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Asymptotic Properties of the Gauge and Power of StepIndicator Saturation

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# Asymptotic Properties of the Gauge and Power of Step-Indicator Saturation 

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#### Abstract

Step-Indicator Saturation (SIS) is an algorithm to address multiple location shifts at unknown dates in time series during model selection. We derive asymptotic theory for tuning parameter choice based on consistency and asymptotic normality of the frequence gauge - the rate of false detections. Simulations suggest that a smaller gauge minimizes bias in post-selection regression estimates. For the small gauge situation, we develop a complementary Poisson theory. We compare the local power of SIS to detect shifts with that of an extant method. We find that SIS excels when breaks are near the sample end or closely spaced. An application to UK labor productivity reveals a growth slowdown after the 2008 financial crisis.


## 1 Introduction

Step-Indicator Saturation (SIS), suggested by Castle et al. (2015), is a model selection algorithm designed to address location shifts in time series without restrictions on their number, date, and distance to each other or sample boundaries. In its most general form, the initial specification has a $k$-variate regressor $x_{i}$, which can be of the exogenous, (trend-)stationary, or random walk type, and as many step indicators as observations:

$$
\begin{equation*}
y_{i}=\beta^{\prime} x_{i}+\sum_{j=1}^{n} \delta_{j} 1_{(i \leq j)}+\varepsilon_{i} \quad \text { for } i=1, \ldots, n \tag{1}
\end{equation*}
$$

If the number of $\delta_{j} \neq 0$ and their location $j$ were small and known, the model could be estimated by least squares. In practice, the nature of location shifts is often unknown,

[^0]and investigators estimate them using block-search algorithms (Doornik, 2009; Hoover \& Perez, 1999; Hendry \& Krolzig, 2005). Such algorithms depend on a tuning parameter, which can be chosen indirectly by controlling the type I error. Castle et al. (2015) measured type I errors in terms of the frequency of falsely detected shifts, which we will refer to as the gauge. We develop an asymptotic theory to understand the gauge of simplified versions of SIS, and show that for conformable values of the gauge, the procedure maintains power to detect shifts.

Location shifts are a common phenomenon in observed time series (Perron, 1989; Andrews, 1993; Bai \& Perron, 1998), and a failure to address them can affect model selection probabilities of variables (Castle \& Hendry, 2014), distort parameter estimation (Hendry \& Mizon, 2011), and result in forecast failure (Clements \& Hendry, 1998). The growing importance of SIS in tackling location shifts is reflected in its applications in fields as varied as economics (Chuffart \& Hooper, 2019; Pellini, 2021; Bernstein \& Martinez, 2021), climate science (Raggad, 2018; Pretis et al., 2018; Koch et al., 2022; O'Callaghan et al., 2022), and public health (Doornik et al., 2022). However, despite its popularity, no study of its asymptotic properties exists. This study fills the gap using theoretical insights to shed light on four pivotal areas for practitioners: First, the control of its tuning parameter with the gauge can be closely aligned with the investigator's preferences without detailed knowledge of the regressor type. Second, the bias in post-selection regression estimates can be addressed by choosing a small gauge, or switching to the Poisson theory for the gauge when it is vanishing. Third, SIS can detect minor shifts after a short period of upheaval and maintains power near the end of the sample. Fourth, SIS has weak regularity conditions for the regressors.

In this paper, we study the split-half SIS algorithm. This is a simplified version of the SIS algorithm as implemented in tools like EViews (2020), gets in R (Pretis et al., 2018; Sucarrat, 2020), and Autometrics in Oxmetrics (Doornik, 2009). Simulations by Castle et al. (2015) indicate that the general SIS has the same gauge properties as splithalf SIS, but can detect a wider range of shifts with more power. Split-half SIS splits the sample into two subsets with $n_{1}$ and $n-n_{1}$ observations. It then applies stylized SIS to both subsamples. Stylized SIS, when applied to the second subsample, excludes the first set of step-indicators. For example, it is imposed that $\delta_{j}=0$ for $j \leq n_{1}$ in (1). The model is then estimated by OLS to determine which of the coefficients $\delta_{j}$ for $j>n_{1}$ are significant. An analysis of split-half SIS can shed light on more general versions of the algorithm and provide mathematical tools for examining related algorithms.

Split-half SIS results in $n$ decisions about the inclusion of step-indicators $1_{(i \leq j)}$. This method requires setting a tuning parameter: a common cut-off $c$ for selecting stepindicators. Drawing inspiration from classical test theory, we aim to determine the cut-off $c$ indirectly from a measure of type I error. Classical testing problems focus on single-decision problems in which the critical value - or the cut-off - is chosen from the size of the test, which is the probability of a type I error of falsely rejecting the hypothesis. In multiple-decision problems, there are many alternative ways of measuring type I error. We study the gauge, which is based on a count of the false rejections. The gauge is also referred to as the expected error rate (Miller, 1981) or the per-comparison error rate (Dudoit \& van der Laan, 2010). A concept similar to the gauge was introduced by Hoover \& Perez (1999). The term gauge originates in Hendry \& Santos (2010) and

Castle et al. (2011), see also Hendry \& Doornik (2014, p. 122).
When comparing the gauge to alternative measures of type I error in the context of multiple-decision testing problems, we note that the gauge is more amenable to asymptotic analysis. These alternatives include the probability error rate (Miller, 1981; Dudoit \& van der Laan, 2010), also called the family-wise error rate (Dudoit \& van der Laan, 2010), and the false discover rate (Benjamini \& Hochberg, 1995). The probability error rate is the probability of at least one false detection. It requires a detailed assessment of the dependence of the individual decisions. In contrast, the gauge ignores this dependence structure. The false discovery rate is the expected value of the proportion of type I errors among the rejected hypotheses. Under our null hypothesis of no location shift in the data generating process, the false discovery rate equals unity.

To formalize the notion of the gauge, consider two equivalent approaches to formulate stylized SIS. We introduced this algorithm by imposing $\delta_{j}=0$ for $j \leq n_{1}$ in (1), estimating the model by least squares and then investigating the significance of the remaining step-indicators. An equivalent alternative formulation is to first regress $y_{i}$ on $x_{i}$ and an intercept for $i \leq n_{1}$. This yields least squares estimators $\hat{\beta}_{1}$ and $\hat{\sigma}_{1}^{2}$. These estimators will be consistent if there are no location shifts in the first subsample. We then compute the scaled residuals in the second subsample. As pointed out by Castle et al. (2015) and as shown in Section 2, we can then inspect the forward differenced residuals for outliers. That is, if there are $n$ observations of (1), compute

$$
\begin{equation*}
\left(\nabla y_{i}-\hat{\beta}_{1}^{\prime} \nabla x_{i}\right) / \sqrt{2} \hat{\sigma}_{1} \quad \text { for } i=n_{1}+1, \ldots, n-1 \tag{2}
\end{equation*}
$$

with $\nabla y_{i}=y_{i}-y_{i+1}$, and where the $\sqrt{2}$-factor arises since the variance of $\nabla \varepsilon_{i}$ is twice the variance of $\varepsilon_{i}$. A location shift is declared if the absolute value of the forward differenced residual exceeds a cut-off, $c$. The frequency of declared location shifts in the stylized SIS algorithm is the frequence gauge:

$$
\begin{equation*}
\hat{\gamma}_{n}=\frac{1}{n-n_{1}-1} \sum_{i=n_{1}+1}^{n-1} 1_{\left(\left|\nabla y_{i}-\hat{\beta}_{1}^{\prime} \nabla x_{i}\right| \geq \sqrt{2} \hat{\sigma}_{1} c\right)} . \tag{3}
\end{equation*}
$$

If the data generating process has no location shifts, then all declarations of shifts are false, so that $\hat{\gamma}_{n}$ is the average type I error. We show the consistency

$$
\begin{equation*}
\hat{\gamma}_{n} \xrightarrow{\mathrm{p}} \gamma=\mathrm{P}\left(\left|\nabla \varepsilon_{i}\right| \geq \sqrt{2} \sigma c\right), \tag{4}
\end{equation*}
$$

for a wide range of time series regressors $x_{i}$ including stationary and non-stationary regressors. We can then choose the cut-off $c$ from the limiting gauge $\gamma$. In simulations, we confirm the consistency result and provide some further analysis.

