# BUREAU OF THE CENSUS <br> STATISTICAL RESEARCH DIVISION REPORT SERIES <br> SRD Research Report Number: CENSUS/SRD/RR-89/08 <br> LIKELIHOOD RATIO PROCEDURES FOR COMPARING NON-NESTED, POSSIBLY INCORRECT REGRESSORS 

by

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#### Abstract

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Report completed: August 10, 1989

Report issued:
August 10, 1989
Report corrected: May 28, 1991

# Beyond Chi-Square: Likelihood Ratio Procedures for Comparing Non-Nested, Possibly Incorrect Regressors. 

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## 1. INTRODUCTION AND OVERVIEW.

Applied work involving statistical modeling frequently leads to situations where models must be compared which are not related to one another by parameter restrictions. In such a situation, log-likelihood ratios of pairs of estimated models do not have a chi-square limiting distribution, and statisticians making model selection decisions frequently resort to rather complicated and subjective comparisons of residuals or other model artifacts to accomplish the selection. In this paper, we give some theoretical background for the use of the usual log-likelihood ratios for non-nested comparisons. The practical importance of this capability is magnified by the fact that maximized likelihood values are usually available from the software used for estimation. Thus comparisons can often be made quickly. This encourages inventiveness and experimentation by the modeler.

In fact, the model selection procedures we examine are the minimum AIC procedure of Akaike (1973) and related procedures like the minimum BIC procedure of Akaike (1977) and Schwarz (1978) and the criteria of Hannan and Quinn (1979) and Rissanen (1986). The contributions of the paper stem from its rather comprehensive analysis of situations where the models are non-nested and not necessarily correct, and from the mathematical completeness of the results presented for Gaussian situations with fixed regressors or with vector autoregressions and their
subregressions. We also provide a revision of the principle of parsimony away from its oversimplified emphasis on counts of parameters.

Our analysis is restricted to linear regression models estimated via least squares, because more intelligible and complete results can be obtained for these models. The regressors can be stochastic. The associated parameter estimates maximize a Gaussian likelihood function. (The true likelihood function could be non-Gaussian.) Some comments about generalizations to other models are given in section 11.

After introducing some terminology in section 2, we illustrate, in section 3, the use of the minimum AIC procedure with a regressor selection problem which arose in the design of a ship autopilot and which involves both nested and nonnested comparisons. Section 4 provides our basic theoretical assumptions and the measure of the coefficient estimation variability, $\operatorname{CVAR}^{(x)}$, associated with a regressor process $x_{t}$, which is central to much of the subsequent discussion. In section 5 , formulas for CVAR $^{(x)}$ are given which show that this quantity is equal to the number of coefficients estimated when $x_{t}$ is complete in the sense that it contains the correct regressor as a subvector. Subsection 5.3 shows that CVAR $^{(x)}$ approximates this number when $x_{t}$ is "almost complete." Subsection 5.1 contain the initial analysis of an important example of two asymptotically equivalent but incomplete autoregressions with the property that the value of $\operatorname{CVAR}^{(\mathrm{x})}$ is larger for the model with fewer estimated coefficients, contradicting the principle of parsimony. Here, asymptotically equivalent means that the difference between the estimated regression functions tends to zero in probability as the sample size increases.

In section 6 some easy results are presented describing situations in which a variety of $\log$-likelihood-ratio based model selection criteria prefer one regressor over another with asymptotic probability 1. Some criteria, like BIC, are seen to consistently prefer a model with fewer estimated coefficients whenever the
log-likelihood ratio is bounded in probability, a preference which is sometimes undesirable as the Example 5.1 shows.

The next several sections analyze the asymptotic behavior of the $\log$-likelihood ratio $\hat{\mathrm{L}}_{\mathrm{N}}^{(1,2)}$ in the only simply defined situation in which the sequence $\hat{\mathrm{L}}_{\mathrm{N}}^{(1,2)}, \mathrm{N} \geq$ $\mathrm{N}_{0}$ is bounded in probability, the situation of asymptotically equivalent regressors. Section 7 investigates the limiting distribution of the log-likelihood ratio and its connection with CVAR $^{(x)}$ values and with the MAIC criterion and a modification thereof. In subsection 7.1 we comment on the use and limitations of the complete-regressor form of the limiting distribution for hypothesis testing with non-nested regressors.
"Section 8 shows that the difference of Kullback-Leibler ("entropy" or "information") numbers of the estimated models overcomes some of the deficiencies of the $\log$-likelihood ratio and motivates the definition of an "ideal" minimum AIC criterion. The generalization (8.11) of a result of Akaike and Shimizu connecting K-L numbers and log-likelihood ratios plays an important role here.

In section 9 , we show that under fairly general circumstances, when two regressors are asymptotically equivalent, one can expect the difference of their $\operatorname{CVAR}^{(x)}$ values to be the limit of the differences of a normalized measure of the mean square prediction errors arising when the estimated regression coefficients are used to predict an independent replicate of the observations. Thus, there is a predictive interpretation of the results of sections 5,7 and 8 . Section 10 completes this discussion by presenting a strategy for showing that finite sample means of the log-likelihood ratios, of AIC differences, of K-L number differences, and of differences of mean square prediction errors, converge as expected. A new lemma on the rate of decrease of the inverse moments of the Wishart distribution makes it possible to verify the assumptions of section 10 for regressors which are fixed or are subvectors
of not necessarily stationary autoregressions with Gaussian noise processes. For these situations, we thereby achieve the first complete demonstration of the bias correction property used by Akaike (1973) to motivate the definition of AIC.

Section 11 contains comments and literature references concerning generalizations of the results of this paper to models different from linear regression models. The Appendices I and II contain proofs omitted from the initial discussion.
2. LINEAR LEAST SQUARES AND MAIC.

Let $y_{t}$ be a $q$-dimensional regressand and $x_{t}$ an $r$-dimensional candidate regressor process for $y_{t}$ satisfying

$$
\begin{equation*}
\sum_{t=1}^{N} x_{t} x_{t}^{\prime}>0, N \geq N_{0} \tag{2.1}
\end{equation*}
$$

with probability one (w.p.1). The coordinate entries of $x_{t}$ can be fixed or random.
Although $\mathrm{y}_{1}, \ldots, \mathrm{y}_{\mathrm{N}}$ need not be Gaussian, the least squares coefficient and error variance estimates for the regression of $y_{t}$ on $x_{t}$,

$$
\begin{equation*}
\hat{A}_{N}^{(x)} \equiv\left(\sum_{t=1}^{N} y_{t} x_{t}^{\prime}\right)\left(\sum_{t=1}^{N} x_{t} x_{t}^{\prime}\right)^{-1} \tag{2.2}
\end{equation*}
$$

and

$$
\begin{equation*}
\hat{\Sigma}_{N}^{(x)} \equiv N^{-1} \sum_{t=1}^{N}\left(y_{t}-\hat{A}_{N^{x}}\right)\left(y_{t}-\hat{A}_{N^{x}}\right)^{\prime}, \tag{2.3}
\end{equation*}
$$

are the maximizers of a Gaussian $\log$-(quasi)likelihood function

$$
\mathrm{L}_{\mathrm{N}}^{(\mathrm{x})}[\Sigma, \mathrm{A}] \equiv-\frac{\mathrm{N}}{2} \log 2 \pi|\Sigma|-\frac{1}{2} \operatorname{tr} \Sigma^{-1} \sum_{\mathrm{t}=1}^{\mathrm{N}}\left(\mathrm{y}_{\mathrm{t}}-\mathrm{Ax} \mathrm{t}_{\mathrm{t}}\right)\left(\mathrm{y}_{\mathrm{t}}-\mathrm{Ax} \mathrm{x}_{\mathrm{t}}\right)^{\prime}
$$

whose maximum value is

$$
\begin{equation*}
\hat{\mathrm{L}}_{\mathrm{N}}^{(\mathrm{x})} \equiv \mathrm{L}_{\mathrm{N}}^{\left.(\mathrm{x})_{\left[\hat{\Sigma}_{\mathrm{N}}\right.}^{(\mathrm{x})}, \hat{\mathrm{A}}_{\mathrm{N}}^{(\mathrm{x})}\right]=-\frac{\mathrm{N}}{2}\left(\log 2 \pi\left|\hat{\Sigma}_{\mathrm{N}}^{(\mathrm{x})}\right|+\mathrm{q}\right) . . . . .} \tag{2.4}
\end{equation*}
$$

(We use tr to denote trace and $\equiv$ to indicate the definition of a symbol.) When two competing regressor processes $x_{t}^{(i)}, i=1,2$ are being considered, we will replace the superscript ${ }^{(x)}$ by the superscript ${ }^{(i)}$ in the preceding notation to indicate quantities associated with $\mathbf{x}_{\mathrm{t}}^{(\mathrm{i})}$. Our investigation focuses on the log-likelihood ratio,

$$
\begin{equation*}
\hat{\mathrm{L}}_{\mathrm{N}}^{(1,2)} \equiv \hat{\mathrm{L}}_{\mathrm{N}}^{(1)}-\hat{\mathrm{L}}_{\mathrm{N}}^{(2)}=-\frac{\mathrm{N}}{2} \log \left(\left|\hat{\mathrm{~L}}_{\mathrm{N}}^{(1)}\right| /\left|\hat{\Sigma}_{\mathrm{N}}^{(2)}\right|\right) \tag{2.5}
\end{equation*}
$$

and several modifications thereof for regressor comparison purposes, the best known of which is due to Akaike (1973, 1974),

$$
\begin{equation*}
\operatorname{AIC}_{\mathrm{N}}^{(1,2)} \equiv(-2) \hat{\mathrm{L}}_{\mathrm{N}}^{(1,2)}+2 \mathrm{q}\left(\mathrm{r}^{(1)}-\mathrm{r}^{(2)}\right) \tag{2.6}
\end{equation*}
$$

with $\mathrm{r}^{(\mathrm{i})} \equiv \operatorname{dim} \mathrm{X}_{\mathrm{t}}^{(\mathrm{i})}, \mathrm{i}=1,2$.
We will write $\mathrm{x}_{\mathrm{t}}^{(1)} \subseteq \mathrm{x}_{\mathrm{t}}^{(2)}$ to indicate that $\mathrm{x}_{\mathrm{t}}^{(1)}=\mathrm{B} \mathrm{x}_{\mathrm{t}}^{(2)}$ for some matrix B . In this case, we will say that $x_{t}^{(1)}$ is nested in $x_{t}^{(2)}$. When this happens, if $x_{t}^{(1)}$ is correct in a strong sense and has certain stability properties (see Lai and Wei (1982) and sections 4 and 5 below), then $(-2) \hat{\mathrm{L}}_{\mathrm{N}}{ }^{(1,2)}$ will have an asymptotic $\chi^{2}\left(\mathrm{q}\left(\mathrm{r}^{(2)}\right.\right.$ $\left.r^{(1)}\right)$ ) distribution. However, in this paper, we are interested in the situation in which the regressors may be non-nested and only approximately correct.

The quantity (2.6) is the difference of the two AIC statistics,

$$
\begin{equation*}
A I C_{\mathrm{N}}^{(\mathrm{i})} \equiv-2 \hat{\mathrm{~L}}_{\mathrm{N}}^{(\mathrm{i})}+2 \mathrm{qr} \mathrm{r}^{(\mathrm{i})} \quad(\mathrm{i}=1,2) \tag{2.7}
\end{equation*}
$$

Akaike's minimum AIC criterion (MAIC) asserts that the regressor associated with the smaller AIC value should be preferred: thus, $\mathrm{x}_{\mathfrak{t}}^{(1)}$ is preferred if $\operatorname{AIC}_{\mathrm{N}}^{(1,2)}<0$. The next section presents an application of MAIC in which this criterion exhibits consistent performance across a range of nested and nonnested comparisons.

## 3. SHIP AUTOPILOT MODELING WITH MAIC: AMERIKA MARU DATA

In Kitagawa and Ohtsu (1976) and Ohtsu et al. (1979) and the papers referenced there, the design and testing of a stochastic-regression-modelbased ship autopilot is described. The success of this experiment influenced the desigin of a new ship (Shoï Maru III) incorporating such an autopilot (K. Ohtsu, personal communication, January 1987). The principal variable to be controlled is yaw (Y), the angular deviation of the ship's forward movement from the intended direction, measured at the bridge. Other less important but useful variables to control include roll (R) and pitch (P). The rudder angle (RU) is the main controller input variable, but measured values of the lateral acceleration (LACC) and vertical acceleration (VACC) of the forepeak may also provide useful information for the controller/autopilot.

Our analysis will seek to determine the situations in which VACC is a useful controller input variable for a specific ship: we consider the problem of choosing between the regressors $\mathbf{x}_{t}^{(m)}$ and $\tilde{\mathbf{x}}_{t}^{(M)}$, these being defined by

$$
\begin{aligned}
& x_{t}^{(m)} \equiv\left(Y_{t-1}, R_{t-1}, P_{t-1}, R U_{t-1}, L A C C_{t-1}, \ldots Y_{t-m}, R_{t-m}, P_{t-m}, R U_{t-m},\right. \\
& \left.L A C C_{t-m}\right)
\end{aligned}
$$

and

$$
\tilde{x}_{t}^{(M)} \equiv\left(x_{t}^{(M)^{\prime}}, V A C C_{t-1}, \ldots, \operatorname{VACC}_{t-M}\right)^{\prime}
$$

for $1 \leq \mathrm{m}, \mathrm{M} \leq 10$. If $\mathrm{M}<\mathrm{m}$, these regressors are non-nested. The modeling will be done with $\mathrm{N}=894$ observations made at 1 second intervals on the container ship Amerika Maru under manual control. These data are discussed in the papers cited above. If $\hat{\mathrm{m}}$ and $\hat{\mathrm{M}}$ denote the lags associated with minimum AIC values for the regressors $\mathrm{x}_{\mathrm{t}}^{(\mathrm{m})}, 1 \leq \mathrm{m} \leq 10$ and $\tilde{\mathrm{x}}_{\mathrm{t}}^{(\mathrm{M})}, \quad 1 \leq M \leq 10$, respectively, then the use of VACC in the autopilot model seems worth considering seriously when

$$
\mathrm{DAIC}_{894} \equiv \mathrm{AIC}_{894}^{(\hat{\mathrm{M}})}-\mathrm{AIC}_{894}^{(\hat{\mathrm{m}})}
$$

is negative. Results obtained from the program MULCON of Akaike et al. (1985) for seven choices of the regressand $y_{t}$ are included in Table 3.1 below. The choices for $y_{t}$ are: $Y_{t}, R_{t}, P_{t},\left(Y_{t}, R_{t}\right)^{\prime},\left(Y_{t}, P_{t}\right)^{\prime},\left(R_{t}, P_{t}\right)^{\prime}$ and $\left(Y_{t}, R_{t}, P_{t}\right)^{\prime}$. In the table, LAG denotes $\hat{\mathrm{M}}$ or $\hat{\mathrm{m}}$, as appropriate, and

$$
\Delta \operatorname{dim} A \equiv q(6 \hat{M}-5 \hat{m})
$$

with $\mathrm{q}=$ dimy $_{\mathrm{t}}$. The results are consistent: the use of VACC is favored only when $P$ is one of the controlled variables. This conclusion has engineering plausibility: VACC is closely related to $P$ but not to the other controlled variables. Thus, MAIC has functioned quite satisfactorily. Note also that in the two cases, $y_{t}=Y_{t}$ and $y_{t}=P_{t}$, the comparison is between non-nested regressors, since $\hat{M}<\dot{m}$. Hypothesis testing based on an asymptotic distribution for $L_{N}^{(1,2)}$ leads to the same conclusions, but this approach has some significant limitations, see subsection 7.1.

Table 3.1. DAIC Values for Various Choices of $y$
With VACC
y $\quad$ LAG $\quad \operatorname{dim} A \quad$ AIC $_{894}$

| Without VACC |  | Difference |  |
| :---: | :---: | :---: | :---: |
| LAG | $\operatorname{dimA}$ | AIC $_{894}$ | $\Delta \operatorname{dimA~DAIC}$ |
| 894 |  |  |  |


| Y,R,P | 8 | 108 | 18683. | 6 | 90 | 18693. | 18 | -10. |
| ---: | ---: | ---: | ---: | ---: | ---: | ---: | ---: | ---: |
| Y,R | 6 | 72 | 12244. | 6 | 60 | 12236. | 12 | 8. |
| Y,P | 7 | 84 | 12468. | 7 | 70 | 12483. | 14 | -15. |
| R,P | 8 | 96 | 13033. | 8 | 80 | 13036. | 16 | -3. |
| Y | 4 | 24 | 6027. | 6 | 30 | 6025. | -6 | 2. |
| R | 5 | 30 | 6574. | 5 | 25 | 6569. | 5 | 5. |
| P | 7 | 42 | 6393. | 8 | 40 | 6491. | 2 | -14. |

An additional analysis with MAIC to determine which components' error processes are uncorrelated is discussed in Findley (1988).

## 4. BASIC ASSUMPTIONS

The fundamental issues we wish to discuss can be described in the context of selecting between two competing regressor processes $\mathrm{x}_{\mathrm{t}}^{(\mathrm{i})}, \mathrm{i}=1,2$. Our minimal assumption beyond (2.1) is that the estimated error variance matrices $\hat{\Sigma}_{\mathbf{N}}^{(i)}$ converge in probability to positive definite limits, $\hat{\mathbf{\Sigma}}_{\mathrm{N}}^{(i)} \xrightarrow{\mathrm{N}} \mathrm{\Sigma}^{(\mathrm{i})}>0$, so that the log-likelihood ratio satisfies

$$
\begin{equation*}
N^{-1} \hat{L}_{N}(1,2) \xrightarrow{N}-(1 / 2) \log \left(\left|\Sigma^{(1)}\right| /\left|\Sigma^{(2)}\right|\right) \tag{4.1}
\end{equation*}
$$

In what follows, $x_{t}$ usually designates either of the regressors $x_{t}(i), j=1,2$. We will assume that a matrix $A^{(x)}$ exists such that $\hat{A}_{N}^{(x)} \xrightarrow[p]{N} A^{(x)}$ holds. Defining $e_{t}^{(x)} \equiv y_{t}-A^{(x)} x_{t}$, we will call the equation

$$
y_{t}=A^{(x)} x_{t}+e_{t}^{(x)}
$$

the model associated with the regressor process $x_{t}$. We note that

$$
\begin{equation*}
\hat{A}_{N}^{(x)}-A^{(x)}=\sum_{t=1}^{N} e_{t}^{(x)} x_{t}^{\prime}\left(\sum_{t=1}^{N} x_{t} x_{t}^{\prime}\right)^{-1} \tag{4.2}
\end{equation*}
$$

and that $\hat{\Sigma}_{N}^{(x)}$ differs from $\Sigma_{N}^{(x)} \equiv N^{-1} \Sigma_{t=1}^{N} e_{t}^{(x)} e_{t}^{(x)^{\prime}}$ by the quantity

$$
\begin{equation*}
\Sigma_{N}^{(x)}-\hat{\Sigma}_{N}^{(x)}=N^{-1}\left(\hat{A}_{N}^{(x)}-A^{(x)}\right) \sum_{t=1}^{N} x_{t} x_{t}^{\prime}\left(A_{N}^{(x)}-A^{(x)}\right)^{\prime} \tag{4.3}
\end{equation*}
$$

We will now introduce a measure of model uncertainty (or variability) due to parameter estimation which is invariant under "scale" transformations $y_{t} \rightarrow \mathrm{By}_{\mathrm{t}}$, $x_{t} \rightarrow C x_{t}$ with nonsingular $B$ and $C$. In the situation of interest in sections $5-7$, where the regressors are asymptotically equivalent, the effects of estimating $\hat{\mathbf{\Sigma}}_{\mathrm{N}}^{(\mathrm{x})}$ are the same for both regressors and cancel in the $\log$-likelihood ratio $\hat{\mathrm{L}}_{\mathrm{N}}^{(1,2)}$, see Proposition 6.3. Our measure will therefore focus on the coefficient estimates. We define

$$
\begin{equation*}
Q_{N}^{(x)} \equiv \operatorname{tr}\left(\Sigma^{(x)}\right)^{-1}\left(\hat{A}_{N}^{(x)}-A^{(x)}\right)\left(\sum_{t=1}^{N} x_{t} x_{t}^{\prime}\right)\left(\hat{A}_{N}^{(x)}-A^{(x)}\right)^{\prime} \tag{4.4}
\end{equation*}
$$

For purposes of interpretation, we note that this reduces to the total squared estimation error of the coefficients, $\operatorname{tr}\left(\hat{\mathrm{A}}_{\mathrm{N}}{ }^{(\tilde{x})}-\mathrm{A}^{(\tilde{\mathrm{x}})}\right)\left(\hat{\mathrm{A}}_{\mathrm{N}}{ }^{(\tilde{x})}-\mathrm{A}^{(\tilde{\mathrm{x}})}\right)^{\prime}$, if the $\mathrm{y}_{\mathrm{t}}$ and $x_{t}$ are transformed in such a way that $\Sigma^{(e)}=I_{q}$ and $\Sigma_{t=1}^{N} \tilde{x}_{t} \tilde{x}_{t}^{\prime}=I_{r} \quad$ Using (4.3), we could also write $Q_{N}^{(x)}=\operatorname{Ntr}\left(\Sigma^{(x)}\right)^{-1}\left(\hat{\Sigma}_{N}^{(x)}-\Sigma_{N}(x)\right.$ and observe that, since $\hat{\Sigma}_{N}(x) \xrightarrow[p]{N} \Sigma^{(x)}$, the variate $Q_{N}^{(x)}$ is asymptotically equivalent to the final term of the decomposition $\hat{L}_{N}^{(x)}=-\frac{N_{N}^{2}}{2}\left\{\log 2 \pi\left|\hat{\Sigma}_{N}^{(x)}\right|+\operatorname{tr}\left(\hat{\Sigma}_{N}^{(x)}\right)^{-1} \Sigma_{N}\right\}+\operatorname{tr}\left(\hat{\Sigma}_{N}^{(x)}\right)^{-1}\left(\Sigma_{N}-\right.$
$\hat{\Sigma}_{N}^{(x)}$ ) obtained from (2.4) via the substitution $q=\operatorname{tr}\left(\hat{\Sigma}_{N}^{(x)}\right)^{-1} \hat{\Sigma}_{N}^{(x)}$. We shall assume that
(A1) $Q_{N}^{(x)} \xrightarrow{N}$ dist. $Q^{(x)}$, and $E Q^{(x)}<\infty$.

