}$. Under asymptotic independence, and if $Z^{0}$ has any positive support, we have $a_{s-s_{0}}(x)<x$, where the bound may only hold asymptotically, i.e., as $x \rightarrow \infty$ (Proposition 3, Appendix A). Focusing on the isotropic case, we propose the general parametric form for $a_{s-s_{0}}$ as

$$
a_{s-s_{0}}(x)=x \alpha\left(s-s_{0}\right)= \begin{cases}x & \left\|s-s_{0}\right\|<\Delta  \tag{15}\\ x \exp \left\{-\left[\left(\left\|s-s_{0}\right\|-\Delta\right) / \lambda\right]^{\kappa}\right\}, & \left\|s-s_{0}\right\| \geq \Delta\end{cases}
$$

Taking $\Delta=0$ provides a flexible model for asymptotic independence, with parameters $(\lambda, \kappa)$ to estimate. Taking $\Delta>0$ but less than the maximum distance between any two sites in the domain would allow for lag-asymptotic dependence; the value of $\Delta$ itself could be estimated directly, although profiling over $\Delta$ on a grid may be preferable. Taking $\Delta$ larger than the maximum distance between any two sites would correspond to asymptotic dependence, in which case $(\lambda, \kappa)$ would not be estimated. Equation (15) covers all forms for $a_{s-s_{0}}$ from lines 1-4 of Table 1, if $\rho$ is an exponential correlation function, and line 5 up to a slowly varying function. Furthermore, condition (2) is satisfied as long as $b_{s-s_{0}}(x) \nrightarrow 0$ as $x \rightarrow \infty$. We note that other monotonically decreasing functions are also candidates for the second line of (15); certain correlation functions and survival functions are natural choices.

For the function $b_{s-s_{0}}$, we propose three forms to achieve different modelling aims. Each of the proposed sets of functions $a_{s-s_{0}}, b_{s-s_{0}}$ satisfies conditions (2).

Model 1 Let $a_{s-s_{0}}$ be given by (15), and

$$
b_{s-s_{0}}(x)=\left[1+\zeta x^{\beta}\right]^{-1}
$$

The rationale behind this suggestion is that, if $\beta<0$ and $\zeta>0$, then $b_{s-s_{0}}(x) ~ \nearrow 1$, with $\zeta$ and $\beta$ controlling the convergence to the constant value. When $a_{s-s_{0}}(x)=x$ and data are (lag-)asymptotically dependent, this permits the model to display some subasymptotic dependence, in the sense that for $\left\|s_{k}-s_{0}\right\|<\Delta$,

$$
\begin{aligned}
\chi_{u}\left(s_{k}-s_{0}\right)=\mathrm{P}\left(X\left(s_{k}\right)>u \mid X\left(s_{0}\right)>u\right) & =\mathrm{P}\left(X\left(s_{0}\right)+\left[1+\zeta X\left(s_{0}\right)^{\beta}\right]^{-1} Z^{0}\left(s_{k}\right)>u \mid X\left(s_{0}\right)>u\right) \\
& =e^{u} \int_{u}^{\infty} \mathrm{P}\left(Z^{0}\left(s_{k}\right)>(u-v) /\left[1+\zeta v^{\beta}\right]^{-1}\right) e^{-v} \mathrm{~d} v \\
& =\int_{0}^{\infty} \mathrm{P}\left(Z^{0}\left(s_{k}\right)>-q /\left[1+\zeta(q+u)^{\beta}\right]^{-1}\right) e^{-q} \mathrm{~d} q \\
& \searrow \int_{0}^{\infty} \mathrm{P}\left(Z^{0}\left(s_{k}\right)>-q\right) e^{-q} \mathrm{~d} q, \quad u \rightarrow \infty .
\end{aligned}
$$

In practice, spatial data nearly always exhibit values of $\chi_{u}\left(s_{k}-s_{0}\right)$ that decrease with the level $u$, although Pareto process models for asymptotic dependence (Ferreira and de Haan, 2014) cannot capture this feature, leading to overestimation of dependence at extreme levels. Taking $b_{s-s_{0}}$ as a constant function here with $a_{s-s_{0}}(x)=x$ would also yield asymptotic dependence with $\chi_{u}$ not varying with $u$. Such a model could be implemented by choice if the estimate of $\beta \ll 0$, for example.

Model 2 Let $a_{s-s_{0}}$ be given by (15) with $\Delta=0$, and

$$
b_{s-s_{0}}(x)=x^{\beta} .
$$

This represents a modelling strategy close to that proposed by Heffernan and Tawn (2004). If the marginal support of $Z^{0}$ includes $(0, \infty)$, we require $\beta<1$ (Proposition 3, Appendix A), whilst conditions (2) imply $\beta \geq 0$. We have

$$
\chi_{u}\left(s_{k}-s_{0}\right)=\int_{0}^{\infty} \mathrm{P}\left[Z^{0}\left(s_{k}\right)>\left\{u-\alpha\left(s_{k}-s_{0}\right)(q+u)\right\} /(q+u)^{\beta}\right] e^{-q} \mathrm{~d} q \searrow 0, \quad u \rightarrow \infty
$$

so that the rate of convergence to zero is controlled by $\beta$, in conjunction with the value of $\alpha\left(s_{k}-s_{0}\right)$ and the distribution of $Z^{0}\left(s_{k}\right)$. Such a model may perform well on data that exhibit asymptotic independence, but positive dependence over the whole region of study. In particular, since $b_{s-s_{0}}(x)$ is still growing even as $a_{s-s_{0}}(x) \rightarrow 0\left(\left\|s-s_{0}\right\| \rightarrow \infty\right)$, independence cannot be achieved as $\left\|s-s_{0}\right\| \rightarrow \infty$. This observation motivates our final model.

Model 3 Let $a_{s-s_{0}}$ be given by (15) with $\Delta=0$, and

$$
b_{s-s_{0}}(x)=1+a_{s-s_{0}}(x)^{\beta} .
$$

With Model 3 , if $a_{s-s_{0}}(x) \rightarrow 0$ as $\left\|s-s_{0}\right\| \rightarrow \infty$, with $\beta>0$, then $b_{s-s_{0}}(x) \rightarrow 1$ as $\left\|s-s_{0}\right\| \rightarrow \infty$. Thus, for $s$ sufficiently far from $s_{0}$, we would have

$$
X(s) \mid X\left(s_{0}\right)>u \stackrel{d}{\approx} Z^{0}(s)
$$

This final observation indicates that, under Model 3 , for $s$ far from $s_{0}$, the margins of the process $Z^{0}$ should be the same as the process $X$.

### 3.2 Process $Z^{0}(s)$

The process $Z^{0}$ must satisfy $Z^{0}\left(s_{0}\right)=0$. Supposing initially that we begin with an arbitrary Gaussian process $Z_{G}$, there are two natural ways to derive a process from $Z_{G}$ with this property:
(i) Set $Z^{0}(s)$ to have the distribution of $Z_{G}(s) \mid Z_{G}\left(s_{0}\right)=0$
(ii) Set $Z^{0}(s)$ to have the distribution of $Z_{G}(s)-Z_{G}\left(s_{0}\right)$,
with the resulting processes both again Gaussian. The initial process $Z_{G}$ might be stationary, and specified by a correlation function $\rho$ and variance $\sigma^{2}$, or have stationary increments, specified by a variogram $\gamma$, as in (5); addition of the drift term $-\gamma\left(s, s_{0}\right) / 2$ to the latter yields the process specified by (6) and (7).

Taking such a Gaussian process may be a natural and parsimonious choice in many situations. Indeed, allowing non-Gaussian dependence in $Z^{0}$ would lead to intractable models for high-dimensions, although we emphasize that the dependence of the modelled process $X$ is not Gaussian. However, the choice of marginal distribution for $Z^{0}$ can impact upon the model; in particular, as noted at the end of Section 3.1, we may wish $Z^{0}$ to have the same margins as $X$ in certain places, e.g., when $\left\|s-s_{0}\right\|$ is large.

We thus propose the following distributional family, that includes both Gaussian and Laplace marginal distributions. We say that a random variable $Z$ has a delta-Laplace distribution, with location parameter $\mu \in \mathbb{R}$ and scale parameter $\sigma>0$, if its density is

$$
\begin{equation*}
f(z)=\frac{\delta}{2 \sigma \Gamma(1 / \delta)} \exp \left\{-\left|\frac{x-\mu}{\sigma}\right|^{\delta}\right\}, \quad \delta>0 \tag{16}
\end{equation*}
$$

with $\Gamma(\cdot)$ the standard gamma function. When $\delta=1$ this is the Laplace distribution. If $X$ has Laplace margins, as suggested in Keef et al. (2013), Model 3 can be completed by allowing the parameter $\delta$ to depend on $s-s_{0}$, such that $\delta\left(s-s_{0}\right) \rightarrow 1$ as $\left\|s-s_{0}\right\| \rightarrow \infty$. When $\delta=2$, density (16) is that of the Gaussian distribution with mean $\mu$ and variance $\sigma^{2} / 2$. In general, if $Z$ has distribution (16), then $\mathrm{E}(Z)=\mu$ and $\operatorname{Var}(Z)=\{\Gamma(3 / \delta) / \Gamma(1 / \delta)\} \sigma^{2}$.

In practice, the location and scale structure implied by the processes (i) and (ii) defined on Gaussian margins are passed through to the delta-Laplace margins via the probability integral transform. In some situations it may be desirable to incorporate alternative mean structures by specifying a particular parametric form for $\mu\left(s-s_{0}\right)$ that is different from those implied by (i) and (ii); further discussion on this is made in Section 6.