The consistency of the frequence gauge for a variety of time series regressors shows that it is possible to control the type I error of SIS without prior knowledge of the detailed time series structure. The regressors do have a second-order effect on this consistency result, which we investigate through an asymptotic expansion of the normalized frequence gauge $n^{1 / 2}\left(\hat{\gamma}_{n}-\gamma\right)$. We find that it is asymptotically zero mean normal, but its variance depends on the correlation structure of $\nabla x_{i}$ and $\nabla \varepsilon_{i}$. Numerical approximations confirm that the asymptotic variance of the frequence gauge is strictly larger for split-half SIS than for split-half IIS. In contrast to split-half IIS, the asymptotic variance
of split-half SIS depends on the temporal persistence of the time series. A small gauge substantially reduces its asymptotic variance.

A challenge to the asymptotic analysis of the frequence gauge for SIS is the temporal and cross-sectional correlation due to the forward differencing of $x_{i}$ and $\varepsilon_{i}$ in (2). For instance, in the autoregression $x_{i+1}=\rho x_{i}+\varepsilon_{i}$ with independent $\varepsilon_{i}$ and $x_{i}$, we get that $\nabla \varepsilon_{i}=\varepsilon_{i}-\varepsilon_{i+1}$ is temporally and cross-sectionally correlated with $\nabla x_{i}=x_{i}-x_{i+1}=$ $(1-\rho) x_{i}-\varepsilon_{i}$. In the related asymptotic analysis of IIS by Hendry et al. (2008) and Johansen \& Nielsen (2009, 2013, 2016a,b), the use of impulse-indicators of the form $1_{(i=j)}$ avoids the temporal and cross-sectional correlation structure. Therefore, IIS can be analyzed using a version of the empirical process theory of (Koul \& Ossiander, 1994), see also Giraitis et al. (2012). Our analysis of the SIS overcomes the correlation problem by combining the empirical process theory with mixingale theory of McLeish (1977).

A simulation study shows that split-half SIS can introduce a bias in the updated estimates for $\beta$ in (1) that does not vanish asymptotically. The bias is largest when regressors are lagged dependent variables with an autoregressive coefficient close to unity, and when the frequence gauge is large. For split-half SIS, the empirical setting after the selection over the step-indicators resembles an unbalanced panel regression with a small temporal and large cross-sectional range. Each interval in between two consecutive retained step-indicators can be interpreted as another $i$ in the panel that introduces a new individual fixed effect. The incidental parameter problem arises because with a non-zero frequence gauge $\gamma$, the number of breaks is approximately $n \gamma$, so that the number of observations in each interval is on average $1 / \gamma$ and therefore finite even as the sample size increases. This matches the situation of a panel data model with large cross-sectional dimension and finite time series dimension, in which biases arise for the dynamic parameter estimators. We conjecture that the bias is due to a combination of the incidental parameter problem (Lancaster, 2000, 2002) and the correlation of the retained step-indicators with the innovations (Arellano \& Bond, 1991).

We suggest two different approaches address the bias in the estimation of $\beta$ under split-half SIS. First, simulations suggest investigators to use a small frequence gauge, as a smaller frequence gauge is associated with a smaller bias. In a sample of 100 observations, we would recommend a frequence gauge of $1 \%$ if one would normally conduct inference at the $5 \%$ level. Second, we develop a theory for shrinking the gauge with increasing sample sizes. For this, we consider the absolute gauge

$$
\begin{equation*}
\hat{\Gamma}_{n}=\sum_{i=n_{1}+1}^{n-1} 1_{\left(\left|\nabla y_{i}-\hat{\beta}_{1}^{\prime} \nabla x_{i}\right| \geq \sqrt{2} \hat{\alpha}_{1} c_{n}\right)} \tag{5}
\end{equation*}
$$

for increasing sequences of the cut-off $c_{n}$ that satisfy, for some $\lambda>0$,

$$
\begin{equation*}
\mathrm{P}\left(\left|\nabla \varepsilon_{i}\right|>\sqrt{2} \sigma c_{n}\right)=\lambda / n \tag{6}
\end{equation*}
$$

As the $c_{n}$ increases with the sample size, the absolute gauge is smaller than the frequence gauge as the sample grows. By modifying the theory of Johansen \& Nielsen (2016b), we show that the absolute gauge $\hat{\Gamma}_{n}$ is asymptotically Poisson distributed. The asymptotic result is the same whether the regressors are stationary or non-stationary. In the proof, we encounter the same dependence issue between $\nabla x_{i}$ and $\nabla \varepsilon_{i}$. We address this using the Poisson limit theorem of Chen (1975).

An alternative to SIS is the Bai \& Perron (1998) procedure. It builds on the Andrews (1993) breaks test and provides estimates for timing and location of breaks. Comparing the power properties of SIS and the Bai-Perron procedure is challenging due to the inherent complexity of both methods. Instead, we compare stylized SIS with the Andrews test. We consider two types of scenarios: Scenarios where the Andrews test is consistent in that power approaches unity while stylized SIS has trivial power approaching the gauge. And scenarios where the Andrews test has trivial power while stylized SIS is consistent. On balance, we find that the Andrews test is preferable if there is one break or two well-separated breaks in the middle of the sample. SIS is preferable for a break near the end of the sample. Such a break is important to discover and address in forecasting contexts (Clements \& Hendry, 1998). SIS is also preferable if two close breaks offset each other, for instance if the growth rate moves from one level to a slightly different level through a short period of upheavel, see Castle et al. (2023) and the empirical illustration. We argue that the results carry over to a comparison with the Bai-Perron procedure.

The proof of the local power results uses convergence on the $D[0,1]$ space of discontinuous functions. We handle the one-break case by the Skorokhod (1956) $J_{1}$-metric discussed by Billingsley (1968). However, in order to establish convergence in the two-close-breaks case, we use Skorokhod's $M_{1}$ metric in line with Whitt (2002).

Our theory for simplified versions of SIS requires knowledge of the innovation distribution. The normal distribution is the standard choice. Just as in a standard regression, the normality assumption will be testable from the residuals once the model has been fitted. With a finite cut-off, the standard cumulant based normality test may have to be adjusted. Indeed, this is the case when applying outlier detection with finite cut-off (Berenguer-Rico \& Nielsen, 2023). In contrast, standard heteroscedasticity tests remain valid after outlier detection with finite cut-off (Berenguer-Rico \& Wilms, 2021). It should be noted that other procedures, such as the Andrews (1993) only require distributional assumptions that are sufficient to apply a Central Limit Theorem. In turn, SIS requires weaker assumptions to the regressors.

We apply our split-half SIS theory to analyze the UK labor productivity from 1980 to 2021. While there is a growing consensus about the decline of productivity growth in the UK (Chadha, 2022), a simple autoregressive model is not rejected by the standard diagnostic tests. This indicates that location shifts are not always obvious to the investigator. Using a $1 \%$ gauge, the split-half SIS algorithm identified multiple shifts in UK productivity growth: $0.56 \%$ before $2000,0.37 \%$ up to 2008 and $0.04 \%$ up to 2020 . These findings also illustrate the ability of SIS to find minor shifts around episodes of upheaval, in our case the 2008 financial crisis and the 2020 Covid pandemic.

Section 2 outlines the model and the SIS algorithm. Sections 3, 4 present the asymptotic results on the frequence gauge for the stylized and split-half SIS respectively, while section 5 presents the Poisson theory for the absolute gauge. Power analysis is found in Section 6. Simulation results are given in Section 7. An empirical illustration follows in Section 8. Section 9 concludes. Proofs follow in an appendix.

## 2 Model and algorithms

We begin by presenting the linear regression model to which we apply the SIS algorithm. Subsequently, we introduce two simplified versions of the SIS: stylized SIS and split-half SIS. Lastly, given that the decisions rules on the retaining of step-indicators pertain to differenced innovations, we discuss of their notable properties.

