# Correlation Testing in Time Series, Spatial and Cross-Sectional Data 

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# Correlation Testing in Time Series, Spatial and Cross-Sectional Data 

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#### Abstract

We provide a general class of tests for correlation in time series, spatial, spatio-temporal and cross-sectional data. We motivate our focus by reviewing how computational and theoretical difficulties of point estimation mount as one moves from regularly-spaced time series data, through forms of irregular spacing, and to spatial data of various kinds. A broad class of computationally simple tests is justified. These specialize Lagrange multiplier tests against parametric departures of various kinds. Their forms are illustrated in case of several models for describing correlation in various kinds of data. The initial focus assumes homoscedasticity, but we also robustify the tests to nonparametric heteroscedasticity. JEL Classifications: C21; C22; C29 Keywords: Correlation; heteroscedasticity; Lagrange multiplier tests.


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## 1 INTRODUCTION

Irregularly-spaced time series, spatial, and spatio-temporal data, and the possibility of cross-sectional correlation, pose considerable difficulties, with respect to modelling, computations and statistical theory. In general, the possibility has to be recognized that there is correlation across time, or space or other relevant dimensions. Rules of inference based on the incorrect assumption of independence will generally be invalidated. Unfortunately, even developing models for dependence can be a far more complicated business than in a regular-spaced time series. Computations can also be more onerous. The development of a satisfactory, useful, asymptotic theory for estimates of both parameters describing the dependence, and parameters of economic interest, such as describing regression effects, can be infeasible. The difficulties arise essentially because of the non-Toeplitz covariance matrix structure that emerges, and the difficulty of separating the regime generating the "location" of observations from that generating the observations themselves, when formulating regularity conditions. Here location can refer to some relevant economic space, not just time or geographical space.

Immense simplification to rules of inference and computations result if there can be assumed to be no dependence. It has been argued (see e.g. Cressie, 1993) that much spatial data can be satisfactorily modelled in terms of mean, regression effects, leaving little to be accounted for by disturbance correlation. Likewise, the common assumption of cross-sectional independence may often be reasonable. This favourable circumstance cannot be taken for granted, but it does further motivate carrying out in the first place tests for independence. If the evidence for independence is strong then we may proceed with simple rules of inference on the remaining parameters of interest. If not, we have to look at developing rules that efficiently take account of dependence, or that are robust. But these tasks are difficult to develop in a very general context. In this paper we focus on testing for independence in such a general context.

This topic has been addressed in a vast time series literature, however little of it permits irregular spacing. It has also been a major, long-standing theme of the spatial literature, with numerous contributions following Moran (1950), Cliff and Ord (1972), but settings have been fairly specific. It seems useful to discuss a general approach which can be applied in a variety of circumstances, under regularity conditions which may shed light on the suitability of the asymptotic theory in specific situations. In a linear regression setting, a general class of statistics is developed that has a chi-square limit distribution under the null hypothesis of independence of disturbances. Special cases can be interpreted as Lagrange multiplier (LM) statistics directed against specified alternatives where they should have good power, though they will have little power against others. It is thus envisaged that in practice several tests may be employed, based on variety of working parametric models.

The tests are developed in Section 3, along with relevant asymptotic theory, of which proofs are left to appendices. In Section 4 they are discussed in some LM examples. First, however, we provide in the following section further
background and motivation by reviewing how difficulties develop as one moves from equally-spaced time series to irregularly-spaced ones, and to spatial and cross-sectionally-correlated data.

## 2 IMPLICATIONS OF IRREGULAR SPACING AND SPATIAL DATA

To fix ideas, and avoid distracting complications, we focus entirely on a linear regression setting, where the regression function is correctly specified, and the covariance matrix is parametric. We will also describe our tests for independence in this setting.

### 2.1 Regression model and Gaussian estimation

We consider the $n \times 1$ vector $y_{n}$ of scalar observations $y_{i n}, i=1, \ldots, n$,

$$
\begin{equation*}
y_{n}=\left(y_{1 n}, \ldots, y_{n n}\right)^{\prime}, \tag{2.1}
\end{equation*}
$$

the prime denoting transposition. The ordering of the $y_{i n}$ is arbitrary, though for time series data it would normally be chronological. The triangular-array aspect of the $y_{i n}$ allows for some asymptotic regimes such as spatial autoregressive (AR) models with row-normalized weight matrices. We suppose that for a given sequence of $n \times q$ matrices $X_{n}, 1<p<n$, and a $q \times 1$ unknown vector $\beta_{0}$,

$$
\begin{equation*}
y_{n}=X_{n} \beta_{0}+u_{n} \tag{2.2}
\end{equation*}
$$

for all sufficiently large $n$, where

$$
\begin{equation*}
u_{n}=\left(u_{1 n}, \ldots, u_{n n}\right)^{\prime} \tag{2.3}
\end{equation*}
$$

is an unobservable vector satisfying

$$
\begin{equation*}
E\left(u_{n}\right)=0, \quad E\left(u_{n} u_{n}^{\prime}\right)=\sigma_{0}^{2} \Omega_{n}\left(\theta_{0}\right) \tag{2.4}
\end{equation*}
$$

where $\sigma_{0}^{2}$ is an unknown positive scalar and $\Omega_{n}(\theta)$ is a given $n \times n$ matrix function of a $p \times 1$ vector parameter $\theta$, and with $\theta_{0}$ being unknown. Lack of correlation in the $u_{i n}$ occurs when $\Omega_{n}\left(\theta_{0}\right)$ is a diagonal. This includes the possibility of heteroscedasticity across $i$, but our main focus is on the implications of nondiagonality.

The main interest may be in $\beta_{0}$, with $\theta_{0}$ and $\sigma_{0}^{2}$ representing nuisance parameters, but in any case their estimation is linked. Conventionally, but conveniently, we consider estimates based on a Gaussian pseudo-likelihood. We have used words such as "independent" and "uncorrelated" rather interchangeably, without drawing a distinction. Of course they are identical if Gaussianity holds, but (2.4) only refers to first- and second-order properties. On the other hand,
stronger conditions than (2.4) would be needed in order to develop asymptotic statistical theory.

The Gaussian pseudo-log-likelihood for $y_{n}$ is given by

$$
\begin{align*}
L_{n}\left(\beta, \sigma^{2}, \theta\right)= & -\frac{n}{2} \log 2 \pi-\frac{n}{2} \log \sigma^{2}-\frac{n}{2} \log \operatorname{det} \Omega_{n}(\theta) \\
& -\frac{1}{2 \sigma^{2}}\left(y_{n}-X_{n} \beta\right)^{\prime} \Omega_{n}(\theta)^{-1}\left(y_{n}-X_{n} \beta\right) \tag{2.5}
\end{align*}
$$

$\beta, \sigma^{2}$ and $\theta$ denoting any admissible values. As is well known, for given $\theta$ $L_{n}\left(\beta, \sigma^{2}, \theta\right)$ is maximized with respect to $\beta, \sigma^{2}$ by

$$
\begin{align*}
\hat{\beta}_{n}(\theta) & =\left(X_{n}^{\prime} \Omega_{n}(\theta)^{-1} X_{n}\right)^{-1} X_{n}^{\prime} \Omega_{n}(\theta)^{-1} y_{n}  \tag{2.6}\\
\hat{\sigma}_{n}^{2}(\theta) & =\frac{1}{n}\left(y_{n}-X_{n} \hat{\beta}_{n}(\theta)\right) \Omega_{n}(\theta)^{-1}\left(y_{n}-X_{n} \hat{\beta}_{n}(\theta)\right) \tag{2.7}
\end{align*}
$$

where it is taken for granted, as it is in (2.5), that the matrix inverses exist. Then

$$
\begin{equation*}
\hat{\theta}_{n}=\arg \min _{\theta} L_{n}\left(\hat{\beta}_{n}(\theta), \hat{\sigma}_{n}^{2}(\theta), \theta\right) \tag{2.8}
\end{equation*}
$$

the maximization conducted over a suitable compact subset of $\mathbb{R}^{q}$ that includes $\theta_{0}$. Equivalently,

$$
\begin{equation*}
\hat{\theta}_{n}=\arg \min _{\theta} Q_{n}(\theta) \tag{2.9}
\end{equation*}
$$

where

$$
\begin{equation*}
Q_{n}(\theta)=\log \hat{\sigma}_{n}^{2}(\theta)+\frac{1}{n} \log \operatorname{det} \Omega_{n}(\theta) \tag{2.10}
\end{equation*}
$$