## 4 Inference

### 4.1 Likelihood

The models described in Section 3 are suitable given an extreme in a single location. However, we would like to combine these models to allow for extremes in any observed location. Assuming stationarity, this motivates the use of a composite likelihood to estimate parameters of $a_{s-s_{0}}, b_{s-s_{0}}$ and $Z^{0}$, all of which do not depend on the conditioning site $s_{0}$, now taken to be one of the observation locations. Specifically, denote these parameters by $\boldsymbol{\theta}=\left(\boldsymbol{\theta}_{a}, \boldsymbol{\theta}_{b}, \boldsymbol{\theta}_{Z}\right)$, where some subsets of parameters may be scalar, and let

$$
L_{j}(\boldsymbol{\theta})=\prod_{i=1}^{n_{j}} f_{Z_{j}}\left(\left\{\left[x_{k}^{i}-a_{s_{k}-s_{j}}\left(x_{j}^{i} ; \boldsymbol{\theta}_{a}\right)\right] / b_{s_{k}-s_{j}}\left(x_{j}^{i} ; \boldsymbol{\theta}_{b}\right)\right\}_{k \in\{1, \ldots, d\} \backslash j} ; \boldsymbol{\theta}_{Z}\right) \prod_{k \in\{1, \ldots, d\} \backslash j} b_{s_{k}-s_{j}}\left(x_{j}^{i} ; \boldsymbol{\theta}_{b}\right)^{-1}
$$

be the likelihood based on the $n_{j}$ points where $X\left(s_{j}\right)>u$, with $f_{Z^{j}}$ the density of $Z^{0}$ at the observation locations excluding $s_{j}$ when $s_{0}=s_{j}$. We estimate $\boldsymbol{\theta}$ from the composite likelihood over all $d$ observed sites $L(\boldsymbol{\theta})=\prod_{j=1}^{d} L_{j}(\boldsymbol{\theta})$. The likelihood is composite, since any realizations of $X$ that are larger than $u$ at multiple sites will be counted more than once in the likelihood, with different conditioning sites. An alternative approach may be to condition on $\left\{X\left(s_{j}\right)>u, X\left(s_{j}\right)=\max _{k}\left(X\left(s_{k}\right)\right)\right\}$, but this leads to different limit processes $Z^{j}$ in general, so we do not adopt this strategy here. Assessment of parameter uncertainty can be undertaken by use of bootstrap techniques; see Section 6 .

### 4.2 Computation time

An advantage of the conditional approach over other models is that the likelihood does not require integrals as in the censored likelihoods commonly used for asymptotically dependent or asymptotically independent models (e.g., Wadsworth and Tawn, 2014; Thibaud and Opitz, 2015; Huser et al., 2017; Huser and Wadsworth, 2018). Such integrals are typically the limiting factor in the number of sites to which one can fit the model, often meaning a reasonable upper limit is $20-30$ sites.

Figure 1 displays example computation times for the evaluation of a single likelihood $L_{j}$, and the full composite likelihood for different dimensions $d$ and different numbers of average repetitions $n_{j}$, both for Model 3. The boxplots are created from 50 repetitions on different data, and the figure is intended as a rough guide to how many dimensions one might reasonably attempt to optimize a likelihood in, based on our R code implementation, which makes use of the mvnfast R library (Fasiolo, 2016). Computation times for the composite likelihood generally slightly exceed $d$ times the computation for the likelihood conditioning at a single site. Based on these findings, it is perfectly feasible to optimize a 100-200 dimensional likelihood, with a moderate number of repetitions, with no special computing power. Higher-dimensional optimization is certainly possible with modest additional computing power or more computationally efficient code.


Figure 1: Left: computation time for a likelihood conditioning on a single site exceeding the threshold, for different numbers of sites, $d$, and average numbers of exceedances at the conditioning site, $n$. Right: computation time for a composite likelihood over all sites.

## 5 Simulation

Inference on quantities of interest, such as the probabilities of different extreme events, can be made via simulation. Given the nature of the model and assumptions, simulation conditional on being extreme at a particular location $s_{0}$ is straightforward, and the algorithm is outlined in Section 5.1. In many circumstances, it is desirable to condition instead on the process being extreme at some part of the domain, but not at a specific location. We address methods for this in Section 5.2.

### 5.1 Simulation given an extreme at a specified site

Under assumption (1) and model (14), simulation of $X(s)$, given an extreme value above a threshold $v \geq u$ at a specific location $s_{0}$, proceeds as follows:

## Algorithm 1.

1. Generate $X\left(s_{0}\right) \mid X\left(s_{0}\right)>v \sim \operatorname{Exp}(1)$
2. Independently of $X\left(s_{0}\right)$, generate $\left\{Z^{0}(s): s \in \mathcal{S}\right\}$ from the model as specified in Section 3.2.
3. Set $\left\{X(s) \mid X\left(s_{0}\right)>v: s \in \mathcal{S}\right\}=\left\{a_{s-s_{0}}\left(X\left(s_{0}\right)\right)+b_{s-s_{0}}\left(X\left(s_{0}\right)\right) Z^{0}(s): s \in \mathcal{S}\right\}$.

In practice, Algorithm 1 is used to simulate at a finite $m$-dimensional collection of sites $\left\{s_{i_{1}}, \ldots, s_{i_{m}}\right\}$, where the observation sites $\left\{s_{1}, \ldots, s_{d}\right\}$ may or may not be included. We denote the $m$-dimensional distribution of $X(s) \mid X\left(s_{0}\right)>v$, obtained via the approximation (14), by $Q_{0}^{v}$. Further, denote by $M=\left\{i_{1}, \ldots, i_{m}\right\}$ the set of indices corresponding to the simulation sites, and by $D=\{1, \ldots, d\}$ the indices of the observation locations,
where we may have $M=D$. Letting $\boldsymbol{X}_{M}=\left\{X\left(s_{i_{1}}\right), \ldots, X\left(s_{i_{m}}\right)\right\}$, Algorithm 1 can be used to estimate quantities of the form

$$
\begin{equation*}
\mathrm{E}\left\{g\left(\boldsymbol{X}_{M}\right) \mid X\left(s_{0}\right)>v\right\}=\mathrm{E}_{Q_{0}^{v}}\left\{g\left(\boldsymbol{X}_{M}\right)\right\}, \tag{17}
\end{equation*}
$$

for some $g: \mathbb{R}^{m} \rightarrow \mathbb{R}^{l}, l \geq 1$. Equation (17) is convenient if one is either interested in conditioning on a specific location, or if $g(\cdot)$ is an indicator function for an event that involves $X\left(s_{0}\right)>v^{\prime} \geq v$, since then the unconditional probability of this event is easily estimated using the exponential distribution of $X\left(s_{0}\right)$. As an illustration, suppose that the simulation sites are the observation sites, let $\mathbb{1}(A)$ be the indicator function for the occurrence of event $A$, and $g(\boldsymbol{X})=\mathbb{1}\left(X\left(s_{1}\right)>v, \ldots, X\left(s_{d}\right)>v\right)$. Then, taking $s_{0}$ as any $s_{j}, j=1, \ldots, d$ we have

$$
\mathrm{P}\left(X\left(s_{1}\right)>v, \ldots, X\left(s_{d}\right)>v\right)=\mathrm{E}_{Q_{0}^{v}}\left\{\mathbb{1}\left(X\left(s_{1}\right)>v, \ldots, X\left(s_{d}\right)>v\right)\right\} \times e^{-v} .
$$

### 5.2 Simulation given an extreme in the domain

In place of conditioning on a single site being extreme, one may often be interested in estimating quantities of the form

$$
\begin{equation*}
\mathrm{E}\left\{g\left(\boldsymbol{X}_{M}\right) \mid \max _{i \in M^{\prime}} X\left(s_{i}\right)>v\right\} \tag{18}
\end{equation*}
$$

where $M^{\prime}$ is a set of indices that could be a subset of, equal to, or completely disjoint from, $M$. The case most likely to be of interest is $M^{\prime}=M$, so that if $m=|M|$ is large and the locations suitably arranged, $\left\{\max _{i \in M^{\prime}} X\left(s_{i}\right)>v\right\} \approx\left\{\sup _{\mathcal{S}} X(s)>v\right\}$. In what follows, we focus on $M^{\prime}=M$, and initially consider the simulation threshold $v=u$.

We denote the distribution of $\boldsymbol{X} \mid \max _{i \in M} X\left(s_{i}\right)>u$ by $P^{u}$, abbreviate $X\left(s_{i}\right)=X_{i}, i \in M$ or $i \in D$, and let $Q^{u}:=\sum_{i \in M} \pi_{i} Q_{i}^{u}$, where $Q_{i}^{u}$ represents $Q_{0}^{u}$ at $s_{0}=s_{i}$ and

$$
\pi_{i}=\frac{\mathrm{P}\left(X_{i}>u\right)}{\sum_{i \in M} \mathrm{P}\left(X_{i}>u\right)}
$$

which equals $1 / m$ if the margins are identical. In other words $Q^{u}$ is the mixture distribution formed by selecting each $Q_{i}^{u}$ with probability $\pi_{i}$. The support of $Q^{u}$ is $\left\{\boldsymbol{x} \in \mathbb{R}^{m}: \max _{i \in M} x_{i}>u\right\}$. We note that because we assume stationarity of $X$, with continuous margins on $\mathbb{R}_{+}$, all margins are equal and some of the following algorithms can be simplified. However, we give the more general formulation below.

### 5.2.1 Rejection sampling

Keef et al. (2013) propose a method to simulate from the distribution of $\boldsymbol{X}_{D} \mid \max _{i \in D} X\left(s_{i}\right)>u$, i.e., where $M=D=\{1, \ldots, d\}$ and the simulation sites are the set of observation locations. Note the partition

$$
\left\{\boldsymbol{x} \in \mathbb{R}^{d}: \max _{i \in D} x_{i}>u\right\}=\cup_{j=1}^{d}\left\{\boldsymbol{x} \in \mathbb{R}^{d}: x_{j}>u, x_{j}=\max _{i \in D} x_{i}\right\}
$$

and that $P^{u}=\sum_{j \in D} \tilde{\pi}_{j}^{u} Q_{j, \max }^{u}$, where $Q_{j, \text { max }}^{u}$ is the distribution of $\boldsymbol{X} \mid X_{j}>u, X_{j}=\max _{i \in D} X_{i}$, and $\tilde{\pi}_{j}^{u}=$ $\mathrm{P}\left(X_{j}=\max _{i \in D} X_{i} \mid \max _{i \in D} X_{i}>u\right)$. Their algorithm is as follows:

## Algorithm 2.

1. Estimate $\tilde{\pi}_{j}^{u}$ using empirical proportions
2. To simulate from $Q_{j, \max }^{u}$, draw from $Q_{j}^{u}$ and reject unless $X_{j}=\max _{i \in D} X_{i}$
3. Simulate from $P^{u}$ by drawing from $Q_{j, \max }^{u}$ with probability $\tilde{\pi}_{j}^{u}$

This rejection method leads directly to draws from (an estimate of) $P^{u}$, but has substantial drawbacks:
(i) One can only condition on being large at the observation locations, rather than any set of locations;
(ii) For high dimensions, the number of rejections to get a single draw from $Q_{j, \text { max }}^{u}$ may be very high;
(iii) If one wishes to simulate from $P^{v}, v \gg u$, one needs to firstly simulate from $P^{u}$ and use the resulting draws to estimate the probabilities $\tilde{\pi}_{j}^{v}$, before proceeding with the rejection sampling approach to simulate $Q_{j, \text { max }}^{v}$;
(iv) If the maximum never occurs at a particular site in the sample, it will not occur there in simulations either.