Our measure of (asymptotic) model uncertainty due to coefficient estimation is defined to be

$$
\begin{equation*}
\operatorname{CVAR}^{(x)} \equiv E Q^{(x)} \tag{4.5}
\end{equation*}
$$

*Explicit formulas for $\mathrm{CVAR}^{(\mathrm{x})}$ will be given in the next section. Its connection with the MAIC procedure will be revealed in sections 5,8 and 11 . In section 9 an alternative measure is described which provides a connection between model uncertainty and prediction error.

Our usual method of verifying (A1) will involve establishing that there is a vector variate $t_{n}^{(x)}$ satisfying

$$
\begin{equation*}
\mathrm{Q}_{\mathrm{N}}^{(\mathrm{x})}=\mathrm{t}_{\mathrm{N}}^{\left(\mathrm{x}^{\prime}\right.} \mathrm{t}_{\mathrm{N}}^{(\mathrm{x})} \tag{4.6}
\end{equation*}
$$

which has a limiting distribution with finite mean and variance. To define this variate, we need some notation. Given a positive definite matrix $\Sigma$, we will use $\Sigma^{1 / 2}$ to denote any matrix $S$ with the property that $\Sigma=S^{\prime}$, providing it is formed continuously, meaning that $\Sigma_{N} \xrightarrow{N} \Sigma$ implies $\Sigma_{N}^{1 / 2} \xrightarrow{N} \Sigma^{1 / 2}$. The Cholesky factorization is an example. For a matrix explicitly of the form CEC', the square root of choice will be $C \Sigma^{1 / 2}$. We will denote the inverse of $\Sigma^{1 / 2}$ by $\Sigma^{-1 / 2}$, this being different from $\left(\Sigma^{-1}\right)^{1 / 2}=\left\{\left(\Sigma^{1 / 2}\right)^{\prime}\right\}^{-1}$ in general. We define
where vec[•] denotes the column vector obtained by stacking the columns of the matrix [•]. This satisfies (4.6). It is easy to check that the variates defined by (4.7) are invariant under nonsingular linear transformations of $x_{t}$ or $y_{t}$.

It follows from (A1) that $Q_{N}$ is bounded in probability ( $\left.0_{p}(1)\right)$. Thus the term on the right in (4.3) converges to 0 in probability, with the result that $\Sigma_{N}(\mathrm{x}) \xrightarrow[\mathrm{p}]{\mathrm{N}} \mathrm{\Sigma}^{(\mathrm{x})}$.
. We will occasionally need to assume
(A2) $\mathrm{N}^{1 / 2}\left(\hat{\mathrm{\Sigma}}_{\mathrm{N}}^{(\mathrm{x})}-\Sigma^{(\mathrm{x})}\right)$ is bounded in probability.

This condition is satisfied when $N^{1 / 2}\left(\hat{\mathbf{\Sigma}}_{\mathrm{N}}^{(\mathrm{x})}-\mathrm{\Sigma}^{(\mathrm{x})}\right.$ ) has a limiting distribution. Two further simplifying assumptions sometimes called upon are
(A3) $\mathrm{Ee}_{\mathrm{t}}^{(\mathrm{x})} \mathrm{x}_{\mathrm{t}}^{\prime}=0$,
(A4) $E e_{t}^{(x)} e_{t}^{(x)}=\Sigma^{(x)}$.

Note that, for non-stochastic regressors, (A3) is equivalent to $E y_{t}=A^{(x)} x_{t}$, meaning that $x_{t}$ has been chosen well enough to capture the mean behavior of $y_{t}$. Shibata (1981) presents results for fixed regressors when (A3) fails, for the case in which $\mathbf{y}_{\mathrm{t}}-E y_{\mathrm{t}}$ is i.i.d. and Gaussian. We discuss his results briefly in section 11.

To obtain formulas for $\operatorname{CVAR}^{(\mathrm{x})}$, we also require

$$
\begin{equation*}
\left\{E\left(\sum_{t=1}^{N} x_{t} x_{t}^{\prime}\right)\right\}^{-1 / 2}\left\{\sum_{t=1}^{N} x_{t} x_{t}^{\prime}\right\}^{1 / 2} \xrightarrow{N} I_{r} \tag{A5}
\end{equation*}
$$

Here $I_{r}$ denotes the identify matrix of order $r=\operatorname{dim} x_{t}$. (A5) is satisfied, for example, if $x_{t}$ is nonstochastic, or if $x_{t}$ is stationary and $N^{-1} \Sigma_{t=1}^{N} x_{t} x_{t}^{\prime} \xrightarrow{N} \Gamma^{(x)} \equiv E x_{t} x_{t}^{\prime}$. It implies that the difference between $t_{N}(x)$
and

$$
\begin{equation*}
Z_{N}^{(x)} \equiv \operatorname{vec}\left[\left(\Sigma^{(x)}\right)^{-1 / 2}\left(\sum_{t=1}^{N} e_{t}^{(x)_{x_{t}^{\prime}}}\right)\left\{\left(E \sum_{t=1}^{N} x_{t} x_{t}^{\prime}\right)^{-1}\right\}^{1 / 2}\right] \tag{4.8}
\end{equation*}
$$

tends to zero in probability, a situation we denote by

$$
\begin{equation*}
t_{N}(x) \sim_{p} Z_{N}^{(x)} \tag{4.9}
\end{equation*}
$$

Hence, under (A5), also, $Z_{N}^{(x)^{\prime}} Z_{N}^{(x)} \xrightarrow{N}$ dist. $^{(x)}$, the limiting distribution in (A1).
5. FORMULAS FOR CVAR ${ }^{(x)}$ WHEN THE LIMITING DISTRIBUTION IS GAUSSIAN.

We will present formulas in subsection 5.1 for the situation in which the limiting distribution of $t_{N}^{(x)}$ in (4.7) is Gaussian with mean zero, and the regressor process $x_{t}$ is not complete, meaning that neither $x_{t}$ nor any subvector is a correct regressor in the sense of subsection 5.2. These results require the joint stationarity
of $x_{t}$ and $y_{t}$ (or $e_{t}$ ), but stationarity is not needed for the familiar formula, $\operatorname{CVAR}^{(\mathrm{x})}=\mathrm{qr}\left(=\operatorname{dim} \mathrm{A}^{(\mathrm{x})}\right.$ ), obtained in subsection 5.2 for complete regressors. We will assume throughout this section that (A3) - (A5) hold and will refer to theorems in the literature verifying (A1).

### 5.1. Stationary Case.

Making joint stationarity assumptions for $x_{t}, y_{t}$, we define $\Gamma^{(x)} \equiv E x_{t} x_{t}^{\prime}$ and


$$
\begin{equation*}
\left.-\quad Z_{N}^{(x)}=N^{-1 / 2} \operatorname{vec}\left[{\underset{\mathrm{E}=1}{\mathrm{~N}}}_{\mathrm{N}}\left\{\left(\Sigma^{(\mathrm{x})}\right)^{-1 / 2} \mathrm{e}_{\mathrm{t}}^{(\mathrm{x})}\right]\left\{\left(\Gamma^{(\mathrm{x})}\right)^{-1 / 2} \mathrm{x}_{\mathrm{t}}\right)\right\}^{\prime}\right] \tag{5.1}
\end{equation*}
$$

is $\mathrm{N}^{1 / 2}$ times the sample mean of the mean zero stationary vector process

$$
\begin{equation*}
\mathbf{M}_{t}^{(x)}=\operatorname{vec}\left[\left\{\left(\Sigma^{(x)}\right)^{-1 / 2} e_{t}(x)\right\}\left\{\left(\Gamma^{(x)}\right)^{-1 / 2} x_{t}\right\}^{\prime}\right] \tag{5.2}
\end{equation*}
$$

Therefore, a variety of Central Limit Theorem results apply to (5.1), see Theorem 5.2 of Brillinger (1969), Hannan (1970, pp. 220-228), Corollary (3.9) of McLeish (1975), Dahlhaus (1985) and Eberlein (1986). Under diverse assumptions, these results yield

$$
\begin{equation*}
\mathrm{z}_{\mathrm{N}}(\mathrm{x}) \xrightarrow[\text { dist. }]{\mathrm{N}} .(0, \mathrm{~V})^{(1)} \tag{5.3}
\end{equation*}
$$

with

$$
\operatorname{CVAR}^{(x)}=\operatorname{trV}=\lim _{N \rightarrow \infty} \operatorname{EZ}_{\mathrm{N}}^{(\mathrm{x})^{\prime}} \mathrm{Z}_{\mathrm{N}}^{(\mathrm{x})}
$$

or an equivalent expression involving integrals of cumulant spectral densities, see Brillinger (1969) and (5.6) below.

The fourth cumulants associated with the fourth moment quantities in (5.4) vanish when $e_{t}$ and $x_{t}$ are jointly Gaussian and also in the not necessarily Gaussian univariate autogression situation, where $y_{t}$ is scalar with mean zero and $x_{t}=$ $\left[y_{t-m_{1}} \ldots y_{t-m_{r}}\right]^{\prime}$ for positive integers $m_{1}<m_{2}<\cdots<m_{r}$, see Remark 3.2 of Hosoya and Taniguchi (1982, p. 138). In these cases, Isserlis' formula (Brillinger, 1975, p. 21) and (A3) can be used to show that (5.4) reduces to

$$
\begin{align*}
\operatorname{CVAR}_{G}^{(x)}= & \sum_{k=-\infty}^{\infty} \operatorname{tr}\left\{\Sigma^{(\mathrm{x})-1} \Gamma^{(\mathrm{e})}(\mathrm{k})\right\} \operatorname{tr}\left\{\Gamma^{(\mathrm{x})-1} \Gamma^{(\mathrm{x})}(\mathrm{k})\right\} \\
& +\sum_{\mathbf{k}=-\infty}^{\infty} \operatorname{tr}\left\{\tilde{\Gamma}^{\mathrm{ex}}(\mathrm{k}) \tilde{\Gamma}^{\mathrm{ex}}(-\mathrm{k})^{\prime}\right\} \tag{5.5}
\end{align*}
$$

where $\Gamma^{(x)}(k) \equiv E x_{t} x_{t+k}^{\prime}, \Gamma^{(e)}(k) \equiv E e_{t}^{(x)} e_{t+k}^{(x)^{\prime}}$, and

$$
\tilde{\Gamma}^{e x}(k) \equiv E\left[\left\{\left(\Sigma^{(x)}\right)^{-1 / 2} e_{t}^{(x)}\right\}\left\{\left(\Gamma^{(x)}\right)^{-1 / 2} x_{t+\mathbf{k}}\right\}^{\prime}\right]
$$

A convenient spectral density form of (5.5) follows via Parseval's formula:

$$
\operatorname{CVAR}_{G}(x)=2 \pi \int_{-\pi}^{\pi} \operatorname{tr}\left\{\Sigma^{(e)-1} f_{f}^{e e}(\lambda)\right\} \operatorname{tr}\left\{\Gamma^{(x)-1} f^{\mathrm{xx}}(\lambda)\right\} \mathrm{d} \lambda
$$

$$
\begin{equation*}
+2 \pi \int_{-\pi}^{\pi} \operatorname{tr}\left\{\Sigma^{(x)-1} \mathrm{f}_{\mathrm{f}}^{\mathrm{ex}}(\lambda) \Gamma^{(\mathrm{x})-1} \mathrm{f}^{\mathrm{xe}}(-\lambda)\right\} \mathrm{d} \lambda, \tag{5.6}
\end{equation*}
$$

where $f^{e e}(\lambda)$ and $f^{x x}(\lambda)$ are the spectral density matrices of $e_{t}$ and $x_{t}$, respectively, $\mathrm{f}^{\mathrm{ex}}(\lambda)=(2 \pi)^{-1} \Sigma_{\mathbf{k}=-\infty}^{\infty} \tilde{\Gamma}^{\mathrm{ex}}(\mathrm{k}) \mathrm{e}^{-\mathrm{i} k}$, and $\mathrm{f}^{\mathrm{xe}}(\lambda)=(2 \pi)^{-1} \Sigma_{\mathbf{k}=-\infty}^{\infty} \tilde{\Gamma}^{\mathrm{ex}}(-\mathrm{k}) \cdot \mathrm{e}^{-\mathrm{ik} \lambda}$.

For scalar processes $\mathrm{y}_{\mathrm{t}}$, set $\rho_{\mathrm{k}} \equiv \mathrm{Ey}_{\mathrm{t}} \mathrm{y}_{\mathrm{t}+\mathrm{k}} / \mathrm{Ey}_{\mathrm{t}}{ }^{2}$. If

$$
\begin{equation*}
\rho_{\mathrm{m}}=\rho_{\mathrm{m}_{1}}=\rho_{\mathrm{m}_{2}}=\rho_{\mathrm{m}_{1}-\mathrm{m}_{2}}=0 \tag{5.7}
\end{equation*}
$$

for distinct lags $m, m_{1}, m_{2}$, two autoregressions, with $x_{t}^{(1)} \equiv y_{t-m}$ and $\mathrm{x}_{\mathrm{t}}^{(2)} \equiv\left[\mathrm{y}_{\mathrm{t}-\mathrm{m}_{1}} \mathrm{y}_{\mathrm{t}-\mathrm{m}_{2}}\right]^{\prime}$, will turn out to be of special interest. The condition (5.7) implies that $A^{(1)}=0$ and $A^{(2)}=0$, so that $e_{t}^{(1)}=e_{t}^{(2)}=y_{t}$. It follows from (5.5) that if $\mathrm{V}_{\mathrm{n}} \equiv \sum_{\mathrm{k}=-\infty}^{\infty}\left\{\rho_{\mathrm{k}}^{2}+\rho_{\mathrm{k}+\mathrm{n}} \rho_{\mathrm{k}-\mathrm{n}}\right\}$, then

$$
\begin{equation*}
\operatorname{CVAR}^{(1)}=\mathrm{V}_{\mathrm{m}} \tag{5.8}
\end{equation*}
$$

and

$$
\begin{equation*}
\operatorname{CVAR}^{(2)}=\mathrm{V}_{\mathrm{m}_{1}}+\mathrm{V}_{\mathrm{m}_{2}} \tag{5.9}
\end{equation*}
$$

Example 5.1. Suppose $y_{t}$ is a stationary autoregressive process of order 6 with variance 1 whose first six partial autocorrelations are $0.0,0.0,0.0, .80,-.41,-.64$. The Levinson-Durbin algorithm (Box and Jenkins (1976, p. 83) can be used to
calculate the autoregressive coefficients and the autocorrelations $\rho_{k}$. The vanishing of the first three partial autocorrelations is equivalent to

$$
\begin{equation*}
\rho_{1}=\rho_{2}=\rho_{3}=0 . \tag{5.10}
\end{equation*}
$$

With $\mathrm{x}_{\mathrm{t}}^{(1)}=\mathrm{y}_{\mathrm{t}-2}$ and $\mathrm{x}_{\mathrm{t}}^{(2)}=\left[\mathrm{y}_{\mathrm{t}-1} \mathrm{y}_{\mathrm{t}-3}\right]^{\text {, the formulas (5.8) and (5.9) yield }}$

$$
\operatorname{CVAR}^{(1)}=26.3
$$

and

$$
\begin{equation*}
\operatorname{CVAR}^{(2)}=2.9+2.4=5.3 \tag{5.11}
\end{equation*}
$$

*Thus, although the regressors $x_{t}^{(1)}$ and $x_{t}^{(2)}$ are asymptotically equivalent in the sense that $e_{t}^{(1)}=e_{t}^{(2)}$, the more parsimonious regressor $x_{t}^{(1)}$ has greater coefficient estimation variability as measured by CVAR. In fact, for 577 out of 1000 models obtained under (5.7) by choosing the partial autocorrelations at lags 4-6 uniformly and independently, it happened that the regression on one of $y_{t-1}, y_{t-2}, y_{t-3}$ had a larger value of CVAR than the regression on the remaining pair of lagged $y$-variates. Some implications of this will be discussed in sections 7-9. This phenomenon does not occur when both regressors are complete in the sense we will now describe.

### 5.2. Complete Regressors.

Let $I_{t}$ denote an information set ( $\sigma$-algebra) containing the information generated by the "past history up to $t$ " of all regressors under consideration, $I_{t} \supseteq$ $\sigma\left(\mathrm{x}_{\mathrm{t}}^{(1)}, \ldots, \mathrm{x}_{1}^{(1)} ; \mathrm{x}_{\mathrm{t}}^{(2)}, \ldots, \mathrm{x}_{1}^{(2)} ; \mathrm{y}_{\mathrm{t}-1}, \ldots, \mathrm{y}_{0} ; \ldots\right)$. It would be natural to say that a regressor $x_{t}$ which is determined by $I_{t}$ (that is, is $I_{t}$-measurable) for each $t=1$, $2, \ldots$, is correct if for some matrix $A^{(x)}$, all of whose columns are non-zero, we have $E\left(y_{t} \mid I_{t}\right)=A^{(x)} x_{t}$, or, equivalently, with $e_{t}^{(x)}=y_{t}-A^{(x)} x_{t}$, if

$$
\begin{equation*}
E\left(e_{t}^{(x)} \mid I_{t}\right)=0 \tag{5.12}
\end{equation*}
$$

holds. However, the case when some columns of $A^{(x)}$ are 0 needs to be considered and additional conditions, such as $\sup _{\mathrm{t}} \mathrm{E}\left\{\mathrm{e}_{\mathrm{t}}^{(\mathrm{x})^{\prime}} \mathrm{e}_{\mathrm{t}}^{(\mathrm{x})}\right\}^{1+\epsilon}<\infty$ for some $\epsilon>0$, need to be imposed to obtain the expected limiting distribution for $t_{N}^{(x)}$. We shall say that $x_{t}$ is a complete regressor (process) for $y_{t}$ if, in addition to (5.12) and

$$
\begin{equation*}
E\left(e_{t}^{(x)} e_{t}^{(x)^{\prime}} \mid I_{t}\right)=\Sigma^{(e)} \tag{5.13}
\end{equation*}
$$

two other conditions hold,

$$
\begin{equation*}
\left\{E \underset{\mathbf{u}=1}{\stackrel{N}{x}} \mathrm{x}_{\mathbf{u}} \mathrm{x}_{\mathbf{u}}^{\prime}\right\}^{-1 / 2} \mathrm{x}_{\mathrm{t}} \xrightarrow{\mathrm{~N}} 0 \tag{5.14}
\end{equation*}
$$

and (A5). It follows then, from a multivariate generalization of Theorem 3 of Lai and Wei (1982), that

$$
\begin{equation*}
\mathrm{t}_{\mathrm{N}}(\mathrm{x}), \mathrm{z}_{\mathrm{N}}(\mathrm{x}) \xrightarrow{\mathrm{N}} \text { dist. } N\left(0, \mathrm{I}_{\mathrm{qr}}\right) \tag{5.15}
\end{equation*}
$$

Hence, for complete regressors,

$$
\begin{equation*}
\operatorname{CVAR}^{(\mathrm{x})}=\mathrm{qr} \tag{5.16}
\end{equation*}
$$

Any two complete regressors are asymptotically equivalent, since $A^{(1)} x_{t}^{(1)}=E\left(y_{t} \mid I_{t}\right)=A^{(2)} x_{t}^{(2)}$. Thus, if we regard CVAR ${ }^{(x)}$ as a cost function, then (5.16) embodies the principle of parsimony ("the fewer coefficients estimated the
better") for complete regressors. The example of subsection 5.1 shows that when asymptotically equivalent but incomplete regressors are considered, the principle of parsimony is no longer valid: the more parsimonious regressor can have greater cost.