### 2.1 The model

Step-Indicator Saturation (SIS) aims to detect location shifts within the model:

$$
\begin{equation*}
y_{i}=\mu+\beta^{\prime} x_{i}+\varepsilon_{i} \quad \text { for } i=1, \ldots, n \tag{7}
\end{equation*}
$$

By saturating with step-indicators of the type $1_{(i \leq j)}$, we obtain equation (1) with $\delta_{n}=\mu$. In practice, one would expect that only a few of the $\delta_{j}$ parameters in (1) are non-zero, but their number and location are unknown. The regressor $x_{i}$ is a $k$-vector, which does not include an intercept. It can include stationary, trend-stationary and random walk variables, but excludes explosive regressors. The innovations $\varepsilon_{i}$ are independent, identically, distributed with a continuous distribution that is known up to the scale. Further, the innovations are independent of current and past regressors $x_{j}$ for $j \leq i$. The coefficient to the intercept is identified when $\mathrm{E} \varepsilon_{i}=0$, but the asymptotic theory does not depend on this constraint.

As a model selection algorithm, the idea of SIS is grounded in the general-to-specific approach to regressor selection of Hoover \& Perez (1999). Its core mechanism revolves around iterative backward elimination: in each step, a regression is estimated, the least significant regressor is eliminated, and the smaller model is re-estimated. The iteration stops when the fit of the model deteriorates too much. While a single backward elimination has poor properties for correlated regressors, Hoover \& Perez (1999) found that multiple backward eliminations with different starting points have better properties in recovering the original data generating process. Algorithms such as PcGets (Hendry \& Krolzig, 2005) and Autometrics (Doornik, 2009) adopt this multi-path approach but search over many more paths to get closer to evaluating all possible subsets of regressors. Autometrics allows situations with more regressors than observations by searching over blocks of regressors. This permits saturation with indicators for each observation as in Impulse-Indicator Saturation (IIS) and Step-Indicator-Saturation (SIS). The saturation approach allows a simultaneous search over regressors $x_{i}$ and indicator variables. The simultaneous search is helpful when there is a high sample correlation between regressors and indicator variables, see Hendry \& Doornik (2014). SIS is implemented in the R package gets (Pretis et al., 2018; Sucarrat, 2020), in EViews (2020), and within a structural time series model in Marczak \& Proietti (2016). It is worth noting that Autometrics employs indicators of the form $1_{(i \leq j)}$ as here, while gets utilizes $1_{(i \geq j)}$. A related algorithm based on sensitivity analysis was presented by Becker et al. (2021).

### 2.2 Split-half estimation and forward differencing

While regression equations with more variables than observations cannot be estimated as a single equation, they can be approached by using a subset - or blocks - of those
variables. The strategy involves experimenting with various blocks to find the relevant regressors. The simplest block search algorithm is the stylized SIS. We apply it to (1). It begins by dividing the observations into two parts: the first $n_{1}$ observations and the remaining $n_{2}=n-n_{1}$ observations. For the first half, we keep only an intercept and otherwise drop the indicator variables. This gives the model equation

$$
\begin{equation*}
y_{i}=\beta^{\prime} x_{i}+\mu 1_{\left(i \leq n_{1}\right)}+\sum_{j=n_{1}+1}^{n} \delta_{j} 1_{(i \leq j)}+\varepsilon_{i} \quad \text { for } i=1, \ldots, n . \tag{8}
\end{equation*}
$$

Since the second half-sample is saturated with indicators, that half will have perfect fit. The consequence of this observation is best seen through reparameterization. Multiply $x_{i}$ by unity, written as a sum of indicators for the first half $\left(i \leq n_{1}\right)$ and for the impulses $(i=\ell)$ for $n_{1}<\ell \leq n$. Decompose the indicator for $(i \leq j)$ likewise. This gives

$$
y_{i}=\beta^{\prime} x_{i} 1_{\left(i \leq n_{1}\right)}+\left(\mu+\sum_{j=n_{1}+1}^{n} \delta_{j}\right) 1_{\left(i \leq n_{1}\right)}+\sum_{\ell=n_{1}+1}^{n} \beta^{\prime} x_{i} 1_{(i=\ell)}+\sum_{j=n_{1}+1}^{n} \delta_{j} \sum_{\ell=n_{1}+1}^{j} 1_{(i=\ell)}+\varepsilon_{i} .
$$

Interchanging summation order in the last $\delta$-term gives the reparameterization

$$
y_{i}=\beta^{\prime} x_{i} 1_{\left(i \leq n_{1}\right)}+\nu 1_{\left(i \leq n_{1}\right)}+\sum_{\ell=n_{1}+1}^{n} \eta_{\ell} 1_{(i=\ell)},
$$

where

$$
\begin{equation*}
\nu=\mu+\sum_{j=n_{1}+1}^{n} \delta_{j}, \quad \eta_{\ell}=\beta^{\prime} x_{\ell}+\sum_{j=\ell}^{n_{1}} \delta_{j} . \tag{9}
\end{equation*}
$$

As the indicators are orthogonal, the least squares estimators for $\beta, \nu$ are found by standard multiple regression on $x_{i}$ and the intercept using the first sample, while $\eta_{\ell}$ is estimated by $\hat{\eta}_{\ell}=y_{\ell}$. Solving the expression for $\eta_{\ell}$ in (9) for $\delta_{\ell}$ shows that

$$
\begin{equation*}
\hat{\delta}_{\ell}=\left(\hat{\eta}_{\ell}-\hat{\beta}^{\prime} x_{\ell}\right)-\left(\hat{\eta}_{\ell+1}-\hat{\beta}^{\prime} x_{\ell+1}\right)=\nabla y_{\ell}-\hat{\beta}^{\prime} \nabla x_{\ell} \quad \text { for } n_{1}<\ell<n, \tag{10}
\end{equation*}
$$

while $\hat{\delta}_{n}=\hat{\eta}_{n}-\hat{\beta}^{\prime} x_{n}=y_{n}-\hat{\beta}^{\prime} x_{n}$. In the subsequent analysis, we will analyze the gauge, which is the count of the significant estimated $\delta_{\ell}$ coefficients. Apart from the last estimate, these are all based on forward differencing. In an asymptotic analysis of the gauge, we can ignore the last estimator without affecting the asymptotic result.

### 2.3 Step-Indicator Saturation algorithms

We present two simplified SIS algorithms in a more formal way. The algorithms involve splitting the sample in two consecutive parts for the $n_{1}$ first observations and the $n_{2}=$ $n-n_{1}$ last observations. When working with differenced variables, one observation is lost from each sub-sample. We define index sets

$$
\begin{equation*}
I_{1}=\left(i \leq n_{1}\right), \quad I_{1}^{\circ}=\left(i<n_{1}\right), \quad I_{2}=\left(n_{1}<i \leq n\right), \quad I_{2}^{\circ}=\left(n_{1}<i<n\right), \tag{11}
\end{equation*}
$$

and counts $n_{1}^{\circ}=n_{1}-1, n_{2}^{\circ}=n_{2}-1$ and $n^{\circ}=n_{1}^{\circ}+n_{2}^{\circ}=n-2$. For each sub-sample $I_{j}$, for $j=1,2$, we estimate the constant intercept regression model $y_{i}=\mu+\beta^{\prime} x_{i}+\varepsilon_{i}$ by least squares regression and get the estimators

$$
\begin{array}{ll}
\bar{x}_{j}=n_{j}^{-1} \sum_{i \in I_{j}} x_{i}, & \hat{\beta}_{j}=\left\{\sum_{i \in I_{j}}\left(x_{i}-\bar{x}_{j}\right)\left(x_{i}-\bar{x}_{j}\right)^{\prime}\right\}^{-1} \sum_{i \in I_{j}}\left(x_{i}-\bar{x}_{j}\right) y_{i}, \\
\bar{y}_{j}=n_{j}^{-1} \sum_{i \in I_{j}} y_{i}, & \hat{\sigma}_{j}^{2}=\frac{1}{n_{j}} \sum_{i \in I_{j}}\left\{\left(y_{i}-\bar{y}_{j}\right)-\hat{\beta}_{j}^{\prime}\left(x_{i}-\bar{x}_{j}\right)\right\}^{2} . \tag{13}
\end{array}
$$