### 2.2 Regularly-spaced time series

For equally-spaced time series, where the $u_{i}=u_{i n}$ are ordered chronologically and stationary, $Q_{n}(\theta)$ can be typically approximated by simpler quantities, which, when minimized produce estimates of $\theta_{0}$ with the same limit distribution as $n^{\frac{1}{2}}\left(\hat{\theta}_{n}-\theta_{0}\right)$. There are two sources of this favourable outcome. One is that $\Omega_{n}(\theta)$ is a Toeplitz matrix, and can thus be approximately diagonalized by a unitary transformation, so that $\hat{\sigma}_{n}^{2}(\theta)$ can be approximated by an integral, or sum across frequency, of the ratio of the periodogram and the parameterized spectral density. Indeed, in many time series models, such as autoregressive moving average (ARMA) ones, the spectral density can be written down by inspection, whereas the elements of $\Omega_{n}(\theta)$ cannot, and can be cumbersome. The second simplification arises when the second term on the right of (2.10) is asymptotically negligible. This occurs in "standard parameterizations" of ARMA models, where the innovations variance is free of the parameters describing autocorrelation. In that case the problem (2.9) can be replaced by minimization of $\hat{\sigma}_{n}^{2}(\theta)$ or a proxy such as described above. This covers the nonlinear least squares procedures recommended by Box and Jenkins (1971) for ARMA models. The computational simplifications are also reflected in a relatively neat asymptotic statistical theory, exemplified by Hannan (1973), Fox and Taqqu (1986). The
estimates of $\theta_{0}$ are root-n-consistent and asymptotically normally distributed under conditions that require a one-sided infinite moving average representation for the $u_{i}$ with innovations that are not necessarily Gaussian or independent and identically distributed, but are homoscedaastic martingale differences with moments of order only 2 required to be finite. Moreover, the covariance matrix in the limiting normal distribution is unaffected by non-Gaussianity of $u_{i}$.

### 2.3 Lattice data

Equally-spaced spatial or spatio-temporal data present additional problems. We consider only the case of "increasing-domain" asymptotics, as implicitly assumed in the preceding discussion. Observations are recorded on a rectangular lattice of dimension $d>1$. Intervals between observations are constant within dimensions, but can vary across dimensions. Here $n$ represents the total number of observations, i.e. $n=\prod_{j=1}^{d} n_{j}$. and asymptotic theory would typically entail $n_{j} \rightarrow \infty$ for all $j$. Looking again at (2.10), when $u_{i}=u_{i n}$ is stationary a generalization of the Toeplitz property described for the time series case means that again $\sigma_{n}^{2}(\theta)$ can be approximated by a weighted periodogram average. However, it is less likely that $\log \operatorname{det} \Gamma_{n}(\theta)$ can be ignored. The problem was first demonstrated by Whittle (1954), the problem occuring in particular when $u_{i}$ depends on "leads" as well as "lags" in one or more dimensions, as seems plausible in a spatial context, by comparison with the unilateral modelling standard in time series analysis. Whittle (1954) also showed that, quite generally, multilateral models have a "half-plane" kind of unilateral moving average representation, extending the Wold representation of time series, whence the $n^{-1} \log \operatorname{det} \Omega_{n}(\theta)$ term in (2.10) can be ignored. However, the half-plane representation typically involves functions of the coefficients in the original multilateral model that cannot be written in closed form. Nor can it necessarily be well approximated by a parsimonious half-plane model, and the curse of dimensionality is a serious potential problem in spatial modelling.

A further difficulty arising with lattice data with dimension $d>1$ is the "edge effect". Estimates of $\theta_{0}$ given by (2.9), and by the usual approximations to this, can be seen as functions of sample autocovariances. In the time series case $d=1$, the lag $-j$ sample autocovariance is the sum of $n-j$ products divided by $n$. The consequent finite-sample bias causes no problem with asymptotic theory for $\hat{\theta}_{n}$. However, when $d>1$ the bias is of greater order, and leads to an asymptotic theory that is not useful. In particular, for $d=2$ the bias is of order at least $n^{-\frac{1}{2}}$ so that $n^{\frac{1}{2}}\left(\hat{\theta}_{n}-\theta_{0}\right)$ does not converge to a zero-mean random variable. For $d>3$ the order of the bias is even greater than $n^{-\frac{1}{2}}$. A solution proposed by Guyon (1982) essentially replaces the usual, biased, sample autocovariances by unbiased ones. However, Dahlhaus and Künsch (1987) noted that this sacrifices the desirable positive definite property of the Gaussian pseudo-likelihood, and can lead to possible numerical difficulties and a covariance matrix estimate that is not necessarily non-negative definite. They overcame this drawback by instead
employing tapering, but thereby introducing ambiguity due to the choice of taper, and due to an additional tapering parameter if asymptotic efficiency is to be claimed. Robinson and Vidal Sanz (2006) proposed an alternative approach, justifying their estimates of a general class of models for any $d>1$. However, they also introduced an element of arbitrariness in implementation in order to cope with the edge effect.

### 2.4 Irregularly-spaced time series

Irregular spacing of data can arise in several ways. Calendar monthly time series data, for example, are not exactly equally-spaced. However, there is evidence that the effects of disregarding this are unlikely to be significant, and in any case this kind of irregular spacing is largely ignored by practitioners. Another phenomenon is a once-and-for-all change in the sampling interval, as when quarterly observation changes to monthly (see, e.g. Sargan and Drettakis, 1974). For a given dynamic model for the monthly observations, a model for the "skip-sampled" quarterly ones can be deduced and the estimation problem addressed in terms of an objective function that combines components from the two regimes.

Observations can be missing from an otherwise regularly-spaced grid in other ways. Periodic sampling, as in case of weekday observations, does disturb the Toeplitz structure of $\Omega_{n}(\theta)$ but not in a way that severely complicates computation; indeed, one can work with a derived model for vector observations (e.g. the five weekday ones). Non-periodic missing can be ignored in case of only a few missing values, but generally $\Omega_{n}(\theta)$ allows no simplified approximation, and nor can the $\log \operatorname{det} \Omega_{n}(\theta)$ term in $(2.10)$ be neglected. Nevertheless, for suitable models, the Kalman filter and EM algorithm can be applied to break up the computations into simple steps. However, whether one treats the regime generating the observation times as deterministic or stochastic, it seems difficult to deduce an asymptotic theory based on reasonably primitive conditions, in particular on ones that separate out the conditions on the process from those on the sampling regime. Dunsmuir (1983) developed a central limit theorem that is perhaps as successful as is possible in this respect, though it requires a condition on the information matrix that depends simultaneously on both features. Moreover he did not treat $\hat{\theta}_{n}$ itself, but rather a one-step Newton approximation commencing from an initial $n^{\frac{1}{2}}$-consistent estimate. This is in order to avoid a consistency proof, a usual preliminary to the central limit theorem for implicitly-defined extremum estimates. Dunsmuir (1983) described the consistency as an open problem, and it still seems to be, indeed it is not clear in general what initial estimate has $n^{\frac{1}{2}}$-consistency requirement. Dunsmuir and Robinson (1983) developed a full asymptotic theory for an alternative estimate employing an equally-spaced "amplitude-modulated" argument introduced by Parzen (1963), but generally this is asymptotically less efficient than $\hat{\theta}_{n}$.

Some forms of irregular spacing of time series are better viewed in the context of an underlying continuous time process. Spacings would typically be represented as real-valued, possibly generated by a point process. The irreg-
ular spacing could be deliberate, in order to avoid loss of identifiability due to aliasing. Again, the Toeplitz structure of $\Omega_{n}(\theta)$ is lost, and it is generally not possible to simply approximate either component of (2.10). An exception is when under the continuous time process as generated by a first order stochastic differential equation driven by white noise. Robinson (1977) deduced a model for the discrete observations, essentially a time-varying first-order autoregression (AR) with heteroscedastic innovations, and consequently approximated (2.10) by a simple form. He established consistency and asymptotic normality of the estimates, but nevertheless in terms of conditions which, to a significant degree, simultaneously restrict the process and the sampling sequence. With more elaborate continuous time models it does not seem possible to deduce a reasonably simple model for the observations, and asymptotic statistical theory would seem difficult to establish under reasonably primitive conditions. See also McDunnough and Wolfson (1979).

### 2.5 Irregular spacing in spatial data

Irregular spacing is a more natural and frequent occurrence with spatial data. In a geographical setting, data are liable to be recorded across heterogenously-sized administrative regions, while economic distances will not correspond to regular spacing. The difficulties reported above will only be compounded, indeed it seems even hard to extend the model and estimate of Robinson (1977). In general there will not be evident computational simplifications, and while it is possible to write down an asymptotic theory in terms of highly unprimitive conditions, it may be difficult to check them in special cases.