For the stationary case, (5.16) follows from (5.4), (5.12) and (5.13): using the formula $E(\cdot)=E\left\{E\left(\cdot \mid I_{t}\right)\right\}$, one sees immediately from (5.12) that the terms in (5.4) with $\mathrm{k} \neq 0$ are 0 , and, from (5.13) one then obtains

$$
\operatorname{CVAR}^{(x)}=\operatorname{tr}\left\{\Sigma^{(x)-1} \operatorname{Ee}_{t}^{(x)} e_{t}^{(x)^{\prime}}\right\} \cdot \operatorname{tr}\left\{\Gamma^{(x)-1} \operatorname{Ex}_{t} x_{t}^{\prime}\right\}=q r
$$

In the next subsection, we shall describe some continuity properties of CVAR ${ }^{(x)}$. These imply that (5.16) holds approximately if $x_{t}$ is "almost complete." We will also give a simple example to show that, although it can be weakened as in Lai and Wei (1982), a condition like (5.13) cannot be completely dispensed with.
5.3. Continuity of CVAR ${ }^{(x)}$ near Complete Regressors for Stationary Regressions. Let us consider $\operatorname{CVAR}_{G}\left({ }^{( }\right)$first. If $x_{t}$ is complete, then, by (5.12) and (5.13), $f^{x e}(\lambda)=f_{-}^{x e}(\lambda) \equiv(2 \pi)^{-1} \Sigma_{k=-\infty}^{-1} \tilde{\Gamma}^{e x}(-k)^{\prime} e^{-i k \lambda}$ and $\left(\Sigma^{(x)}\right)^{-1} f^{e e}(\lambda)=(2 \pi)^{-1} I_{q}$. Clearly, $\operatorname{CVAR}_{G}(\mathrm{x})$ will be close to qr if $\mathrm{f}^{\mathrm{xe}}(\lambda)-\mathrm{f}_{-}^{\mathrm{xe}}(\lambda)$ and $\mathrm{f}^{\mathrm{ee}}(\lambda)-(2 \pi)^{-1} \Sigma^{(x)}$ are close to zero in any of a variety of senses. For example, if the entries of the spectral density matrices in (5.6) are square integrable over $2 \pi$, one can obtain such a result from the fact that the left-hand side of

$$
\int_{0}^{2 \pi}|g(\lambda) \mathrm{h}(\lambda)| \mathrm{d} \lambda \leq\left\{\int_{0}^{2 \pi}|\mathrm{~g}(\lambda)|^{2} \mathrm{~d} \lambda\right\}^{1 / 2}\left\{\int_{0}^{2 \pi}|\mathrm{~h}(\lambda)|^{2} \mathrm{~d} \lambda\right\}^{1 / 2}
$$

will be small if the integral of $|g(\lambda)|^{2}$ is small enough.

We will now indicate how, if fourth moments exist, the Cauchy-Schwarz inequality for expectations can be applied to obtain an analogous result for CVAR ${ }^{(x)}$ via (5.4). For a random $h$-vector $w=\left[w_{1} \cdots w_{h}\right]^{\prime}$, we will use $\|w\|_{2}$ to denote $\max _{1 \leq j \leq h}\left\{E w_{j}^{2}\right\}^{1 / 2}$.

With $\mathrm{M}_{\mathrm{t}}^{(\mathrm{x})}$ as in (5.2) and $\mathrm{c}_{2} \equiv\left\|\mathrm{M}_{\mathrm{t}}^{(\mathrm{x})}\right\|_{2}$, we will first show, following an approach suggested by Madga Peligrad, that, for all N ,

$$
\begin{equation*}
\left|E Z_{N}^{(x)^{\prime}} Z_{N}^{(x)}-E M_{t}^{(x)^{\prime}} M_{t}^{(x)}\right| \leq 2 c_{1} c_{2} \tag{5.17}
\end{equation*}
$$

where $c_{1}$ is a measure of the $t$-dependencies among the entries $m_{t}$ of $M_{t}^{(x)}$, $\mathrm{t}==1,2, \ldots$, which is described below. Observe that if (5.12) holds, then for each entry $m_{t}$, the quantity

$$
\Delta_{N}=N^{-1} E\left\{\left(\Sigma_{t=1}^{N} m_{t}\right)^{2}-\Sigma_{t=1}^{N} m_{t}^{2}\right\}
$$

is zero, as is also the expected value of $S_{n}(p) \equiv \Sigma_{t=n+1}^{n+p} m_{t}$ conditional on $I_{n+1}$. One verifies as in Eberlein (1986) that

$$
\begin{equation*}
\sup _{n, p}| | E\left(S_{n}(p) \mid I_{n+1}\right)| |_{2} \leq c_{1} \tag{5.18}
\end{equation*}
$$

where $c_{1}$ is the maximum over the components of $M_{t}^{(x)}$ of the sum of the mixingale coefficients as defined in McLeish (1975). Since $\Delta_{N}=(2 / N) \Sigma_{t=1}^{N} E\left\{m_{t} S_{t}(N-t)\right\}$, and since $\left|E\left\{m_{t} E\left(S_{t}(N-t) \mid I_{t+1}\right)\right\}\right| \leq c_{1} c_{2}$ by Cauchy-Schwarz, (5.17) follows.

The left hand side of (5.17) will be small if $c_{1}$ is small enough. We will complete our examination of $\operatorname{CVAR}^{(\mathrm{x})}$ - qr by showing that $\delta \equiv \mathrm{EM}_{\mathrm{t}}^{(\mathrm{x})^{\prime}} \mathrm{M}_{\mathrm{t}}^{(\mathrm{x})}-\mathrm{qr}$ is neglible if $\left\|E\left(e_{t}^{(x)^{\prime}} \Sigma^{(x)-1} e_{t}^{(x)} \mid I_{t}\right)-q\right\| \|_{2}$ is small enough. Noting that
$E e_{t}^{(x)^{\prime}} \Sigma^{(x)-1} e_{t}^{(x)}=q$ and $E x_{t}^{\prime} \Gamma^{(x)-1} x_{t}=r$, this assertion follows from the identity $\delta=E\left[x_{t}^{\prime} \Gamma^{(x)-1} x_{t}\left\{E\left(e_{t}^{(x)^{\prime}} \Sigma^{(x)-1} e_{t}^{(x)} \mid I_{t}\right)-q\right\}\right]$ via the Cauchy-Schwarz inequality. This last argument is clearly related to (5.13). We close this section with an elementary stationary example for which (A1)-(A5), (5.3), (5.12) and (5.14) hold, but not (5.13), and for which (5.16) does not hold, because the asymptotic variance matrix of $t_{N}^{(x)}$ and $Z_{N}^{(x)}$ is different from the identity matrix indicated in (5.15). The basic construction is due to Andrew Siegel (personal communication, March 1987). Let F be the distribution on the eight number pairs $\pm(\sqrt{3 / 2}, \sqrt{1 / 2}), \pm(\sqrt{3 / 2}$, $-\sqrt{1 / 2}), \pm(\sqrt{1 / 2}, \sqrt{3 / 2}), \pm(\sqrt{1 / 2},-\sqrt{3 / 2})$, which assigns probability $1 / 8$ to each pair. Let $\left(x_{t}, e_{t}\right), t=0,1, \ldots$ be an i.i.d. sequence with distribution $F$. If $y_{t}=a x_{t}+e_{t}$ for some $a$, we have a regression with $q=r=1$ and $e_{t}^{(x)}=e_{t}$. Also, $E e_{t}^{2}=E x_{t}^{2}=$ 1 , and $E x_{t}=E e_{t}=E e_{t} x_{t}=0$. If $I_{t} \equiv \sigma\left(x_{1}, \ldots, x_{t}, y_{0}, \ldots, y_{t-1}\right)$, then $E\left(e_{t} \mid I_{t}\right)=E e_{t}$ $=0$, but $\mathrm{E}\left(\mathrm{e}_{\mathrm{t}}^{2} \mid \mathrm{x}_{\mathrm{t}}^{2}=3 / 2\right)=1 / 2$, whereas $\mathrm{E}\left(\mathrm{e}_{\mathrm{t}}^{2} \mid \mathrm{x}_{\mathrm{t}}^{2}=1 / 2\right)=3 / 2$, so (5.13) fails. Finally, (5.3) holds, with $V=\operatorname{Ee}_{\mathrm{t}}^{2} \mathrm{x}_{\mathrm{t}}^{2}=3 / 4$, so that $\mathrm{CVAR}^{(\mathrm{x})}=3 / 4$. Thus $\operatorname{CVAR}^{(\mathrm{x})} \neq \mathrm{qr}(=1)$ in contrast to (5.16).

## 6. SOME REGRESSOR SELECTION CRITERIA AND THEIR CONSISTENCY

 PROPERTIES.To obtain a broader perspective on MAIC and the role of CVAR, we now consider additional adjusted log-likelihood ratios,

$$
\begin{equation*}
\mathrm{D}_{\mathrm{N}}^{(1,2)}\left[\mathrm{c}_{\mathrm{N}}^{(1,2)}\right]=-2 \hat{\mathrm{~L}}_{\mathrm{N}}^{(1,2)}+\mathrm{c}_{\mathrm{N}}^{(1,2)} \tag{6.1}
\end{equation*}
$$

and their allied criteria, according to which the regressor process $\mathbf{x}_{\mathbf{t}}^{(1)}$ is preferred if $\mathrm{D}_{\mathrm{N}}(1,2)\left[\mathrm{c}_{\mathrm{N}}^{(1,2)}\right]<0$. For fixed regressors, all such criteria are admissible in the
decision-theoretic sense, see Takada (1982). Some examples of $c_{N}^{(1,2)}$ and the names of their associated criteria are given in (6.2):


The minimum Ideal-AIC criterion is not implementable because the quantities CVAR $^{(i)}, \mathrm{i}=1,2$ are unknown. The following proposition, an immediate consequence of (4.1), shows that all the criteria of (6.2) consistently prefer $x_{t}^{(1)}$ if it provides a better fit asymptotically, in the sense that $\left|\Sigma^{(1)}\right|<\left|\Sigma^{(2)}\right|$.

Proposition 6.1. If $\left|\Sigma^{(1)}\right|<\left|\Sigma^{(2)}\right|$ and if $c_{N}^{(1,2)} / N \xrightarrow{N} 0$, then $\mathrm{P}\left(\mathrm{D}_{\mathrm{N}}^{\left.\left.(1,2)_{\left[\mathrm{c}_{\mathrm{N}}\right.}^{(1,2)}\right]<0\right) \xrightarrow{\mathrm{N}} 1 .}\right.$

For example, if $\mathbf{x}_{\mathbf{t}}^{(1)}$ is a complete regressor process, as defined in section 5 , and $X_{t}^{(2)}$ is not complete, then $A^{(1)} x_{t}^{(1)}-A^{(2)} x_{t}^{(2)} \neq 0$, but $E e_{t}^{(1)}\left(A^{(1)} x_{t}^{(1)}-\right.$ $\left.A^{(2)} x_{t}^{(2)}\right)^{\prime}=0$, by (5.12). Hence (A4) and the decomposition $e_{t}^{(2)}=e_{t}^{(1)}+$ $\left\{A^{(1)} x_{t}^{(1)}-A^{(2)} x_{t}^{(2)}\right\}$ yield $\Sigma^{(1)} \leq \Sigma^{(2)}$ and $\Sigma^{(1)} \neq \Sigma^{(2)}$, from which it follows that $\left|\Sigma^{(1)}\right|<\left|\Sigma^{(2)}\right|$. Proposition 6.1 shows, therefore, in particular, that the criteria defined by (6.1) and (6.2) consistently prefer a complete regressor over an incomplete regressor. The Cox tests discussed briefly in subsection 7.1 are intended to provide a more traditional model selection approach for this situation.

Another immediate result applies to the BIC and Hannan and Quinn criteria and also to the criteria of Rissanen (1986). We will call a criterion of the form (6.1) Strongly parsimonious if $\mathrm{c}_{\mathrm{N}}^{(1,2)} \xrightarrow{\mathrm{N}}-\infty$ whenever $\mathrm{r}^{(1)}<\mathrm{r}^{(2)}$.

Proposition 6.2. If $\mathrm{r}^{(1)}<\mathrm{r}^{(2)}$, and if $\mathrm{D}_{\mathrm{N}}^{(1,2)}\left[\mathrm{c}_{\mathrm{N}}^{(1,2)}\right]$ is strongly parsimonious, then
 probability.

The next result shows that Proposition 6.2 applies to asymptotically equivalent regressors. Its proof is given in Appendix I.

Proposition 6.3. Under assumption (A2), if the regressor processes $x_{t}^{(1)}$ and $x_{t}^{(2)}$ are asymptotically equivalent in the sense that their error processes coincide, $e_{t}^{(1)}=$ $e_{t}^{(2)}$ (w. p. 1), then

$$
\begin{equation*}
(-2) L_{N}^{(1,2)} \sim_{p} Q_{N}^{(2)}-Q_{N}^{(1)} \tag{6.3}
\end{equation*}
$$

Therefore, if (A1) holds for both regressors, as well as (A2), then $\hat{\mathrm{L}}_{\mathrm{N}}^{(1,2)}$ is bounded in probability.

These last two propositions show that when asymptotically equivalent regressors are being compared (and (A1) and (A2) hold), then the minimum BIC and Hannan-Quinn criteria, among others, consistently prefer the regressor with smaller dimension (fewer estimated coefficients). Example 5.1 reveals that this preference can be undesirable. The deep results of $\operatorname{Shibata}(1980,1981)$ also show that the strongly parsimonious criteria can perform poorly relative to MAIC when the regressors are
not complete. Our Example 5.1 is simpler and more accessible than Shibata's results. It too has implications for prediction, see section 9.

The discussion after Proposition 7.3 below shows that, in typical situations involving asymptotically equivalent regressors, each regressor has a non-zero probability of selection by MAIC, so this criterion does not have a consistency property.

A theoretical prototype for $\operatorname{AIC}_{\mathrm{N}}^{(1,2)}$ and Ideal-AIC ${ }_{\mathrm{N}}^{(1,2)}$ which has a more focused consistency property than the strongly parsimonions criteria is investigated in section 8.

Except in Corollary 7.3, we will not establish any further theoretical properties of the strongly parsimonious criteria. These criteria are as easily calculated as AIC's and, for certain applications, might be preferable on the basis of experiments and subject-matter considerations, see Franke et al.(1985). If the model selection need not be done automatically, most users will want to examine several criteria.

## 7. LIMITING DISTRIBUTIONS OF $\hat{L}_{N}^{(1,2)}$ AND AIC ${ }_{\mathrm{N}}^{(1,2)}$ FOR ASYMPTOTICALLY EQUIVALENT REGRESSORS.

We would like to conclude from (6.3) that the asymptotic mean of $(-2) \hat{\mathrm{L}}_{\mathrm{N}}(1,2)$ is $\operatorname{CVAR}^{(2)}-\operatorname{CVAR}^{(1)}$. In section 10 , we will obtain this quantity as the limit of the finite-sample means $E\left\{-2 \hat{\mathrm{~L}}_{\mathrm{N}}^{(1,2)}\right\}$. Here we ignore the finite-sample means and invoke less restrictive assumptions than those of section 10 . Our goal is to obtain the existance of a random variable $Q^{(1,2)}$ such that (7.1) holds,

This property will follow from the additive property of means if we can show that $Q_{N}^{(1)}$ and $Q_{N}^{(2)}$ (or $t_{N}^{(1)}$ and $t_{N}^{(2)}$ ) have a joint asymptotic distribution, which we will do for a rather broad class of regressors in Proposition 7.1.

The mean formula in (7.1) reveals that, with asymptotically equivalent regressors, $\hat{\mathrm{L}}_{\mathrm{N}}^{(1,2)}$ will have an asymptotic tendency to be positive if $\mathrm{x}_{\mathfrak{t}}^{(1)}$ is the less desirable regressor (that is, if $\operatorname{CVAR}^{(1)}>\operatorname{CVAR}^{(2)}$ ). This is the opposite of what happens if $\left|\Sigma^{(1)}\right| \neq\left|\Sigma^{(2)}\right|$, when, according to Proposition 6.1, a positive tendency of $\hat{L}_{N}^{(1,2)}$ means that $\left|\Sigma^{(1)}\right|<\left|\Sigma^{(2)}\right|$, so that $x_{t}^{(1)}$ is the better regressor. It is this - dichotomous behavior for which a log-likelihood-ratio-based regressor selection procedure must compensate.
*We pointed out in section 4 that the $\mathrm{t}_{\mathrm{N}}$-variates are invariant under non-singular transformations of the regressor processes. We will make frequent use of this property now. Throughout this section, the coinciding regression error processes $e_{t}^{(1)}$ and $e_{t}^{(2)}$ will be denoted by $e_{t}$.

The simplest path to (7.1) uses the familiar device of isolating the effect of the factor $\left(\Sigma_{t=1}^{N} x_{t}^{(i)} x_{t}^{(i)}\right)^{-1}$ in ${ }_{N}^{(i)}$ by suitably normalizing $x_{t}^{(i)}$. We assume that there exist nonsingular matrices $C_{N}^{(i)}, N \geq N_{0}$ such that the transformed variates $x_{t, N}^{(i)}=$ $C_{N}^{(i)-1} x_{t}^{(i)}$ satisfy

$$
\begin{equation*}
\sum_{t=1}^{N} x_{t, N}^{(i)} x_{t, N}^{(i)} \longrightarrow_{p} V^{(i)}>0 \tag{7.2}
\end{equation*}
$$

with $\mathrm{V}^{(\mathrm{i})}$ a non-stochatistic, positive definite matrix, for $\mathrm{i}=1,2$, and such that an appropriate joint limiting distribution exists for the fundamental sums,

$$
\begin{equation*}
\operatorname{vec}\left[\sum_{t=1}^{N} e_{t}\left[x_{t, N}^{(1)}{ }^{\prime} x_{t, N}^{(2)^{\prime}}\right]\right] \longrightarrow \text { dist. } U, \text { with } E U^{\prime} U<\infty \tag{7.3}
\end{equation*}
$$

(In the most familiar cases, $\mathrm{C}_{\mathrm{N}}^{(\mathrm{i})}=\mathrm{N}^{1 / 2} \mathrm{I}_{\mathrm{r}}(\mathrm{i})$ ). Using the property $\operatorname{vec}\left[\mathrm{BC}^{\prime}\right]=$ (C®I)vecB, it follows from (7.2) and (7.3) that the joint distribution of $t_{N}^{(1)}$ and $\mathrm{t}_{\mathrm{N}}^{(2)}$ is obtained by left multiplying U by the block diagonal matrix

$$
\left[\begin{array}{cc}
\left(\mathrm{V}^{(1)}\right)^{-1 / 2} \mathrm{I}_{\mathrm{r}}(1) & 0 \\
0 & \left(\mathrm{~V}^{(2)}\right)^{-1 / 2_{\mathrm{r}}(2)}
\end{array}\right]
$$

- Now, from Proposition (6.3), we obtain

Proposition 7.1. Under assumption (A2) of section 4, if (7.2) and (7.3) hold, so does (7.1).

The approach taken above to (7.1) obscures the role of components common to $x_{t}^{(1)}$ and $x_{t}^{(2)}$ : if $A^{(1)} x_{t}^{(1)}=A^{(2)} x_{t}^{(2)}$ is non-zero, the distribution of $U$ in (7.3) will be singular (have a singular variance matrix, etc.). It also does not make clear the form of the asymptotic distribution of $-2 \hat{L}_{\mathrm{N}}^{(1,2)}$; this can have a rather simple form that does not depend on the nature of common components, as we shall demonstrate in Proposition 7.3. To motivate this result, we consider the simplifications which occur when $x_{t}^{(1)}$ and $x_{t}^{(2)}$ are jointly stationary. The following propostion is proved in Appendix $I$.

Proposition 7.2. Suppose that $x_{t}^{(1)}$ and $x_{t}^{(2)}$ are non-nested, iointly covariance stationary regressor processes for $y_{t}$, with mean 0 and nonsingular variance matrices, and with the property that the variance matrix of the joint process $\left[\mathrm{x}_{\mathrm{t}}^{(1)^{\prime}} \mathrm{x}_{\mathrm{t}}^{(2)^{\prime}}\right.$ ] is singular. Suppose also that the sample variance matrices $N^{-1} \Sigma_{t=1} N x_{t}^{(i)} x_{t}^{(j)^{\prime}}$ converge
in probability to $\operatorname{Ex}_{0}{ }^{(\mathrm{i})} \mathrm{x}_{0}(\mathrm{j})^{\prime}, \mathrm{i}, \mathrm{j}=1,2$. Then there exist non-singular matrices $B^{(1)}$ and $B^{(2)}$ such that

$$
\mathrm{B}^{(\mathrm{i})} \mathrm{x}_{\mathrm{t}}^{(\mathrm{i})}=\left[\begin{array}{c}
\mathrm{x}_{\mathrm{t}}^{\mathrm{c}}  \tag{7.4}\\
z_{\mathrm{t}}^{(\mathrm{i})}
\end{array}\right], \mathrm{i}=1,2
$$

and such that the combined process $x_{t} \equiv\left[x_{t}{ }^{c^{\prime}} z_{t}^{(1)^{\prime}}{ }_{z}(2)^{\prime}{ }^{\prime}\right]^{\prime}$ satisfies
with $W^{c}\left(\equiv E x_{t}^{c} x_{t}^{c^{\prime}}\right)$ and $W^{z}$ both positive definite. Consequently, the regression model equations can be put in the form

$$
y_{t}=A x_{t}^{c}+\bar{A}^{(i)} z_{t}^{(i)}+e_{t}^{(i)}, i=1,2
$$

The regressor processes $x_{t}^{(1)}$ and $x_{t}^{(2)}$ are asymptotically equivalent if and only if $\overline{\mathrm{A}}^{(1)} \mathrm{z}_{\mathrm{t}}^{(1)}=\overline{\mathrm{A}}^{(2)_{z_{\mathrm{t}}}(2)}=0$ (w-p.1). In this case, the combined process $\mathrm{x}_{\mathrm{t}}$ is also asymptotically equivalent to $x_{t}^{(1)}$ and $x_{t}^{(2)}$.

Without assuming the regressors are jointly stationary, we will, for the remainder of this section, suppose that matrices $B^{(1)}$ and $B^{(2)}$ exist such that (7.4) holds and such that there is a sequence of invertible, block-diagonal "weighting" matrices,

$$
\mathrm{D}_{\mathrm{N}}=\left[\begin{array}{ccc}
\mathrm{D}_{\mathrm{N}}^{\mathrm{c}} & 0 & 0 \\
0 & \mathrm{D}_{\mathrm{N}}^{(1)} & 0 \\
0 & 0 & \mathrm{D}_{\mathrm{N}}^{(2)}
\end{array}\right],\left(\mathrm{N} \geq \mathrm{N}_{0}\right)
$$

with the property that

$$
D_{N}^{-1} \sum_{t=1}^{N} x_{t} x_{t}^{\prime}\left(D_{N}^{\prime}\right)^{-1} \longrightarrow p\left[\begin{array}{ll}
W^{c} & 0  \tag{7.5}\\
0 & W^{z}
\end{array}\right]
$$

 nonstochastic matrix and $\mathrm{W}^{\mathrm{C}}$ is positive definite w.p.1.