The final point can be addressed by simulating from the model to estimate $\tilde{\pi}_{j}^{u}$, at the cost of adding complexity to the algorithm. We propose an approach to estimate quantities of the form (18) via importance sampling, that avoids all of these drawbacks.

### 5.2.2 Importance sampling

Compared to $P^{u}$, the distribution $Q^{u}$ samples more frequently in regions where multiple variables are extreme. However, this frequency is tractable: in the region where exactly $k$ variables are larger than $u, Q^{u}$ samples $k$ times more observations than $P^{u}$, because there are $k$ distributions $Q_{j}^{u}$ covering this area.

To formalize this connection, let $P$ be a probability measure on $\mathcal{B}\left(\mathbb{R}^{m}\right)$, the Borel sigma algebra of $\mathbb{R}^{m}$, for which $P\left(X_{i}>u\right)>0$ for all $i \in M$. Letting $R_{\max }=\left\{\boldsymbol{x} \in \mathbb{R}^{m}: \max _{i \in M} x_{i}>u\right\}$, and $\mathrm{E}, \mathrm{E}_{P^{u}}, \mathrm{E}_{Q_{j}^{u}}$ denote expectation with respect to $P, P^{u}$ and $Q^{u}$ respectively, we can express

$$
\begin{equation*}
\mathrm{E}_{P^{u}}\left\{g\left(\boldsymbol{X}_{M}\right)\right\}=\frac{\mathrm{E}\left\{g\left(\boldsymbol{X}_{M}\right) \mathbb{1}\left(\boldsymbol{X}_{M} \in R_{\max }\right)\right\}}{\mathrm{E}\left\{\mathbb{1}\left(\boldsymbol{X}_{M} \in R_{\max }\right)\right\}}, \quad \quad \mathrm{E}_{Q_{i}^{u}}\left\{g\left(\boldsymbol{X}_{M}\right)\right\}=\frac{\mathrm{E}\left\{g\left(\boldsymbol{X}_{M}\right) \mathbb{1}\left(X_{i}>u\right)\right\}}{\mathrm{E}\left\{\mathbb{1}\left(X_{i}>u\right)\right\}} \tag{19}
\end{equation*}
$$

and $\mathrm{E}_{Q^{u}}\{g(\boldsymbol{X})\}=\sum_{i \in M} \pi_{i} \mathrm{E}_{Q_{i}^{u}}\{g(\boldsymbol{X})\}$. Further note that $R_{\max }=\cup_{K} R_{K}$, where

$$
R_{K}=\left\{\boldsymbol{x} \in \mathbb{R}^{m}: x_{k}>u, \text { for all } k \in K, \text { and } x_{l} \leq u \text { for all } l \in M \backslash K\right\}, \quad K \subseteq M,
$$

and the union is over all $K$ in the power set of $M$, minus the empty set, with all $R_{K}$ disjoint.
Proposition 2. For any random vector $\boldsymbol{X}_{M} \in \mathbb{R}^{m}$, for which $P\left(X_{i}>u\right)>0$ for all $i \in M$, and any function $g: \mathbb{R}^{m} \rightarrow \mathbb{R}^{l}$
$E_{P^{u}}\left\{g\left(\boldsymbol{X}_{M}\right)\right\}=\frac{\sum_{i \in M} P\left(X_{i}>u\right)}{P\left(\max _{i \in M} X_{i}>u\right)} \sum_{K} E_{Q^{u}}\left\{g\left(\boldsymbol{X}_{M}\right) \mathbb{1}\left(\boldsymbol{X}_{M} \in R_{K}\right) /|K|\right\}=\frac{\sum_{K} E_{Q^{u}}\left\{g\left(\boldsymbol{X}_{M}\right) \mathbb{1}\left(\boldsymbol{X}_{M} \in R_{K}\right) /|K|\right\}}{\sum_{K} E_{Q^{u}}\left\{\mathbb{1}\left(\boldsymbol{X}_{M} \in R_{K}\right) /|K|\right\}}$.
Proof. We have

$$
\begin{equation*}
\mathrm{E}_{P^{u}}\left\{g\left(\boldsymbol{X}_{M}\right)\right\}=\frac{\mathrm{E}\left\{g\left(\boldsymbol{X}_{M}\right) \mathbb{1}\left(\boldsymbol{X}_{M} \in R_{\max }\right)\right\}}{\mathrm{E}\left\{\mathbb{1}\left(\boldsymbol{X}_{M} \in R_{\max }\right)\right\}}=\frac{\sum_{K} \mathrm{E}\left\{g\left(\boldsymbol{X}_{M}\right) \mathbb{1}\left(\boldsymbol{X}_{M} \in R_{K}\right)\right\}}{P\left(\max _{i \in M} X_{i}>u\right)} \tag{20}
\end{equation*}
$$

whilst for any $j \in K$ we also have

$$
\mathrm{E}\left\{g\left(\boldsymbol{X}_{M}\right) \mathbb{1}\left(\boldsymbol{X}_{M} \in R_{K}\right)\right\}=\frac{\mathrm{E}\left\{g\left(\boldsymbol{X}_{M}\right) \mathbb{1}\left(\boldsymbol{X}_{M} \in R_{K}\right) \mathbb{1}\left(X_{j}>u\right)\right\}}{\mathrm{E}\left\{\mathbb{1}\left(X_{j}>u\right)\right\}} \mathrm{E}\left\{\mathbb{1}\left(X_{j}>u\right)\right\} .
$$

Averaging over each $j \in K$ we get

$$
\begin{align*}
\mathrm{E}\left\{g\left(\boldsymbol{X}_{M}\right) \mathbb{1}\left(\boldsymbol{X}_{M} \in R_{K}\right)\right\} & =\frac{1}{|K|} \sum_{j \in K} \frac{\mathrm{E}\left\{g\left(\boldsymbol{X}_{M}\right) \mathbb{1}\left(\boldsymbol{X}_{M} \in R_{K}\right) \mathbb{1}\left(X_{j}>u\right)\right\}}{\mathrm{E}\left\{\mathbb{1}\left(X_{j}>u\right)\right\}} \mathrm{E}\left\{\mathbb{1}\left(X_{j}>u\right)\right\} \\
& =\frac{1}{|K|} \sum_{j \in M} \frac{\mathrm{E}\left\{g\left(\boldsymbol{X}_{M}\right) \mathbb{1}\left(\boldsymbol{X}_{M} \in R_{K}\right) \mathbb{1}\left(X_{j}>u\right)\right\}}{\mathrm{E}\left\{\mathbb{1}\left(X_{j}>u\right)\right\}} \pi_{j} \sum_{i \in M} \mathrm{P}\left(X_{i}>u\right), \tag{21}
\end{align*}
$$

the second line following since for $j \in M \backslash K, \mathrm{E}\left\{g\left(\boldsymbol{X}_{M}\right) \mathbb{1}\left(\boldsymbol{X}_{M} \in R_{K}\right) \mathbb{1}\left(X_{j}>u\right)\right\}=0$, and using the definition of $\pi_{j}$. Putting (20) and (21) together

$$
\begin{aligned}
\mathrm{E}_{P^{u}}\left\{g\left(\boldsymbol{X}_{M}\right)\right\} & =\frac{\sum_{i \in M} \mathrm{P}\left(X_{i}>u\right)}{P\left(\max _{i \in M} X_{i}>u\right)} \sum_{K} \frac{1}{|K|} \sum_{j \in M} \pi_{j} \frac{\mathrm{E}\left\{g\left(\boldsymbol{X}_{M}\right) \mathbb{1}\left(\boldsymbol{X}_{M} \in R_{K}\right) \mathbb{1}\left(X_{j}>u\right)\right\}}{\mathrm{E}\left\{\mathbb{1}\left(X_{j}>u\right)\right\}} \\
& =\frac{\sum_{i \in M} \mathrm{P}\left(X_{i}>u\right)}{P\left(\max _{i \in M} X_{i}>u\right)} \sum_{K} \mathrm{E}_{Q^{u}}\left\{g\left(\boldsymbol{X}_{M}\right) \mathbb{1}\left(\boldsymbol{X}_{M} \in R_{K}\right) /|K|\right\}
\end{aligned}
$$

which gives the first equality. For the second equality, note that

$$
1=\mathrm{E}_{P^{u}}\left\{\mathbb{1}\left(\boldsymbol{X}_{M} \in R_{\max }\right)\right\}=\frac{\sum_{i \in M} \mathrm{P}\left(X_{i}>u\right)}{P\left(\max _{i \in M} X_{i}>u\right)} \sum_{K} \mathrm{E}_{Q^{u}}\left\{\mathbb{1}\left(\boldsymbol{X}_{M} \in R_{\max }\right) \mathbb{1}\left(\boldsymbol{X}_{M} \in R_{K}\right) /|K|\right\}
$$

and dividing through by this gives the result.
We note that there is nothing specific to our general assumptions or extreme-value theory in this proposition: it holds for any random vector and any $u$ for which $P\left(X_{j}>u\right)$ for all $j$, whether $u$ is extreme or not. For our purposes, $u$ is a high marginal threshold and the distribution $Q^{u}$ is obtained through asymptotic theory via $Q_{j}^{u}$, which can be simulated from as outlined in Section 5.1. Since we have continuous margins on $\mathbb{R}_{+}$, $\mathrm{P}\left(X_{j}>v\right)>0$ for all $j$ and all $v \geq u$, so the result can be applied at higher thresholds.