Some of these difficulties can be circumvented by a different approach to modelling which is formally covered by our set-up, namely spatial AR and related models. Indeed, when there is no geographical aspect, the methods reviewed above are unsuitable. Rules of inference for much microeconomic data routinely take for granted cross-sectional independence, at least at some level, yet there is also an awareness that this can be inappropriate. In some circumstances it is natural to envisage that correlation varies with relevant measures of economic distance, such as differences in household income. Econometricians are familiar with the notion of leads and lags from time series models, and spatial AR models have had considerable appeal. They rely on specification of an $n \times n$ "weight matrix", which essentially embodies in a simple way notions of irregular spacing. Lee (2004) has developed asymptotic theory for $\hat{\theta}_{n}$. In general the $\log \operatorname{det} \Omega_{n}(\theta)$ term in (2.10) cannot be neglected, though for a related model Lee (2002) has shown that this is possible (so least squares works) under suitable conditions on the weight matrix. Under similar conditions, Robinson (2006) has developed asymptotic theory for efficient estimates when the innovations in the spatial AR model are not necessarily normally distributed, both in case of a parametric model for their distribution, and a nonparametric one.

Though asymptotic theory under the null of independence is relatively simple with respect to any test statistic, the computational difficulties of point estimation described in the preceding section make LM tests more appealing
than Wald or likelihood-ratio ones. These serve to motivate a general class of statistic treated in the following section. It is introduced without reference to LM testing because versions of it may lack such an interpretation. Moreover, this will be lost in any case in another statistic also investigated, which nonparametrically robustifies to heteroscedasticity in the $u_{i n}$.

An alternative type of model is motivated by a different form of asymptotics from the "increasing domain" asymptotics usually employed in time series and many spatial settings. This is "fixed domain", or "infill", asymptotics, where the observations are regarded as becoming denser on a bounded region (see e.g. Cressie, 1993, Stein, 1991, Lahiri, 1996). While seemingly more natural in many circumstances, nonstandard results that are not practically useful often emerge, for example estimates may not be consistent, converging instead to a nondegenerate random variable.

## 3 A GENERAL CLASS OF TEST STATISTICS

We present a class of test statistics that has a limiting $\chi^{2}$ distribution under the null hypothesis that the $u_{i n}$ in (2.2), (2.3) are independently (and homoscedastically) distributed. For a given $\Omega_{n}(\theta)$ in (2.4), there is a member of the class that has an LM interpretation, and thus can be expected to have optimal power against local alternatives in directions implied by $\Omega_{n}(\theta)$. However, such an interpretation is not necessary for the asymptotic validity.

### 3.1 Testing assuming homoscedasticity

Choose the $p \times 1$ vectors $\psi_{i j n}, i, j=1, \ldots, n, n \geq 1$, such that $\psi_{i i n}=0$, $\psi_{j i n}=\psi_{i j n}$ for all $i, j, n$. Define $\hat{\beta}_{n}=\hat{\beta}_{n}(0), \hat{\sigma}_{n}^{2}=\hat{\sigma}_{n}^{2}(0)$ and the least squares residuals

$$
\begin{equation*}
\hat{u}_{n}=\left(\hat{u}_{1 n}, \ldots, \hat{u}_{n n}\right)^{\prime}=y_{n}-X_{n} \hat{\beta}_{n} . \tag{3.1}
\end{equation*}
$$

Define

$$
\begin{align*}
\hat{a}_{n} & =\sum_{i, j=1}^{n} \psi_{i j n} \hat{u}_{i n} \hat{u}_{j n}  \tag{3.2}\\
A_{n} & =\sum_{i, j=1}^{n} \psi_{i j n} \psi_{i j n}^{\prime}  \tag{3.3}\\
\zeta_{n} & =\hat{\sigma}_{n}^{-4} \hat{a}_{n}^{\prime} A_{n}^{-1} \hat{a}_{n} \tag{3.4}
\end{align*}
$$

There is no loss of generality in taking $\psi_{i j n}=\psi_{j i n}$ because if it were not so we could redefine $\hat{a}_{n}$ with $\left(\psi_{i j n}+\psi_{j i n}\right) / 2$ in place of $\psi_{i j n}$.

Denote by $x_{i}$ the $i$-th column of $X_{n}^{\prime}$. We allow the $x_{i}$ to be either deterministically or stochastically generated, but independent of the $u_{i n}$. Likewise, the $\psi_{i j n}$ can also be deterministically or stochastically generated, possibly dependent on the $x_{i}$, but again independent of the $u_{i n}$. This is relevant if, say, in a
spatial AR model, the weight matrix reflects economic distances between observations measured by the distance between respective stochastically-generated explanatory variables, for example the $(i, j)$-th element might be proportional to $\left\|x_{i}-x_{j}\right\| /\left(1+\left\|x_{i}-x_{j}\right\|^{2}\right)$, where the factor of proportionality might vary across rows.

Assumption 1 The $u_{i}=u_{i n}, i=1,2, \ldots$ are independent with zero mean, constant variance $\sigma^{2}$, and, for some $\delta>0$,

$$
\begin{equation*}
\max _{i \geq 1} E\left|u_{i}\right|^{2+\delta}<\infty \tag{3.5}
\end{equation*}
$$

Assumption $2\left\{x_{i}, i \geq 1\right\}$ is independent of $\left\{u_{i}, i \geq 1\right\}$ and for some $n \geq q$, $X_{n}$ has full column rank.

Define

$$
\begin{equation*}
D_{n}=\operatorname{diag}\left\{d_{1 n}, \ldots, d_{p n}\right\} \tag{3.6}
\end{equation*}
$$

where, with $\psi_{i j h n}$ denoting the $h$-th element of $\psi_{i j n}$,

$$
\begin{equation*}
d_{h n}=\sum_{i, j=1}^{n} \psi_{i j h n}^{2}, \quad h=1, \ldots, p \tag{3.7}
\end{equation*}
$$

Assumption $3\left\{\psi_{i j n}, i, j=1, . ., n, n \geq 1\right\}$ is independent of $\left\{u_{i}, i \geq 1\right\}$, and, as $n \rightarrow \infty$,

$$
\begin{equation*}
D_{n}^{-\frac{1}{2}} A_{n} D_{n}^{-\frac{1}{2}} \rightarrow_{p} R \tag{3.8}
\end{equation*}
$$

for some positive definite matrix $R$, and

$$
\begin{gather*}
d_{h n} \rightarrow_{p} \infty, \quad h=1, \ldots, p,  \tag{3.9}\\
\frac{\max _{1 \geq i \leq n} \sum_{j=1}^{n}\left|\psi_{i j h n}\right|}{d_{h n}^{\frac{1}{2}}} \rightarrow_{p} 0, \quad h=1, \ldots, p,  \tag{3.10}\\
\frac{\sum_{i, j=2}^{n}\left(\sum_{k=1}^{m i n}(i, j)-1\right.}{\sum_{\ell n} d_{m n}} \rightarrow_{p} 0, \quad \ell, m=1, \ldots, p \tag{3.11}
\end{gather*}
$$

In time series settings independence in Assumption 1 can be replaced by a martingale difference assumption, but in spatial configurations there may be no natural ordering. The final two parts of Assumption 3 appear to heavily restrict the $\psi_{i j n}$, but are satisfied in particular when there is a degree of sparseness, with many elements zero, and both parts can be checked in case of LM examples. Asymptotic analysis of a similar class of statistic was considered by Pinkse (1999,
2004), improving on an earlier treatment of Sen (1976). In some ways his focus was broader, mainly in that his statistics permits investigation also of correlation between two different sets of random variables. Also, he operated in the setting of a more general nonlinear model (see also Kelejian and Prucha, 2001). In this, his regressors are independent of the disturbances, as in Assumption 2 and earlier in the treatment in Robinson (1991) of LM tests in a general class of time series models for regularly-spaced data. As there, we exploit the linear regression structure to enable a treatment under relatively primitive conditions; note also the generality of the last part of Assumption 2, which permits different rates of growth of elements of $x_{i}$. Pinke (1999) did not allow his weights corresponding to $\psi_{i j n}$ to be stochastic, and took $p=1$. Our allowance for $p>1$ follows the time series asymptotic treatment of Robinson (1991), and reflects LM statistics against $\mathrm{AR}(p)$ and $\mathrm{MA}(p)$ time series alternatives (see Godfrey, 1978), and against generalizations of spatial AR models (Anselin, 1998).

Theorem 1 Let (2.2) hold for all sufficiently large $n$, and Assumptions 1-3. Then as $n \rightarrow \infty, \zeta_{n} \rightarrow{ }_{d} \chi_{p}^{2}$.

The proof is in Appendix 1.