In the stationary case, $\mathrm{D}_{\mathrm{N}}=\mathrm{N}^{1 / 2}$ times the identity matrix of appropriate order. For other types of regressors, including sinusoids and polynomials, see Theorems 10.2.6-7 and pages 581-584 of Anderson (1971) and the discussion of Grenander's conditions in Hannan (1970). For complete, unstable autoregressors of the form $\left(y_{t-1}, \ldots, y_{t-p}\right)^{\prime}$ (no lags missing, dimy $y_{t}=1$ ), see Chan and Wei (1988).

Theorem 10.2.11 of Anderson (1971) and Theorem VII. 10 of Hannan (1970) describe somewhat different conditions under which $t_{N}^{(1)}$ and $t_{N}^{(2)}$ have (nonsingular) Gaussian limiting distributions. In Chan and Wei (1988), the common component $\mathrm{x}_{\mathrm{t}}^{\mathrm{C}}$ is nonstationary (and $W^{\mathbf{C}}$ is random), but their results show that

$$
\begin{equation*}
t_{n}^{c}=\operatorname{vec}\left[\sum_{t=1}^{N} e_{t} x_{t}^{c}\left\{\left(\sum_{t=1}^{N} x_{t}^{c} x_{t}^{c^{\prime}}\right)^{-1}\right\}^{1 / 2}\right] \sim 0_{p}(1) \tag{7.6}
\end{equation*}
$$

and, with $z_{t}=\left[z_{t}^{(1)^{\prime}}{ }_{z_{t}}^{(2)^{\prime}}\right]$, that

For simplicity, we shall assume that $\tilde{W}>0$, so $\operatorname{var}\left(e_{t} z_{t}\right)$ must be non-singular.
These conditions suffice to establish the results we are after. In fact, setting $\tilde{t}_{N}^{(i)} \equiv \Sigma_{t=1}^{N} e_{t} z_{t}^{(i)^{\prime}}\left\{\left(\Sigma_{t=1}^{N} z_{t}^{(i)} z_{t}^{(i)^{\prime}}\right)^{-1}\right\}^{1 / 2}$, it follows from (7.5), and the fact that the $t_{N}$ - statistics are bounded in probability, that $t_{N}^{(i)}{ }^{\prime} t_{N}^{(i)} \sim_{p} t_{N} c^{\prime} t_{N}^{c}+\tilde{t}_{N}^{(i)}{ }^{\prime} \tilde{t}_{N}^{(i)}$. Therefore,

$$
\begin{equation*}
Q_{N}^{(2)}-Q_{N}^{(1)} \sim_{p} \tilde{t}_{N}^{(2)^{\prime}} \tilde{t}_{N}^{(2)}-\tilde{t}_{N}(1)^{\prime} \tilde{t}_{N}^{(1)} \tag{7.8}
\end{equation*}
$$

Now set $d^{(i)} \equiv q\left(r^{(i)}-r^{c}\right)$, where $r^{c} \equiv \operatorname{dim} x_{t}^{c}$. Since $\tilde{W}$ is positive definite, the matrix

$$
\tilde{\mathrm{W}}^{1 / 2}\left[\begin{array}{cc}
\mathrm{I}_{\mathrm{d}}(2) & 0 \\
0 & -\mathrm{I}_{\mathrm{d}}(1)
\end{array}\right]\left[\tilde{\mathrm{W}}^{1 / 2}\right]^{\prime}
$$

will have $\mathrm{d}^{(2)}$ positive eigenvalues $\lambda_{1}^{2}, \ldots, \lambda_{\mathrm{d}}^{2}(2)$ and $\mathrm{d}^{(1)}$ negative eigenvalues $-\mu_{1}^{2}, \ldots,-\mu_{d}^{2}(1)$, see Noble(1969, p. 419). A standard argument applied to the right hand side of (7.8) leads to

Proposition 7.3 If (7.4) - (7.7) hold, then

$$
\begin{equation*}
\left.(-2) \hat{\mathrm{L}}_{\mathrm{N}}^{(1,2)} \xrightarrow{\mathrm{N}} \text { dist. } \sum_{\mathrm{j}=1}^{\mathrm{d}}{ }^{(2)} \lambda_{\mathrm{j}}^{2} \chi_{\mathrm{j}}^{2}(1)-\sum_{\mathrm{k}=1}^{\mathrm{d}} \mu_{\mathrm{k}}^{(1)} \chi_{\{\mathrm{d}}^{2}(2)+\mathrm{k}\right\}(1) \tag{7.9}
\end{equation*}
$$

where $\chi_{j}^{2}(1), j=1, \ldots, \mathrm{~d}^{(1)}+\mathrm{d}^{(2)}$ are independent chi-square variates with one d.f. In particular, if both regressor processes are complete (in the sense of subsection 5.2), then

$$
\begin{equation*}
(-2) \hat{\mathrm{L}}_{\mathrm{N}}^{(1,2)} \xrightarrow[\text { dist. }]{\mathrm{N}}^{\chi^{2}\left(\mathrm{~d}^{(2)}\right)-\chi^{2}\left(\mathrm{~d}^{(1)}\right), ~} \tag{7.10}
\end{equation*}
$$

$\underline{\text { a difference }} \underline{\text { of }} \underline{\text { independent }}$ chi-square variates with d.f.'s $\mathrm{d}^{(2)} \underline{\text { and }} \mathrm{d}^{(1)}$ respectively.

Davies(1980) describes an algorithm suitable for calculating values of the distributions in (7.9) and (7.10). The variance of the distribution in (7.10) is $2\left\{\mathrm{~d}^{(2)}\right.$ $\left.+d^{(1)}\right\}$. This is greater than the variance $2\left|d^{(2)}-d^{(1)}\right|$ of $\chi^{2}\left(\left|d^{(2)}-d^{(1)}\right|\right)$, which is the limiting distribution of $(-2) \hat{\mathrm{L}}_{\mathrm{N}}^{(1,2)}$ for the comparison of complete, nested regressors when the parameter excess of the larger regressor is $\left|d^{(2)}-d^{(1)}\right|$. In this sense, non-nested comparisons are more problematic than nested comparisons. The instability of $(-2) \hat{\mathrm{L}}_{\mathrm{N}}^{(1,2)}$ is further increased when the non-nested regressors are weakly equivalent but not strongly equivalent, see the discussions below (8.7) and (8.12).

Remark When the $\mathrm{D}_{\mathrm{N}}$ in (7.5) are multiples of the identity matrix, say $\mathrm{D}_{\mathrm{N}}=$ $\mathrm{N}^{1 / 2} \mathrm{I}$, as in the case of bounded regressors which do not decrease too rapidly, then the block diagonal form in (7.5) can be achieved starting from the weaker requirement
through replacement of $z_{t}^{(i)}$ by $z_{t}^{(i)}-W^{c, i}\left(W^{c}\right)^{-1} x_{t}^{c}$.
Example 5.1 continued. For this example, the distribution in (7.9) is

$$
\begin{equation*}
5.1 \chi_{1}^{2}(1)+0.03 \chi_{2}^{2}(1)-26.0 \chi_{3}^{2}(1) \tag{7.11}
\end{equation*}
$$

If the variate on the right in (7.9) is denoted by $\delta(\underline{\lambda}, \underline{\mu})$, then the limiting distribution of $\operatorname{AIC}_{\mathrm{N}}^{(1,2)}$ is that of $\delta(\underline{\lambda}, \underline{\mu})+2 \mathrm{q}\left(\mathrm{r}^{(1)}-\mathrm{r}^{(2)}\right)$. If, say, $\mathrm{r}^{(1)}<\mathrm{r}^{(2)}$, it follows that $\lim _{N \rightarrow \infty} \mathrm{P}\left(\operatorname{AIC}_{\mathrm{N}}^{(1,2)}<0\right)$ is always non-zero. The same is true also - of $\lim _{\mathrm{N} \rightarrow \infty} \mathrm{P}\left(\operatorname{AIC}_{\mathrm{N}}^{(1,2)}>0\right)$. That is, each regressor has a positive asymptotic probability of being chosen by MAIC. Some tables related to (7.10) and further discüssion of Example 5.1 are given in subsection 8.3.

If $x_{t}^{(1)} \subseteq x_{t}^{(2)}$ we can arrange (7.2) so that $x_{t}^{c}=x_{t}^{(1)}$. Then $z_{t}^{(1)}$ and all terms related to it should be removed from the formulas and discussion above. In this case, (7.10) reduces to the familiar assertion of a limiting $\chi^{2}\left(q\left(r^{(2)}-r^{(1)}\right)\right)$ distribution for $(-2) \hat{\mathrm{L}}_{\mathrm{N}}(1,2)$.

These results yield corresponding results for $\operatorname{AIC}_{\mathrm{N}}^{(1,2)}$ and Ideal-AIC ${ }_{\mathrm{N}}^{(1,2)}$ by the addition of an appropriate constant. We note the following corollaries.

Corollary 7.2. If the assumptions of Proposition 7.3 hold, then the mean of the asymptotic distribution of $\operatorname{Ideal-AIC}{ }_{\mathrm{N}}{ }^{(1,2)}$ is $\operatorname{CVAR}^{(1)}-\operatorname{CVAR}^{(2)}$.

Corollary 7.3. If the Assumptions of Proposition 7.2 are satisfied, and if the regressors $x_{t}^{(1)}$ and $x_{t}^{(2)}$ are complete and have the same dimension, $r^{(1)}=r^{(2)}$, then all of the statistics $D_{N}^{(1,2)}\left[\mathrm{c}_{\mathrm{N}}^{(1,2)}\right]$ defined in (6.2) coincide with $(-2) \mathrm{L}_{\mathrm{N}}^{(1,2)}$ and have a symmetric limiting distribution. Thus, in this situation these criteria give equal preference to both regressors, asymptotically.
7.1. Use of (7.10) for Hypothesis Testing.

To complete this section, we remark that if the dimension $r^{c}$ of the common component $x_{t}^{c}$ in (7.4) is known or can be estimated reliably, and if it is assumed that at least one of the regressors $x_{t}^{(1)}$ and $x_{t}^{(2)}$ is complete (and that other appropriate assumptions described above hold), then (7.10) can be used to test the null hypothesis that both regressors are complete, and therefore the regressor $x_{t}^{c}$ should be preferred, against the alternative that one regressor, presumably the one with smaller maximized likelihood value (larger $\left|\hat{\Sigma}_{N}^{(i)}\right|$ ), is not complete, see the discussion after Proposition 6.1. This is a generalization of the familiar test of the nested model against the nesting model. Consider the autopilot design problem of section 3, for example. With $\mathrm{x}_{\mathrm{t}}^{(1)} \equiv \tilde{\mathrm{x}}_{\mathrm{t}}^{(\hat{\mathrm{M}})}$ and $\mathrm{x}_{\mathrm{t}}^{(2)} \equiv \mathrm{x}_{\mathrm{t}}^{(\dot{\mathrm{m}})}$, values of $(-2) \hat{\mathrm{L}}_{\mathrm{N}}{ }^{(1,2)}$ can be calculated from Table 3.1 as DAIC $_{894}-2 \Delta \operatorname{dimA}$. Thus, the asymptotic p-value associated with the observed value of $(-2) \hat{L}_{N}^{(1,2)}$ under (7.10) for the regression with $y_{t}=P_{t}$, where $r^{c}=7 \cdot 5=35, d^{(1)}=7$ and $d^{(2)}=5$, is

$$
\begin{equation*}
\mathrm{p}_{\mathrm{P}}=\mathrm{P}\left(\chi^{2}(5)-\chi^{2}(7) \leq-18 .\right)=0.003 \tag{7.12}
\end{equation*}
$$

The asymptotic $p$-value for the regression with $y_{t}=Y_{t}$, where $r^{c}=4 \cdot 5=20, d^{(1)}$ $=4$ and $\mathrm{d}^{(2)}=10$, is

$$
\begin{equation*}
p_{Y}=P\left(\chi^{2}(10)-\chi^{2}(4) \geq 14 .\right)=0.07 \tag{7.13}
\end{equation*}
$$

The value of $\mathrm{p}_{\mathrm{P}}$ would cause the null hypothesis to be rejected in favor of a preference for the regressor including VACC values at the significance levels usually used. The value of $p_{Y}$ would lead to acceptance of the null hypothesis at some popular significance levels. The decision to reject the null hypothesis coincides in
these examples with the decision reached via MAIC. However, in the case of (7.13), acceptance of the null hypothesis, favoring the "intersection" regressor $\mathrm{x}_{\mathrm{t}}^{\mathrm{c}}$ containing just the variables which both regressors share, would also exclude VACC. Thus (7.10) does not provide a way of testing for the inclusion of a specific variable unless this variable only occurs in the regressor process having greater dimension.

There are other modified-log-likelihood-ratio procedures for doing regressor selection via hypothesis-testing with non-nested models deriving from Cox (1961, 1962). These have been extensively developed in the theoretical econometrics literature, but not widely used, it appears. Their adjustments to the log-likelihood ratio are more difficult to calculate than those we have discussed because they requite an approximation to the expected value of the log-likelihood ratio under the null hypothesis. The test statistic requires a consistent estimate of the corresponding asymptotic variance as well. It seems to happen rather frequently in applications of Cox tests that each model of the pair under consideration is rejected in favor of the other. For a survey of the econometrics literature concerned with these procedures see Judge et al.(1984, pp. 883-888) and White (1989), and their references. In some limited simulation experiments by Tsurumi and Wago (1987), MAIC did a more satisfactory job of regressor selection than the Cox-test procedure they investigated.

Hypothesis testing would seem to be an appropriate tool for vindicating one theory over another when the theories specify competing regression models. It is a less natural procedure for trying to decide which of two possibly incomplete regressors has greater predictive power.

## 8. AN IDEAL CRITERION: MAXIMIZING THE KULLBACK-LEIBLER NUMBER

We will now investigate some properties relevant for regressor selection which are possessed by the expected log-likelihood function,

$$
\mathcal{E}_{\mathrm{N}}^{(\mathrm{x})}[\Sigma, \mathrm{A}] \equiv \mathrm{E}\left\{\mathrm{~L}_{\mathrm{N}}^{(\mathrm{x})}[\Sigma, \mathrm{A}]\right\}
$$

a quantity we will call the Kullback-Leibler (K-L) number, or cross-entropy, associated with $A$ and $\Sigma$ and the regressor $x_{t}$. We will assume that (A3) and (A4) of section $\underline{2}$ hold for the regressor processes under consideration. Then, from the decomposition,

$$
e_{t}^{(A)} \equiv y_{t}-A x_{t}=e_{t}^{(x)}+\left(A^{(x)}-A\right) x_{t}
$$

we obtain

$$
E e_{t}^{(A)} e_{t}^{(A)^{\prime}}=\Sigma^{(x)}+\left(A-A^{(x)}\right) E x_{t} x_{t}^{\prime}\left(A-A^{(x)}\right)^{\prime}
$$

which leads via the definition of $L_{N}[\Sigma, A]$ to the basic formula for $\mathcal{E}_{N}^{(x)}[\Sigma, A]$,

$$
\begin{align*}
\mathcal{E}_{\mathrm{N}}^{(\mathrm{x})}[\Sigma, \mathrm{A}] & =-\frac{N}{2}\left(\log 2 \pi|\Sigma|+\operatorname{tr} \Sigma^{-1} \Sigma^{(x)}\right) \\
& -\frac{1}{2} \operatorname{tr} \Sigma^{-1}\left(\mathrm{~A}-\mathrm{A}^{(\mathrm{x})}\right)\left\{\mathrm{E}{\left.\underset{\mathrm{t}=1}{N} \mathrm{x}_{\mathrm{t}} \mathrm{x}_{\mathrm{t}}^{\prime}\right\}}^{\mathrm{N}}\left(\mathrm{~A}-\mathrm{A}^{(\mathrm{x})}\right)^{\prime} .\right. \tag{8.1}
\end{align*}
$$

If $N \geq N_{0}$ as in (2.1), then $E \Sigma_{t=1}^{N} x_{t} x_{t}^{\prime}$ is positive definite and the quadratic expression in $A-A^{(x)}$ in (8.1) takes on its maximum value, 0 , only when $A=$ $A^{(x)}$. An elementary analysis of the eigenvalues (see the proof of Lemma 3.1 of Hosoya and Taniguchi (1982)) shows that the other term in (8.1) is uniquely maximized at $\Sigma=\Sigma^{(x)}$. Thus, $\mathcal{E}_{N}^{(x)}[\Sigma, A]$ is uniquely maximized at $\Sigma=\Sigma^{(\mathrm{x})}, \mathrm{A}=\mathrm{A}^{(\mathrm{x})}$ :

$$
\begin{equation*}
\Sigma^{(x)}, A^{(x)}: \mathcal{E}_{N}^{(x)}[\Sigma, A] \rightarrow \max !\quad\left(N \geq N_{0}\right) \tag{8.2}
\end{equation*}
$$

From (8.1) we obtain that, for all $N=1,2, \ldots$,

$$
\begin{equation*}
N^{-1} \mathcal{E}_{N}^{(x)}\left[\Sigma^{(x)}, A^{(x)}\right]=-\frac{1}{2}\left\{\log 2 \pi\left|\Sigma^{(x)}\right|+q\right\} \tag{8.3}
\end{equation*}
$$

Given regressor processes $x_{t}^{(i)}, i=1,2$, we define $\hat{\mathcal{E}}_{N}^{(i)} \equiv \mathcal{E}_{N}^{(i)}\left[\hat{\Sigma}_{N}^{(i)}, \hat{A}_{N}^{(i)}\right], \mathcal{E}_{N, \infty}^{(i)} \equiv$ $\mathcal{E}_{N}^{(i)}\left[\Sigma^{(i)}, A^{(i)}\right]$, and $\hat{\mathcal{E}}_{N}^{(1,2)} \equiv \hat{\mathcal{E}}_{N}^{(1)}-\hat{\mathcal{E}}_{\mathrm{N}}^{(2)}$. Note, from (8.3) that

$$
\begin{equation*}
\mathcal{E}_{N, \infty}^{(1)}>\mathcal{E}_{N, \infty}^{(2)} \text { if and only if }\left|\Sigma^{(1)}\right|<\left|\Sigma^{(2)}\right| \tag{8.4}
\end{equation*}
$$

From (8.1) we also obtain

$$
\begin{gather*}
\hat{\varepsilon}_{N}^{(i)}=-\frac{N}{2}\left\{\log 2 \pi\left|\hat{\Sigma}_{N}^{(i)}\right|+\operatorname{tr}\left(\hat{\Sigma}_{N}^{(i)}\right)^{-1} \Sigma^{(i)}\right\} \\
-\frac{1}{2} \operatorname{tr}\left[\left(\hat{\Sigma}_{N}^{(i)}\right)^{-1}\left(\hat{A}_{N}^{(i)}-A^{(i)}\right)\left\{E \sum_{t=1}^{N} x_{t}^{(i)} x_{t}^{(i)}\right\}\left(\hat{A}_{N}^{(i)}-A^{(i)}\right)^{\prime}\right] \tag{8.5}
\end{gather*}
$$

The regression analogue of the entropy maximization principle of Akaike (1977) asserts that the regressor $\mathrm{x}_{\mathrm{t}}^{(\mathrm{i})}$ with larger $\hat{\mathcal{E}}_{\mathrm{N}}^{(\mathrm{i})}$ is to be preferred:
(EMP) Prefer $\mathrm{x}_{\mathrm{t}}^{(1)} \underline{\text { if }} \hat{\mathcal{E}}_{\mathrm{N}}^{(1,2)}>0$.

In this section, we shall show that this principle favors the desired regressor both when $\left|\Sigma^{(1)}\right| \neq\left|\Sigma^{(2)}\right|$ and when $x_{t}^{(1)}$ and $x_{t}^{(2)}$ are asymptotically equivalent. In other words, $\hat{\mathcal{E}}_{\mathrm{N}}^{(1,2)}$ does not exhibit the dichotomous behavior of $\hat{\mathrm{L}}_{\mathrm{N}}^{(1,2)}$ discussed in - section 5 . We will give separate analyses for the situations

$$
\begin{equation*}
\left|\Sigma^{(1)}\right|<\left|\Sigma^{(2)}\right| \tag{8.6}
\end{equation*}
$$

and

$$
\begin{equation*}
\left|\Sigma^{(1)}\right|=\left|\Sigma^{(2)}\right| \tag{8.7}
\end{equation*}
$$

The situation $\left|\Sigma^{(2)}\right|<\left|\Sigma^{(1)}\right|$ is covered by interchanging indices, (1,2) $\longrightarrow(2,1)$, in the discussion of (8.6).