Thanks to Proposition 2, we have the following algorithm to estimate $\mathrm{E}\left\{g\left(\boldsymbol{X}_{M}\right) \mid \max _{i \in M} X_{i}>v\right\}=$ $\mathrm{E}_{P^{v}}\left\{g\left(\boldsymbol{X}_{M}\right)\right\}:$

## Algorithm 3.

1. With probability $\pi_{j}=1 / m$, draw $\boldsymbol{X}_{M}$ from $Q_{j}^{v}, j \in M$
2. Repeat step $1 n$ times to get $n$ draws $\boldsymbol{X}_{M}^{1}, \ldots, \boldsymbol{X}_{M}^{n}$ from $Q^{v}$
3. Estimate the expectation $\mathrm{E}_{P^{v}}\left\{g\left(\boldsymbol{X}_{M}\right)\right\}$ by

$$
\begin{equation*}
\frac{\sum_{i=1}^{n} g\left(\boldsymbol{X}_{M}^{i}\right) /\left|\left\{j: X_{j}^{i}>v\right\}\right|}{\sum_{i=1}^{n} 1 /\left|\left\{j: X_{j}^{i}>v\right\}\right|}, \quad \boldsymbol{X}_{M}^{i} \sim Q^{v} \tag{22}
\end{equation*}
$$

We note that the natural estimate of $\mathrm{E}_{P^{v}}\left\{g\left(\boldsymbol{X}_{M}\right)\right\}$ from equation (19) would look like

$$
\frac{\sum_{K} \sum_{i=1}^{n} g\left(\boldsymbol{X}_{M}^{i}\right) \mathbb{1}\left(\boldsymbol{X}_{M}^{i} \in R_{K}\right) /|K|}{\sum_{K} \sum_{i=1}^{n} \mathbb{1}\left(\boldsymbol{X}_{M}^{i} \in R_{K}\right) /|K|}, \quad \boldsymbol{X}^{i} \sim Q^{v}
$$

but since it is only the size of $K$ that matters we do not need to sum over all $K$ but simply divide by the number of variables exceeding the threshold. Furthermore, if it is desired to have realizations from the distribution conditioning upon the maximum being large, i.e., of $\boldsymbol{X}_{M} \mid \max _{i \in M} X_{i}>u$, this can be achieved approximately by using the importance weights in expression (22). That is, we sub-sample $n^{\prime}$ realizations from the collection $\left\{\boldsymbol{X}^{i}\right\}_{i=1}^{n}$ with probabilities proportional to $\left\{1 /\left|\left\{j: X_{j}^{i}>v\right\}\right|\right\}_{i=1}^{n}$.

### 5.3 Conditional infill simulation of an existing event

Given the observation of a process $X(s)$ at a collection of sites $\left\{s_{1}, \ldots, s_{d}\right\}$, we may wish to simulate $X$, conditionally upon the values of the observed process when extreme for at least one site, at an alternative collection of sites $\left\{s_{i_{1}}, \ldots, s_{i_{l}}\right\}$, where we let $L=\left\{k_{1}, \ldots, k_{l}\right\}$ with $L \cap D=\emptyset$. One could also condition on the sites in $D^{\prime} \subset D$ take $L=D \backslash D^{\prime}$, if there are missing data, for example, or for checking model fit. In contrast to purely Gaussian models, conditional simulation is more challenging from models tailored to spatial extremes, although it is possible for models based on elliptical processes such as those described in Huser et al. (2017). An algorithm was proposed for max-stable processes by Dombry et al. (2013).

Consider a realization $X_{i}(s)$, with $X_{i}\left(s_{j}\right)>u$ for all $j \in J_{i} \subseteq D$, and $X_{i}\left(s_{j}\right)<u$ for all $j \notin J_{i}$. For each $j \in J_{i}$, we have

$$
Z_{i}^{j}(s)=\frac{X_{i}(s)-a_{s-s_{j}}\left(X_{i}\left(s_{j}\right)\right)}{b_{s-s_{j}}\left(X_{i}\left(s_{j}\right)\right)}
$$

which, following the discussion in Section 3.2, is modelled as a (marginally-transformed) Gaussian process. Simulation of $\left\{Z^{j}\left(s_{i_{1}}\right), \ldots, Z^{j}\left(s_{i_{l}}\right)\right\} \mid\left\{Z^{j}\left(s_{1}\right), \ldots, Z^{j}\left(s_{d}\right)\right\}$ can thus be achieved exploiting conditional simulation from a Gaussian process. Simplifying notation slightly, let $\left(\boldsymbol{Z}_{L}, \boldsymbol{Z}_{D}\right)$ represent an $(l+d)$-dimensional random vector from the multivariate Gaussian with mean $\boldsymbol{\mu}$ and covariance matrix $\Sigma$, and transformed to have deltaLaplace margins with mean vector $\boldsymbol{\mu}^{\delta}$ and scale parameter vector $\boldsymbol{\sigma}^{\delta}$. Partition $\boldsymbol{\mu}=\left(\boldsymbol{\mu}_{L}, \boldsymbol{\mu}_{D}\right)$, with $\boldsymbol{\mu}_{L} \in \mathbb{R}^{l}$, $\boldsymbol{\mu}_{D} \in \mathbb{R}^{d}$, and similarly for other vectors, whilst $\Sigma_{L D} \in \mathbb{R}^{l \times d}$ etc., represent the partitioned $\Sigma$. Let $F_{\delta}, \Phi, f_{\delta}$ and $\phi$ represent the univariate distribution functions and densities, respectively, of the delta-Laplace and Gaussian distributions, where $F_{\delta}\left(\boldsymbol{z}_{L} ; \boldsymbol{\mu}_{L}^{\delta}, \boldsymbol{\sigma}_{L}^{\delta}\right)=\left(F_{\delta}\left(z_{L, k_{1}} ; \mu_{L, k_{1}}^{\delta}, \sigma_{L, k_{1}}^{\delta}\right), \ldots, F_{\delta}\left(z_{L, k_{l}} ; \mu_{L, k_{l}}^{\delta}, \sigma_{L, k_{l}}^{\delta}\right)\right)^{\top}$ etc., and denote by $\phi_{l}(\cdot ; \boldsymbol{m}, \Omega)$ the $l$-dimensional multivariate Gaussian density with mean $\boldsymbol{m}$ and covariance $\Omega$.

The conditional density of $\left(\boldsymbol{Z}_{L} \mid \boldsymbol{Z}_{D}=\boldsymbol{z}_{D}\right)$, denoted $f_{Z_{L} \mid Z_{D}}$, can be expressed
$f_{Z_{L} \mid Z_{D}}\left(\boldsymbol{z}_{L} \mid \boldsymbol{z}_{D}\right)=\phi_{l}\left[\Phi^{-1}\left\{F_{\delta}\left(\boldsymbol{z}_{L} ; \boldsymbol{\mu}_{L}^{\delta}, \boldsymbol{\sigma}_{L}^{\delta}\right) ; \boldsymbol{\mu}_{L}, \boldsymbol{\sigma}_{L}\right\} ; \boldsymbol{\mu}_{L \mid D}, \Sigma_{L \mid D}\right] \prod_{k \in L} \frac{f_{\delta}\left(z_{L, k} ; \mu_{k}^{\delta}, \sigma_{k}^{\delta}\right)}{\phi\left(\Phi^{-1}\left(F_{\delta}\left(z_{L, k} ; \mu_{L, k}^{\delta}, \sigma_{L, k}^{\delta}\right) ; \mu_{L, k}, \sigma_{L, k}\right) ; \mu_{L, k}, \sigma_{L, k}\right)}$,
with

$$
\begin{aligned}
& \boldsymbol{\mu}_{L \mid D}=\boldsymbol{\mu}_{L}-\Sigma_{L D} \Sigma_{D, D}^{-1}\left(\boldsymbol{\mu}_{D}-\Phi^{-1}\left\{F_{\delta}\left(\boldsymbol{z}_{D} ; \boldsymbol{\mu}_{D}^{\delta}, \boldsymbol{\sigma}_{D}^{\delta}\right) ; \boldsymbol{\mu}_{D}, \boldsymbol{\sigma}_{D}\right\}\right) \\
& \Sigma_{L \mid D}=\Sigma_{L L}-\Sigma_{L D} \Sigma_{D, D}^{-1} \Sigma_{D L} .
\end{aligned}
$$

To simulate from density (23):

## Algorithm 4.

1. Given $\boldsymbol{z}_{D}, \boldsymbol{\mu}, \Sigma, \boldsymbol{\mu}_{\delta}, \boldsymbol{\sigma}_{\delta}$, calculate $\boldsymbol{\mu}_{L \mid D}$ and $\Sigma_{L \mid D}$
2. Simulate $\boldsymbol{Y}_{L}$ from the $l$-dimensional Gaussian with mean $\boldsymbol{\mu}_{L \mid D}$ and covariance $\Sigma_{L \mid D}$


Figure 2: Example of conditional simulation. A realization of $X$ (black solid line) is displayed at 10 sites but the information from three sites (depicted by circles) is treated as missing. The red lines depict conditional simulations based on the other seven values, and being extreme at one of the sites above the modelling threshold. The model is the fitted model from Section 6, fitted to all observations.
3. Apply the transformation $\boldsymbol{Z}_{L}=F_{\delta}^{-1}\left\{\Phi\left(\boldsymbol{Y}_{L} ; \boldsymbol{\mu}_{L}, \boldsymbol{\sigma}_{L}\right) ; \boldsymbol{\mu}_{L}^{\delta}, \boldsymbol{\sigma}_{L}^{\delta}\right\}$.

Note the final step of Algorithm 4 is not a typical application of the probability integral transform, since the margins of $\boldsymbol{Y}_{L}$ do not have location and scale parameters $\boldsymbol{\mu}_{L}, \boldsymbol{\sigma}_{L}$, but rather $\boldsymbol{\mu}_{L \mid D}, \boldsymbol{\sigma}_{L \mid D}$ where the latter is the square root of the diagonal of $\Sigma_{L \mid D}$. When $\delta=2$ and $\boldsymbol{\sigma}_{\delta}=\boldsymbol{\sigma}\{\Gamma(1 / 2) / \Gamma(3 / 2)\}^{1 / 2}=2^{1 / 2} \boldsymbol{\sigma}$, then this is just standard conditional simulation from the Gaussian.

Once draws from $\left\{Z^{j}\left(s_{k_{1}}\right), \ldots, Z^{j}\left(s_{k_{l}}\right)\right\} \mid\left\{Z^{j}\left(s_{1}\right), \ldots, Z^{j}\left(s_{d}\right)\right\}$ have been made, then the process $X(s) \mid X_{i}\left(s_{j}\right)>$ $u$ is recovered by setting

$$
X\left(s_{k}\right) \mid X_{i}\left(s_{j}\right)>u=a_{s_{k}-s_{j}}\left(X_{i}\left(s_{j}\right)\right)+b_{s_{k}-s_{j}}\left(X_{i}\left(s_{j}\right)\right) Z^{j}\left(s_{k}\right), \quad k \in L
$$

and this can be done for each $j \in J_{i}$. That is, we are conditioning both on the values at observed sites, and on the process being extreme at a specific site $s_{j} \in J_{i}$.

The conditional simulation using Algorithm 4 is illustrated in Figure 2, using a subset of the Australian temperature data analyzed in Section 6. For these data, which are gridded and complete, conditional simulation may serve to help understand the fit of the model, but for other datasets it could be useful to deal with missing observations or to simulate at unobserved sites. The intended use of the conditional simulation may dictate whether one of these sites is more interesting or whether inference should be combined across all sites in $J_{i}$.