### 3.2 Testing robust to heteroscedasticity

While Assumption 1 does not assume identity of distribution, and limits constancy of moments to the mean and variance, homoscedasticity seems an unreasonable assumption in many kinds of spatial data, where for example, observations are based on aggregation over administrative regions that differ considerably in size. Versions of $\zeta_{n}$ designed to test for correlation may be significant due to unanticipated heteroscedasticity. In fact, formally, versions of $\zeta_{n}$ can be interpreted as LM tests of (conditional or unconditional) heteroscedasticity, not just correlation, though we do not stress this aspect because asymptotically Gauss-Markov efficient weighted least squares estimation of $\beta_{0}$, treating either parametric or nonparametric heteroscedassticity, is entirely feasible. Instead we robustify $\zeta_{n}$ to heteroscedasticity.

Define

$$
\begin{align*}
\hat{B}_{n} & =\sum_{i, j=1}^{n} \psi_{i j n} \psi_{i j n}^{\prime} \hat{u}_{i n}^{2} \hat{u}_{j n}^{2}  \tag{3.12}\\
\xi_{n} & =\hat{a}_{n}^{\prime} \hat{B}_{n}^{-1} \hat{a}_{n} \tag{3.13}
\end{align*}
$$

This weighting by squared raw residuals is in the spirit of heteroscedasticityconsistent variance estimation first introduced by Eicker (1963), and much employed since by econometricians. We modify two of our previous assumptions accordingly. Define

Assumption $1^{*}$ Assumption 1 holds, with $\delta=2$ in (3.5) but without the requirement that the variance of $u_{i}$, now denoted $\sigma_{i}^{2}$, be constant over $i ; \min _{i \geq 1} \sigma_{i}^{2}>$ 0.

Assumption 3 ${ }^{*}$ Assumption 3 holds with (3.8) replaced by

$$
\begin{equation*}
D_{n}^{-\frac{1}{2}} B_{n} D_{n}^{-\frac{1}{2}} \rightarrow_{p} S, \quad \text { as } n \rightarrow \infty \tag{3.14}
\end{equation*}
$$

for some positive definite matrix $S$, where

$$
\begin{equation*}
B_{n}=\sum_{i, j=1}^{n} \psi_{i j n} \psi_{i j n}^{\prime} \sigma_{i}^{2} \sigma_{j}^{2} \tag{3.15}
\end{equation*}
$$

The fourth moment condition on $u_{i}$ seems unavoidable, indeed some care is needed in the proof to avoid something stronger. Notice it implies, via Hölder's inequality, that

$$
\begin{equation*}
\max _{i \geq 1} \sigma_{i}^{2}<\infty \tag{3.16}
\end{equation*}
$$

Theorem 2 Let (2.2) hold for all sufficiently large n, and Assumptions 1*, 2 and $3^{*}$. Then as $n \rightarrow \infty, \xi_{n} \rightarrow_{d} \chi_{p}^{2}$.

## 4 LAGRANGE MULTIPLIER-MOTIVATED SPECIAL CASES

The present section illustrates how our general statistics apply when the $\psi_{i j n}$ are motivated by the LM principle. Of course the consequent optimality will apply only to $\zeta_{n}$, and not to $\xi_{n}$.

Referring to (2.4), we identify the null hypothesis $H_{0}$ that the $u_{i n}$ are independent (and homoscedastic) with $\theta_{0}$ taking a particular value, which with no loss of generality we take to be the null vector, 0 :

$$
\begin{equation*}
H_{0}: \theta_{0}=0 \tag{4.1}
\end{equation*}
$$

where

$$
\begin{equation*}
\Omega_{n}(0)=I_{n}, \text { all sufficiently large } n . \tag{4.2}
\end{equation*}
$$

Assuming $\Omega_{n}(\theta)$ is differentiable at $\theta=0$, a version of the LM statistic is given by (3.4) with

$$
\begin{equation*}
\psi_{i j n}=\frac{\partial}{\partial \theta} \omega_{i j n}(0) \tag{4.3}
\end{equation*}
$$

where $\omega_{i j n}(\theta)$ is the $(i, j)$-th element of $\Omega_{n}(\theta)$ and the derivative in (4.3) is zero for $i=j$. We consider the following examples.

### 4.1 Missing data in time series

Here $\left\{y_{t}\right\}$ are the consecutive, un-missed observations from a regularly-spaced time series. Correspondingly $u_{n}=\left(u\left(t_{1}\right), \ldots, u\left(t_{n}\right)\right)^{\prime}$, where the $t_{i}$ are integers, $t_{1}<t_{2}<\ldots<t_{n}$, and $u(t)$ is stationary with zero mean and lag- $j$ autocovariance $\gamma(j ; \theta)$, a known function of $j, \theta$. Thus $\psi_{i j n}=(\partial / \partial \theta) \gamma\left(t_{i}-t_{j} ; 0\right)$. The $\psi_{i j n}$ are thus functions of $\left\{t_{i}\right\}$, which may be deterministically or stochastically generated, as Assumptions 3 and $3^{*}$ permit.

One special case not previously considered is a missing-data version of the test of Robinson (1991) against long memory alternatives. Here $d=1$ and $(1-L)^{\theta} u_{i}=\varepsilon_{i}$, where $L$ is the lag operator $d$, the $\varepsilon_{i}$ are independent and heteroscedastic, and $|\theta|<\frac{1}{2}$. Then $\psi_{i j n}=\left|t_{i}-t_{j}\right|^{-1}$, for $i \neq j$. Part (3.9) of Assumption 3 is satisfied if $d_{1 n}=\sum_{i, j, i \neq j}^{n}\left|t_{i}-t_{j}\right|^{-2} \rightarrow_{p} \infty$. In case there is no missing, or with periodic or roughly periodic missing, $d_{1 n}$ increases at rate $n$, but a slower rate with missing is possible, permitting observations to "peter out". With respect to (3.10), we have $\Sigma_{j=1, j \neq i}^{n}\left|t_{t}-t_{j}\right|^{-1} \leq \Sigma_{i=1}^{t_{n}} i^{-1} \sim \log t_{n}$, uniformly in $i$, so that (3.10) is equivalent to

$$
\begin{equation*}
t_{n}^{2} / d_{n} \rightarrow_{p} 0 \tag{4.4}
\end{equation*}
$$

The previous conditions imply (3.11) also, as we now show. The numerator of its left side is

$$
\begin{equation*}
\left.\sum_{i, j=2}^{n} \sum_{k<i, j}\left|t_{i}-t_{k}\right|^{-1}\left|t_{j}-t_{k}\right|^{-1}\right)^{2} \tag{4.5}
\end{equation*}
$$

The contribution from $i=j$ is

$$
\begin{equation*}
\sum_{i=j}^{n}\left(\sum_{k<i}\left|t_{i}-t_{k}\right|^{-1}\right)^{2} \leq C \sum_{i}\left|t_{i}-t_{i+1}\right|^{-2}=O_{p}(1) \tag{4.6}
\end{equation*}
$$

where $C$ denotes a generic finite constant. The contribution from $i \neq j$ is bounded by

$$
\begin{equation*}
2 \sum_{i<j} \sum_{i}\left|t_{i}-t_{j}\right|^{-2}\left(\sum_{k<i}\left|t_{i}-t_{k}\right|^{-1}\right)^{2}=O_{p}\left(d_{1 n} t_{n}^{2}\right) \tag{4.7}
\end{equation*}
$$

Another leading alternative is the $A R(p)$ hypothesis already considered by Robinson (1986), who obtained a missing-data version of the Box and Pierce (1970) statistic. We have $\psi_{i j k n}=1\left(\left|t_{i}-t_{j}\right|=k\right), d_{1 n}=\sum_{i, j=1, i \neq j}^{n} 1\left(\left|t_{i}-t_{j}\right|=k\right)$, $k=1, \ldots, p$, where $1($.$) is the indicator function. With \ell \leq m$ the numerator of the left side of (3.11) is

$$
\begin{equation*}
\sum_{i, j=2}^{n}\left(\sum_{k=1}^{\min (1, j)-1} 1\left(t_{i}-t_{k}=\ell\right) 1\left(t_{j}-t_{k}=m\right)\right)^{2} \leq 2 \sum_{\substack{i, j=2 \\ i \leq j}}^{n} 1\left(t_{j}-t_{i}=m-\ell\right) \tag{4.8}
\end{equation*}
$$

This is bounded by $d_{m-\ell, n}$ for $\ell<m$, and by $n-1$ for $\ell=m$. Thus if $n / d_{\ell n}^{2} \rightarrow_{p} 0$, (3.11) is satisfied for $\ell=m=1, \ldots, p$. It is satisfied for $\ell<m$ if $d_{m-\ell, n} / d_{\ell n} d_{m n} \rightarrow_{p} 0$. Clearly because $d_{\ell n} \leq n$ a sufficient condition for (3.11) is that $n^{\frac{1}{2}} / d_{\ell n} \rightarrow_{p} 0$. This implies (3.9) and (3.10), the numerator of the latter being 2 .