We will use the estimates from the first sub-sample to predict location shifts in the second sub-sample. This corresponds to predicting outliers for the differenced series using $\nabla y_{i}-\hat{\beta}_{1} \nabla x_{i}$. This gives the forecast correction factors

$$
\begin{equation*}
\omega_{1, i}^{2}=1+\left(\nabla x_{i}\right)^{\prime}\left\{2 \sum_{k \in I_{1}^{\circ}}\left(x_{k}-\bar{x}_{j}\right)\left(x_{k}-\bar{x}_{j}\right)^{\prime}\right\}^{-1} \nabla x_{i} \quad \text { for } i \in I_{2}, \tag{14}
\end{equation*}
$$

and we define $\omega_{2, i}^{2}$ vice versa when replacing the index sets $I_{2}^{\circ}, I_{2}$ by $I_{1}^{\circ}, I_{1}$. The factors arise as follows. First, rewrite $\nabla y_{i}-\hat{\beta}_{1}^{\prime} \nabla x_{i}=\nabla \varepsilon_{i}-\left(\hat{\beta}_{1}-\beta\right)^{\prime} \nabla x_{i}$ by applying equation (7). Then, assuming fixed regressors and independent normal $\mathrm{N}\left(0, \sigma^{2}\right)$ innovations we get that $\nabla y_{i}-\hat{\beta}_{1}^{\prime} \nabla x_{i}$ is normal $\mathrm{N}\left(0,2 \sigma^{2} \omega_{1, i}^{2}\right)$. Later we show that under mild regularity conditions $\omega_{2, i}^{2}$ is uniformly close to unity and it can indeed be replaced by unity for asymptotic purposes. We define the stylized SIS algorithm, which searches for location shifts in the second sub-sample.

## Algorithm 2.1. The stylized Step-Indicator Saturation algorithm.

1. Choose a cut-off value $c>0$ to select breakpoints.
2. Calculate the least squares estimators $\left(\hat{\beta}_{1}, \hat{\sigma}_{1}^{2}\right)$ based on sample $I_{1}$.
3. Calculate forecast correction factors $\omega_{1, i}^{2}$ for $i \in I_{2}$.
4. Declare a location shift at $i+1$ if

$$
\begin{equation*}
\left|\nabla y_{i}-\hat{\beta}_{1}^{\prime} \nabla x_{i}\right| \geq \sqrt{2} \hat{\sigma}_{1} \omega_{1, i} c \quad \text { for } i \in I_{2}^{\circ} \tag{15}
\end{equation*}
$$

The frequency of location shifts declared by Algorithm 2.1 is

$$
\begin{equation*}
\hat{\gamma}_{n}^{\text {stylized }}=\frac{1}{n_{2}^{\circ}} \sum_{i \in I_{2}^{\circ}} 1_{\left(\left|\nabla y_{i}-\hat{\beta}_{1}^{\prime} \nabla x_{i}\right| \geq \sqrt{2} \hat{\sigma}_{1} \omega_{1, i} c\right)} . \tag{16}
\end{equation*}
$$

When the data generating process has no location shifts, so that $\mu_{i}=\mu$, the expression $\hat{\gamma}_{n}$ is the frequence gauge of the algorithm, which is the object of interest in this paper.

Castle et al. (2015) refer to a split-half SIS algorithm, which is a symmetrized version of the above algorithm. For reference, we define that algorithm including a statement on how to update the estimators for $\beta, \sigma^{2}$ in light of the identified location shifts. We allow the sub-samples to be of unequal size, but retain the split-half descriptor.

## Algorithm 2.2. The split-half Step-Indicator Saturation algorithm.

1. Choose a cut-off value $c>0$ to select breakpoints.
2. Calculate the least squares estimators $\left(\hat{\beta}_{j}, \hat{\sigma}_{j}^{2}\right)$ based on sample $I_{j}$ for $j=1,2$.
3. Calculate forecast correction factors $\omega_{j, i}^{2}$ for $i \notin I_{j}$ and $j=1,2$.
4. Declare a location shift at $i+1$ if

$$
\begin{equation*}
\left|\nabla y_{i}-\hat{\beta}_{j}^{\prime} \nabla x_{i}\right| \geq \sqrt{2} \hat{\sigma}_{j} \omega_{j, i} c \quad \text { for } i \in I_{3-j}^{\circ} \text { and } j=1,2 . \tag{17}
\end{equation*}
$$

For notational simplicity, we do not consider the possibility of a location shift from $i=n_{1}$ to $i=n_{1}+1$. The split-half SIS algorithm of Castle et al. (2015) continues to re-estimate $\beta, \sigma$ on the full sample while taking the detected location shifts into account.

The frequency of declared location shifts by Algorithm 2.2 is

$$
\begin{equation*}
\hat{\gamma}_{n}^{\text {split }}=\frac{1}{n^{\circ}}\left\{\sum_{i \in I_{1}^{\circ}} 1_{\left(\left|\nabla y_{i}-\hat{\beta}_{2}^{\prime} \nabla x_{i}\right| \geq \sqrt{2} \hat{\sigma}_{2} \omega_{2, i} c\right)}+\sum_{i \in I_{2}^{\circ}} 1_{\left(\left|\nabla y_{i}-\hat{\beta}_{1}^{\prime} \nabla x_{i}\right| \geq \sqrt{2} \hat{\sigma}_{1 \omega_{1}, i c}\right)}\right\} . \tag{18}
\end{equation*}
$$

### 2.4 Properties of the differenced innovations

The scaled innovations $\varepsilon_{i} / \sigma$ have density f . In applications, we often assume f to be the normal density. The forward differenced innovations are denoted

$$
\begin{equation*}
\nabla \varepsilon_{i}=\varepsilon_{i}-\varepsilon_{i+1}, \quad \chi_{i}=\nabla \varepsilon_{i} /(\sqrt{2} \sigma) \tag{19}
\end{equation*}
$$

The scaled forward differenced innovations $\chi_{i}$ have the convolution density

$$
\begin{equation*}
\mathrm{h}(x)=\sqrt{2} \int_{-\infty}^{\infty} \mathrm{f}(y) \mathrm{f}(\sqrt{2} x+y) \mathrm{d} y \tag{20}
\end{equation*}
$$

and distribution function H . Following (4) let

$$
\begin{equation*}
\gamma=\mathrm{P}\left(\left|\chi_{i}\right| \geq c\right) \tag{21}
\end{equation*}
$$

We highlight four properties of the density $h$.
Theorem 2.3. Assume $\varepsilon_{i} / \sigma$ are i.i.d. and continuous with density f . The density h then satisfies the following properties:
(a) Symmetry: $\mathrm{h}(x)=\mathrm{h}(-x)$;
(b) Suppose f has second moment. Then $\mathrm{f}=\mathrm{h}$ if and only if f is standard normal;
(c) for $k \in \mathbb{N}_{0}: \sup _{v \in \mathbb{R}}|v|^{k} \mathrm{f}(v)<\infty \Rightarrow \sup _{v \in \mathbb{R}}|v|^{k} \mathrm{~h}(v)<\infty$;
(d) for $k \in \mathbb{N}_{0}$ : $\sup _{v \in \mathbb{R}}\left(1+|v|^{k}\right)|\dot{\mathrm{f}}(v)|<\infty$ and $\mathrm{E}\left|\varepsilon_{i}^{k}\right|<\infty \Rightarrow \sup _{v \in \mathbb{R}}\left|v^{k} \dot{\mathrm{~h}}(v)\right|<\infty$.

Theorem 2.3 implies that when the reference distribution f for $\varepsilon$ is standard normal so is the distribution $h$ for $\chi_{i}$. Thus, the gauge $\gamma$ is associated with a cut-off $c$ chosen as the normal $(1-\gamma / 2)$ quantile.

## 3 The main results for stylized SIS

We present an asymptotic theory for the frequence gauge of stylized SIS. The first-order result is consistency. This allows us to choose the cut-off $c$ indirectly from the gauge. We obtain consistency for a wide range of stationary and non-stationary regressors. We will also develop a second-order expansion of the gauge with a view to understand
how uniform the consistency result is. In this section, we give an asymptotic expansion, which is developed into an asymptotic theory for split-half SIS in the subsequent section. We then find that the asymptotic distribution is normal for a wide range of regressors, but with an asymptotic variance depending on the type of regressors.

We require the following time series structure for innovations $\varepsilon_{i}$ and regressors $x_{i}$.
Assumption 3.1. Let $\mathcal{F}_{i}$ be a filtration so that $\varepsilon_{i-1}$ and $x_{i}$ are $\mathcal{F}_{i-1}$-adapted, and $\varepsilon_{i} / \sigma$ has unit variance and is independent of $\mathcal{F}_{i-1}$ with distribution function F and positive density $f$ on $\mathbb{R}$ with derivative $\dot{f}$.