In the nested case, $\mathrm{x}_{\mathrm{t}}^{(1)}=\mathrm{Bx}_{\mathrm{t}}^{(2)}$, it follows from the uniqueness of the maximizer of $\mathcal{E}_{\mathrm{N}}^{(2)}[\Sigma, \mathrm{A}]$ that (8.7) is equivalent to $\mathrm{A}^{(2)}=\mathrm{A}^{(1)} \mathrm{B}$, which is equivalent to $e_{t}^{(1)}=e_{t}^{(2)}$. In the non-nested case, however, the condition (8.7), defining what we shall call weak equivalence of regressors, is weaker than $e_{t}^{(1)}=e_{t}^{(2)}$. For instance, let $y_{t}$ be a mean zero stationary process for which non-zero autocorrelations at two different lags coincide, $\rho_{\mathbf{k}}(1)=\rho_{\mathbf{k}}(2)=\rho \quad\left(\mathbf{k}^{(1)} \neq \mathbf{k}^{(2)} ; 0<|\rho|<1\right)$; an example would be a stationary third-order autoregressive process with $\rho_{1}=\rho_{3}=0.2$
and $\rho_{2}=0.5$. Then, for $x_{t}^{(i)} \equiv y_{t-k}(i)$, the error processes $e_{t}^{(i)}=y_{t}-\rho x_{t}^{(i)} \quad(i=$ 1,2 ) are distinct, but $\Sigma^{(1)}=\Sigma^{(2)}=\left(E y_{o}^{2}\right)\left(1-\rho^{2}\right)$.
8.1 A Consistency Result for EMP when $\left|\Sigma^{(1)}\right|<\left|\Sigma^{(2)}\right|$.

If the trace term on the right in (8.5) is bounded in probability, or even just $\mathrm{o}_{\mathrm{p}}(\mathrm{N})$, for $\mathrm{i}=1,2$, then it follows from $\hat{\Sigma}_{\mathrm{N}}^{(\mathrm{i})} \xrightarrow{\mathrm{N}} \mathrm{p}^{(\mathrm{i})}$ and (4.1) that, under (8.6),

$$
\begin{equation*}
\lim _{N \rightarrow \infty} P\left(\hat{\mathcal{E}}_{N}^{(1,2)}>0\right)=1 \tag{8.8}
\end{equation*}
$$

showing that $x_{t}^{(i)}$ is preferred by EMP with asymptotic probability one. For example, if (A1) and (A5) hold, then

$$
\begin{equation*}
R_{N}^{(i)} \equiv \operatorname{tr}\left(\hat{\Sigma}_{N}^{(i)}\right)^{-1}\left(\hat{A}_{N}^{(i)}-A^{(i)}\right)\left\{E \sum_{t=1}^{N} x_{t}^{(i)} x_{t}^{(i)^{\prime}}\right\}\left(\hat{A}_{N}^{(i)}-A^{(i)}\right)^{\prime} \tag{8.9}
\end{equation*}
$$

is easily seen to satisfy $R_{N}^{(i)} \sim_{p} Q_{N}^{(i)}$, so $R_{N}^{(i)}$ is bounded in probability. Hence the same is true of the trace term in (8.5) and (8.8) applies.
8.2 Results for the Cases $\left|\Sigma^{(1)}\right|=\left|\Sigma^{(2)}\right|$ and $e_{t}^{(1)}=e_{t}^{(2)}$. When (8.7) holds, then $\mathcal{E}_{N, \infty}^{(1)}=\mathcal{E}_{N, \infty}^{(2)}$ and we have

$$
\begin{equation*}
\hat{\varepsilon}_{N}^{(1,2)}=\left\{\mathcal{E}_{N, \infty}^{(2)}-\hat{\mathcal{E}}_{\mathrm{N}}^{(2)}\right\}-\left\{\mathcal{E}_{\mathrm{N}, \infty}^{(1)}-\hat{\mathcal{E}}_{\mathrm{N}}^{(1)}\right\} \tag{8.10}
\end{equation*}
$$

a decomposition in which the bracketed terms are nonnegative, by (8.2). In Appendix I, we will demonstrate that, under (A1) - (A5), the likelihoods and K-L numbers deviate from their maximum values in the same way,

$$
\begin{equation*}
\hat{L}_{N}^{(x)}-L_{N, \infty}^{(x)} \sim_{p} \mathcal{E}_{N, \infty}^{(x)}-\hat{\mathcal{E}}_{N}^{(x)} \tag{8.11}
\end{equation*}
$$

a phenomenon first examined by Akaike and Shimizu in the case of overparameterized autoregressions, see Shimizu (1978). When (8.11) holds, then (8.10) yields

$$
\begin{equation*}
\hat{\mathcal{E}}_{\mathrm{N}}^{(1,2)} \sim_{p}\left\{\hat{\mathrm{~L}}_{\mathrm{N}}^{(2)}-\mathrm{L}_{\mathrm{N}, \infty}^{(2)}\right\}-\left\{\hat{\mathrm{L}}_{\mathrm{N}}^{(1)}-\mathrm{L}_{\mathrm{N}, \infty}^{(1)}\right\} . \tag{8.12}
\end{equation*}
$$

It follows from (8.12) and the analysis of the expressions (I.5)-(I.7) in Appendix I that, when Assumptions (A1)-(A5) are satisfied, then $\hat{\mathcal{E}}_{\mathrm{N}}^{(1,2)}$ is bounded in probatility if (and only if) (8.7) holds, a cleaner result than is possible for $\hat{\mathrm{L}}_{\mathrm{N}}(1,2)$. In fact,

$$
\begin{equation*}
\hat{\mathrm{L}}_{\mathrm{N}}^{(1,2)}=\left\{\hat{\mathrm{L}}_{\mathrm{N}}^{(1)}-\mathrm{L}_{\mathrm{N}, \infty}^{(1)}\right\}-\left\{\hat{\mathrm{L}}_{\mathrm{N}}^{(2)}-\mathrm{L}_{\mathrm{N}, \infty}^{(2)}\right\}+\left\{\mathrm{L}_{\mathrm{N}, \infty}^{(1)}-\mathrm{L}_{\mathrm{N}, \infty}^{(2)}\right\} \tag{8.13}
\end{equation*}
$$

and, under (8.7),

$$
L_{N, \infty}^{(1)}-L_{N, \infty}^{(2)}=(1 / 2) \sum_{t=1}^{N}\left\{e_{t}^{(2)^{\prime}} \Sigma^{(2)-1} e_{t}^{(2)}-e_{t}^{(1)^{\prime}} \Sigma^{(1)-1} e_{t}^{(1)}\right\}
$$

which has mean zero but magnitude $0_{p}\left(N^{1 / 2}\right)$ when $N^{-1 / 2}$ times the right hand sum has a nondegenerate limiting distribution, as in the example discussed after (8.7).

Since $L_{N, \infty}^{(1)}=L_{N, \infty}^{(2)}$ when $e_{t}^{(1)}=e_{t}^{(2)}$, we obtain the following fundamental result, from which regressor comparison properties of EMP follow.

Proposition 8.1. If the regressor processes $x_{t}^{(1)}$ and $x_{t}^{(2)}$ for $y_{t}$ are asymptotically equivalent and satisfy assumptions (A1)-(A5) of section 4, then $\hat{\mathrm{E}}_{\mathrm{N}}^{(1,2)}$ and $\hat{\mathrm{L}}_{\mathrm{N}}^{(1,2)}$ behave oppositely for large $N$, in the precise sense of (8.14):

$$
\begin{equation*}
\hat{\mathcal{E}}_{\mathrm{N}}^{(1,2)} \sim_{\mathrm{p}}-\hat{\mathrm{L}}_{\mathrm{N}}^{(1,2)} . \tag{8.14}
\end{equation*}
$$

 then (8.14) implies

$$
\lim _{\mathrm{N} \rightarrow \infty} \mathrm{P}\left(\hat{\mathcal{E}}_{\mathrm{N}}^{(1,2)}>0\right)=1 .
$$

Thus, among nested, asymptotically equivalent regressors satisfying (7.9) and (8.14), EMP consistently prefers the regressor with smallest dimension. This consistency property is more limited in scope than that of the strongly parsimonious criteria, see Proposition 6.2. In light of Example 5.1, this more limited scope may be desirable. What are the large sample properties of $\hat{\mathcal{E}}_{\mathrm{N}}^{(1,2)}$ for this example? The result (8.12) and some results of section 7 combine to provide the basis for an answer to this question.

Proposition 8.2. If (A1)-(A4) and the other assumptions of Proposition 7.1 or Proposition 7.3 are satisfied by the asymptotically equivalent regressors $x_{t}^{(1)}$ and $x_{t}^{(2)}$, then $2 \hat{\varepsilon}_{\mathrm{N}}^{(1,2)}$ has an asymptotic distribution whose mean is $\mathrm{CVAR}^{(2)}-\mathrm{CVAR}^{(1)}$.

Thus, asymptotically, (EMP) has a tendency to favor the regressor with smaller CVAR.

Example $\underline{5} . \underline{1}$ continued. For this example, $\operatorname{CVAR}^{(2)}<\mathrm{CVAR}^{(1)}$. The asymptotic probability that $x_{t}^{(2)}$ is selected by EMP can be obtained from (7.11):

$$
\lim _{N \rightarrow \infty} P\left(\hat{\mathcal{E}}_{\mathrm{N}}^{(1,2)}<0\right)=0.73
$$

Similarly, the asymptotic probabilities that $\operatorname{AIC}_{\mathrm{N}}^{(1,2)}$ and Ideal-AIC $\mathrm{N}_{\mathrm{N}}^{(1,2)}$ lead to the selection of $x_{t}^{(2)}$ are

$$
\lim _{\mathrm{N} \rightarrow \infty} \mathrm{P}\left(\operatorname{AIC}_{\mathrm{N}}^{(1,2)}>0\right)=0.17
$$

and

$$
=\quad \lim _{N \rightarrow \infty} P\left(\operatorname{Ideal}-\operatorname{AIC}_{N}^{(1,2)}>0\right)=0.81
$$

respectively. See (6.2) for the definition of Ideal-AIC ${ }_{N}^{(1,2)}$.

It can be shown that if $x_{t}^{(1)}$ and $x_{t}^{(2)}$ are only weakly equivalent and $\Sigma^{(1)} \neq$ $\Sigma^{(2)}$, then the asymptotic distribution of $2 \hat{\mathcal{E}}_{\mathrm{N}}^{(1,2)}$ (which exists rather generally) has a mean which can involve additional terms related to the variances of the estimates of $\Sigma^{(1)}$ and $\Sigma^{(2)}$, see (10.16) below. Arguments like those of subsection 5.3 can be used to show that the additional terms will be negligible if the regressors are nearly complete, see Findley (1985).
8.3 Ideal- $^{-I I C_{N}}(1,2)$ and $\operatorname{AIC}_{\mathrm{N}}^{(1,2)}$.

The attractive properties of EMP described in subsection 8.2 provide the motivation for our definition in section 6 of

$$
\operatorname{Ideal}-\operatorname{AIC}_{\mathrm{N}}^{(1,2)}=(-2) \hat{\mathrm{L}}_{\mathrm{N}}^{(1,2)}+2\left\{\operatorname{CVAR}^{(1)}-\operatorname{CVAR}^{(2)}\right\}
$$

This was defined in such a way that it has the same asymptotic mean as $\hat{\mathcal{E}}_{\mathrm{N}}^{(1,2)}$, see Proposition 8.2 and Corollary 7.2. Similarly, $\operatorname{AIC}_{\mathrm{N}}^{(1,2)}$ is motivated by the special case in which the regressors are complete in the sense of subsection 5.2. In these contexts, our results show that Ideal-AIC ${ }_{\mathrm{N}}^{(1,2)}$ and $\operatorname{AIC}_{\mathrm{N}}^{(1,2)}$ are asymptotically unbiased estimators of $\hat{\mathcal{E}}_{\mathrm{N}}^{(1,2)}$, the property emphasized by Akaike $(1973,1977)$, who discusses $\operatorname{AIC}_{\mathrm{N}}(\mathrm{i})$ as a bias-corrected estimate of $\hat{\mathcal{E}}_{\mathrm{N}}^{(\mathrm{i})}$. It is clear from (8.14) that these estimators are not consistent.

One of us (D.F.F.) will report elsewhere on simulation experiments concerning the estimation of $\mathrm{CVAR}^{(\mathrm{x})}$ for scalar autoregressions, in order to directly estimate Ideal-AIC ${ }_{\mathbf{N}}^{(1,2)}$. Lacking such estimates, it is properties of $\operatorname{AIC}_{\mathrm{N}}^{(1,2)}$ which are of practical interest.

To get some sense of the asymptotic behavior of MAIC for non-nested models, we will now look at the case of complete regressors, where (7.10) holds. We assume that $d \equiv \mathrm{q}\left(\mathrm{r}^{(2)}-\mathrm{r}^{(1)}\right)>0$. Let $\mathrm{m} \equiv \mathrm{q}\left(\mathrm{r}^{(1)}-\mathrm{r}^{\mathrm{c}}\right)$, where $\mathrm{r}^{\mathrm{c}}$ is the dimension of the shared regressor $x_{t}^{c}$ in (7.4). Then the variate on the right in (7.10) becomes

$$
\begin{equation*}
\delta(\mathrm{m}, \mathrm{~d}) \equiv \chi^{2}(\mathrm{~m}+\mathrm{d})-\chi^{2}(\mathrm{~m}) \tag{8.15}
\end{equation*}
$$

In Table 8.1 below, three sets of $\delta(\mathrm{m}, \mathrm{d})$-probabilities are given for a range of values of $m$ and $d$. These are asymptotic probabilities of selection of the more parsimonious regressor $\mathrm{x}_{\mathrm{t}}^{(1)}$, which has the smaller CVAR value, see (5.16):

$$
\lim _{N \rightarrow \infty} P\left(\operatorname{ArC}_{N}^{(1,2)}<0\right)=P(\delta(m, d)<2 d)
$$

$$
\lim _{\mathrm{N} \rightarrow \infty} \mathrm{P}\left(\hat{\mathcal{E}}_{\mathrm{N}}^{(1,2)}>0\right)=\mathrm{P}(\delta(\mathrm{~m}, \mathrm{~d})>0)
$$

and

$$
\lim _{\mathrm{N} \rightarrow \infty} \mathrm{P}\left(\mathrm{AIC}_{\mathrm{N}}^{(1,2)}<0, \hat{\varepsilon}_{\mathrm{N}}^{(1,2)}>0\right)=\mathrm{P}(0<\delta(\mathrm{m}, \mathrm{~d})<2 \mathrm{~d})
$$

the last being the asymptotic probability that MAIC and EMP agree on the choice of $X_{t}^{(1)}$. For this situation, the low probabilities which arise when $m$ is larger than d are a consequence of (8.14).

Table 8.1 Asymptotic Probabilities of Parsimonious Choice Between Complete, Non-Nested Regressors, by AIC, EMP and Both Simultaneously. d $\equiv \operatorname{dimA}{ }^{(2)}-\operatorname{dimA}{ }^{(1)}$; $m$ is the number of estimated coefficients for variables in $x_{t}^{(1)}$ which are not linear combinations of those in $\mathrm{x}_{\mathrm{t}}^{(2)} ; \delta(\mathrm{m}, \mathrm{d})$ is defined in (8.15).

| $\mathrm{P}(\delta(\mathrm{m}, \mathrm{d})<2 \mathrm{~d})$ |  |  |  |  |  |  |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| m/d | 1 | 2 | 6 | 12 | 18 | $\infty$ |
| 0 | . 84 | . 87 | . 94 | . 98 | . 99 | 1.00 |
| 1 | . 74 | . 81 | . 92 | . 98 | . 99 | 1.00 |
| 2 | . 68 | . 77 | . 90 | . 97 | . 99 | 1.00 |
| 6 | . 59 | . 67 | . 85 | . 95 | . 98 | 1.00 |
| 12 | . 56 | . 62 | . 79 | . 92 | . 97 | 1.00 |
| 18 | . 55 | . 59 | . 75 | . 89 | . 95 | 1.00 |
| $\infty$ | . 50 | . 50 | . 50 | . 50 | . 50 |  |
| $\mathrm{P}(\delta(\mathrm{m}, \mathrm{d})>0)$ |  |  |  |  |  |  |
| $\mathrm{m} / \mathrm{d}$ | 1 | 2 | 6 | 12 | 18 | $\infty$ |
| 0 | 1.00 | 1.00 | 1.00 | 1.00 | 1.00 | 1.00 |
| 1 | . 71 | . 82 | . 97 | 1.00 | 1.00 | 1.00 |
| 2 | . 65 | . 75 | . 94 | . 99 | 1.00 | 1.00 |
| 6 | . 58 | . 65 | . 86 | . 97 | . 99 | 1.00 |
| 12 | . 56 | . 61 | . 79 | . 93 | . 98 | 1.00 |
| 18 | . 55 | . 59 | . 75 | . 90 | . 96 | 1.00 |
| $\infty$ | . 50 | . 50 | . 50 | . 50 | . 50 |  |

$$
\mathrm{P}(0<\delta(\mathrm{m}, \mathrm{~d})<2 \mathrm{~d})
$$

| $\mathrm{m} / \mathrm{d}$ | 1 | 2 | 6 | 12 | 18 | $\infty$ |
| ---: | ---: | ---: | ---: | ---: | ---: | ---: |
| 0 | .84 | .87 | .94 | .98 | .99 | 1.00 |
| 1 | .45 | .63 | .89 | .98 | .99 | 1.00 |
| 2 | .33 | .52 | .84 | .96 | .99 | 1.00 |
| 6 | .17 | .32 | .71 | .92 | .97 | 1.00 |
| 12 | .12 | .23 | .58 | .85 | .95 | 1.00 |
| 18 | .10 | .10 | .50 | .79 | .91 | 1.00 |
| $\infty$ | 0.00 | 0.00 | 0.00 | 0.00 | 0.00 |  |

## 9. A SECOND COST FUNCTION: NORMALIZED MEAN SQUARE PREDICTION ERROR WITH INDEPENDENT REPLICATES.

One would expect that, between asymptotically equivalent regressors, one important consequence of greater coefficient estimation variability would be diminished predictive performance. In this section, we establish a connection, between $\operatorname{CVAR}^{(2)}-$ CVAR $^{(1)}$ and the corresponding difference of a measure of mean square prediction error in two situations: predicting independent replicates of the data used to estimate the regression coefficients; and predicting the observation set used for estimation.

Let $\hat{A}_{N}^{(x)}$ denote the least squares coefficient estimate of $A^{(x)}$ in the model

$$
y_{t}=A^{(x)} x_{t}+e_{t}^{(x)}
$$

from data $y_{t}, x_{t}, t=1, \ldots, N$. We assume that (A1)-(A5) hold. As before, $E e_{t}^{(x)} e_{t}^{(x)}{ }^{\prime}$ is denoted by $\Sigma^{(x)}$. Let $\bar{y}_{t}, \bar{x}_{t}, t=1, \ldots, N$ denote an independent replicate of the data which were used to determine $\hat{A}_{N}^{(x)}$ and let $\bar{E}$ denote the expectation operator for this replicate. Consider the normalized mean square prediction error measure defined by

$$
\begin{equation*}
\operatorname{MSPE}_{N}^{(x)} \equiv \operatorname{EE}\left[\sum_{t=1}^{N}\left(\bar{y}_{t}-\hat{A}_{N}\left(x_{\bar{x}_{t}}\right)^{\prime} \Sigma^{(x)-1}\left(\bar{y}_{t}-\hat{A}_{N}(x)_{\bar{x}_{t}}\right)\right] .\right. \tag{9.1}
\end{equation*}
$$

 we obtain

$$
\operatorname{MSPE}_{N}^{(x)}=N q+\operatorname{trE}\left[\left(\Sigma^{(x)}\right)^{-1}\left(\hat{A}_{N}^{(x)}-A^{(x)}\right)\left[\bar{E}_{\mathrm{E}}^{\mathrm{N}=1} \mathrm{~N} \bar{x}_{t} \bar{x}_{t}\right]\left(\hat{A}_{N}^{(x)}-A^{(x)}\right)^{\prime}\right]
$$

Hence, for $\mathrm{R}_{\mathrm{N}}^{(\mathrm{i})}$ defined in (8.9), we have

$$
=\quad \operatorname{MSPE}_{\mathrm{N}}^{(2)}-\operatorname{MSPE}_{\mathrm{N}}^{(1)}=\mathrm{E}\left\{\mathrm{R}_{\mathrm{N}}^{(2)}-\mathrm{R}_{\mathrm{N}}^{(1)}\right\}
$$

Since $R_{N}^{(i)}{ }_{\sim}{ }_{p} Q_{N}^{(i)}$ under our assumptions, we would expect to have

$$
\begin{gather*}
\lim _{\mathrm{N} \rightarrow \infty}\left\{\operatorname{MSPE}_{\mathrm{N}}^{(2)}-\operatorname{MSPE}_{\mathrm{N}}^{(1)}\right\}=\lim _{\mathrm{N} \rightarrow \infty} \mathrm{E}\left\{\mathrm{Q}_{\mathrm{N}}^{(2)}-\mathrm{Q}_{\mathrm{N}}^{(1)}\right\} \\
=\operatorname{CVAR}^{(2)}-\operatorname{CVAR}^{(1)} . \tag{9.2}
\end{gather*}
$$

It follows from taking expectations in (4.3) that $E\left\{Q_{N}^{(2)}-Q_{N}^{(1)}\right\}$ corresponds to the difference of the normalized mean square prediction error obtained if, instead of the independent replicate in (9.1), the data used to estimate the $\hat{A}_{\mathrm{N}}^{(\mathrm{i})}$ are predicted. The equalities in (9.2) establish a connection between estimation variability and prediction error. In section 10 , we shall describe how (9.2) can be verified for some important classes of models.