## $4.2 d$-dimensional lattice

Introduce the $d$-dimensional lattice $\mathcal{L}_{d}=\left\{I: I=\left(i_{1}, \ldots, i_{d}\right), i_{j}=0, \pm 1, \ldots, j=1, \ldots, d\right\}$, for $d>1$. We observe $Y_{I}$, for $I \in N=\left\{I: i_{j}=1, \ldots, n_{j}, j=1, \ldots, d\right\}$, and take $n=\Pi_{j=1}^{d} n_{j}$. Corresponding to (2.2) we have $Y_{I}=\beta_{0}^{\prime} x_{I}+u_{I}, I \in N$. Identifying the $i$-th element of $u_{n}$ with $U_{I}$ (possibly with lexicographic ordering) correspondingly denote the $i$-th element of $\hat{u}_{n}$ by $\hat{U}_{I n}$. Suppose $U_{I}$ is stationary with autocovariance $\operatorname{Cov}\left(U_{I}, U_{I+J}\right)=\gamma\left(J ; \theta_{0}\right)$ for $J \in \mathcal{L}_{d}$, where $\gamma(J ; \theta)$ is boundedly differentiable in $\theta$ but $\theta_{0}$ is unknown. Denote $\Psi_{I}=(\partial / \partial \theta) \gamma(I ; 0)$. Thus $\zeta_{n}$ and $\xi_{n}$ are given by (3.4) and (3.13) with

$$
\begin{align*}
\hat{\sigma}_{n}^{2} & =n^{-1} \sum_{I \in N} \hat{U}_{I n}^{2}, \quad \hat{a}_{n}=\sum_{I, J \in N} \Psi_{I-J} \hat{U}_{I n} \hat{U}_{J n},  \tag{4.9}\\
A_{n} & =\sum_{I, J \in N} \Psi_{I-J} \Psi_{I-J}^{\prime}, \quad \hat{B}_{n}=\sum_{I, J \in N} \Psi_{I-J} \Psi_{I-J}^{\prime} \hat{U}_{I n}^{2} \hat{U}_{J n}^{2} . \tag{4.10}
\end{align*}
$$

One example tests against long memory, taking $d=1$ and $\Pi_{j=1}^{d}\left(1-L_{j}\right)^{\theta_{0}} U_{I}=$ $\varepsilon_{I}$ where $L_{j}$ is the lag-operator is the $j$-th dimension only, the $\varepsilon_{I}$ are independent and homoscedastic, and $\left|\theta_{0}\right|<\frac{1}{2}$. Then $\Psi_{I}=\Sigma_{j=1}^{d} i_{j}^{-1}$.

Tests against AR alternatives are also available. Let $P$ be a set of $p$ distinct $I$ indices, such that $I \neq\{0, \ldots, 0\}$, and consider the model

$$
\begin{equation*}
U_{I}-\sum_{J \in P} \Theta_{O J} \prod_{\substack{i=1 \\\left(j_{1}, \ldots, j_{d}\right)=J}}^{d} L^{j_{i}} U_{I-J}=\varepsilon_{I} \tag{4.11}
\end{equation*}
$$

with $\varepsilon_{I}$ as before. Now $\theta_{0}$ consists of scalars $\Theta_{O J}$ (which must satisfy stationarity conditions if $\theta_{0} \neq 0$ ). A typical element of $\Psi_{1}$ is

$$
\begin{equation*}
\Psi_{I J}=\frac{\partial}{\partial \Theta_{0 J}} \gamma(I ; 0)=1(I=J), \quad J \in P \tag{4.12}
\end{equation*}
$$

However there is a restriction on $P$ which affects multilateral modelling motivated by a lack of natural ordering in one or more of the dimensions; in spatiotemporal data there is a natural ordering in the time dimension, but typically not in the others. There is thus a temptation to include $J$ in (4.1) that contain some negative indices, as well as $J$ with all non-negative ones. This can present identification problems as recently reviewed by Robinson and Vidal Sanz
(2006). We encounter a corresponding problem. From (4.12) and the symmetry property $\gamma(I ; \theta)=\gamma(-I ; \theta)$ it is clear that for $K=-J$

$$
\begin{equation*}
\Psi_{I K}=\left(\frac{\partial}{\partial \Theta_{0 K}}\right) \gamma(I ; 0)=-1(-I=K)=-1(I=J) \tag{4.13}
\end{equation*}
$$

Thus $A_{n}$ (and $\hat{B}_{n}$ ) will not be invertible. We might also think of including such mirror-image $J$ but constraining their coefficients to be equal. This avoids the identifiability condition but it is easily seen to produce the same statistic as if we included only one of them. Altogether, taking, say, all $J$ to have non-negative elements, we get

$$
\begin{equation*}
\zeta_{n}=\sum_{J \in P}\left(n^{-1} \sum_{I, I+J \in N} \hat{U}_{I} \hat{U}_{I+J}\right)^{2} / \hat{\sigma}_{n}^{4} \tag{4.14}
\end{equation*}
$$

a natural extension of the Box and Pierce (1970) statistic for time series. With respect to potential "edge effect", the discrepancy between the numbers of summands over $I$ and $N$ has no asymptotic effect under the null hypothesis because there is no bias, due to $E\left(U_{I} U_{I+J}\right)=0, J \neq\{0, \ldots, 0\}$.

Tests can also be based on more parsimonious models that have the property of isometry. For example take $\gamma(I ; \theta)=\theta^{\|I\|}$, for scalar $\theta \in(-1,1)$. Then

$$
\begin{equation*}
\zeta_{n}=\left(\sum_{I \in N} U_{I} \sum_{J:\|J\|=1} U_{I+J}\right)^{2} /\left(\sum \sum 1(\|I-J\|=1)\right) \tag{4.15}
\end{equation*}
$$

It is straightforward to extend the above statistics to allow for missing observations, in the manner of the previous sub-section.

### 4.3 Spatial autoregressive models

Spatial AR models are especially convenient when there is irregular spacing that cannot be handled in the framework of missing values in an otherwise regular time series or lattice, or when the space is economic rather than geographic. Consider the model

$$
\begin{equation*}
\left(I_{n}-\sum_{k=1}^{p} \theta_{0 k} W_{k n}\right) u_{n}=\varepsilon_{n}, \tag{4.16}
\end{equation*}
$$

where $\varepsilon_{n}$ is a vector of independent, homoscedastic variables, and the $W_{k n}$ are $n \times n$ weight matrices, possibly stochastically generated and possibly $X_{n}$ dependent. The most familiar version of (4.16) has $p=1$. Anselin (2001) discussed LM tests for spatial independence against a related model where instead of combining (2.2) with (4.16), one incorporates spatially lagged $y$ 's in (2.2). The null model is the same in both cases.

Testing for spatial independence in (4.16) and related models has been widely considered, and our purpose here is not to present new tests but to discuss conditions in the $W_{k n}$ for asymptotic validity, and consider the connection with infill asymptotics. The identifiability problem in (4.16) is similar to that discussed by Anselin (2001) in his model. We have

$$
\begin{equation*}
\psi_{i j k n}=\frac{\partial}{\partial \theta_{k}} \omega_{i j n}(0)=2 W_{i j k n} \tag{4.17}
\end{equation*}
$$

where $W_{i j k n}$ is the $(i, j)$-th element of $W_{k n}$. Then

$$
\begin{equation*}
A_{n}=\left(4 \sum_{i, j=1}^{n} W_{i j k n} W_{i j \ell n}\right) \tag{4.18}
\end{equation*}
$$

indicating the $(k, \ell)$-th element. It is obvious that (2.5) requires in particular that all the $W_{k n}$ must differ. Anselin (2001) assumed that

$$
\begin{equation*}
\sum_{j=1}^{n} W_{i j k n} W_{i j \ell n}=0 \tag{4.19}
\end{equation*}
$$

for $k \neq \ell$, which implies that $A_{n}$ is diagonal; a special case is where the $n$ observations are sub-divided into subsets such that $W_{k n}$ has zero elements corresponding to the non- $k$ th subsets, so $\Sigma_{k=1}^{p} W_{k n}$ is block diagonal. Indeed (4.19) requires existence of some negative weights unless $W_{i j k n}=0$ or $W_{i j \ell n}=0$ for each $i, j$ and each $k \neq \ell$.