## 6 Australian temperature extremes

### 6.1 Data

We analyze temperature data in Australia from the HadGHCND global gridded dataset (Caesar et al., 2006). These are observed data translated onto a relatively coarse $2.5^{\circ} \times 3.75^{\circ}$ grid, with 72 points covering Australia; see the top-left panel of Figure 3 for an illustration. A complete record of daily maximum temperatures is available from $1 / 1 / 1957-31 / 12 / 2014$, and we focus on the summer temperatures recorded in December, January and February. This yields a total of 5234 days of observations at the 72 locations. The same data, recorded until the end of 2011, were analyzed in Winter et al. (2016) using non-spatial models, where the focus lay on investigating the effects of the El Niño Southern Oscillation (ENSO).

We transform the marginal distributions to unit Laplace, using the empirical distribution function. The choice of the double exponential-tailed Laplace distribution is motivated by the fact that we may anticipate independence at long range, and by allowing the margins of $Z^{0}$ to follow distribution (16), this feature can be captured. An alternative to empirical transformation is to use a semi-parametric marginal transformation, modelling the upper tail with a univariate generalized Pareto distribution. In high-dimensional datasets this becomes more laborious as one needs to select a threshold and check the model at each site; failing to do so can lead to poor marginal fits that may impact upon dependence structure estimation. On the other hand, the semi-parametric transformation is more useful for back-transforming simulations to the original scale in the extremes.

### 6.2 Exploratory analysis

Exploratory analysis indicates anisotropy and some spatial non-stationarity. This latter issue has received relatively little attention in an extreme value context, although Huser and Genton (2016) consider non-stationary max-stable processes. In more classical spatial modelling, one approach to dealing with non-stationarity is to deform the coordinate system, as suggested by Sampson and Guttorp (1992). The essential idea of this technique is to transform from one coordinate system (the "G-plane") in which non-stationarity is detected, to another configuration (the "D-plane"), where stationarity might be more reasonably assumed. We choose here to adopt such an approach, using the methodology of Smith (1996). Denote by $\tau: \mathbb{R}^{2} \rightarrow \mathbb{R}^{2}$ the mapping from the G-plane to D-plane. The technique of Smith (1996) is designed for estimation on all data and is driven by the covariance matrix: the task is to find a deformation function $\tau$ such that $\operatorname{Cov}\left\{Y\left(\tau\left(s_{1}, s_{2}\right)\right), Y\left(\tau\left(s_{1}, s_{2}\right)\right)\right\}$ only depends on $s_{1}, s_{2}$ through $s_{1}-s_{2}$. However, if patterns of non-stationarity are similar in the extremes to the body then this procedure should improve estimation in the extremes nonetheless. Assuming $d$ locations in the G-plane, Smith (1996) parameterizes

$$
\begin{equation*}
\tau\left(s_{1}, s_{2}\right)=\left(\kappa^{2} s_{1}+\psi \kappa \lambda s_{2}+\sum_{i=1}^{d} \omega_{1, i} \xi_{i}\left(s_{1}, s_{2}\right), \quad \psi \kappa \lambda s_{1}+\lambda^{2} s_{2}+\sum_{i=1}^{d} \omega_{2, i} \xi_{i}\left(s_{1}, s_{2}\right)\right), \quad \kappa, \lambda>0, \psi \in \mathbb{R} \tag{24}
\end{equation*}
$$

with $\xi_{i}\left(s_{1}, s_{2}\right)=\frac{1}{2}\left\{\left(s_{1}-s_{1, i}\right)^{2}+\left(s_{2}-s_{2, i}\right)^{2}\right\} \log \left\{\left(s_{1}-s_{1, i}\right)^{2}+\left(s_{2}-s_{2, i}\right)^{2}\right\}$, and $\omega_{r, i}, r=1,2$, weights that satisfy $\sum_{i=1}^{d} \omega_{r, i}=\sum_{i=1}^{d} \omega_{r, i} s_{1, i}=\sum_{i=1}^{d} \omega_{r, i} s_{2, i}=0$. In practice, following Smith (1996), we set most $\omega_{r, i}$ to zero, and focus on using a smaller number of geographically dispersed anchor sites to estimate the deformation. Figure 3 displays the G-plane and D-plane, highlighting the sites used in the estimation. Also displayed are estimates of the pairwise coefficient of tail dependence, $\eta$, (Ledford and Tawn, 1996) against distance in each coordinate system. The value of $\eta \in(0,1]$ is higher for stronger dependence; when two sites are asymptotically dependent $\eta=1$, and when two sites are independent, $\eta=1 / 2$. The more stationary the data, the less variability one expects to observe in $\eta$ for a given spatial lag $\|h\|$ : some improvement certainly arises using the D-plane over the G-plane. A number of factors account for this: reduced anisotropy, accounting for differences in distance between $1^{\circ}$ of latitude and longitude, as well as non-stationarity itself. Slightly different D-planes would be estimated using different anchor sites, and there is no clear way to optimize this aspect given the combinatorial possibilities. We proceed using the D-plane coordinates displayed in Figure 3. We also include geometric anisotropy in the models, to pick up any residual effects of anisotropy not accounted for already; see Appendix B for details.

### 6.3 Fitted model and investigation of El Niño effect

After further exploration, Model 3 from Section 3, with $Z^{0}$ defined by a Gaussian process $Z_{G} \mid Z_{G}\left(s_{0}\right)=0$, for which $Z_{G}$ is stationary with powered exponential correlation, marginally transformed to delta-Laplace scale, was deemed to fit the data best. The threshold for observations being taken as extreme at a given site was set at the $97.5 \%$ quantile of the Laplace distribution, which yields an average of 130 exceedances per site. The fit of the model indicates asymptotic independence of the data and independence at long range, which is supported by the estimates of $\eta$ in Figure 3, as well as estimates from a pairwise fit of the model, displayed in Figure 7 of Appendix B. Note that in the bottom-left panel of Figure 3, we observe that $\eta \approx 1 / 2$ after a distance of approximately 4 in the D-plane, which is about the same distance at which $\alpha\left(s-s_{0}\right) \approx 0$ in Figure 7, indicating that this is the approximate range for independence between two sites. Figure 7 also supports parameterizing $\delta$ to decrease from 2 to 1 as distance increases, and so we set $\delta\left(s-s_{0}\right)=1+\exp \{-(\| s-$ $\left.\left.s_{0} \| / \delta_{1}\right)^{\delta_{2}}\right\}$. Parameter estimates are displayed in Table 2; Table 3 in Appendix B recalls the definition of each parameter for convenience. Estimates of $\beta$ were constrained to be less than one, but are very close. The uncertainty in the parameter estimates is visualized in Figure 8 of Appendix B. To produce this figure, we used a stationary bootstrap procedure (Politis and Romano, 1994), resampling entire fields with block lengths following a geometric distribution with mean length ten. The maximum composite likelihood estimates from the full dataset have then been subtracted to create a common scale. The angle for the anisotropy was restricted to be between $(-\pi / 2,0]$ for identifiability.

The assumed model implies that the collection of residual processes $\left\{Z^{j}(s)\right\}_{j=1}^{72}$ have a distribution defined by

1. Dependence: $Z_{G}(s) \mid Z_{G}\left(s_{j}\right)=0$, where $Z_{G}$ is a stationary Gaussian process with mean $\mu$, and covariance function $\sigma^{2} \exp \left\{-(\|h\| / \phi)^{\nu}\right\}$;
2. Margins: delta-Laplace distribution with shape $\delta\left(s-s_{0}\right)$, and location and scale parameter to match that of the conditional field $Z_{G}(s) \mid Z_{G}\left(s_{j}\right)=0$.
One test of model fit therefore is to examine how closely the residual processes obtained by

$$
Z^{j}(s)=\left\{X(s)-\hat{\alpha}\left(s-s_{j}\right) X\left(s_{j}\right)\right\} /\left[1+\left\{\hat{\alpha}\left(s-s_{j}\right) X\left(s_{j}\right)\right\}^{\hat{\beta}}\right]
$$



Figure 3: Top: G-plane and D-plane, with anchor sites for which $\delta_{r, i} \neq 0$ in (24) highlighted. Bottom: estimates of the coefficient of tail dependence, $\eta$, plotted against distance in the coordinate system.
follow this structure. Upon examination the empirical $Z^{j}$, we find a lack of fit, which translates into inadequate reproduction of extreme events. This emanates from an incorrect mean structure. For the possibilities outlined in Section 3.2, the mean of $Z_{G}(s) \mid Z_{G}\left(s_{j}\right)=0$ is either increasing or decreasing from zero, or identically zero, as $\left\|s-s_{j}\right\|$ increases. However, in practice here the mean increases then decreases again towards zero; see Figure 7 . This is indicative of the long-range independence in this dataset, since if $X\left(s_{k}\right)$ is independent of $X\left(s_{j}\right)$, with $\alpha\left(s_{k}-s_{j}\right) \approx 0$, then $Z^{j}\left(s_{k}\right)=X\left(s_{k}\right)$, which has delta-Laplace margins with $\delta=1, \sigma=1, \mu=0$. Possibilities for dealing with this include attempting to parameterize the non-monotonic form, or changing the marginal parameterization of $X$ to have a different mean, which would in turn affect the mean of $Z^{j}$ at long range, and which would not impact the motivation for the model. A further possibility in the current context of gridded data, is to extract the fitted $Z^{j}$ and re-fit the model for these residual processes using the empirical means $\left(1 / n_{j}\right) \sum_{i=1}^{n_{j}} Z_{i}^{j}\left(s_{k}\right), k \in\{1, \ldots, 72\} \backslash j$. A disadvantage of this approach is that it would not allow simulation at a new location without placing further spatial structure on the means. Where this is not an issue, as here, an advantage is that it can help alleviate symptoms of non-stationarity, and we thus adopt this approach. Table 2 also displays the parameter estimates for $Z^{j}$ where the Gaussian process model with delta-Laplace margins has been re-fitted, using the empirical means.

As a further check on the modelling assumptions, we investigate the independence of $X\left(s_{0}\right) \mid X\left(s_{0}\right)>u$ and $Z^{0}(s)$. Figure 5 in Appendix B displays $X\left(s_{j}\right) \mid X\left(s_{j}\right)>u$ plotted against the mean and variance of the corresponding $Z^{j}(s)$ for all conditioning sites $s_{j}$, whilst Figure 6 displays a summary of the Kendall's $\tau$ coefficients. Based on these diagnostics, whilst not a perfect assumption, independence of $X\left(s_{0}\right) \mid X\left(s_{0}\right)>u$ and $Z^{0}(s)$ seems a reasonable working hypothesis.