The results of Kunitomo and Yamamoto (1985) show that the analogue of (9.2) for the same-realization forecast error quantities $N^{1 / 2}\left(y_{N+1}-\hat{A}^{(x)} x_{N}\right)\left(\mathbf{\Sigma}^{(x)-1}\right)^{1 / 2}$
contains additional terms. The examples in their Table 3 can be rescaled by ${\left(\Sigma^{(x)}\right.}^{(1)}$ as in (9.1) to show that $\operatorname{MSPE}_{N}^{(x)}$ and $\operatorname{CVAR}^{(x)}$ can be smaller for an incorrect regressor than for the correct regressor. Thus these quantities by themselves (when they can be adequately estimated) do not provide completely satisfactory regressor selection criteria. In theory, they can be used to discriminate between weakly equivalent regressors, as defined in the preceding section, a different situation from that of Table 3 of the above reference.
10. CONVERGENCE OF FINITE-SAMPLE MEANS TO THE ASYMPTOTIC MEANS.

To increase our confidence in the relevence to the moderate sample size situation of the asymptotic results given in sections 5,7 and 8 , we would like to know that convergence in distribution or probability leads to convergence of the means, for example,

$$
\begin{equation*}
\lim _{\mathrm{N} \rightarrow \infty} \mathrm{E}\left\{-2 \hat{\mathrm{~L}}_{\mathrm{N}}^{(1,2)}\right\}=\operatorname{CVAR}^{(2)}-\mathrm{CVAR}^{(1)} \tag{10.1}
\end{equation*}
$$

This is the same issue that arose with (9.2). This chapter shows how such results can be obtained, including complete verifications for two important Gaussian cases: non-stochastic regressors; and subregressions of full-rank autoregressive processes. The Gaussian version of the Example 5.1 will be encompassed by our discussion.

Let $Q$ denote a matrix with stochastic entries and let $|\cdot|$ denote a convenient matrix norm, see Noble (1969). For any $\beta \geq 1$, define the $\beta$-mean (or $\mathrm{L}^{\beta}$-) norm of $Q$ by $|Q|_{\beta}=\left\{E|Q|^{\beta}\right\}^{1 / \beta}$. This will be finite if and only if all the entries $Q_{i j}$ of $Q$ satisfy $E\left|Q_{i j}\right|^{\beta}<\infty$. Our basic strategy can be summarized in two lemmas.

Lemma 10.1. (Billingsley (1985, p. 348)) If $Q_{N} \xrightarrow{N}$ dist. $Q$, and also, for some $N_{0}$ and some $\epsilon>0$, the condition, $\sup _{N \geq N_{0}}\left\|Q_{N}\right\|_{1+\epsilon}<\infty$ is satisfied, then $E Q_{N} \xrightarrow{N} E Q$.

Using the matrix norm inequality $\left\|Q_{1} Q_{2} \ldots Q_{m}\right\| \leq\left\|Q_{1}\right\|\left\|Q_{2}\right\| \ldots \| Q_{m}$ and
Hölder's inequality, it is easy to verify

Lemma 10.2. Given $\epsilon>0$ and $\beta_{\mathrm{j}} \geq 1, \mathrm{j}=1, \ldots, \mathrm{~m}$ such that $\beta_{1}^{-1}+\cdots+\beta_{\mathrm{m}}^{1}=1$,

$$
\text { then }\left\|\mathrm{Q}_{1} \mathrm{Q}_{2} \ldots \mathrm{Q}_{\mathrm{m}}\right\|_{1+\epsilon} \leq\left\|\mathrm{Q}_{1}\right\|_{(1+\epsilon) \beta_{1}} \cdots\left\|\mathrm{Q}_{\mathrm{m}}\right\|_{(1+\epsilon) \beta_{\mathrm{m}}}
$$

.We are investigating equivalent regressors, with $e_{t}^{(1)}=e_{t}^{(2)}=e_{t}$. All of the quantities we wish to examine, $\hat{\mathrm{L}}_{\mathrm{N}}(1,2), \mathrm{Q}_{\mathrm{N}}{ }^{(\mathrm{i})}$, etc. are unchanged by the transformation $e_{t} \rightarrow\left(\Sigma^{(i)}\right)^{-1 / 2} e_{t}$, so we will assume for the remainder of our discussion that

$$
\begin{equation*}
\Sigma^{(\mathrm{i})}=\mathrm{I}_{\mathrm{q}} \tag{10.2}
\end{equation*}
$$

Then the eigenvalues $\hat{\lambda}_{j, N}^{(i)}$ of $\hat{\Sigma}_{N}^{(i)}$ converge to 1 in probability. We will only consider the mean of $\hat{\mathrm{L}}_{\mathrm{N}}(1,2)$; the arguments for the other quantities are similar. Using a first degree Taylor expansion of $\log \lambda$ about $\lambda=1$, we show in Appendix I that

$$
\begin{align*}
&\left|(-2) \hat{\mathrm{L}}_{\mathrm{N}}^{(1,2)}\right| \leq\left|\operatorname{Ntr}\left(\hat{\Sigma}_{\mathrm{N}}^{(2)}-\hat{\Sigma}_{\mathrm{N}}^{(1)}\right)\right| \\
&+\quad \sum_{\mathrm{i}=1}^{2} \operatorname{tr}\left[\left\{\left(\hat{\Sigma}_{\mathrm{N}}^{(\mathrm{i})}\right)^{-2}+\mathrm{I}_{\mathrm{q}}\right\}\left\{\mathrm{N}^{1 / 2}\left(\hat{\Sigma}_{\mathrm{N}}^{(\mathrm{i})}-\mathrm{I}_{\mathrm{q}}\right)\right\}^{2}\right] \tag{10.3}
\end{align*}
$$

Set $\Sigma_{N} \equiv N^{-1} \Sigma_{t=1}^{N} e_{t} e_{t}^{\prime} . \quad$ Now,

$$
\begin{equation*}
\left|N \operatorname{tr}\left(\hat{\Sigma}_{N}^{(2)}-\hat{\Sigma}_{\mathrm{N}}^{(1)}\right)\right| \leq \sum_{\mathrm{i}=1}^{2} \mathrm{Ntr}\left(\Sigma_{\mathrm{N}}-\hat{\Sigma}_{\mathrm{N}}^{(\mathrm{i})}\right) \tag{10.4}
\end{equation*}
$$

and

$$
\begin{equation*}
\hat{\Sigma}_{N}^{(i)}-I_{q}=\left(\hat{\Sigma}_{N}^{(i)}-\Sigma_{N}\right)+\left(\Sigma_{N}-I_{q}\right) \tag{10.5}
\end{equation*}
$$

- Using (4.3), we can rewrite $\Sigma_{N}-\hat{\Sigma}_{N}^{(i)}$ as a product of analyzable factors,

$$
\begin{equation*}
\Sigma_{N}-\hat{\Sigma}_{N}^{(i)}=N^{-1}\left\{\sum_{t=1}^{N} e_{t} x_{t}^{(i)} C_{N}^{\prime}-1\right\}\left\{C_{N}^{\prime}\left(\sum_{t=1}^{N} x_{t}^{(i)} x_{t}^{(i)^{\prime}}\right)^{-1} C_{N}\right\}\left\{\sum_{t=1}^{N} e_{t} x_{t}^{(i)} C_{N}^{\prime}\right\}^{-1} \tag{10.6}
\end{equation*}
$$

By substituting (10.4) - (10.6) into the right-hand side of (10.3), one obtains an upper bound for $\left|(-2) \hat{\mathrm{L}}_{\mathrm{N}}^{(1,2)}\right|$ which is a sum of products involving up to eight factors. Since $\left|(-2) L_{N}^{(1,2)}\right|_{1+1 / 8}$ will be less than the sum of the ( $1+1 / 8$ )-norms of each of the products, we can establish

$$
\sup _{N \geq \mathrm{N}_{0}}\left|(-2) \hat{\mathrm{L}}_{\mathrm{N}}^{(1,2)}\right|_{1+1 / 8}<\infty
$$

by verifying the moment conditions (AM1) - (AM4) below and applying Lemma 10.2. Then, if $\mathrm{C}_{\mathrm{N}}^{(1)}$ and $\mathrm{C}_{\mathrm{N}}^{(2)}$ are matrices such that (7.2) and (7.3) hold, or if (7.4) (7.7) hold, and

$$
\mathrm{C}_{\mathrm{N}}^{(\mathrm{i})} \equiv\left[\begin{array}{ll}
\mathrm{D}_{\mathrm{N}}^{\mathrm{c}} & 0 \\
0 & \mathrm{D}_{\mathrm{N}}^{(\mathrm{i})}
\end{array}\right] \mathrm{B}^{(\mathrm{i})}
$$

we can use Propositions 7.1 or 7.3 and Lemma 10.1 to obtain (10.1).
The remaining subsections are devoted to describing the situations in which we have been able to verify the following ninth moment conditions:

$$
\begin{equation*}
\sup _{N \geq N_{0}} \mid \sum_{t=1}^{N} e_{t} x_{t}^{\prime} C_{N}^{\prime}-1 \|_{9}<\infty \tag{AM1}
\end{equation*}
$$

$$
\begin{equation*}
\sup _{N \supseteq N_{0}} \mid N^{1 / 2}\left(\Sigma_{N}-I_{q}\right) \|_{9}<\infty, \tag{AM2}
\end{equation*}
$$

$$
\begin{gather*}
\sup _{N \geq N_{0}}\left\|\left(\hat{\Sigma}_{N}^{(x)}\right)^{-1}\right\|_{9}<\infty,  \tag{AM3}\\
\sup _{N \geq N_{0}}\left\|C_{N}^{\prime}\left(\sum_{t=1}^{N} x_{t} x_{t}^{\prime}\right)^{-1} C_{N}\right\|_{9}<\infty, \tag{AM4}
\end{gather*}
$$

for some sufficiently large $\mathrm{N}_{0}$.
Substantially greater generality can be achieved for (AM1) and (AM2). We start with these.
10.1. Regressions Satisfying (AM1) and (AM2).

The easiest results deal with the case in which $e_{t}$ is independent of $x_{t}$ and $e_{s}$, $\mathrm{s}<\mathrm{t}$, and, in addition, the moment conditions

$$
\begin{equation*}
\sup _{\mathrm{t}}\left|e_{\mathrm{t}}\right|_{\mathrm{k}}<\infty \tag{10.7}
\end{equation*}
$$

and

$$
\begin{equation*}
\sup _{N \geq N_{0}}\left|C_{N}^{-1} \sum_{t=1}^{N} x_{t} x_{t} C_{N}^{\prime}{ }^{-1}\right|_{k / 2}<\infty \tag{10.8}
\end{equation*}
$$

hold for some $k \geq 9$. Indeed, if $\tilde{e}_{t}$ and $\tilde{x}_{t}\left(=\tilde{x}_{t}(N)\right)$ denote entries of $e_{t}$ and $\mathrm{C}_{\mathrm{N}}{ }^{1} \mathrm{x}_{\mathrm{t}}$, respectively, then it follows from Burkholder's inequality, see Hall and Heyde (1980, p. 23), that there is a constant $K_{0}$ such that

$$
\mid \sum_{t=1}^{N} \tilde{e}_{t} \tilde{x}_{t}\left\|_{k} \leq K_{0}\right\|\left(\sum_{t=1}^{N} \tilde{e}_{t}^{2} \tilde{x}_{t}^{2}\right) \|_{k / 2}^{1 / 2}
$$

$$
\begin{aligned}
& \leq K_{0}\left(\sum_{t=1}^{N}\left|\tilde{e}_{t}^{2} \tilde{x}_{t}\right|_{k / 2}\right)^{1 / 2} \text { (triangle inequality) } \\
& \quad \leq K_{0}\left(\sup _{t}\left|\tilde{e}_{t}\right|_{k}\right)\left(\sum_{t=1}^{N}\left|\tilde{x}_{t}\right|_{k / 2}^{2}\right)^{1 / 2}
\end{aligned}
$$

since $E\left|\tilde{e}_{t} \tilde{x}_{t}\right|^{\mathbf{k}}=E\left|\tilde{e}_{t}\right|^{\mathbf{k}} E\left|\tilde{x}_{t}\right|^{\mathbf{k}}$. Thus the boundedness of $\left|\Sigma_{t=1}^{N} e_{t} x_{t}^{\prime} C_{N}^{\prime}\right|_{k}, N=$ $1,2, \ldots$, which implies (AM1), follows from (10.7) and (10.8). The condition (10.8) is satisfied, for example, if $x_{t}$ is non-stochastic and if $C_{N} x_{t} x_{t}^{\prime} C_{N}^{\prime}-1$ is convergent, as we assumed in section 7 , or if $x_{t}$ has stationary $k$-th order moments and $C_{N}=$ $N^{1 / 2} I_{r}$.

A variety of results are available which lead to (AM1) and (AM2) without independence assumptions, using instead either a linear representation assumption, see Lemma 3.3 of Bhansali (1981), or mixing assumptions, see Yokoyama (1980), Theorem 5.1 of Brillinger (1969) and sections 3 and 4 of Chapter 1 of Zhurbenko (1986). Yokoyama's, Bhansali's and Brillinger's results require that $e_{t}$ be stationary
with mean 0 or, for (AM1), that $e_{t}$ and $x_{t}$ be jointly stationary with mean 0 and $\mathrm{Ee}_{\mathrm{t}} \mathrm{x}_{\mathrm{t}}^{\prime}=0$ (our (A4)), and they cover the stationary Gaussian subregressions considered in the next subsection. Zhurbenko's results do not require stationarity and could be used when $x_{t}$ is a bounded, non-stochastic regressor sequence. Brillinger's and Zhurbenko's results establish the boundedness of cumulants. Since a moment of order $k$ can be obtained from sums of products of cumulants of orders $k$ and less, see McCullagh(1987), the boundedness of $k$-th moments follows. The interested reader may consult these references for further details.

For (AM3) and (AM4), we utilize Gaussian assumptions.
$10.2^{*}$ Regressions Satisfying (AM3) and (AM4).
If $\epsilon_{\mathrm{t}}$ is a sequence of independent $M(0, \Sigma)$ random h-vectors with $\Sigma>0$ (positive definite), then $W_{N}=\sum_{t=1}^{N} \epsilon_{t} \epsilon_{t}^{\prime}$ has the Wishart distribution $W_{h}(\Sigma, N)$. If $\lambda_{\min }\left(W_{N}\right)$ denotes the minimum eigenvalue of $W_{N}$, then $\lambda_{\min }^{-1}\left(W_{N}\right)$ is the maximum eigenvalue of $W_{N}^{-1}$, which is a convenient matrix norm for $W_{N}^{-1}$. The following lemma concerning the Wishart distribution is fundamental to our investigation of (AM3) and (AM4). It appears to be new. We will use $\sim$ for "is distributed as."

Lemma 10.3 If $W_{N} \sim W_{h}(\Sigma, N), N=1,2 \ldots$, then for every $k \geq 1$,

$$
\begin{equation*}
\sup _{N \geq h+2 k} N^{k_{E}\left\{\lambda_{\min }^{-k}\left(W_{N}\right)\right\}<\infty . . . ~ . ~} \tag{10.9}
\end{equation*}
$$

The proof of this lemma, and of the Propositions 10.1 and 10.2 below, are given in Appendix II.

Remark. Complete, Non-Stochastic Regressors with Gaussian Errors. We use the notation N.I.D. $(0, \Sigma)$ to indicate an i.i.d. $M(0, \Sigma)$ process. Observe that if the regressors $x_{t}$ are non-stochastic and the errors $e_{t}$ are N.I.D. $\left(0, \Sigma^{(x)}\right)$, then $N \hat{\Sigma}_{N}(x){ }_{\sim}$ $\mathrm{W}_{\mathrm{r}}\left(\mathrm{\Sigma}^{(\mathrm{x})}, \mathrm{N}-\mathrm{qr}\right.$ ), and (AM3) follows from this lemma. The condition (AM4) holds if the sequence $C_{N}^{-1}{ }_{t=1}^{N} x_{t} x_{t}^{\prime} C_{N}^{\prime}{ }^{-1}$ has a nonsingular limit, as in the examples in Hannan(1970) and Anderson(1971) referred to after (7.5). Then (10.8) also holds, as well as (10.7), and (10.1) follows.

For the case of stochastic regressors, we will obtain our most general verification of (AM3) in the Corollary of the following result.

Proposition 10.1. Suppose that $\tilde{x}_{\mathrm{t}} \underline{\text { is }} \underline{\mathrm{a}} \tilde{\mathrm{r}}$-dimensional, not necessarily stationary, autoregressive process of order $p$, which satisfies

$$
\begin{equation*}
\tilde{x}_{t}=A_{1} \tilde{x}_{t-1}+\cdots+A_{p} \tilde{x}_{t-p}+a_{t} \tag{10.10}
\end{equation*}
$$

where $a_{t} \sim$ N.I.D. $(0, \Sigma)(\Sigma>0)$ is independent of $\tilde{x}_{s},-p+1 \leq s<t$. Let the initializing values $\tilde{x}_{-p+1}, \ldots, \tilde{x}_{0}$ have $\underline{a}$ joint density function, $f\left(\tilde{x}_{-p+1}, \ldots, \tilde{x}_{0}\right)$. Then, if $\hat{A}_{1}, \ldots, \hat{\mathrm{~A}}_{p}$ denote the least squares estimates of $A_{1}, \ldots, A_{p}$ from the data $\tilde{x}_{-p+1}, \ldots, \tilde{x}_{N}$, the error variance matrix estimate defined by

$$
\begin{equation*}
\left.\hat{\Sigma}_{N}=N^{-1}{\underset{\Sigma}{\mathrm{\Sigma}=1}}_{\mathrm{N}}^{\tilde{x}_{t}}-\hat{A}_{1} \tilde{x}_{t-1}-\cdots-\hat{A}_{p} \tilde{x}_{t-p}\right)\left(\tilde{x}_{t}-\hat{A}_{1} \tilde{x}_{t-1}-\cdots-\hat{A}_{p} \tilde{x}_{t-p}\right)^{\prime} \tag{10.11}
\end{equation*}
$$

has the property that for every $k=1,2, \ldots$, the $k$-th moments of $\hat{\Sigma}_{N}^{-1}$ are ultimately bounded: that is, there is an $N(k)$ such that

$$
\begin{equation*}
\sup _{N \geq \mathrm{N}(\mathbf{k})} \mathrm{E}\left\{\lambda_{\min }^{-\mathbf{k}}\left(\hat{\mathcal{L}}_{\mathrm{N}}\right)\right\}<\infty . \tag{10.12}
\end{equation*}
$$

Now we introduce the concept of a subregression and show how (10.12) can be applied to subregressions of (10.10). Suppose that $\tilde{\mathrm{x}}_{\mathrm{t}}$ in (10.10) has the form $\tilde{\mathrm{x}}_{\mathrm{t}}=$ [ $\left.y_{t}^{\prime} v_{t}^{\prime}\right]^{\prime}$. Then the residual variance matrix $\bar{\Sigma}_{\mathrm{N}}$ of the regression of $\mathrm{y}_{\mathrm{t}}$ on $\bar{x}_{\mathrm{t}}=$ $\left[\tilde{x}_{t-1}^{\prime}, \ldots, \tilde{x}_{t-p}^{\prime}\right]^{\prime}$ is a submatrix of $\hat{\Sigma}_{N}$, and therefore $\lambda_{\text {min }}\left(\hat{\Sigma}_{N}\right) \leq \lambda_{\text {min }}\left(\Sigma_{N}\right)$. If the regressor $x_{t}$ of interest for $y_{t}$ is a subvector of $\bar{x}_{t}$, we will further have $\Sigma_{N} \leq \hat{\Sigma}_{N}(x)$, which leads to $\lambda_{\text {min }}\left(\bar{\Sigma}_{\mathrm{N}}\right) \leq \lambda_{\text {min }}\left(\hat{\Sigma}_{\mathrm{N}}^{(\mathrm{x})}\right)$ and, therefore, finally to $\lambda_{\text {min }}^{-\mathrm{k}}\left(\hat{\Sigma}_{\mathrm{N}}^{(\mathrm{x})}\right) \leq$ $\lambda_{\min }^{-\mathbf{k}}\left(\hat{\Sigma}_{\mathrm{N}}\right)$ for $\mathbf{k}=1,2, \ldots$. We will summarize the regressor situation just described by saying that the regression of $y_{t}$ on $x_{t}$ is a subregression of (10.10), or alternatively, is a subautoregression. For example, the regressors $x_{t}^{(1)}$ and $x_{t}^{(2)}$ of Example 5.1 are subregressions of the correct $\operatorname{AR}(6)$ autoregression for $y_{t}$. In general, dim $\tilde{x}_{t}$ and $p$ could be unknown and quite large relative to $\operatorname{dimy}_{t}$ and $\operatorname{dimx}_{t}$. Then the regression seeks to approximate the dynamics of a small subsystem $y_{t}$ of the complex process $\tilde{x}_{t}$.

The following result is apparent.

Corollary 10.1. If the regression of $y_{t}$ on $x_{t}$ is a subregression of (10.10), then (AM3) is satisfied.

Our main result establishing (AM4) generalizes a result of Fuller and Hasza (1981) which concerned the more restricted situation of (stationary) univariate autoregressions. We now suppose that $\overline{\bar{x}}_{\mathrm{t}}$ is an $\overline{\mathrm{r}}$-vector process satisfying

$$
\begin{equation*}
\bar{x}_{t}=A \bar{x}_{t-1}+\bar{e}_{t} \tag{10.13}
\end{equation*}
$$

with $\bar{e}_{\mathrm{t}} \sim$ N.I.D. $(0, \bar{\Sigma})$, where, although $\overline{\mathrm{\Sigma}}$ may be singular,

$$
\begin{equation*}
\sum_{\mathrm{j}=0}^{\overline{\mathrm{r}}} A^{\mathrm{j}}{ }_{\Sigma}\left(A^{j}\right)^{\prime} \text { is nonsingular. } \tag{10.14}
\end{equation*}
$$

Full-rank autoregressions of order greater than one, such as (10.10), can be rewritten in the form (10.13) in such a way that (10.14) is satisfied. In Appendix II, we will prove the following result.