The preceding discussion applies only to $\hat{B}_{n}$ in $\xi_{n}$. Kelejian and Robinson (2004) combined heteroscedasticity in a spatial AR context but applied it to $\varepsilon_{n}$ and adopted a different approach to the problem.

With respect to both $\zeta_{n}$ and $\xi_{n}$, conditions (3.9)-(3.11) become

$$
\begin{gather*}
d_{h n}=\sum_{i, j=1}^{n} W_{i j h n}^{2} \rightarrow_{p} \infty, \quad h=1, \ldots, p,  \tag{4.20}\\
\frac{\max _{1 \leq i \leq n} \sum_{j=1}^{n}\left|W_{i j h n}\right|}{d_{h n}^{\frac{1}{2}}} \rightarrow_{p} 0, \quad h=1, \ldots, p,  \tag{4.21}\\
\frac{\sum_{i, j=2}^{n}\left(\sum_{h=1}^{\min (i, j)-1}\left|W_{i k \ell n} W_{j k m n}\right|\right)^{2}}{d_{\ell n} d_{m n}} \rightarrow_{p} 0, \quad \ell, m=1, \ldots, p \tag{4.22}
\end{gather*}
$$

(cf. Sen, 1976, Pinkse, 1999, 2004). Given (4.20), a sufficient condition for (4.21) is that $W_{h n}$ have non-negative elements and are row-normalized. The
left side of (4.22) has numerator bounded by

$$
\begin{equation*}
\sum_{i, j=2}^{n} \sum_{k} \max _{k} W_{j k m n}^{2}\left(\sum_{k=1}^{n}\left|W_{i k \ell n}\right|\right)^{2} / d_{\ell n} d_{m n}=o_{p}\left(n \sum_{j=2}^{n} \max _{k} W_{j k m n}^{2} / d_{m n}\right) \tag{4.23}
\end{equation*}
$$

Suppose $W_{j k m n}=O_{p}\left(h_{m}^{-1}\right)$ uniformly (cf. Lee, 2002). Then (4.23) $=o_{p}\left(n^{2} /\left(h_{m}^{2} d_{m n}\right)\right)$. Thus (3.11) entails

$$
\begin{equation*}
\frac{n^{2}}{h_{m}^{2} d_{m n}}=O_{p}(1) \tag{4.24}
\end{equation*}
$$

Given (4.24), (3.9) is satisfied by

$$
\begin{equation*}
n / h_{m} \rightarrow \infty, \quad m=1, \ldots, p \tag{4.25}
\end{equation*}
$$

Indeed, $d_{m n}=O_{p}\left(n^{2} / h_{m}^{2}\right)$ so (4.25) is necessary for (3.9).
Some formal comparison is possible between our asymptotic discussion, and (4.25) in particular, and infill asymptotics. Consider for simplicity in place of (4.21) the first order spatial MA

$$
\begin{equation*}
u_{n}=\left(I_{n}+\theta_{0} W_{1 n}\right) \varepsilon_{n} \tag{4.26}
\end{equation*}
$$

On the other hand consider a process $u(t), t \in(0,1]$ such that

$$
\begin{equation*}
u(t)=\varepsilon_{t}+\frac{1}{n} \sum_{s=1}^{n} \alpha\left(t-\frac{s}{n} ; \theta_{0}\right) \varepsilon_{s}, \quad t \in(0,1] \tag{4.27}
\end{equation*}
$$

for a function $\alpha(t ; \theta),|t| \leq 1$, that is boundedly differentiable in $\theta$. (Extension to a process defined on a finite region in $d$ dimensions is immediate.) For example, $\alpha(t ; \theta) \equiv \theta$, where there is a close formal similarity with (4.26). Consider sampling $u(t)$ at intervals $1 / n$. Thus taking $u_{n}=(u(1 / n), \ldots, u(1-1 / n))^{\prime}$ and applying the LM principle for testing $\theta_{0}=0$ we find that (3.9) is violated. Likewise we cannot take $h_{n} \sim n^{-1}$ in (4.26).

## APPENDIX 1: PROOF OF THEOREM 1

Proof. We write $\psi_{i j}$ for $\psi_{i j n}$ throughout. The limit distribution is independent of the $x_{i}$ and $\psi_{i j}$, so it suffices to show that the result holds conditionally on $\left\{x_{i}, i \geq 1\right\}$ and $\left\{\psi_{i j}, i, j=1, \ldots, n, n \geq 1\right\}$; correspondingly, all expectations in what follows will thereby be conditional, though we suppress reference to this. Define $a_{n}=\Sigma_{i, j} \psi_{i j} u_{i} u_{j}$, writing $\psi_{i j}=\psi_{i j n}$ and unqualified summation over $i$ covering $i=1, \ldots, n$. The result follows from

$$
\begin{gather*}
\hat{\sigma}_{n}^{2} \rightarrow_{p} \sigma^{2},  \tag{A.1}\\
A_{n}^{-\frac{1}{2}}\left(\hat{a}_{n}-a_{n}\right) \rightarrow_{p} 0, \tag{A.2}
\end{gather*}
$$

and (conditionally)

$$
\begin{equation*}
A_{n}^{-\frac{1}{2}} a_{n} \rightarrow_{p} N\left(0, I_{p}\right) \tag{A.3}
\end{equation*}
$$

We omit the proof of (A.1), as it is essentially implied by that of (A.2). To consider this, write $\hat{u}_{i}=\hat{u}_{i n}, \hat{v}_{i}=\hat{u}_{i}-u_{i}=\Sigma_{j} u_{j} b_{i j}$, where $b_{i j}=x_{i}^{\prime}\left(\Sigma_{h} x_{h} x_{h}^{\prime}\right)^{-1} x_{j}$. Thus

$$
\begin{equation*}
\hat{a}_{n}-a_{n}=\sum_{i, j} \psi_{i j}\left(\hat{v}_{i} u_{j}+u_{i} \hat{u}_{j}+\hat{v}_{i} \hat{v}_{j}\right) \tag{A.4}
\end{equation*}
$$

Since

$$
\begin{equation*}
\left\|A_{n}^{-\frac{1}{2}} D_{n}^{\frac{1}{2}}\right\| \leq \operatorname{tr}\left\{D_{n}^{\frac{1}{2}} A_{n}^{-1} D_{n}^{\frac{1}{2}}\right\} \rightarrow_{p} \operatorname{tr}\left(R^{-1}\right) \tag{A.5}
\end{equation*}
$$

we can prove (A.2) with $A_{n}^{-\frac{1}{2}}$ replaced by $D_{n}^{-\frac{1}{2}}$. We consider an arbitrary element, and so to avoid additional subscripting $\psi_{i j}$ for the time being represents a scalar.

We have

$$
\begin{equation*}
\sum_{i, j} \psi_{i j} \hat{v}_{i} u_{j}=\sum_{i, j} \psi_{i j} u_{j}^{2} b_{i j}+\sum_{i, j} \psi_{i j} u_{j} \sum_{h \neq j} u_{h} b_{i h} \tag{A.6}
\end{equation*}
$$

The modulus of the first term on the right has expectation bounded by

$$
\begin{equation*}
C \sum_{i, j}\left|\psi_{i j} b_{i j}\right| \leq C \sum_{i, j}\left|\psi_{i j}\right|\left(b_{i i}+b_{j j}\right) \tag{A.7}
\end{equation*}
$$

because by the Cauchy and elementary inequalities $\left|b_{i j}\right| \leq b_{i i}^{\frac{1}{2}} b_{j j}^{\frac{1}{2}} \leq b_{i i}+b_{j j}, C$ denoting throughout a generic constant. Because $\Sigma_{i} b_{i i}=q$ and $\psi_{i j}=\psi_{j i}$, this is bounded by $C \max _{i} \Sigma_{j}\left|\psi_{i j}\right|$. The second term has mean zero and variance bounded by

$$
\begin{equation*}
C \sum_{i, j, \ell, m} \psi_{i j} \psi_{\ell m} b_{i m} b_{\ell j}+C \sum_{h, i, j, \ell} \psi_{i j} \psi_{\ell j} b_{i j} b_{\ell h} \tag{A.8}
\end{equation*}
$$

$$
\begin{align*}
& \leq C\left(\sum_{i, j}\left|\psi_{i j}\right| b_{i i}^{\frac{1}{2}} b_{j j}^{\frac{1}{2}}\right)^{2}+C \sum_{i, j, \ell}\left|\psi_{i j}\right|\left|\psi_{\ell j}\right| b_{i i}^{\frac{1}{2}} b_{\ell \ell}^{\frac{1}{2}} \\
& \leq C\left(\sum_{i, j}\left|\psi_{i j}\right|\left(b_{i i}+b_{j j}\right)\right)^{2}+C \sum_{i, j, \ell}\left|\psi_{i j}\right|\left|\psi_{\ell j}\right|\left(b_{i i}+b_{\ell \ell}\right) \\
& \leq C\left(\max _{i} \sum_{j}\left|\psi_{i j}\right|\right)^{2} . \tag{A.9}
\end{align*}
$$