Using the parameter estimates from the re-fitted $Z^{j}$, and the original estimates for parameters not relating to $Z^{j}$, we employ the importance sampling techniques of Section 5.2 .2 to estimate the expected number of grid locations exceeding a certain quantile, given that at least one location exceeds that quantile. Figure 4 displays results for quantiles ranging from $97.5 \%-99.9999 \%$, with empirical values where available; the in-

Table 2: Parameter estimates to two decimal places. 'Angle' and 'Stretch' refer to geometric anisotropy; see Appendix B for details.

|  | $\kappa$ | $\lambda$ | $\beta$ | $\phi$ | $\nu$ | $\sigma$ | $\mu$ | $\delta_{1}$ | $\delta_{2}$ | Angle |
| ---: | ---: | ---: | ---: | ---: | ---: | ---: | ---: | ---: | ---: | ---: |
| All | 1.81 | 1.63 | 1.00 | 2.47 | 1.88 | 0.88 | -0.09 | 1.30 | 1.71 | -0.81 |
| $Z$ only | - | - | - | 2.55 | 1.86 | 0.96 | - | 1.72 | 2.23 | -0.75 |
| El Niño | 1.78 | 1.59 | 1.00 | 2.59 | 1.86 | 0.88 | -0.06 | 1.12 | 1.51 | -0.96 |
| $Z$ only | - | - | - | 2.66 | 1.84 | 0.97 | - | 1.62 | 2.08 | -0.77 |
| La Niña | 1.79 | 1.62 | 1.00 | 2.52 | 1.86 | 0.87 | -0.04 | 1.25 | 1.66 | -0.94 |
| $Z$ only | - | - | - | 2.62 | 1.83 | 0.97 | - | 1.77 | 2.32 | -0.97 |
| La Nada | 1.82 | 1.65 | 1.00 | 2.47 | 1.87 | 0.87 | -0.11 | 1.35 | 1.76 | -0.75 |
| $Z$ only | - | - | - | 2.57 | 1.85 | 0.96 | - | 1.76 | 2.26 | -0.70 |



Figure 4: Expected number of exceedances of a particular marginal quantile given that there is at least one exceedance of that quantile somewhere over Australia. Points marked $\times$ are this quantity estimated directly from the data, with bars giving approximate $95 \%$ confidence intervals; there are no data to estimate directly above the $99.95 \%$ quantile. Left: thick solid line represents model-based estimation, with estimates from bootstrap samples in grey. Right: different colours denote estimates from the models fitted to all data, as well as that stratified by ENSO covariate.
sample agreement appears good, although there is some question about whether the expected number decreases rapidly enough. Such a summary could not easily be calculated using the pairwise methods in Winter et al. (2016). To do so, one would have to fit $72 \times 71=5112$ pairwise models, calculate the $Z^{j}$ by concatenating empirical residuals, and then using these with the 10,224 estimated parameters to implement the rejection scheme described in Section 5.2.1. Furthermore, the use of empirical residuals only restricts the shape of new events: where sites are effectively independent from the conditioning site, i.e., $a_{s-s_{0}}(x) \approx 0$, simulated new events will look just like past events in these areas.

To investigate the possible effect of ENSO on the spatial extent of high temperature events, separate models were fitted to data from El Niño, La Niña and "La Nada" seasons. El Niño (respectively La Niña) events are defined here as those for which the sea surface temperature (SST) anomaly from 1980-2010 mean levels is higher than $+1^{\circ}$ (respectively lower than $-1^{\circ}$ ); La Nada events are the remaining ones. The anomalies were taken from https://www.esrl.noaa.gov/psd/gcos_wgsp/Timeseries/Data/nino34.long.anom.data. The covariate was defined by averaging monthly SST anomalies over the summer, so that each summer season is either El Niño, La Niña or La Nada. Marginal transformations were made separately for the three categories, whilst dependence parameter estimates are given in Table 2. There is some modest deviation from the combined parameter estimates particularly for the El Niño years where $\delta$ decays more rapidly, indicating slightly increased variability in these years. Figure 4 displays estimates of the expected number of exceedances from the models fit to the different ENSO regimes: La Niña years are slightly higher and El Niño slightly lower, but differences do not appear significant. However, differences in marginal quantiles make the practical interpretation less straightforward. In the South-East of Australia particularly, the high quantiles represent hotter temperatures under El Niño than La Niña, and a temperature that represents an exceedance of a marginal $99.5 \%$ quantile in La Niña conditions - which from Figure 4 occur in around 3.5 grid squares on average - might be closer to a $97.5 \%$ marginal quantile, or even lower, in El Niño conditions, which Figure 4 suggests might affect 5 or more grid squares on average.

## 7 Discussion

The conditional approach to spatial extreme value analysis offers a number of advantages: flexible, asymptoticallymotivated dependence structures that can capture asymptotic (in)dependence; the models can be fit in reasonably high dimensions; and conditional simulation at unobserved locations is simple. The principal drawback of the approach lies in the more complicated interpretation of what constitutes "the model" for a given dataset. When conditioning only on a single site being extreme, interpretation is straightforward. When wishing to condition on any of a set of sites being extreme, we need to combine these individual models in a way that entails no clearly defined overall model for the process given that it is extreme somewhere over space. However, for the purpose of inference on many quantities of interest, the importance sampling techniques ensure that this is not an issue.

Owing to computational limitations, there have been relatively few attempts at high-dimensional inference for extremes. In general, higher-dimensional problems do not only present computational issues, but they are often accompanied by additional modelling challenges. The spatial domain of interest is likely to be larger, leading to potentially more diverse behaviour in the dependence. In the Australian temperature example, we observed non-stationarity, and independence at longer range. These issues are less likely to arise when focusing on smaller areas, and these complexities should be kept in mind as spatial extremes moves in to a higher-dimensional phase.

A natural next step is to extend this approach in to a multivariate or space-time setting. For example, convergence (1) can be generalized by replacing all locations $s \in \mathcal{S}$ by $(s, t) \in \mathcal{S} \times \mathcal{T} \subset \mathbb{R}^{2} \times \mathbb{R}$, with the conditioning location ( $s_{0}, t_{0}$ ). For modelling purposes, there are additional considerations of the dependence regime in both space and time, and how these interact. However, in principle one can have asymptotic (in)dependence in both space and time, with potentially different behaviour in the different dimensions. A simple approach, assuming separability of space and time, would be to take $a_{s-s_{0}, t-t_{0}}(x)=\alpha_{1}\left(s-s_{0}\right) \alpha_{2}\left(t-t_{0}\right)$, with $\alpha_{1}, \alpha_{2}$ as in (15) with different parameter values.

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## Data and code

The data and code for the analysis of Section 6 are available as Supplementary Material. The most recent data can be obtained from https://www.metoffice.gov.uk/hadobs/hadghend/, subject to the conditions detailed at the URL.

## A Additional results and proofs

Before the proof of Proposition 1, Lemma 1 specifies conditions under which there is some flexibility in the normalization leading to convergence.

Lemma 1. Suppose that for $\boldsymbol{a}(v)$ and $\boldsymbol{b}(v)$ with twice-differentiable components $a_{l}, b_{l}$ satisfying $a_{l}^{\prime}(v) \sim \alpha_{l}$, $a_{l}^{\prime \prime}(v)=o(1), b_{l}^{\prime}(v) / b_{l}(v)=o(1)$, as $v \rightarrow \infty$, for $l=1, \ldots, d$,

$$
P\left(\left.\frac{\boldsymbol{V}-\boldsymbol{a}\left(V_{0}\right)}{\boldsymbol{b}\left(V_{0}\right)} \leq \boldsymbol{z} \right\rvert\, V_{0}=v\right)=P\left(\left.\frac{\boldsymbol{V}-\boldsymbol{a}(v)}{\boldsymbol{b}(v)} \leq \boldsymbol{z} \right\rvert\, V_{0}=v\right) \rightarrow G(\boldsymbol{z}), \quad v \rightarrow \infty
$$

and that all first and second order partial derivatives converge, i.e.

$$
\begin{aligned}
\frac{\partial}{\partial z_{k}} P\left(\left.\frac{\boldsymbol{V}-\boldsymbol{a}(v)}{\boldsymbol{b}(v)} \leq \boldsymbol{z} \right\rvert\, V_{0}=v\right) & \rightarrow \frac{\partial}{\partial z_{k}} G(\boldsymbol{z}), \\
\frac{\partial^{2}}{\partial z_{k} \partial z_{l}} P\left(\left.\frac{\boldsymbol{V}-\boldsymbol{a}(v)}{\boldsymbol{b}(v)} \leq \boldsymbol{z} \right\rvert\, V_{0}=v\right) & \rightarrow \frac{\partial^{2}}{\partial z_{k} \partial z_{l}} G(\boldsymbol{z}), \quad v \rightarrow \infty .
\end{aligned}
$$

Then
(i) for $\boldsymbol{h}^{1}(v)=\left(h_{1}^{1}(v), \ldots, h_{d}^{1}(v)\right)^{\top}, \boldsymbol{h}^{2}(v)=\left(h_{1}^{2}(v), \ldots, h_{d}^{2}(v)\right)^{\top}$, with $h_{l}^{1}(v)=o(1), h_{l}^{2}(v)=o(1)$, for all $l=1, \ldots, d$,

$$
P\left(\left.\frac{\boldsymbol{V}-\boldsymbol{a}(v)}{\boldsymbol{b}(v)\left[1+\boldsymbol{h}^{1}(v)\right]}+\boldsymbol{h}^{2}(v) \leq \boldsymbol{z} \right\rvert\, V_{0}=v\right) \rightarrow G(\boldsymbol{z}), \quad v \rightarrow \infty
$$

(ii) for $\boldsymbol{h}^{1}$ as above but $h_{l}^{2}(v)=O(1)$, i.e., some components may be asymptotically non-zero constants, whilst some may converge to zero,

$$
P\left(\left.\frac{\boldsymbol{V}-\boldsymbol{a}(v)}{\boldsymbol{b}(v)\left[1+\boldsymbol{h}^{1}(v)\right]}+\boldsymbol{h}^{2}(v) \leq \boldsymbol{z} \right\rvert\, V_{0}=v\right) \rightarrow G\left(\boldsymbol{z}-\lim _{v \rightarrow \infty} \boldsymbol{h}^{2}(v)\right), \quad v \rightarrow \infty
$$