Assumption 3.1 rules out endogeneity of the form, $\operatorname{Cov}\left(x_{i}, \varepsilon_{i}\right) \neq 0$, but allows predetermined time series regressors. The innovations need not have zero mean as Theorem $2.3(a)$ implies $\mathrm{E} \nabla \varepsilon_{i}=0$ even if $\mathrm{E} \varepsilon_{i} \neq 0$. Intriguingly, Jiao (2019) exploits the techniques developed here to analyze situations with endogeneity.

The theory results allow for stationary and non-stationary regressors. For this purpose, we introduce normalization matrices $N_{j}$ for each sub-sample $j=1,2$. This yields normalized regressors

$$
\begin{equation*}
x_{i n}=N_{j}^{\prime} x_{i}, \quad \nabla x_{i n}=N_{j}^{\prime}\left(x_{i}-x_{i+1}\right) \quad \text { for } i \in I_{j}^{\circ}, \tag{22}
\end{equation*}
$$

where we have suppressed the index $j$ in the definition of the normalized regressor $x_{i n}$. We choose the normalizations depending on the stochastic properties of $x_{i}$ so that

$$
\begin{equation*}
\widehat{\Sigma}_{j n}=\sum_{i \in I_{j}} N_{j}^{\prime}\left(x_{i}-\bar{x}_{j}\right)\left(x_{i}-\bar{x}_{j}\right)^{\prime} N_{j} \quad \text { where } \quad \widehat{\Sigma}_{j n}^{-1}=\mathrm{O}_{\mathrm{P}}(1) . \tag{23}
\end{equation*}
$$

In the asymptotic theory we will require that

$$
\begin{equation*}
\widehat{V}_{j n}=\sum_{i \in I_{j}} N_{j}^{\prime}\left(x_{i}-\bar{x}_{j}\right)\left(\varepsilon_{i}-\mathrm{E} \varepsilon_{i}\right)=\mathrm{O}_{\mathrm{P}}(1) ; \quad \mathrm{E} \sum_{i \in I_{j}^{\circ}}\left|\nabla x_{i n}\right|^{2}=\mathrm{O}(1) \tag{24}
\end{equation*}
$$

For the practitioner it will be possible to choose the cut-off $c$ without precise knowledge of the type of regressors and hence the normalization. The knowledge of the type is only needed for the second-order theory.

We give some examples of normalizations. If $x_{i}$ is stationary, then $\nabla x_{i}$ is also stationary. Thus, we let $N_{j}=n_{j}^{-1 / 2} I_{\operatorname{dim} x}$ and find that $\widehat{\Sigma}_{1 n}, \widehat{V}_{1 n}$ and $\mathrm{E} \sum_{i \in I_{2}^{\circ}}\left|\nabla x_{i n}\right|^{2}$ converge under mild regularity conditions. If $x_{i}$ is a random walk, then $\nabla x_{i}$ is i.i.d. and we let $N_{j}=n_{j}^{-1} I_{\operatorname{dim} x}$. Then, under mild regularity conditions, $\widehat{\Sigma}_{1 n}, \widehat{V}_{1 n}$ converge, while $\mathrm{E} \sum_{i \in I_{2}^{\circ}}\left|\nabla x_{i n}\right|^{2}$ vanishes. Thus, the asymptotic expansions simplify in the latter case. As an example of cointegrated regressors, we could have

$$
N_{1}=\left(\begin{array}{cc}
n^{-1 / 2} & 0 \\
0 & n^{-1}
\end{array}\right)\left(\begin{array}{cc}
1 & -1 \\
0 & 1
\end{array}\right) \quad \text { if } \quad x_{i}=\binom{1}{1} \sum_{j=1}^{i-1} \varepsilon_{j}+z_{i}
$$

for some stationary, bivariate process $z_{i}$. We note that this $N_{1}$ is non-diagonal.
In most applications, the density of the innovations $\varepsilon_{i}$ will be normal. However, the density needs neither be centered at zero nor be symmetric as the theory results will only depend on the implied density for the differenced innovations $\nabla \varepsilon_{i}=\varepsilon_{i}-\varepsilon_{i+1}$.

Our theory does require that the density f of the innovations $\varepsilon_{i}$ and its derivative are bounded. The condition is satisfied for a wide range of densities, including the normal density. Moreover, the differenced innovations' conditional density, given the differenced regressors and the past, should also be bounded. If the regressors are pre-determined, this reduces to boundedness of the density of the differenced innovations and follows from the boundedness of the density f of the innovations $\varepsilon_{i}$ due to Theorem 2.3.

Assumption 3.2. Suppose that
(i) the density f satisfies $(a) \sup _{v \in \mathbb{R}} \mathrm{f}(v)<\infty$, (b) $\sup _{v \in \mathbb{R}}\left(1+v^{2}\right)|\dot{\mathrm{f}}(v)|<\infty$;
(ii) the conditional density $\mathrm{m}_{i}(y \mid x)$ of $\chi_{i}$ given $\nabla x_{i}$ and $\mathcal{F}_{i-1}$ exists for $i=n_{1}+1, \ldots, n$, it is differentiable in $y$ and satisfies $\max _{n_{1}+1 \leq i \leq n} \sup _{y \in \mathbb{R}, x \in \mathbb{R}^{p}}(1+|y|)\left|\dot{\mathrm{m}}_{i}(y \mid x)\right|<\infty$;
(iii) the regressors $x_{i}$ satisfy, with $\widehat{\Sigma}_{1 n}, \widehat{V}_{1 n}$ defined in (23), (24):
(a) $\widehat{\Sigma}_{1 n}^{-1}=\mathrm{O}_{\mathrm{P}}(1)$, (b) $\widehat{V}_{1 n}=\mathrm{OP}(1),(c) \mathrm{E} \sum_{i \in I_{2}^{\circ}}\left|\nabla x_{i n}\right|^{2}=\mathrm{O}(1)$;
(iv) the sub-sample lengths satisfy $\left(n_{2} / n_{1}\right)^{1 / 2}, N_{2}^{-1} N_{1}=\mathrm{o}\left(n_{2}^{1 / 4-\eta}\right)$ for some $\eta>0$.

We start by showing that the forecast correction factor $\omega_{1, i}^{2}$ can be replaced by unity with negligible asymptotic consequences.

Theorem 3.3. Consider the gauge of the stylized SIS Algorithm 2.1. Suppose Assumptions 3.1, 3.2(ia,iii,iv) apply and that no locations shifts are present so that $\mu_{i}=\mu$. Then, we get for fixed $c \in \mathbb{R}$ that

$$
\hat{\gamma}_{n}^{\text {stylized }}=\frac{1}{n_{2}^{\circ}} \sum_{i \in I_{2}^{\circ}} 1_{\left(\left|\nabla y_{i}-\hat{\beta}_{1}^{\prime} \nabla x_{i}\right| \geq \sqrt{2} \hat{\sigma}_{1} c\right)}=\frac{1}{n_{2}^{\circ}} \sum_{i \in I_{2}^{\circ}} 1_{\left(\left|\nabla y_{i}-\hat{\beta}_{1}^{\prime} \nabla x_{i}\right| \geq \sqrt{2} \hat{\sigma}_{1} \omega_{1, i} c\right)}+\mathrm{o}_{\mathrm{P}}\left(n_{2}^{-1 / 2}\right) .
$$

The next result presents the expansion for the frequence gauge $\hat{\gamma}_{n}^{\text {stylized }}$ of stylized SIS as defined in (16) around the population gauge $\gamma=\mathrm{P}\left(\left|\chi_{i}\right| \geq c\right)=\mathrm{P}\left(\left|\nabla \varepsilon_{i}\right| \geq \sqrt{2} \sigma c\right)$. The data generating process is assumed to have no location shifts.