Next,

$$
\begin{align*}
\sum_{i, j} \psi_{i j} \hat{v}_{i} \hat{v}_{j} & =\sum_{i, j} \psi_{i j}\left(\sum_{\ell} u_{\ell} b_{i \ell}\right)\left(\sum_{m} u_{m} b_{j m}\right) \\
& =\sum_{i, j, \ell} \psi_{i j} b_{i \ell} b_{j \ell} u_{\ell}^{2}+\sum_{i, j, k} \sum_{i j} \psi_{i k} b_{i k} \sum_{\ell \neq k} b_{i \ell} u_{k} u_{\ell} . \tag{A.10}
\end{align*}
$$

The first term on the right has modulus with expectation bounded by

$$
\begin{equation*}
C \sum_{i, j}\left|\psi_{i j}\right|\left(b_{i i}+b_{j j}\right) \leq C \max _{i} \sum_{i}\left|\psi_{i j}\right| \tag{A.11}
\end{equation*}
$$

as before. The second term has mean zero and variance bounded by

$$
\begin{gather*}
\sum_{h, i, j, k, \ell, m} \psi_{i j} \psi_{h m} b_{j \ell}\left(b_{i k} b_{m \ell}+b_{h \ell} b_{m k}\right) \\
\leq C\left(\sum_{i} \sum_{j} \psi_{i j} b_{i i}^{\frac{1}{2}} b_{j j}^{\frac{1}{2}}\right) \leq C\left(\max _{i} \sum_{j}\left|\psi_{i j}\right|\right)^{2} \tag{A.12}
\end{gather*}
$$

as before. Then (A.2) follows from Assumption 3. To prove (A.3) we show that for all $p \times 1$ vectors $\lambda$, such that $\|\lambda\|^{2}=1$,

$$
\begin{equation*}
\lambda^{\prime} A_{n}^{-\frac{1}{2}} a_{n} \rightarrow{ }_{d} N(0,1) \tag{A.13}
\end{equation*}
$$

conditionally. The left side can be written $\Sigma_{i} z_{i n}$, where

$$
\begin{equation*}
z_{i n}=2 u_{i} \lambda^{\prime} A_{n}^{-\frac{1}{2}} \sum_{j<i} \psi_{i j} u_{j} . \tag{A.14}
\end{equation*}
$$

Clearly $\Sigma_{i} z_{i n}$ has mean zero and variance 1 , so (A.3) follows from Theorem 2 of Scott (1973) on showing that conditionally

$$
\begin{equation*}
\sum_{i} E\left(z_{i n}^{2} \mid u_{i}, j<i\right) \rightarrow_{p} 1 \tag{A.15}
\end{equation*}
$$

$$
\begin{equation*}
\sum_{i} E\left\{z_{i n}^{2} 1\left(\left|z_{i n}\right| \geq \varepsilon\right)\right\} \rightarrow_{p} 0, \quad \forall \varepsilon>0 \tag{A.16}
\end{equation*}
$$

It is easily seen that (A.15) follows if

$$
\begin{equation*}
D_{n}^{-\frac{1}{2}}\left\{\sum_{i}\left(\sum_{j<i} \psi_{i j} u_{j}\right)\left(\sum_{j<i} \psi_{i j} u_{j}\right)^{\prime}-\sigma^{2} \sum_{j<i} \psi_{i j} \psi_{i j}^{\prime}\right\} D_{n}^{-\frac{1}{2}} \rightarrow_{p} 0 \tag{A.17}
\end{equation*}
$$

Again we consider a typical element, and again identify a scalar $\psi_{i j}$ with this; strictly speaking the differential norming needs to be taken account of, but this is a routine aspect. Thus in place of the expression in braces in (A.17) we consider

$$
\begin{equation*}
2 \sum_{i} \sum_{j<i} \psi_{i j}^{2}\left(u_{j}^{2}-\sigma^{2}\right)+\sum_{i} \sum_{\substack{j, k<i \\ j \neq k}} \sum_{i j} \psi_{i k} \psi_{j} u_{k} \tag{A.18}
\end{equation*}
$$

By inequalities of Jensen and of von Bahr and Esseen (1965), the modulus of the first term has mean bounded by

$$
\begin{align*}
C\left\{\sum_{i}\left|\sum_{j<i} \psi_{i j}^{2}\right|^{1+\delta / 2}\right\}^{2 /(2+\delta)} & \leq C\left\{\left(\max _{i} \sum_{j} \psi_{i j}^{2}\right)^{\delta / 2} \sum_{i, j} \psi_{i j}^{2}\right\}^{2 /(2+\delta)} \\
& \leq C\left(\max _{i} \sum_{j}|\psi|\right)^{\delta}\left\{\sum_{i, j} \psi_{i j}\right\}^{2 /(2+\delta)} \\
& =o_{p}\left(\sum_{i, j} \psi_{i j}^{2}\right) \tag{A.19}
\end{align*}
$$

as desired. The second term in (A.18) has zero mean and variance bounded by

$$
\begin{align*}
& C \sum_{\substack{h, i \\
j, k<i, h}}\left|\psi_{i j} \psi_{i k}\right|\left(\left|\psi_{h j} \psi_{h k}\right|\right) \\
\leq & C \sum_{i, j}\left(\sum_{k<i, j}\left|\psi_{i k} \psi_{k j}\right|\right)^{2}=o_{p}\left(\sum_{i, j} \psi_{i j}^{2}\right) \tag{A.20}
\end{align*}
$$

by Assumption 3. This completes the proof of (A.15). To prove (A.16) we check the sufficient Lyapunov condition

$$
\begin{equation*}
\sum_{i} E\left|z_{i n}\right|^{2+\delta} \rightarrow_{p} 0 \tag{A.21}
\end{equation*}
$$

We can instead check the condition with $z_{i n}$ replaced by $\left(\Sigma_{i, j} \psi_{i j}^{2}\right)^{-\frac{1}{2}} u_{i} \Sigma_{j<i} \psi_{i j n} u_{j}$,
again treating $\psi_{i j}$ as a generic element. Thus we need

$$
\begin{equation*}
\left(\sum_{i, j} \psi_{i j}^{2}\right)^{-1-\delta / 2} \sum_{i} E\left(\sum_{j<i} \psi_{i j}^{2} u_{j}^{2}\right)^{1+\delta / 2} \rightarrow_{p} 0 \tag{A.22}
\end{equation*}
$$

by the Marcinkiewicz-Zygmund inequality. The sum is bounded by

$$
\begin{align*}
& C \sum_{i} E\left(\sum_{j}\left|\psi_{i j}\right|^{2+\delta} E\left|u_{j}\right|^{2+\delta}\right) \leq C \sum_{i}\left(\sum_{j} \psi_{i j}^{2}\right)^{\delta / 2} \sum_{j} \psi_{i j}^{2} E\left|u_{j}\right|^{2+\delta} \\
\leq & C\left(\sum_{i, j} \psi_{i j}^{2}\right)^{\frac{1}{2}} \max _{i}\left\{\sum_{j}\left|\psi_{i j}\right|^{2}\right\}^{\delta / 2}, \tag{A.23}
\end{align*}
$$

as desired.

## APPENDIX 2: PROOF OF THEOREM 2

Proof. In place of (A.1)-(A.3) we need that, as $n \rightarrow \infty$,

$$
\begin{align*}
& D_{n}^{-\frac{1}{2}} \hat{B}_{n} D_{n}^{-\frac{1}{2}} \rightarrow_{p} S  \tag{B.1}\\
& B_{n}^{-\frac{1}{2}}\left(\hat{a}_{n}-a_{n}\right) \rightarrow_{p} 0  \tag{B.2}\\
& B_{n}^{-\frac{1}{2}} a_{n} \rightarrow_{d} N\left(0, I_{p}\right) \tag{B.3}
\end{align*}
$$

To prove (B.1) we define

$$
\begin{equation*}
B_{n}=\sum_{i, j} \psi_{i j} \psi_{i j}^{\prime} \sigma_{i}^{q} \sigma_{j}^{q} \tag{B.4}
\end{equation*}
$$

and prove

$$
\begin{align*}
& D_{n}^{-\frac{1}{2}}\left(\tilde{B}_{n}-B_{n}\right) D_{n}^{-\frac{1}{2}} \rightarrow_{p} 0  \tag{B.5}\\
& D_{n}^{-\frac{1}{2}}\left(\hat{B}_{n}-\tilde{B}_{n}\right) D_{n}^{-\frac{1}{2}} \rightarrow_{p} 0 \tag{B.6}
\end{align*}
$$

In both cases, as in part of the proof of Theorem 1 , it clearly suffices to give the proof as if $p=1$, and shows that $\tilde{B}_{n}-B_{n}$ and $\hat{B}_{n}-\hat{B}_{n}$ are both $o_{p}\left(\Sigma_{i, j} \psi_{i j}^{2}\right)$.