Proof. (i) Write

$$
\begin{gathered}
\mathrm{P}\left(\left.\frac{\boldsymbol{V}-\boldsymbol{a}(v)}{\boldsymbol{b}(v)} \leq \boldsymbol{z} \right\rvert\, V_{0}=v\right)=F_{V \mid V_{0}}(\boldsymbol{b}(v) \boldsymbol{z}+\boldsymbol{a}(v) \mid v) \\
\mathrm{P}\left(\left.\frac{\boldsymbol{V}-\boldsymbol{a}(v)}{\boldsymbol{b}(v)\left[1+\boldsymbol{h}^{1}(v)\right]}+\boldsymbol{h}^{2}(v) \leq \boldsymbol{z} \right\rvert\, V_{0}=v\right)=F_{V \mid V_{0}}\left(\boldsymbol{b}(v) \boldsymbol{z}+\boldsymbol{a}(v)+\boldsymbol{b}(v) \boldsymbol{h}^{1}(v) \boldsymbol{z}-\boldsymbol{b}(v) \boldsymbol{h}^{2}(v)+O\left(\boldsymbol{b}(v) \boldsymbol{h}^{1}(v) \boldsymbol{h}^{2}(v)\right) \mid v\right)
\end{gathered}
$$

Now write $\boldsymbol{g}(v)=\boldsymbol{b}(v) \boldsymbol{h}^{1}(v) \boldsymbol{z}+\boldsymbol{b}(v) \boldsymbol{h}^{2}(v)+O\left(\boldsymbol{b}(v) \boldsymbol{h}^{1}(v) \boldsymbol{h}^{2}(v)\right)$, with $g_{l}(v)=o\left(b_{l}(v)\right)$, and consider the Taylor expansion
$F_{V \mid V_{0}}(\boldsymbol{b}(v) \boldsymbol{z}+\boldsymbol{a}(v)+\boldsymbol{g}(v) \mid v)=F_{V \mid V_{0}}(\boldsymbol{b}(v) \boldsymbol{z}+\boldsymbol{a}(v) \mid v)+\nabla F_{V \mid V_{0}}(\boldsymbol{b}(v) \boldsymbol{z}+\boldsymbol{a}(v) \mid v)^{\top} \boldsymbol{g}(v)+O\left(\max \left(V_{l} h_{1}^{l}(v), \vee_{l} h_{2}^{l}(v)\right)^{2}\right)$.

The components of $\nabla F_{V \mid V_{0}}(\boldsymbol{b}(v) \boldsymbol{z}+\boldsymbol{a}(v) \mid v)^{\top}$ are $F_{V \mid V_{0}}^{(l)}(\boldsymbol{b}(v) \boldsymbol{z}+\boldsymbol{a}(v) \mid v)$, where $F_{V \mid V_{0}}^{(l)}(\boldsymbol{x} \mid v)=\left.\frac{\partial}{\partial y_{l}} F_{V \mid V_{0}}(\boldsymbol{y} \mid v)\right|_{\boldsymbol{y}=\boldsymbol{x}}$.
The convergence

$$
\frac{\partial}{\partial z_{l}} \mathrm{P}\left(\left.\frac{\boldsymbol{V}-\boldsymbol{a}\left(V_{0}\right)}{\boldsymbol{b}\left(V_{0}\right)} \leq \boldsymbol{z} \right\rvert\, V_{0}=v\right) \rightarrow \frac{\partial}{\partial z_{l}} G(\boldsymbol{z})
$$

is equivalent to

$$
F_{V \mid V_{0}}^{(l)}(\boldsymbol{b}(v) \boldsymbol{z}+\boldsymbol{a}(v) \mid v) b_{l}(v) \rightarrow \frac{\partial}{\partial z_{l}} G(\boldsymbol{z}),
$$

hence

$$
\begin{equation*}
F_{V \mid V_{0}}^{(l)}(\boldsymbol{b}(v) \boldsymbol{z}+\boldsymbol{a}(v) \mid v)=\frac{\partial}{\partial z_{l}} G(\boldsymbol{z})\left\{b_{l}(v)\right\}^{-1}[1+o(1)] \tag{26}
\end{equation*}
$$

with a similar approach for the mixed partial derivatives. Substituting (26) into (25) yields that the second term in the expansion is $O\left(\max \left(\vee_{l} h_{l}^{1}(v), \vee_{l} h_{l}^{2}(v)\right)\right)$, whilst the next order term is $O\left(\max \left(\vee_{l} h_{l}^{1}(v), \vee_{l} h_{l}^{2}(v)\right)^{2}\right)$.
(ii) The proof is very similar to part (i), except the Taylor expansion is about $\boldsymbol{b}(v)\left(\boldsymbol{z}-\lim _{v \rightarrow \infty} \boldsymbol{h}^{2}(v)\right)+\boldsymbol{a}(v)$.

Proof of Proposition 1. For brevity, write $c=-\log (1-\lambda)$ and let $q_{x}^{\star}=\left(\left[c-v_{*}+\min \left\{\min _{l}\left(a_{l}(x)+b_{l}(x) z_{l}\right), x\right\}\right] / \lambda\right)_{+}$ with $y_{+}=\max (y, 0)$. We have

$$
\begin{aligned}
& \lim _{x \rightarrow \infty} \mathrm{P}\left(\left.\frac{\boldsymbol{X}-\boldsymbol{a}\left(X_{0}\right)}{\boldsymbol{b}\left(X_{0}\right)} \leq \boldsymbol{z} \right\rvert\, X_{0}=x\right)= \lim _{x \rightarrow \infty} \int_{0}^{q_{x}^{\star}} \mathrm{P}\left(\frac{\boldsymbol{X}-\boldsymbol{a}\left(X_{0}\right)}{\boldsymbol{b}\left(X_{0}\right)} \leq \boldsymbol{z}, Q=q \mid X_{0}=x\right) \mathrm{d} q \\
&= \lim _{x \rightarrow \infty} \int_{0}^{q_{x}^{\star}} \mathrm{P}\left(\left.\frac{\boldsymbol{X}-\boldsymbol{a}(x)}{\boldsymbol{b}(x)} \leq \boldsymbol{z} \right\rvert\, Q=q, X_{0}=x\right) f_{Q \mid X_{0}}(q \mid x) \mathrm{d} q \\
&= \lim _{x \rightarrow \infty} \int_{0}^{q_{x}^{\star}} \mathrm{P}\left(\left.\frac{\boldsymbol{V}+\lambda q-c-\boldsymbol{a}(v(x)+\lambda q-c)}{\boldsymbol{b}(v(x)+\lambda q-c)} \leq \boldsymbol{z} \right\rvert\, V_{0}=v(x), Q=q\right) \\
& \times f_{Q \mid X_{0}(q \mid x) \mathrm{d} q} \\
&=\lim _{x \rightarrow \infty} \int_{0}^{q_{x}^{\star}} \mathrm{P}\left(\left.\frac{\boldsymbol{V}-\boldsymbol{a}(v(x))+(1-\boldsymbol{\alpha})(\lambda q-c)+o(1)}{\boldsymbol{b}(v(x))[1+O(\nabla \boldsymbol{b}(v(x)) / \boldsymbol{b}(v(x)))]} \leq \boldsymbol{z} \right\rvert\, V_{0}=v(x), Q=q\right) \\
& \times f_{Q \mid X_{0}(q \mid x) \mathrm{d} q}
\end{aligned}
$$

with $v(x)=x-\lambda q+c$ and $f_{Q \mid X_{0}}(q \mid x)$ the conditional density of $Q \mid X_{0}=x$. The integrand is dominated by $f_{Q \mid X_{0}}(q \mid x)$, and
$f_{Q \mid X_{0}}(q \mid x)=\frac{f_{Q, V_{0}}(q, x-\lambda q+c)}{f_{X_{0}}(x)} \sim \frac{e^{-q} e^{-(x-\lambda q+c)} \mathbb{1}(q>0) \mathbb{1}\left(x-\lambda q+c>v_{*}\right)}{e^{-x}} \rightarrow(1-\lambda) e^{-(1-\lambda) q} \mathbb{1}(q>0), \quad x \rightarrow \infty$.
By Lemma 1, as $v \rightarrow \infty$,

$$
\mathrm{P}\left(\left.\frac{\boldsymbol{V}-\boldsymbol{a}(v)+(1-\boldsymbol{\alpha})(\lambda q-c)+o(1)}{\boldsymbol{b}(v)[1+O(\nabla \boldsymbol{b}(v) / \boldsymbol{b}(v))]} \leq \boldsymbol{z} \right\rvert\, V=v, Q=q\right) \rightarrow G\left(\boldsymbol{z}+\frac{(\boldsymbol{\alpha}-1)(\lambda q+\log (1-\lambda))}{\lim _{v \rightarrow \infty} \boldsymbol{b}(v)}\right)
$$

which simplifies when $\lim _{v \rightarrow \infty} b_{l}(v)=\infty$ for all $l$. Dominated convergence then yields the result stated.