Theorem 3.4. Consider the gauge of the stylized SIS Algorithm 2.1. Suppose Assumptions 3.1, 3.2 apply and that no locations shifts are present so that $\mu_{i}=\mu$. Let

$$
\xi_{2 n}(c)=n_{2}^{-1 / 2} \sum_{i \in I_{2}^{\circ}} \mathrm{E}_{i-1}\left(\nabla x_{i n} \mid \chi_{i}=c\right)=\mathrm{O}_{\mathrm{P}}(1)
$$

Then, we get for fixed $c \in \mathbb{R}$ that

$$
\begin{align*}
n_{2}^{1 / 2}\left(\hat{\gamma}_{n}^{\text {stylized }}-\gamma\right)= & n_{2}^{-1 / 2} \sum_{i \in I_{2}^{\circ}}\left\{1_{\left(\left|x_{i}\right| \geq c\right)}-\mathrm{E} 1_{\left(\left|x_{i}\right| \geq c\right)}\right\}  \tag{25}\\
& -\operatorname{ch}(c)\left(n_{2} / n_{1}\right)^{1 / 2} n_{1}^{-1 / 2} \sum_{i \in I_{1}}\left(\varepsilon_{i}^{2} / \sigma^{2}-1\right) \\
& -\mathrm{h}(c)(\sqrt{2} \sigma)^{-1}\left\{\xi_{2 n}(c)-\xi_{2 n}(-c)\right\}^{\prime} N_{2}^{-1} N_{1} \widehat{\Sigma}_{1 n}^{-1} \widehat{V}_{1 n}+\mathrm{op}_{\mathrm{P}}(1)
\end{align*}
$$

Finally, $\hat{\gamma}_{n}^{\text {stylized }}$ is consistent in that $\hat{\gamma}_{n}^{\text {stylized }} \rightarrow \gamma$ in probability and in mean.
The consistency statement in Theorem 3.4 for the stylized SIS algorithm is nuisance parameter-free. It can be used for calibrating the SIS algorithm. The result provides
the rationale for choosing $c$ to match the desired population gauge $\gamma$ : We specify our tolerance for false positives expressed by $\gamma$. Given the innovation density f we obtain a selection quantile $c$. For example, if the innovations $\varepsilon_{i}$ are normal, then the forward differenced innovations $\chi_{i}$ are standard normal by Theorem 2.3. If the sample is $n=100$ and $\gamma=1 \%$, we choose $c=2.58$, which is the normal $99.5 \%$ quantile.

The expansion in Theorem 3.4 has three components. The first component is a binomial term. The next two components relate to the estimation uncertainty from the initial estimation. They involve factors $n_{2} / n_{1}$ and $N_{2}^{-1} N_{1}$, respectively, where $N_{2}^{-1} N_{1}$ is an increasing function of $n_{2} / n_{1}$. These factors are allowed to diverge at an o $\left(n^{1 / 4-\eta}\right)$ rate. This means that the expansion would apply if we choose, in a stationary context, $n_{1}=n^{7 / 8}$ and $n_{2}=n-n_{1}$, so that $n_{2} / n_{1}=\mathrm{O}\left(n^{1 / 8}\right)$, which requires that $\eta<1 / 8$ in Assumption $3.2(i v)$. In other words, the length of the sub-sample used for the initial estimation may be of a lower order of magnitude than the sub-sample used to search for location shifts. This feature is implicitly exploited in more complicated versions of the algorithm, which search for small sub-sets of observations without location shifts.

The third term in the Theorem 3.4 expansion involves the nuisance quantity $\xi_{2 n}(c)$. It vanishes in two distinct cases. First, if the regressors are strictly exogenous, then $\mathrm{E}_{i-1}\left(\nabla x_{i} \mid \chi_{i}=c\right)=\mathrm{E}_{i-1} \nabla x_{i}$ does not depend on $c$ so that $\xi_{2 n}(c)-\xi_{2 n}(-c)=0$. Second, for random walk type regressors with stationary $\nabla x_{i}$ the normalization is $N_{2}=n^{-1}$ so that $\xi_{2 n}(c)$ vanishes. The third term simplifies if the sequence $\nabla x_{i}, \chi_{i}$ is stationary. In this case, we let $N_{2}=n_{2}^{-1 / 2}$ and get $\xi_{2 n}(c)=n_{2}^{-1} \sum_{i \in I_{2}^{\circ}} \mathrm{E}_{i-1}\left(\nabla x_{i} \mid \chi_{i}=c\right)$. Under regularity conditions, this converges in probability to $\mathrm{EE}_{0}\left(\nabla x_{1} \mid \chi_{1}=c\right)=\mathrm{E}\left(\nabla x_{1} \mid \chi_{1}=c\right)$. Under a normality assumption, this can be computed explicitly. Thus, suppose that $\left(\nabla x_{1}, \chi_{1}\right)$ are normal given $\mathcal{F}_{0}$ with conditional mean $\left(v_{0}, 0\right)$, where $v_{0}$ is $\mathcal{F}_{0}$-measurable with expectation $\mathrm{E} v_{0}=0$. Noting that $\chi_{1}$ has unit variance, we have

$$
\binom{\nabla x_{1}}{\chi_{1}} \left\lvert\, \mathcal{F}_{0} \stackrel{\mathrm{D}}{=} \mathrm{N}\left\{\binom{v_{0}}{0},\left(\begin{array}{cc}
\sigma_{\nabla \nabla} & \sigma_{\nabla \chi}  \tag{26}\\
\sigma_{\chi \nabla} & 1
\end{array}\right)\right\} .\right.
$$

Then, $\xi_{2 n}(c) \rightarrow \mathrm{E} v_{0}+c \sigma_{\nabla \chi}=c \sigma_{\nabla \chi}$ in probability, while $\xi_{2 n}(c)-\xi_{2 n}(-c) \rightarrow 2 c \sigma_{\nabla \chi}$. For example, in the autoregression $y_{i}=\mu+\alpha y_{i-1}+\varepsilon_{i}$ so that $x_{i}=y_{i-1}$, we find that $\sigma_{\nabla \chi}=\mathrm{E}_{0}\left(x_{1}-x_{2}\right)\left(\varepsilon_{1}-\varepsilon_{2}\right) /(\sqrt{2} \sigma)=\mathrm{E}_{0}\left(y_{0}-y_{1}\right)\left(\varepsilon_{1}-\varepsilon_{2}\right) /(\sqrt{2} \sigma)=-\sigma / \sqrt{2}$.

In the statement of Theorem 3.4, the initial least squares estimation is based on observations with indices $I_{1}=\left(i \leq n_{1}\right)$, while the search for location shifts is based on observations with indices $I_{2}=\left(i>n_{1}\right)$. The consecutive nature of the sets $I_{1}$ and $I_{2}$ are convenient in the proof to simplify notation. However, the result extends to situations where the sets $I_{1}$ and $I_{2}$ are more complicated. Indeed, this is possible because Theorem 3.4 is derived under the hypothesis of no location shifts. It would be possible to choose $I_{1}$ as all odd and $I_{2}$ as all even indices. In that case, all observations will be involved when computing the forward differences arising from the set $I_{2}$.

## 4 The main results for split-half SIS

We provide an asymptotic expansion for split-half SIS and analyze the asymptotic distribution of the frequence gauge for stationary and for random walk regressors.

### 4.1 Expansion of the gauge for split-half SIS

We expand the gauge for split-half SIS by applying Theorem 3.4 to each sub-sample. This requires a symmetrized version of Assumption 3.2.