We have

$$
\begin{equation*}
\tilde{B}_{n}-B_{n}=\sum_{i, j} \psi_{i j}^{2}\left\{2 \sigma_{i}^{2}\left(u_{j}^{2}-\sigma_{j}^{2}\right)+\left(u_{i}^{2}-\sigma_{i}^{2}\right)\left(u_{j}^{2}-\sigma_{j}^{2}\right)\right\} \tag{B.7}
\end{equation*}
$$

The contribution from the first term in braces has absolute value with expectation bounded by

$$
\begin{equation*}
C\left\{\sum_{i}\left(\sum_{j} \psi_{i j}^{2}\right)^{1+\delta / 2}\right\}^{2 /(2+\delta)}=o_{p}\left(\sum_{i, j} \psi_{i j}^{2}\right) \tag{B.8}
\end{equation*}
$$

as in (A.19). The contribution from the second term has mean zero (because $\left.\psi_{i i}=0\right)$ and variance bounded by

$$
\begin{equation*}
C \sum_{i, j} \psi_{i j}^{4} \leq C\left(\max _{i} \sum_{j}\left|\psi_{i j}\right|\right)^{2} \sum_{i, j} \psi_{i j}^{2} \tag{B.9}
\end{equation*}
$$

as desired.
With respect to (B.6) routine development indicates that it suffices to show that each of the following expansions is $o_{p}\left(\Sigma_{i, j} \psi_{i j}^{2}\right)$ :

$$
\begin{align*}
s_{1} & =\sum_{i, j} \psi_{i j}^{2} u_{i}^{2} u_{j} \hat{v}_{j}, \quad s_{2}=\sum_{i, j} \psi_{i j}^{2} u_{i}^{2} \hat{v}_{j}^{2}  \tag{B.10}\\
s_{3} & =\sum_{i, j} \psi_{i j}^{2} u_{i} u_{j} \hat{v}_{i} \hat{v}_{j}, \quad s_{4}=\sum_{i, j} \psi_{i j}^{2} u_{i} \hat{v}_{i} \hat{v}_{j}^{2}  \tag{B.11}\\
s_{5} & =\sum_{i, j} \psi_{i j}^{2} \hat{v}_{i}^{2} \hat{v}_{j}^{2} \tag{B.12}
\end{align*}
$$

We have

$$
\begin{equation*}
s_{1}=\sum_{i \neq j} \psi_{i j}^{2} u_{i}^{2} u_{j} b_{i j}+\sum_{i, j} \psi_{i j}^{2} u_{i}^{2} u_{j}\left(\sum_{h \neq j} b_{h j} u_{h}\right) \tag{B.13}
\end{equation*}
$$

Feom previous calculations, the modulus of the first term is easily seen to be $O\left(\left(\max _{i} \Sigma_{j}\left|\psi_{i j}\right|\right)^{2}\right)$. The second term can be written

$$
\begin{equation*}
\sum_{i, j} \psi_{i j}^{2} u_{i}^{3} u_{j} b_{i j}+\sum_{i, j} \psi_{i j}^{2} u_{i}^{2} u_{j}\left(\sum_{h \neq i, j} b_{h j} u_{h}\right) \tag{B.14}
\end{equation*}
$$

The modulus of the first term has expectation $O\left(\max _{i}\left(\Sigma_{j}\left|\psi_{i j}\right|\right)^{2}\right)$. The second term has mean zero and variance bounded by

$$
\begin{equation*}
C \sum_{i, j, \ell, h} \psi_{i j}^{2} \psi_{i \ell}^{2} b_{h j}^{2}+C \sum_{i, j, k, \ell} \psi_{i j}^{2} \psi_{k \ell}^{2} b_{\ell j}^{2} \tag{B.15}
\end{equation*}
$$

$$
\begin{align*}
& \leq C \sum_{i, j, \ell} \psi_{i j}^{2} \psi_{i \ell}^{2} b_{j j}+\sum_{i, j, k, \ell} \psi_{i j}^{2} \psi_{k \ell}^{2}\left(b_{\ell \ell}+b_{j j}\right) \\
& \leq C \max _{i} \sum_{\ell} \psi_{i \ell}^{2} \sum_{i, j} \psi_{i j}^{2} b_{j j}+\sum_{i, j} \psi_{i j}^{2} \max _{\ell} \sum_{k}\left|\psi_{k \ell}^{2}\right| \\
& \leq C\left(\max _{i} \sum_{j}\left|\psi_{i j}\right|\right)^{4}+C\left(\max _{i} \sum_{j}\left|\psi_{i j}\right|\right)^{2}\left(\sum_{i, j}\left|\psi_{i j}^{2}\right|\right) \\
& =o\left(\left(\sum_{i, j} \psi_{i j}^{2}\right)^{2}\right) . \tag{B.16}
\end{align*}
$$

Next,

$$
\begin{align*}
E\left|s_{2}\right| & \leq C \sum_{i, j} \psi_{i j}^{2}\left(E \hat{v}_{j}^{4}\right)^{\frac{1}{2}} \leq \sum_{i, j} \psi_{i j}^{2}\left\{\sum_{h} b_{h j}^{2}+\left(\sum_{h} b_{h j}^{4}\right)^{\frac{1}{2}}\right\} \\
& \leq C \sum_{i, j} \psi_{i j}^{2}\left(b_{i i}+b_{j j}\right) \leq C\left(\max _{i} \sum_{j}\left|\psi_{i j}\right|\right)^{2} . \tag{B.17}
\end{align*}
$$

The remaining terms are dealt with similarly. Indeed application of Hölder's and elementary inequalities gives the same bound for $E\left|s_{3}\right|$ as $E\left|s_{2}\right|$, while

$$
\begin{align*}
E\left|s_{4}\right| & \leq C \sum_{i, j} \psi_{i j}^{2}\left(E \hat{v}_{i}^{4}\right)^{\frac{1}{4}}\left(E \hat{v}_{j}^{4}\right)^{\frac{1}{2}} \\
& \leq C \sum_{i, j} \psi_{i j}^{2}\left(\sum_{i} b_{h i}^{2}\right)^{\frac{1}{2}}\left(\sum_{h} b_{h j}^{2}\right) \\
& \leq C \sum_{i, j} \psi_{i j}^{2} b_{i i}^{\frac{1}{2}} b_{j j} \\
& \leq C\left(\max _{i} \sum_{j}\left|\psi_{i j}\right|\right)^{2}, \tag{B.18}
\end{align*}
$$

$$
\begin{align*}
E\left|s_{5}\right| & \leq C \sum_{i, j} \psi_{i j}^{2}\left(E \hat{v}_{i}^{4} E \hat{v}_{j}^{4}\right)^{\frac{1}{2}} \\
& \leq C \sum_{i, j} \psi_{i j}^{2}\left(\sum_{h} b_{h i}^{2}\right)\left(\sum_{h} b_{h j}^{2}\right) \\
& \leq C \sum_{i, j} \psi_{i j}^{2} b_{i i} b_{j j} \\
& \leq C\left(\max _{i} \sum_{j}\left|\psi_{i j}\right|\right)^{2} \tag{B.19}
\end{align*}
$$

in both cases using $b_{i i} \leq 1$. Thus (B.6) is proved.
The proofs of (B.2) and (B.3) hardly differ from those of (A.2) and (A.3). With respect to (B.2), after replacing $B_{n}$ by $D_{n}$ there is no difference due to the uniform bound on relevant moments. The latter is also relevnt to (B.3); we only note that in place of (A.8) we need to establish

$$
\begin{equation*}
D_{n}^{-\frac{1}{2}}\left\{\sum_{n}\left(\sum_{j<i} \psi_{i j} u_{j}\right)\left(\sum_{j<i} \psi_{i j} u_{j}\right)^{\prime}-\sum_{j<i} \psi_{i j} \psi_{i j}^{\prime} \sigma_{j}^{2}\right\} D_{n}^{-\frac{1}{2}} \rightarrow_{p} 0 \tag{B.20}
\end{equation*}
$$

but the details differ only trivially.

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