Lemma 2 (Convergence of Gaussian partial derivatives). Suppose $\left(\boldsymbol{Y}, Y_{0}\right)^{\top}$ follows a $(d+1)$-dimensional Gaussian distribution, and let $\left(\boldsymbol{V}, V_{0}\right)^{\top}=T\left(\left(\boldsymbol{Y}, Y_{0}\right)^{\top}\right) \in \mathbb{R}_{+}^{d+1}$ be a random vector with unit exponential margins and Gaussian copula. Denote by $\boldsymbol{\rho}_{0}>\mathbf{0}$ the d-vector of correlation parameters between $\boldsymbol{Y}$ and $Y_{0}$. Then for $\boldsymbol{a}(v)=\boldsymbol{\rho}_{0}^{2} v, \boldsymbol{b}(v)=1+\left(\boldsymbol{\rho}_{0} v\right)^{1 / 2}$ and any $r \leq d$,

$$
\frac{\partial^{r}}{\partial z_{1} \cdots \partial z_{r}} P\left(\left.\frac{\boldsymbol{V}-\boldsymbol{a}(v)}{\boldsymbol{b}(v)} \leq \boldsymbol{z} \right\rvert\, V_{0}=v\right) \rightarrow \frac{\partial^{r}}{\partial z_{1} \cdots \partial z_{r}} G(\boldsymbol{z})
$$

Proof. We can express

$$
\begin{aligned}
\mathrm{P}\left(\left.\frac{\boldsymbol{V}-\boldsymbol{a}(v)}{\boldsymbol{b}(v)} \leq \boldsymbol{z} \right\rvert\, V_{0}=v\right) & =\mathrm{P}\left[\boldsymbol{Y} \leq T^{-1}\{\boldsymbol{b}(T(y)) \boldsymbol{z}+\boldsymbol{a}(T(y))\} \mid T\left(Y_{0}\right)=T(y)\right] \\
& =\Phi_{d}\left[T^{-1}\{\boldsymbol{b}(T(y)) \boldsymbol{z}+\boldsymbol{a}(T(y))\}-\boldsymbol{\rho}_{0} y ; \Sigma_{0}\right]
\end{aligned}
$$

where $\Phi_{d}\left(\cdot ; \Sigma_{0}\right)$ is the cdf of the centred $d$-variate Gaussian with covariance matrix $\Sigma_{0}=\left(\rho_{k, l}-\rho_{k, 0} \rho_{l, 0}\right)_{1 \leq k, l \leq d}$ Taking the derivative yields

$$
\begin{align*}
& \frac{\partial^{r}}{\partial z_{1} \cdots \partial z_{r}} \mathrm{P}\left[\boldsymbol{Y} \leq T^{-1}\{\boldsymbol{b}(T(y)) \boldsymbol{z}+\boldsymbol{a}(T(y))\} \mid T\left(Y_{0}\right)=T(y)\right] \\
& \quad=\Phi_{d}^{(1: r)}\left[T^{-1}\{\boldsymbol{b}(T(y)) \boldsymbol{z}+\boldsymbol{a}(T(y))\} ; \Sigma_{0}\right] \prod_{k=1}^{r} \frac{\partial}{\partial z_{k}} T^{-1}\left\{b_{k}(T(y)) z_{k}+a_{k}(T(y))\right\} \tag{27}
\end{align*}
$$

where $\Phi_{d}^{(1: r)}\left(\cdot \mid \Sigma_{0}\right)$ is the $r$ th-order mixed partial derivative of $\Phi_{d}$. The components

$$
\begin{equation*}
\frac{\partial}{\partial z_{k}} T^{-1}\left\{b_{k}(T(y)) z_{k}+a_{k}(T(y))\right\}=\frac{b_{k}(T(y))}{T^{\prime}\left[T^{-1}\left\{b_{k}(T(y)) z_{k}+a_{k}(T(y))\right\}\right]} \tag{28}
\end{equation*}
$$

Now $T(y)=-\log \{1-\Phi(y)\}=y^{2} / 2+O(\log y), y \rightarrow \infty$, whilst $T^{-1}(x)=(2 x)^{1 / 2}+O\left(\log x / x^{1 / 2}\right), x \rightarrow \infty$, and $T^{\prime}(y)=\phi(y) /\{1-\Phi(y)\} \sim y, y \rightarrow \infty$. Further, $b_{k}(T(y))=1+\rho_{0, k} y / \sqrt{2}+O(\log y)$, whilst $a_{k}(T(y))=$ $\rho_{0, k}^{2} y^{2} / 2+O(\log y)$. Combining these,

$$
T^{-1}\left\{b_{k}(T(y)) z_{k}+a_{k}(T(y))\right\}=\rho_{0, k} y+z_{k} / \sqrt{2}+o(1)
$$

and equation (28) converges to $1 / \sqrt{2}$, whilst (27) converges to

$$
\Phi_{d}^{(1: r)}\left(\boldsymbol{z} / \sqrt{2} ;\left.\Sigma\right|_{0}\right) 2^{-r / 2}
$$

which is the $r$ th-order mixed partial derivative of the Gaussian limit distribution.
Remark 1. For application of Proposition 1, we also require convergence of the second derivative $\partial^{2} /\left(\partial z_{k}\right)^{2}$. Iteration of the above manipulations leads to a conclusion that the derivative converges to $\Phi_{d}^{(k k)}\left(\boldsymbol{z} / \sqrt{2} ; \Sigma_{0}\right) / 2$, with $\Phi_{d}^{(k k)}(\boldsymbol{z})=\partial^{2} \Phi_{d}(\boldsymbol{y}) /\left.\left(\partial y_{k}\right)^{2}\right|_{\boldsymbol{y}=\boldsymbol{z}}$.
Proposition 3. Suppose that $\left(X_{1}, X_{2}\right)$ have identical margins with infinite upper endpoint and

1. $\lim _{x \rightarrow \infty} P\left(X_{1}>x \mid X_{2}>x\right)=0$
2. $\lim _{t \rightarrow \infty} P\left[\left\{X_{1}-a\left(X_{2}\right)\right\} / b\left(X_{2}\right)>z \mid X_{2}>t\right]=G(z)$,
with $G$ a non-degenerate distribution function for which $\lim _{z \rightarrow \infty} G(z)=1$. Then if there exists $t_{0}$ such that $a(t)>0$ is monotonically non-decreasing in $t$ for all $t>t_{0}$
(i) If $G(z) \in(0,1)$ for some $z \in(0, \infty), a(t)<t$ for all sufficiently large $t$.
(ii) If $G(z) \in(0,1)$ for all $z \in(0, \infty)$, then $b(t)=o(t), t \rightarrow \infty$.

Proof. (i). When $z>0$,

$$
0 \leq \mathrm{P}\left(X_{1}>a\left(X_{2}\right)+b\left(X_{2}\right) z \mid X_{2}>t\right) \leq \mathrm{P}\left(X_{1}>a\left(X_{2}\right) \mid X_{2}>t\right) \leq \mathrm{P}\left(X_{1}>a(t) \mid X_{2}>t\right)
$$

for sufficiently large $t$. But since $\mathrm{P}\left(X_{1}>t \mid X_{2}>t\right) \rightarrow 0$, we must have $a(t)<t$ for all large $t$ to get convergence to $G(z) \in(0,1)$.
(ii). If $b(t)$ is constant or decreasing then clearly the statement holds, so suppose it is increasing as $t \rightarrow \infty$. Similarly to above

$$
0 \leq \mathrm{P}\left(X_{1}>a\left(X_{2}\right)+b\left(X_{2}\right) z \mid X_{2}>t\right) \leq \mathrm{P}\left(X_{1}>b\left(X_{2}\right) z \mid X_{2}>t\right) \leq \mathrm{P}\left(X_{1}>b(t) z \mid X_{2}>t\right)
$$

for $z>0$ and large $t$. Now if $b(t) \sim c t$ for some $c>0$ then for any $z>1 / c$, and sufficiently large $t$, $0<\mathrm{P}\left(X_{1}>b(t) z \mid X_{2}>t\right) \leq \mathrm{P}\left(X_{1}>t \mid X_{2}>t\right) \rightarrow 0$. But since $G(z) \in(0,1)$ for any $z \in(0, \infty)$, we must have $b(t)=o(t)$.

## B Supporting information for Section 6

## B. 1 Model information

Table 3: Description of parameters in fitted model

| Parameter | Description |
| :--- | :--- |
| $\kappa$ | Shape parameter in $\alpha\left(s-s_{0}\right)$ (eq. (15)) |
| $\lambda$ | Scale parameter in $\alpha\left(s-s_{0}\right)$ (eq. (15)) |
| $\beta$ | Power in $b_{s-s_{0}}=1+a_{s-s_{0}}(x)^{\beta}$ |
| $\phi$ | Scale parameter in Gaussian correlation |
| $\nu$ | Shape parameter in Gaussian correlation |
| $\sigma$ | Scale parameter of $Z_{G}$ used to determine structure of $Z^{j}$ |
| $\mu$ | Mean parameter of $Z_{G}$ used to determine structure of $Z^{j}$ |
| $\delta_{1}$ | Scale parameter in $\delta\left(s-s_{0}\right)=1+\exp \left\{-\left(\left\\|s-s_{0}\right\\| / \delta_{1}\right)^{\delta_{2}}\right\}$ |
| $\delta_{2}$ | Shape parameter in $\delta\left(s-s_{0}\right)=1+\exp \left\{-\left(\left\\|s-s_{0}\right\\| / \delta_{1}\right)^{\left.\delta_{2}\right\}}\right.$ |
| Angle | Angle for geometric anisotropy |
| Stretch | Stretch for geometric anisotropy |

For geometric anisotropy, the coordinate system is changed from $\left(s^{1}, s^{2}\right)^{\top}$ to $\left(\tilde{s}^{1}, \tilde{s}^{2}\right)^{\top}$, via

$$
\binom{\tilde{s}^{1}}{\tilde{s}^{2}}=\left(\begin{array}{cc}
1 & 0 \\
0 & 1 / \text { Stretch }
\end{array}\right)\left(\begin{array}{cc}
\cos (\text { Angle }) & -\sin \text { (Angle) } \\
\cos \text { (Angle) } & \sin \text { (Angle) }
\end{array}\right)\binom{s^{1}}{s^{2}} .
$$

## B. 2 Diagnostic plots and parameter uncertainty



Figure 5: $X^{i}\left(s_{j}\right) \mid X^{i}\left(s_{j}\right)>u$ against the mean (left) and variance (right) of $Z_{i}^{j}(s), i=1, \ldots, n_{j}$. The samples were transformed empirically to a standard uniform scale for each $j=1, \ldots, 72$.


Figure 6: Histograms (red) of Kendall's $\tau$ coefficients for: $X^{i}\left(s_{j}\right) \mid X^{i}\left(s_{j}\right)>u$ and the mean (left) and variance (right) of $Z_{i}^{j}(s), i=1, \ldots, n_{j}$ for the 72 sites, with average sample size $n_{j} \approx 130$. For comparison, in blue is the histogram of the approximate null Kendall's $\tau$ distribution obtained from 10000 samples of 130 independent bivariate datapoints to give an impression of the null distribution.


Figure 7: Parameter estimates from pairwise model fits using the same structure as Model 3, i.e., $a(x)=\alpha x$, $b(x)=1+(\alpha x)^{\beta}$ and $Z$ following a delta Laplace distribution. Here $\beta$ was constrained between $(-1,1)$. Fits were made using composite likelihoods where a single set of parameters was assumed to apply for each pair, whichever the conditioning variable. Distance is in the transformed coordinate space, after accounting for the additional estimated anisotropy. Solid red lines display implied values from the full fitted model; dashed blue lines display implied estimates from the model refitted to the extracted residual processes $Z^{j}$.


Figure 8: Distribution of estimates from 100 bootstrap repetitions, with population MLE subtracted. Left: full model fit; right: fit to extracted $Z^{j}$ using empirical means. $\mathrm{A}=$ Angle, $\mathrm{S}=$ Stretch.

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