Assumption 4.1. Suppose that
(i) the density f satisfies $\sup _{v \in \mathbb{R}} \mathrm{f}(v)<\infty$, $\sup _{v \in \mathbb{R}}\left(1+v^{2}\right)|\dot{\mathrm{f}}(v)|<\infty$;
(ii) the conditional density $\mathrm{m}_{i}(y \mid x)$ of $\chi_{i}$ given $\nabla x_{i}$ and $\mathcal{F}_{i-1}$ exists for $i=1, \ldots, n$, it is differentiable in $y$ and satisfies $\max _{1 \leq i \leq n} \sup _{y \in \mathbb{R}, x \in \mathbb{R}^{p}}(1+|y|)\left|\dot{\mathrm{m}}_{i}(y \mid x)\right|<\infty$;
(iii) the regressors $x_{i}$ satisfy for $j=1,2$, with $\widehat{\Sigma}_{j n}, \widehat{V}_{j n}$ defined in (23), (24):
(a) $\widehat{\Sigma}_{1 n}^{-1}=\mathrm{O}_{\mathrm{P}}(1)$, (b) $\widehat{V}_{j n}=\mathrm{O}_{\mathrm{P}}(1)$, and (c) $\mathrm{E} \sum_{i \in I_{j}^{\circ}}\left|\nabla x_{i n}\right|^{2}=\mathrm{O}(1)$;
(iv) the sub-sample lengths satisfy $\left(n_{2} / n_{1}\right)^{1 / 2}, N_{2}^{-1} N_{1}=\mathrm{o}\left(n_{2}^{1 / 4-\eta}\right)$, and $\left(n_{1} / n_{2}\right)^{1 / 2}$, $N_{1}^{-1} N_{2}=\mathrm{o}\left(n_{1}^{1 / 4-\eta}\right)$ for some $\eta>0$.
Theorem 4.2. Consider the gauge of the split-half SIS Algorithm 2.2. Suppose Assumptions 3.1, 4.1 apply and that no locations shifts are present so that $\mu_{i}=\mu$. Let $\xi_{j n}(c)=n_{j}^{-1 / 2} \sum_{i \in I_{j}} \mathrm{E}_{i-1}\left(N_{j}^{\prime} \nabla x_{i} \mid \chi_{i}=c\right)$ for $j=1,2$. Then, we get for fixed $c \in \mathbb{R}$ that

$$
\begin{aligned}
& \sqrt{n}\left(\hat{\gamma}_{n}^{s p l i t}-\gamma\right)= n^{-1 / 2} \sum_{i=1}^{n-1}\left\{1_{\left(\left|\chi_{i}\right| \geq c\right)}-\mathrm{E} 1_{\left(\left|\chi_{i}\right| \geq c\right)}\right\} \\
&- \operatorname{ch}(c) n^{-1 / 2} \sum_{i=1}^{n}\left\{n_{2} n_{1}^{-1} 1_{\left(i \in I_{1}\right)}+n_{1} n_{2}^{-1} 1_{\left(i \in I_{2}\right)}\right\}\left(\varepsilon_{i}^{2} \sigma^{-2}-1\right) \\
&-\mathrm{h}(c)(\sqrt{2} \sigma)^{-1}\left[\left(n_{1} / n\right)^{1 / 2}\left\{\xi_{1 n}(c)-\xi_{1 n}(-c)\right\}^{\prime} N_{1}^{-1} N_{2} \widehat{\Sigma}_{2 n}^{-1} \widehat{V}_{2 n}\right. \\
&\left.\quad+\left(n_{2} / n\right)^{1 / 2}\left\{\xi_{2 n}(c)-\xi_{2 n}(-c)\right\}^{\prime} N_{2}^{-1} N_{1} \widehat{\Sigma}_{1 n}^{-1} \widehat{V}_{1 n}\right]+\mathrm{op}(1)
\end{aligned}
$$

Finally, $\hat{\gamma}_{n}^{\text {split }}$ is consistent in that $\hat{\gamma}_{n}^{\text {split }} \rightarrow \gamma$ in probability and in mean.
Once again, the consistency statement in Theorem 4.2 for the split-half SIS algorithm is nuisance parameter-free.

### 4.2 Asymptotic distribution in the stationary case

We now consider the expansion of split-half SIS when the regressors $x_{j}$ are stationary. We start by introducing some notations for various moments for the innovations $\varepsilon_{i}$ :

$$
\begin{align*}
\varkappa_{1} & =\mathrm{E} \varepsilon_{i} / \sigma, \quad \varkappa_{2}=\mathrm{E} \varepsilon_{i}^{2} / \sigma^{2}=1, \quad \varkappa_{4}=\mathrm{E} \varepsilon_{i}^{4} / \sigma^{4}  \tag{27}\\
\varsigma_{0} & =\mathrm{E}\left\{1_{\left(\left|\chi_{i}\right| \geq c\right)} 1_{\left(\left|\chi_{i+1}\right| \geq c\right)}\right\}  \tag{28}\\
\varsigma_{2} & =\mathrm{E}\left\{1_{\left(\left|\chi_{i}\right| \geq c\right)}\left(\varepsilon_{i+1}^{2} / \sigma^{2}-1\right)\right\}=\mathrm{E}\left\{1_{\left(\left|\chi_{i}\right| \geq c\right)}\left(\varepsilon_{i}^{2} / \sigma^{2}-1\right)\right\} . \tag{29}
\end{align*}
$$

Further, for the stationary regressor $x_{i}$, we denote

$$
\begin{equation*}
\mu_{x}=\mathrm{E} x_{i}, \quad \Sigma_{x}=\operatorname{Var} x_{i} \tag{30}
\end{equation*}
$$

and finally, for a cross moment for innovations and regressors, we denote

$$
\begin{align*}
\varsigma_{1 x} & =\mathrm{E}\left\{\nabla x_{i}\left(1_{\left(\left|\chi_{i}\right| \geq c\right)}-\gamma\right)\left(\varepsilon_{i} / \sigma-\varkappa_{1}\right)\right\}  \tag{31}\\
\xi_{c} & =\mathrm{E}\left(\nabla x_{i} \mid \chi_{i}=c\right)=\mathrm{E}\left(\nabla x_{i} \mid \chi_{i}=-c\right) \tag{32}
\end{align*}
$$

Then, the vector $s_{i}=\left\{1_{\left(\left|\chi_{i}\right| \geq s\right)}-\gamma, \varepsilon_{i}^{2} / \sigma^{2}-1,\left(\varepsilon_{i} / \sigma-\varkappa_{1}\right)\left(x_{i}-\mu_{x}\right)^{\prime} \Sigma_{x}^{-1}\right\}^{\prime}$ has variance and first-order autocovariance of the form

$$
\Omega_{0}=\left(\begin{array}{ccc}
\gamma(1-\gamma) & \varsigma_{2} & 0  \tag{33}\\
\varsigma_{2} & \varkappa_{4}-1 & 0 \\
0 & 0 & \Sigma_{x}^{-1}\left(1-\varkappa_{1}^{2}\right)
\end{array}\right), \quad \Omega_{1}=\left(\begin{array}{ccc}
\varsigma_{0}-\gamma^{2} & \varsigma_{2} & \varsigma_{1 x}^{\prime} \Sigma_{x}^{-1} \\
0 & 0 & 0 \\
0 & 0 & 0
\end{array}\right)
$$

Finally, we define long-run variances for the summands of the frequence gauge in (18). Let $(j, k)$ be $(1,2)$ or $(2,1)$ and define with $n_{j} / n \rightarrow \lambda_{j}>0$ for $j=1,2$

$$
d_{j}=\left(\begin{array}{c}
1  \tag{34}\\
-c \mathrm{~h}(c)\left(\lambda_{k} / \lambda_{j}\right) \\
-\mathrm{h}(c) \xi_{c}\left(\lambda_{k} / \lambda_{j}\right) / \sqrt{2}
\end{array}\right), \quad \omega_{j}^{2}=d_{j}^{\prime} \Omega_{0} d_{j}+2 e_{1}^{\prime} \Omega_{1} d_{j}
$$

The long-run variances $\omega_{j}^{2}$ will be assumed to be positive in order to exploit the Functional Central Limit Theorem for non-stationary mixingales in McLeish (1977).

Example 4.1. If $\varepsilon_{i} / \sigma$ has standard normal density $\varphi$ and distribution function $\Phi$, then $\mathrm{h}(x)=\varphi(x)$, while $\varkappa_{1}=0$ and $\varkappa_{4}=3$. It is argued in Appendix A. 9 that

$$
\begin{equation*}
\varsigma_{0}=2 \gamma-4\{T(c, 1 / \sqrt{3})+T(c, \sqrt{3})\}, \quad \varsigma_{2}=c \varphi(c) \tag{35}
\end{equation*}
$$

Here, $T(c, a)=\int_{c}^{\infty} \varphi(x) \int_{0}^{a x} \varphi(y) d y d x$ following Owen (1980, 2.2; 2.8). In particular, $T(c, a)$ is positive and decreasing in $c$ with $T(0,1 / \sqrt{3})=1 / 12$ and $T(0, \sqrt{3})=1 / 6$. Finally, if $\nabla x_{1}, \chi_{1}$ are jointly normal given $\mathcal{F}_{0}$ as in (26) then $\xi_{c}=2 c \sigma_{\nabla \chi}$.