Proposition 10.2. If $\bar{x}_{\mathrm{t}}$ is $\underline{\text { a }}$ process satisfying the conditions above, then for every $k=1,2, \ldots$, there is an $N(k)$ such that

From an argument used to establish Corollary 10.1, we obtain our result for (AM4).

Corollary 10.2. If $x_{t}$ is an r-dimensional subvector of a process satisfying (10.13) (10.14), then (AM4) holds with $C_{N}=N^{1 / 2} I_{r}$.

In summary, since the matrices $\mathrm{C}_{\mathrm{N}}$ must be the same in (AM1) and (AM4), the only stationary stochastic regressors for which we have completely verified (AM1) - (AM4) and (10.1) are the Gaussian subautoregressions. For non-stochastic regressors, see the Remark above.

Results analogous to (10.6) hold for $\mathrm{K}-\mathrm{L}$ number differences and for $2\left\{\mathcal{E}_{\mathrm{N}, \infty}^{(\mathrm{x})}\right.$ $\left.\hat{\mathcal{E}}_{\mathrm{N}}^{(\mathrm{x})}\right\}$ and $2\left\{\hat{\mathrm{~L}}_{\mathrm{N}}^{(\mathrm{x})}-\mathrm{L}_{\mathrm{N}, \infty}^{(\mathrm{x})}\right\}$ under the same assumptions The latter variates have the same asymptotic mean, $\operatorname{CVAR}^{(x)}+\operatorname{\Sigma VAR}^{(x)}$, where $\Sigma V A R^{(x)}$ is the trace of
the variance of the asymptotic distribution of $(N / 2)^{1 / 2}\left(\operatorname{vec} \tilde{\Sigma}_{N}^{(x)}-I_{q}\right)$, with $\tilde{\Sigma}_{N}^{(x)}=$ $\Sigma^{(\mathrm{x})-1 / 2} \Sigma_{\mathrm{N}}(\mathrm{x})\left(\Sigma^{(\mathrm{x})-1 / 2)^{\prime}}\right.$. This follows from applying the Taylor expansion (I.2) to the bracketed term on the right in (I.6) of Appendix I. As a consequence, for weakly equivalent regressors $\left(\left|\Sigma^{(1)}\right|=\left|\Sigma^{(2)}\right|\right)$, one obtains from (8.13) and from $\mathrm{E}\left\{\mathrm{L}_{\mathrm{N}, \infty}^{(2)}-\mathrm{L}_{\mathrm{N}, \infty}^{(2)}\right\}=0$ that

$$
\begin{align*}
& \lim _{N \longrightarrow \infty} E\left\{-2 \hat{\mathrm{~L}}_{\mathrm{N}}^{(1,2)}\right\}=\lim _{\mathrm{N} \longrightarrow \infty} \mathrm{E}\left\{2 \hat{\mathcal{E}}_{\mathrm{N}}^{(1,2)}\right\}= \\
& \left\{\mathrm{CVAR}^{(2)}+\Sigma \mathrm{VAR}^{(2)}\right\}-\left\{\mathrm{CVAR}^{(1)}+\Sigma \operatorname{VAR}^{(1)}\right\} \tag{10.16}
\end{align*}
$$

*Finally, we note that the conditions (AM3) and (AM4) are easily verified for some special non-Gaussian situations, such as the example at the end of section 5 , where the processes $e_{t}$ and $x_{t}$ are bounded away from zero.

## 11. GENERALIZATIONS

In this section, we will briefly describe some elements of a natural conceptual framework for generalizations of the main results of the previous sections to model comparison problems different from regressor selection. For additional details, see Findley(1985) for time series models and Findley(1989) for models for independent observations. For an interesting application, see Ogata (1988). Suppose $\mathrm{L}_{\mathrm{N}}[\theta]$ denotes a $\log$-likelihood function for N observations with parameter vector $\theta$, having the property that $\mathrm{N}^{-1} \mathrm{~L}_{\mathrm{N}}[\theta]$ and also its first and second partial $\theta$-derivatives converge in probability as $\mathrm{N} \rightarrow \infty$ uniformly on compact subsets of the convex parameter space. The limit function $\mathcal{E}[\theta] \equiv \lim _{N \rightarrow \infty} E\left\{N^{-1} L_{N}[\theta]\right\}$ is a type of Kullback-Leibler number for the model defined by $\theta . \quad \mathcal{E}[\theta]$ is assumed to have a
unique maximum in the interior of the parameter space at a point $\theta_{\infty}$ where the matrix of second partial derivatives, $\mathcal{E}^{\prime \prime}[\theta]$, is non-singular. Then, under rather general circumstances, see Pollard(1985) or White(1989), maximum likelihood estimates $\hat{\theta}_{\mathrm{N}}$ satisfying $\partial \mathrm{L}_{\mathrm{N}}\left[\hat{\theta}_{\mathrm{N}}\right] / \partial \theta=0$ will converge to $\theta_{\infty}$ in such a way that $\mathrm{N}^{1 / 2}\left(\hat{\theta}_{\mathrm{N}}-\theta_{\infty}\right)$ has a Gaussian limiting distribution. In this situation, the Taylor expansions

$$
2\left\{\mathrm{~L}_{\mathrm{N}}\left[\theta_{\infty}\right]-\mathrm{L}_{\mathrm{N}}\left[\hat{\theta}_{\mathrm{N}}\right]\right\}=\left(\hat{\theta}_{\mathrm{N}}-\theta_{\infty}\right)^{\prime} \mathrm{L}_{\mathrm{N}}\left[\bar{\theta}_{\mathrm{N}}\right]\left(\hat{\theta}_{\mathrm{N}}-\theta_{\infty}\right)
$$

and

$$
2\left\{\mathcal{E}_{\mathrm{N}}\left[\hat{\theta}_{\mathrm{N}}\right]-\mathcal{E}_{\mathrm{N}}\left[\theta_{\infty}\right]\right\}=\left(\hat{\theta}_{\mathrm{N}}-\theta_{\infty}\right)^{\prime} \mathcal{E}_{\mathrm{N}}^{\prime \prime}\left[\tilde{\theta}_{\mathrm{N}}\right]\left(\hat{\theta}_{\mathrm{N}}-\theta_{\infty}\right)
$$

with $\bar{\theta}_{\mathrm{N}}$ and $\bar{\theta}_{\mathrm{N}}$ on the line segment between $\dot{\theta}_{\mathrm{N}}$ and $\theta_{\infty}$, motivate a generalization of the $Q_{N}$-statistic (4.4), namely

$$
\mathrm{Q}_{\mathbf{N}} \equiv\left(\hat{\theta}_{\mathbf{N}}-\theta_{\infty}\right)^{\prime} \mathrm{L}_{\mathrm{N}}^{\prime \prime}\left[\theta_{\infty}\right]\left(\hat{\theta}_{\mathrm{N}}-\theta_{\infty}\right)
$$

and lead to a generalization of the Akaike-Shimizu relation (8.11),

$$
\mathrm{L}_{\mathrm{N}}\left[\hat{\theta}_{\mathrm{N}}\right]-\mathrm{L}_{\mathrm{N}}\left[\theta_{\infty}\right] \sim_{\mathrm{p}} \mathrm{~N}\left\{\mathcal{E}\left[\theta_{\infty}\right]-\mathcal{E}\left[\hat{\theta}_{\mathrm{N}}\right]\right\}
$$

Two competing families of $\log$-likelihoods $L_{N}{ }_{N}^{(i)}\left[\theta^{(i)}\right]$ with these properties, with m.l.e.'s $\hat{\theta}_{\mathrm{N}}^{(\mathrm{i})} \rightarrow \mathrm{p} \theta_{\infty}^{(\mathrm{i})}, \mathrm{i}=1,2$, are said to be asymptotically equivalent if $\mathrm{L}_{\mathrm{N}}^{(1)}\left[\theta_{\infty}^{(1)}\right]$ $=\mathrm{L}{ }_{\mathrm{N}}{ }^{(2)}\left[\theta_{\infty}^{(2)}\right]\left(\mathrm{w}\right.$. p. 1) for $\mathrm{N} \geq \mathrm{N}_{0} . \quad$ Distributional results like those of section 7 can be obtained if twice differentiable, nonsingular parameter transformations $g^{(i)}$
exist such that $g^{(i)}\left(\theta^{(i)}\right)=\left[\eta^{\mathrm{c}^{\prime}} \eta^{(\mathrm{i})^{\prime}}\right]^{\prime}, \mathrm{i}=1,2$, and if the model defined by the log-likelihood $\mathrm{L}_{\mathrm{N}}[\eta]$ with $\eta=\left[\eta^{\mathrm{c}^{\prime}} \eta^{(1)^{\prime}} \eta^{(2)^{\prime}}\right]^{\prime}$ is asymptotically equivalent to those defined by the $L_{N}(i)\left[\theta^{(i)}\right], i=1,2$. To establish results like

$$
\begin{equation*}
\lim _{\mathrm{N} \rightarrow \infty} 2 \mathrm{E}\left\{\mathrm{~L}_{\mathrm{N}}^{(\mathrm{i})}\left[\hat{\theta}_{\mathrm{N}}^{(\mathrm{i})}\right]-\mathrm{L}_{\mathrm{N}}^{(\mathrm{i})}\left[\theta_{\infty}^{(\mathrm{i})}\right]\right\}=\operatorname{CVAR}^{(\mathrm{i})} \tag{11.1}
\end{equation*}
$$

(as before, $\operatorname{CVAR}^{(i)}=E Q^{(i)}$, where $Q_{N}^{(i)} \xrightarrow{N}{ }_{\text {dist. }} Q^{(i)}$ ) following the strategy of section 10 , it is necessary to have explicit formulas for the m.l.e.'s $\hat{\theta}_{\mathrm{N}}^{(\mathrm{i})}$. Also, conditional expectations must sometimes be used. In Findley (1989), the formula (11.1) is established for some models related to the multinomial distribution (histögrams, contingency tables). For this analysis, the expectation operator E , when applied to $\mathrm{L}_{\mathrm{N}}[\theta]$, was taken to be the conditional expectation conditioned on cells with non-zero probability having at least one observation, in order to have $\mathcal{E}_{\mathrm{N}}\left[\hat{\theta}_{\mathrm{N}}\right]>$ $-\infty$. We mention this to illustrate that there are a variety of ways, depending on the models under consideration, of filling in the theoretical structure outlined in this section. For the case of density models estimated from i.i.d. data, a formula for CVAR ${ }^{(i)}$ can be obtained from Takeuchi (1976), see also Härdle (1987) and Findley (1989).

Shibata (1981) considers the case of fixed regressors, with $\operatorname{dimy}_{t}=1$ and with $y_{t}-E y_{t}$ being i.i.d. and Gaussian. He takes a very interesting and different approach from ours. He assumes that the correct regressor $x_{t}$ has infinite dimension, but the not necessarily nested regressors $x_{t}^{(i)}$ under consideration are finite dimensional subvectors of $x_{t}$ whose dimension increases with $N$ (the range of $i$ can increase also). For a modified version of the mean square prediction error criterion of section 9 , he shows that MAIC is optimally efficient and that the strongly parsimonious criteria we discussed in section 6 are not. Härdle (1987) has considered
the extension to the case of i.i.d. non-Gaussian $y_{t}-E y_{t}$, allowing the error density to be misspecified. The extension of Shibata's and Härdle's results to the case of stochastic regressors appears to be very difficult, see Shibata (1980), where increasing-order autoregressions are considered.

Finally, it should be mentioned that various cross-validation procedures for model selection are asymptotically equivalent to MAIC or simple variants thereof, see Stoica et al. (1986) and the references given there.

## 12. CONCLUDING REMARKS

- Our goal in this paper has been to provide a coherent theory supporting the use of the ordinary $\log$-likelihood ratio for making non-nested regressor comparisons. We were motivated to do this by the importance for applications of the non-nested comparison problem and by a desire to understand the substantial industrial successes of Akaike's MAIC procedure (which, for linear regressions, is asymptotically equivalent to the minimum FPEC criterion of Akaike(1971)). Some of these successes are described in Akaike and Nakagawa(1988), Nakamura et al.(1986), Otomo et al.(1972) and Ohtsu et al.(1979). (There are many industrial applications which are not publicly documented for company confidentiality reasons: Mr. K. Toki of System Sougou Kaihatsu in Tokyo kindly told one of the authors in 1987, in response to a query, that his company has implemented more than sixty statistical model-based controllers using the regressor selection procedures described here and in these references.) Akaike developed AIC as an asymptotically unbiased approximation to the Kullback-Leibler number, see This Week's Citation Classic (1981) and Akaike (1985). The results of sections 8 and 10 reveal attractive properties of $K-L$ numbers for model comparison and clarify the nature of the connection with MAIC .

We hope that the results presented here will stimulate further research on non-nested model comparisons and the role of the likelihood ratio therein. It would be attractive to have generalizations of our results, or a reasonably comprehensive alternative theory, for the situation in which the number of variables in each regressor, $\operatorname{dim} x_{t}^{(i)}$, is permitted to increase as the sample size increases: there are circumstances in which the number of estimated variables must increase if the sequence of log-likelihood ratios, $\hat{\mathrm{L}}_{\mathrm{N}}^{(1,2)}, \mathrm{N} \geq \mathrm{N}_{0}$, is to be bounded in probability. (This is the relevent situation because, in practice, statisticians are only concerned about the interpretation of small-to-moderate values of $\hat{\mathrm{L}}_{\mathrm{N}}^{(1,2)}$ ). Such results should shed light on finite-sample properties.
*The authors wish to gratefully acknowledge the excellent computing support they received from E. Arahata and M. Pugh for the calculations presented in this paper. Some of the results presented here were obtained by the first-named author while he was a Visiting Professor at the Institute of Statistical Mathematics in Tokyo. He wishes to express his gratitude for the support and hospitality he received during this visit, especially from professor G. Kitagawa, who provided the data analyzed in section 3, along with insightful comments and advice.

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## APPENDIX I: PROOFS FOR SECTIONS 6-10.1.

.We begin with the arguments leading to Propositions 6.3 and the results (10.3) and (8.11), which concern $\hat{\mathrm{L}}_{\mathrm{N}}^{(1,2)}, \hat{\mathrm{L}}_{\mathrm{N}}^{(\mathrm{x})}-\mathrm{L}_{\mathrm{N}, \infty}^{(\mathrm{x})}$ and $\mathcal{E}_{\mathrm{N}, \infty}^{(\mathrm{x})}-\hat{\mathcal{E}}_{\mathrm{N}}^{(\mathrm{x})}$. These quantities are unchanged by the transformation $e_{t}^{(x)} \rightarrow\left(\Sigma^{(x)}\right)^{-1 / 2} e_{t}^{(x)}$, so we shall assume that $\Sigma^{(x)}=I_{q}\left(=\Sigma^{(1)}, \Sigma^{(2)}\right.$ etc.). Then the assumption (A2) implies that the eigenvalues $\hat{\lambda}_{N, 1}^{(x)}, \ldots, \hat{\lambda}_{N, q}^{(x)}$ of $\hat{\Sigma}_{N}^{(x)}$ converge in probability to 1 at the rate $\mathrm{N}^{-1 / 2}$,

$$
\begin{equation*}
N^{1 / 2}\left(\hat{\lambda}_{N, j}^{(x)}-1\right) \sim O_{p}(1),(1 \leq j \leq q) \tag{I.1}
\end{equation*}
$$

In general, if $\Sigma$ is a positive definite matrix of order $q$ with eigenvalues $\lambda_{1}, \ldots, \lambda_{q}$, then it follows from Taylor's formula that there exists an $\alpha=\alpha\left(\lambda_{1}, \ldots, \lambda_{q}\right)$ between 0 and 1 such that, with $\tilde{\lambda}_{j}=1+\alpha\left(\lambda_{j}-1\right), 1 \leq j \leq q$, we have

$$
\begin{equation*}
\sum_{j=1}^{q} \log \lambda_{j}=\sum_{j=1}^{q}\left(\lambda_{j}-1\right)-\frac{1}{2} \sum_{j=1}^{q} \tilde{\lambda}_{j}^{-2}\left(\lambda_{j}-1\right)^{2} \tag{I.2}
\end{equation*}
$$

From (I.1) and (I.2), we obtain

$$
\begin{equation*}
N \log \left|\hat{\Sigma}_{\mathbf{N}}^{(x)}\right| \sim_{p} N \operatorname{tr}\left(\hat{\Sigma}_{N}^{(x)}-I_{q}\right)-(N / 2) \operatorname{tr}\left(\hat{\Sigma}_{N}(x)-I_{q}\right)^{2} \tag{I.3}
\end{equation*}
$$

Proof of Proposition 6.3: Set $e_{t} \equiv e_{t}^{(1)}=e_{t}^{(2)}$ and $\Sigma_{N} \equiv N^{-1} \Sigma_{t=1}^{N} e_{t} e_{t}^{\prime}$. Observe via the identity (4.3) that $\operatorname{Ntr}\left(\Sigma_{N}-\hat{\Sigma}_{N}^{(i)}\right)=Q_{N}^{(i)}$, so that

$$
\begin{equation*}
\operatorname{Ntr}\left(\hat{\Sigma}_{\mathbf{N}}^{(2)}-\hat{\Sigma}_{\mathrm{N}}^{(1)}\right)=\mathrm{Q}_{\mathrm{N}}^{(1)}-\mathrm{Q}_{\mathrm{N}}^{(2)} \tag{I.4}
\end{equation*}
$$

- It follows from assumption (A1) that $\operatorname{Ntr}\left(\hat{\mathbf{\Sigma}}_{\mathrm{N}}^{(2)}-\hat{\mathrm{\Sigma}}_{\mathrm{N}}^{(1)}\right)$ is bounded in probability. Consequently,

$$
\begin{gathered}
N \operatorname{tr}\left(\hat{\Sigma}_{N}^{(1)}-I_{q}\right)^{2}-N \operatorname{tr}\left(\hat{\Sigma}_{N}^{(2)}-I_{q}\right)^{2}= \\
\operatorname{Ntr}\left[\left(\hat{\Sigma}_{N}^{(1)}-\hat{\Sigma}_{N}^{(2)}\right)\left\{\left(\hat{\Sigma}_{N}^{(1)}-I_{q}\right)+\left(\hat{\Sigma}_{N}^{(2)}-I_{q}\right)\right\}\right]
\end{gathered}
$$

converges to zero in probability, and Proposition 6.3 follows immediately from (I.3) and (I.4).

Proof of (10.3). If $0 \leq \alpha \leq 1$ and $\lambda \geq 0$, then $(1+\alpha(\lambda-1))^{-2} \leq \lambda^{-2}+1$. Hence, the remainder term in the expansion (I.2) of $\log \left|\hat{\Sigma}_{\mathrm{N}}^{(i)}\right|$ has the upper bound

$$
\begin{align*}
& \sum_{j=1}^{q}\left\{\left[\hat{\lambda}_{N, j}^{(i)}\right]^{-2}+1\right\}\left\{\hat{\lambda}_{N, j}^{(i)}-1\right\}^{2} \\
& =\operatorname{tr}\left\{\hat{\Sigma}_{N}^{(i)-2}+I_{q}\right\}\left\{\hat{\Sigma}_{N}^{(i)}-I_{q}\right\}^{2} \tag{I.5}
\end{align*}
$$

The inequality (10.3) follows easily from (I.5) and the expansion (I.2) of $\log \left|\hat{\Sigma}_{\mathrm{N}}^{(2)}\right|$ and $\log \left|\hat{\mathbf{\Sigma}}_{\mathbf{N}}^{(1)}\right|$.

Proof of (8.11). Using (4.3) and $\Sigma^{(x)}=I_{q}$, one sees that

$$
\begin{equation*}
(-2)\left\{\hat{L}_{N}^{(x)}-L_{N}(x) \infty\right\}=N\left\{\log \left|\hat{\Sigma}_{N}^{(x)}\right|-\operatorname{tr}\left[\hat{\Sigma}_{N}^{(x)}-I_{q}\right]\right\}-Q_{N}^{(x)} \tag{I.6}
\end{equation*}
$$

and, from (8.1), that

$$
\begin{equation*}
(-2)\left\{\varepsilon_{N, \infty}^{(x)}-\dot{\mathcal{E}}_{N}^{(x)}\right\}=N\left\{\log \left|\dot{\Sigma}_{N}^{(x)-1}\right|-\operatorname{tr}\left[\hat{\Sigma}_{N}^{(x)-1}-I_{q}\right]\right\}-R_{N}(x) \tag{I.7}
\end{equation*}
$$

with $\mathrm{Q}_{\mathrm{N}}^{(\mathrm{x})}$ defined by (4.4) and $\mathrm{R}_{\mathrm{N}}^{(\mathrm{x})}$ defined as in (8.9). Since $\mathrm{Q}_{\mathrm{N}}^{(\mathrm{x})} \sim_{\mathrm{p}} \mathrm{R}_{\mathrm{N}}^{(\mathrm{x})}$, see section 8 , it remains to show the asymptotic coincidence of the terms in curly brackets in (I.6) and (I.7). From the expansions (I.2) for $\hat{\Sigma}_{N}^{(x)}$ and $\hat{\Sigma}_{N}^{(x)-1}$, we deduce that the bracketed terms differ by

$$
\begin{align*}
& N \sum_{j=1}^{q}\left[\left\{1-\left(\hat{\lambda}_{N, j}^{(x)}\right)^{-1}\right\}^{2}\left(\bar{\lambda}_{N, j}^{(x)}\right)^{2}-\left\{1-\hat{\lambda}_{N, j}^{(x)}\right\}^{2}\left(\tilde{\lambda}_{N}^{(x), j}\right)^{-2}\right] \\
& \quad=\sum_{j=1}^{q} N\left(1-\hat{\lambda}_{N, j}^{(x)}\right)^{2}\left\{\left(\bar{\lambda}_{N, j}^{(x)} / \hat{\lambda}_{N, j}^{(x)}\right)^{2}-\left(\tilde{\lambda}_{N, j}^{(x)}\right)^{-2}\right\}, \tag{I.8}
\end{align*}
$$

where $\bar{\lambda}_{N}(x)$ and $\tilde{\lambda}_{N, j}^{(x)}$ are between $\hat{\lambda}_{N, j}^{(x)}$ and 1 , for $j=1, \ldots, q$. Since the factors in curly brackets on the right hand side of (I.8) tend to zero in probability while their multipliers $N(1-\hat{\lambda}(x), j)^{2}$ are bounded, by (I.1), both sides of (I.8) tend to 0 , and (8.11) follows.

Now we turn to the proof required for section 7.

Proof of Proposition 7.2: Let $\mathrm{V}_{\mathrm{t}}^{(\mathrm{i})}$ denote the vector space of linear combinations of the $r^{(i)}$ entries of $x_{t}^{(i)}, i=1,2$. If $V_{t} \equiv V_{t}^{(1)} \cap V_{t}^{(2)}$ has dimension $r^{c}$, let $x_{t, j}^{c}$, $j=1, \ldots, r^{c}$ denote a basis for $V_{t}$ and let $z_{t, j}^{(i)}, j=1, \ldots, r^{(i)}-r^{c}$ denote a basis for the orthogonal complement of $\mathrm{V}_{\mathrm{t}}$ in $\mathrm{V}_{\mathrm{t}}^{(\mathrm{i})}$, orthogonal in the sense of the inner product defined by covariance. By stationarity, the coefficients of the linear combinations of the entries of $x_{t}^{(i)}$ used to produce these bases can be chosen independently of $t$. If we do this, and define $x_{t}^{c}=\left[x_{t, 1}^{c}, \ldots, x_{t, r}^{c}\right]^{\prime}$ and $z_{t}^{(i)}=$ - $\left[\mathrm{z}_{\mathrm{t}, 1}^{(\mathrm{i})}, \ldots, \mathrm{z}_{\mathrm{t}, \mathrm{r}}^{(\mathrm{i})}(\mathrm{i})_{-\mathrm{r}} \mathrm{c}^{\prime}\right.$, then the assertions of Proposition 7.2 follow easily.

## APPENDIX II: PROOFS FOR SUBSECTION 10.2

Proof of Lemma 10.3: The proof involves a sequence of reductions to simpler cases. Reduction to the case $N=m n, n=1,2, \ldots$, for any fixed $m$. Given $m$, we choose $n$ so that $(\mathrm{m}-1) \mathrm{n} \leq \mathrm{N} \leq \mathrm{mn}$. With $\mathrm{W}_{\mathrm{N}}=\mathrm{E}_{\mathrm{t}=1}^{\mathrm{N}} \epsilon_{\mathrm{t}} \epsilon_{\mathrm{t}}^{\prime}$ as in the subsection, we have $\lambda_{\text {min }}\left(\mathrm{W}_{(\mathrm{m}-1) \mathrm{n}}\right) \leq \lambda_{\min }\left(\mathrm{W}_{\mathrm{N}}\right) \leq \lambda_{\min }\left(\mathrm{W}_{\mathrm{mn}}\right)$, so

$$
(m-1) n \lambda_{\min }^{-1}\left(W_{m n}\right) \leq N \lambda_{\min }^{-1}\left(W_{N}\right) \leq m n \lambda_{\min }^{-1}\left(W_{(m-1) n}\right)
$$

holds. This reveals that (10.9) will follow provided we can show that the sequence $\mathbf{n}^{\mathbf{k}} E \lambda_{\min }^{-\mathbf{k}}\left(\mathrm{W}_{\mathrm{mn}}\right), \mathbf{n}=1,2, \ldots$, is bounded whenever $\mathrm{m} \geq \mathrm{h}+2 \mathbf{k}$.

Reduction to the case $n=1$. First, we observe that if $M_{1}, \ldots, M_{n}$, are positive definite matrices of order $h$, then from the arithmetic-geometric mean inequality, for any $h$-vector x ,

$$
n^{-1} x^{\prime}\left(\sum_{j=1}^{n} M_{j}\right) x \geq \prod_{j=1}^{n}\left(x^{\prime} M_{j} x\right)^{1 / n}
$$

which implies that

$$
\begin{equation*}
n \lambda_{\min }^{-1}\left(\sum_{j=1}^{n} M_{j}\right) \leq \prod_{j=1}^{n} \lambda_{\min }^{-1 / n}\left(M_{j}\right) \tag{II.1}
\end{equation*}
$$

Now set $M_{j}=\Sigma_{t=m(j-1)+1}^{m} \epsilon_{t} \epsilon_{t}^{\prime}$, so that $M_{j} \sim W_{h}(\Sigma, m)$ and $W_{m n}=\Sigma_{j=1}^{n} M_{j}$. Then from (II.1),

$$
\begin{aligned}
& \quad n^{k} E \lambda_{\min }^{-k}\left(W_{m n}\right) \leq E \prod_{j=1}^{n} \lambda_{\min }^{-k / n}\left(M_{j}\right) \\
& \leq \prod_{j=1}^{n}\left\{E \lambda_{\min }^{-k}\left(M_{j}\right)\right\}^{1 / n} \quad \text { (Hölder's inequality) } \\
& =E \lambda_{\min }^{-k}\left(M_{1}\right) \quad \text { (identical distribution). }
\end{aligned}
$$

Reduction to the case $\Sigma=I_{h}$. Let $\Sigma^{1 / 2}$ denote a symmetric square root of $\Sigma$. Then $\Sigma^{-1 / 2} W_{m} \Sigma^{-1 / 2} \sim W_{h}\left(I_{h}, m\right)$, and we will show that

$$
\begin{equation*}
\lambda_{\min }^{-1}\left(W_{m}\right) \leq \lambda_{\min }^{-1}\left(\Sigma^{-1 / 2} W_{m} \Sigma^{-1 / 2}\right) \lambda_{\min }^{-1}(\Sigma) \tag{II.2}
\end{equation*}
$$

In fact, for any $h$-vector $\mathbf{x}$,

$$
x^{\prime} W_{m} x=\left(\Sigma^{1 / 2} x\right)^{\prime} \Sigma^{-1 / 2} W_{m} \Sigma^{-1 / 2}\left(\Sigma^{1 / 2} x\right)
$$

$$
\geq \lambda_{\min }\left(\Sigma^{-1 / 2} W_{m} \Sigma^{-1 / 2^{\prime}}\right) x^{\prime} \Sigma x
$$

so that $\lambda_{\text {min }}\left(\mathrm{W}_{\mathrm{m}}\right) \geq \lambda_{\text {min }}\left(\Sigma^{-1 / 2} \mathrm{~W}_{\mathrm{m}} \Sigma^{-1 / 2}\right) \lambda_{\text {min }}(\Sigma)$, from which (II.2) follows. Our proof of (10.9) will therefore be complete if we verify (II.3):

$$
\begin{equation*}
\text { If } W_{m} \sim W_{h}\left(I_{h}, m\right) \text { and } m \geq h+2 k \text {, then } E \lambda_{\min }^{-k}\left(W_{m}\right)<\infty . \tag{II.3}
\end{equation*}
$$

In fact, the density function of the joint distribution of the eigenvalues $\lambda_{1} \geq \lambda_{2} \geq \ldots \geq \lambda_{h}$ - (see formula (11) of Anderson (1984, p. 534)) is bounded above by
$=\quad C_{h, m} \prod_{i=1}^{h} \lambda_{i}(m+h-2 i-1) \exp \left(-\lambda_{i} / 2\right)$
for some constant $C_{h, m}$. Hence, for some constant $C_{h, m}$,

$$
E \lambda_{\min }^{-k}\left(W_{m}\right) \leq C_{h, m} \int_{0}^{\infty} \lambda_{h}(m-h-1-2 k) / 2 \exp \left(-\lambda_{h} / 2\right) d \lambda_{h}
$$

The exponent of $\lambda_{h}$ in this integral is greater than -1 when $m \geq h+2 k$, so the integral is finite, and (II.3) follows. This shows that (10.9) holds.

Proof of Proposition 10.1: Set $\tilde{\mathrm{S}}_{\mathrm{N}} \equiv \mathrm{N} \hat{\Sigma}_{\mathrm{N}} . \quad$ Then $\lambda_{\min }\left(\tilde{\mathrm{S}}_{\mathrm{N}}\right)=\mathrm{N} \lambda_{\min }\left(\hat{\Sigma}_{\mathrm{N}}\right)$. Our proof of (10.12) is based mainly on an elaboration of an argument sketched in Appendix I of Fuller and Hasza (1981). Let the sample residual variance matrix of the least squares regression of $\tilde{x}_{m(p+1)}$ on $\tilde{x}_{m(p+1) \pm 1}, \ldots, \tilde{x}_{m(p+1)} \pm p, m=1, \ldots, n$, be denoted by $n^{-1} S_{n(p+1)}$. Then

$$
\begin{aligned}
& S_{n(p+1)^{\leq}} \sum_{m=1}^{n}\left(\tilde{x}_{m(p+1)^{-}} \hat{A}_{1} \tilde{x}_{m(p+1)-1}-\cdots-\hat{A}_{p} \tilde{x}_{m(p+1)-p}\right) \\
& \left(\tilde{x}_{\left.m(p+1)^{-\hat{A}_{1}} \tilde{x}_{m}(p+1)-1^{-\cdots-\hat{A}_{p}} \tilde{x}_{m(p+1)-p}\right)^{\prime} \leq \tilde{S}_{n(p+1)}} .\right.
\end{aligned}
$$

Thus, to establish (10.12), it is sufficient to prove that for any $k \geq 1$, there is an $n(k)$ such that

$$
\begin{equation*}
\sup _{\mathrm{n} \supseteq \mathrm{n}(\mathrm{k})} \mathrm{n}^{\mathbf{k}} \mathrm{E}_{\mathrm{min}}-\mathbf{k}\left(\mathrm{S}_{\mathrm{n}(\mathrm{p}+1)}\right)<\infty . \tag{II.4}
\end{equation*}
$$

We shall establish the existence of coefficient matrices $C_{j}, j= \pm 1, \ldots \pm p$, such that, conditional on $\tilde{x}_{m}(p+1) \pm 1, \ldots, \tilde{\mathbf{x}}_{m}(p+1) \pm p, m=1, \ldots, n$, the random variables $\tilde{e}_{m}$ defined by

$$
\begin{equation*}
\tilde{\mathrm{e}}_{\mathrm{m}} \equiv \tilde{\mathrm{x}}_{\mathrm{m}(\mathrm{p}+1)}-\sum_{|j|=1}^{\mathrm{p}} C_{j} \tilde{\mathrm{x}}_{\mathrm{m}}(\mathrm{p}+1)-\mathrm{j} \tag{II.5}
\end{equation*}
$$

are N.I.D. $(0, \tilde{\Sigma})$ for some $\tilde{\Sigma}>0$ which does not depend on the values of the conditioning variables. This result, which follows from Lemma II. 1 below, implies that for $\mathrm{n}>2 \mathrm{p} \tilde{\mathrm{r}}$, the conditional distribution of $\mathrm{S}_{\mathrm{n}(\mathrm{p}+1)}$ given $\mathrm{u}_{\mathrm{m}}=$ $\left(\tilde{\mathbf{x}}_{\mathrm{m}}^{\prime}(\mathrm{p}+1)-\mathrm{p}^{\prime}, \ldots, \tilde{\mathrm{x}}_{\mathrm{m}}^{\prime}(\mathrm{p}+1)-1\right)^{\prime}, \mathrm{m}=1, \ldots, \mathrm{n}+1$, is $\mathrm{W}_{\tilde{\mathrm{r}}}(\tilde{\mathrm{\Sigma}}, \mathrm{n}-2 \mathrm{p} \tilde{\mathrm{r}})$, by Theorem 8.22 of Anderson (1984). Therefore, from (10.9), there are constants $C$ and $n(k)$ such that the conditional ( -k )-th moments of $\lambda_{\text {min }}\left(\mathrm{S}_{\mathrm{n}(\mathrm{p}+1)}\right)$ satisfy

$$
\begin{equation*}
\sup _{n \geq n(k)} n^{k} E\left\{\lambda_{\min }^{-k}\left(S_{n(p+1)}\right) \mid u_{1}, \ldots, u_{n+1}\right\} \leq C . \tag{II.6}
\end{equation*}
$$

Since

$$
\left.E\left\{\lambda_{\min }^{-\mathbf{k}}\left(S_{n(p+1)}\right)\right\}=E_{u_{1}, \ldots, u_{n+1}}\left(E\left\{\lambda_{\min }^{-k} S_{n(p+1)}\right) \mid u_{1}, \ldots, u_{n+1}\right\}\right)
$$

property (II.4) follows from (II.6).

We return to the discussion of (II.5) to complete the proof. Let $g(a)$ denote the density of $a \sim V(0, \Sigma)$. The result needed is the following lemma.

- Lemma II.1. The conditional distribution of $z_{m}=\tilde{x}_{m(p+1)}, 1 \leq m \leq n$, given $\mathrm{u}_{1}, \ldots, \mathrm{u}_{\mathrm{n}+1}$, has the form

$$
\begin{equation*}
f\left(z_{1}, \ldots, z_{n} \mid u_{1}, \ldots, u_{n+1}\right)=\prod_{m=1}^{n}\left\{h\left(z_{m}, u_{m}, u_{m+1}\right) / \int_{\mathbb{R}^{\tilde{\Gamma}}} h\left(z_{m}, u_{m}, u_{m+1}\right) d z_{m}\right\} \tag{II.7}
\end{equation*}
$$

where $h\left(z_{m}, u_{m}, u_{m+1}\right)$ is the $p$-fold product function defined by $h\left(z_{m}, u_{m}, u_{m+1}\right) \equiv$

$$
\begin{equation*}
g\left(z_{m}-A_{1} \tilde{x}_{m}(p+1)-1 \cdots-A_{p} \tilde{x}_{m}(p+1)-p\right) \cdots g\left(\tilde{x}_{(m+1)}(p+1)-1^{\left.\left.-A_{1} \tilde{x}_{(m+1}\right)(p+1)-2^{\cdots-A_{p} z_{m}}\right) . . . . .}\right. \tag{II.8}
\end{equation*}
$$

Proof: Let $f\left(\tilde{\mathrm{x}}_{1-\mathrm{p}}, \ldots, \tilde{\mathrm{x}}_{0}\right)$ denote the joint density of $\tilde{\mathrm{x}}_{1-\mathrm{p}}, \ldots \tilde{\mathrm{x}}_{0}$. Then the joint density of $\tilde{\mathbf{x}}_{1-\mathrm{p}}, \ldots, \tilde{\mathrm{x}}_{(\mathrm{n}}(\mathrm{n})(\mathrm{p}+1)-1$ is

$$
f\left(\tilde{x}_{1-p}, \cdots, \tilde{x}_{0}\right) g\left(\tilde{x}_{1}-\sum_{j=1}^{p} A_{j} \tilde{x}_{1-j}\right) \cdots g\left(\tilde{x}_{(n+1)(p+1)-1}-\sum_{j=1}^{p} A_{j} \tilde{x}_{(n+1)}(p+1)-1-j\right)
$$

So, with the function $\mathrm{K}_{0}$ defined by

$$
K_{0} \equiv \int_{R} p \tilde{r} f\left(\tilde{x}_{1-p}, \cdots, \tilde{x}_{0}\right) g\left(\tilde{x}_{1}-\sum_{j=1}^{p} A_{j} \tilde{x}_{1-j}\right) \cdots g\left(\tilde{x}_{p}-\sum_{j=1}^{p} A_{j} \tilde{x}_{p-j}\right) d \tilde{x}_{1-p} \cdots d \tilde{x}_{0}
$$

the joint density of $z_{1}, \ldots, z_{n}, u_{1}, \ldots, u_{n+1}$ is given by

$$
\begin{equation*}
f\left(z_{1}, \ldots, z_{n}, u_{1}, \ldots u_{n+1}\right) \equiv K_{0} \cdot \stackrel{(n+1)}{\prod_{t=p+1}^{(p+1)-1}} g\left(\tilde{x}_{t}-\sum_{j=1}^{p} A_{j} \tilde{x}_{t-j}\right) \tag{II.9}
\end{equation*}
$$

Integrating over $z_{1}, \ldots z_{n}$, we obtain the joint density of $u_{1}, \ldots, u_{n+1}$,

$$
\mathrm{f}\left(\mathrm{u}_{1}, \ldots, \mathrm{u}_{\mathrm{n}+1}\right)=\mathrm{K}_{0} \cdot \prod_{\mathrm{m}=1}^{\mathrm{n}}\left\{\int_{\mathrm{R}^{\tilde{r}}} \mathrm{~h}\left(\mathrm{z}_{\mathrm{m}}, \mathrm{u}_{\mathrm{m}}, \mathrm{u}_{\mathrm{m}+1}\right) \mathrm{d} z_{\mathrm{m}}\right\}
$$

and the assertation of (II.7) follows from dividing (II.9) by this expression.
By adding the exponents of the $\mathrm{N}(0, \Sigma)$ density g-functions in (II.8), one sees from (II.7) that, conditioned on $u_{1}, \ldots u_{n+1}$, the random variables $z_{m}$ are independently normally distributed with means of the form $\sum_{|j|=1}^{p} C_{j} \tilde{x}_{m}(p+1)-j$ and nonsingular variance matrix $\tilde{\Sigma}=\left\{\Sigma^{-1}+\sum_{j=1}^{p} A_{j} \Sigma^{-1} A_{j}^{\prime}\right\}^{-1}$. Hence, the $\tilde{e}_{m^{\prime}}, 1 \leq m \leq n$ of (II.5) are N.I.D. $(0, \tilde{\Sigma})$, which is the result needed to complete the verification of (10.12).

Proof of Proposition 10.2: We shall obtain (10.15) from the special case (10.9). Let $\operatorname{det}\left(A-\lambda I_{p}\right)=\lambda^{\bar{r}}+a_{1} \lambda^{\bar{r}-1}+\cdots+a_{\bar{r}}$ and $\tilde{z}_{t}=\bar{x}_{t}+\Sigma_{j=1}^{\bar{r}} a_{j} \bar{x}_{t-j}$. Then $\tilde{\mathbf{z}}_{\mathrm{t}}=\tilde{e}_{\mathrm{t}}+\left(\mathrm{A}+\mathrm{a}_{1} \mathrm{I}_{\overline{\mathrm{r}}}\right) \tilde{e}_{\mathrm{t}-1}+\cdots+\left(\mathrm{A}^{\overline{\mathrm{r}}-1}+\mathrm{a}_{1} \mathrm{~A}^{\overline{\mathrm{r}}-2}+\cdots+a_{\overline{\mathbf{r}}-1} I_{\overline{\mathrm{r}}}\right) \tilde{\mathrm{e}}_{\mathrm{t}-\overline{\mathrm{r}}+1}$, see Lai and Wei (1985), equation (3.17). Clearly $\tilde{\mathbf{z}}_{\mathrm{t}} \sim \mathrm{N}(0, \Sigma)$, with

$$
\Sigma=\bar{\Sigma}+\left(A+a_{1} I_{\overline{\mathbf{r}}}\right) \bar{\Sigma}\left(A+a_{1} I_{\overline{\mathbf{r}}}\right)^{\prime}+\cdots+\left(A^{\overline{\mathbf{r}}-1}+\cdots+a_{\overline{\mathbf{r}}-1} I_{\overline{\mathbf{r}}}\right) \bar{\Sigma}\left(A^{\overline{\mathbf{r}}-1}+\cdots+a_{\overline{\mathbf{r}}-1} I_{\overline{\mathbf{r}}}\right)^{\prime}
$$

It follows from (10.14) that $\Sigma$ is nonsingular, see Lai and Wei (1985, p. 381). Let $\mathrm{a}_{\mathrm{o}}=1$. Then, by (3.20) of the same reference,

$$
\begin{equation*}
\lambda_{\min }\left(\sum_{t=\bar{r}+1}^{N} \tilde{z}_{t} \tilde{z}_{t}^{\prime}\right) \leq \overline{\mathrm{r}}\left(\sum_{j=0}^{\overline{\mathrm{r}}} \mathrm{a}_{\mathrm{j}}^{2}\right) \lambda_{\min }\left(\sum_{t=1}^{N} \bar{x}_{t} \bar{x}_{t}^{\prime}\right) \tag{II.10}
\end{equation*}
$$

Consider the time series $e_{t}$ obtained by observing every $\overline{\mathrm{r}}$-th value of $\tilde{\mathrm{z}}_{\mathrm{t}}$ : $e_{t} \equiv \mathbb{*}_{\overline{\mathrm{z}} \mathrm{t}}, \mathrm{t}=1,2, \ldots$. Then $e_{\mathrm{t}}$ is N.I.D. $(0, \Sigma)$ and

$$
\begin{equation*}
\lambda_{\min }\left(\sum_{t=2}^{N} e_{t} e_{t}^{\prime}\right) \leq \lambda_{\min }\left(\sum_{t=\bar{r}+1}^{N \bar{r}} \tilde{z}_{t^{\prime}} \tilde{z}_{t}^{\prime}\right) \tag{II.11}
\end{equation*}
$$

Now (10.15) follows immediately from (II.10), (II.11), and (10.9) via a reduction argument of the sort used at the start of the proof of Lemma 10.3 in Appendix I.