Assumption 4.3. Suppose
(i) the density f satisfies $\sup _{v \in \mathbb{R}}|v| \mathrm{f}(v)<\infty$ and $\int_{\mathbb{R}} v^{4+} \mathrm{f}(v) d v<\infty$;
(ii) the pairs $x_{i}, \varepsilon_{i}$ are stationary with $\mathrm{E}\left|x_{i}^{2+}\right|<\infty$;
(iii) $\omega_{1}^{2}, \omega_{2}^{2}>0$;
(iv) let $z_{i}$ be either of $x_{i}, x_{i} x_{i}^{\prime}$ or $\nabla x_{i} 1_{\left(\left|x_{i}\right| \geq c\right)}\left(\varepsilon_{i} / \sigma-\varkappa_{1}\right)$ and suppose
$\mathrm{E}\left|\mathrm{E}_{k-m} n^{-1} \sum_{i=k+1}^{k+n}\left(z_{i}-\mathrm{E} z_{i}\right)\right| \rightarrow 0$ as $\min (k, m, n) \rightarrow \infty ;$
(v) $n^{-1} \sum_{i=1}^{n} x_{i}=\mu_{x}+\mathrm{op}_{\mathrm{P}}(1)$.

Theorem 4.4. Consider the gauge of the split-half SIS Algorithm 2.2 with $n_{j} / n \rightarrow$ $\lambda_{j}>0$ for $j=1,2$, so that $\lambda_{1}+\lambda_{2}=1$. Suppose Assumptions 3.1, 4.1, 4.3 apply and that no locations shifts are present so that $\mu_{i}=\mu$. Then, for fixed $c \in \mathbb{R}$, we get $n^{1 / 2}\left(\hat{\gamma}_{n}^{\text {split }}-\gamma\right) \xrightarrow{\mathrm{D}} \mathrm{N}(0, B)$, where

$$
\begin{align*}
B= & \lambda_{1} \omega_{1}^{2}+\lambda_{2} \omega_{2}^{2} \\
= & \gamma(1-\gamma)+2\left(\varsigma_{0}-\gamma^{2}\right)-4 c \mathrm{~h}(c) \varsigma_{2}-\sqrt{2} \mathbf{h}(c) \varsigma_{1 x}^{\prime} \Sigma_{x}^{-1} \xi_{c} \\
& +\left(\lambda_{1}^{2} / \lambda_{2}+\lambda_{2}^{2} / \lambda_{1}\right) \mathbf{h}^{2}(c)\left\{c^{2}\left(\varkappa_{4}-1\right)+\left(1-\varkappa_{1}^{2}\right) \xi_{c}^{\prime} \Sigma_{x}^{-1} \xi_{c} / 2\right\} . \tag{36}
\end{align*}
$$

Example 4.2. Let $y_{i}=\mu+\alpha y_{i-1}+\varepsilon_{i}$ be stationary so that $|\alpha|<1$ and $\varepsilon_{i} / \sigma$ is standard normal. Then $x_{i}=y_{i-1}$ has mean $\mu_{x}=\mu /(1-\alpha)$ and variance $\Sigma_{x}=\sigma^{2} /\left(1-\alpha^{2}\right)$. It is argued in Appendix A.9 that $\sigma_{\nabla \chi}=-\sigma / \sqrt{2}$ in (26), that $\varsigma_{1 x}=-\sigma \varsigma_{2}$ and that condition (iv) of Assumption 4.3 holds.

Example 4.3. We consider the asymptotic distribution of the gauge for standard normally distributed error terms $\varepsilon_{i} / \sigma$, so that $\varkappa_{1}=0$ and $\varkappa_{4}=3$. Further, assume that the sample size in the two sub-samples is equal and that the regressors $x_{i}$ are stationary. The asymptotic variance (36) in Theorem 4.4 then reduces to

$$
\begin{align*}
B= & \gamma(1-\gamma)+2\left(\varsigma_{0}-\gamma^{2}\right)-4 c h(c) \varsigma_{2}-\sqrt{2} \mathbf{h}(c) \varsigma_{1 x}^{\prime} \Sigma_{x}^{-1} \xi_{c} \\
& +\mathrm{h}^{2}(c)\left(2 c^{2}+\xi_{c}^{\prime} \Sigma_{x}^{-1} \xi_{c} / 2\right) . \tag{37}
\end{align*}
$$

Recall that if, in addition, $\nabla x_{1}, \chi_{1}$ are normal given $\mathrm{F}_{0}$ as in (26) then $\xi_{c}=2 c \sigma_{\nabla \chi}$. Further, in a first-order autoregression $y_{i}=\mu+\alpha y_{i-1}+\varepsilon_{i}$ the conditional covariance $\sigma_{\nabla \chi}$ equals $-\sigma / \sqrt{2}$, while $\varsigma_{1 x}=-\varsigma_{2}$ and $\Sigma_{x}=\operatorname{Var} x_{i}=\sigma^{2} /\left(1-\alpha^{2}\right)$.

### 4.3 Distribution of split-half SIS when $\xi_{n j}$ vanishes

The Theorem 4.2 expansion for the split-half SIS's frequence gauge simplifies when the term $\xi_{n j}$ vanishes so that the third term in the expansion falls away. As remarked after Theorem 3.4, this happens for strictly exogenous or random walk regressors. The limiting long-run variance simplifies so that

$$
\begin{equation*}
\tilde{\omega}_{j}^{2}=\gamma(1-\gamma)+2\left(\varsigma_{0}-\gamma^{2}\right)-4 c h(c)\left(\lambda_{k} / \lambda_{j}\right) \varsigma_{2}+c^{2} h^{2}(c)\left(\lambda_{k} / \lambda_{j}\right)^{2}\left(\varkappa_{4}-1\right) . \tag{38}
\end{equation*}
$$

We will require that $\tilde{\omega}_{j}^{2}>0$.
Theorem 4.5. Consider the gauge of the split-half SIS Algorithm 2.2 with $n_{j} / n \rightarrow \lambda_{j}>$ 0 for $j=1,2$, so that $\lambda_{1}+\lambda_{2}=1$. Let $\xi_{j n}=\mathrm{OP}(1)$. Suppose Assumptions 3.1, 4.1, 4.3(i) apply, $\tilde{\omega}_{j}^{2}>0$ for $j=1,2$ and that no locations shifts are present so that $\mu_{i}=\mu$. Then, for fixed $c \in \mathbb{R}$, we get $n^{1 / 2}\left(\hat{\gamma}_{n}^{\text {split }}-\gamma\right) \xrightarrow{\mathrm{D}} \mathrm{N}(0, \tilde{B})$, where

$$
\begin{align*}
\tilde{B} & =\lambda_{1} \tilde{\omega}_{1}^{2}+\lambda_{2} \tilde{\omega}_{2}^{2} \\
& =\gamma(1-\gamma)+2\left(\varsigma_{0}-\gamma^{2}\right)-4 c h(c) \varsigma_{2}+c^{2} h^{2}(c)\left(\lambda_{1}^{2} / \lambda_{2}+\lambda_{2}^{2} / \lambda_{1}\right)\left(\varkappa_{4}-1\right) \tag{39}
\end{align*}
$$

## 5 Poisson approximation

We present a theory for a vanishing frequence gauge. We set the cut-off so as to control the absolute gauge, the number of falsely discovered outliers. This gives a Poisson exceedance theory. For stylized and split-half SIS, the absolute gauges are

$$
\begin{align*}
\hat{\Gamma}_{n}^{s t y l i z e d} & =\sum_{i \in I_{2}^{\circ}} 1_{\left(\left|\nabla y_{i}-\hat{\beta}_{1}^{\prime} \nabla x_{i}\right| \geq \sqrt{2} \hat{\sigma}_{1} \omega_{1, i} c_{n}\right)},  \tag{40}\\
\hat{\Gamma}_{n}^{s p l i t} & =\sum_{i \in I_{2}^{\circ}} 1_{\left(\left|\nabla y_{i}-\hat{\beta}_{1}^{\prime} \nabla x_{i}\right| \geq \sqrt{2} \hat{\sigma}_{1} \omega_{1, i} c_{n}\right)}+\sum_{i \in I_{1}^{\circ}} 1_{\left(\left|\nabla y_{i}-\hat{\beta}_{2}^{\prime} \nabla x_{i}\right| \geq \sqrt{2} \hat{\sigma}_{2} \omega_{2, i} c_{n}\right)} . \tag{41}
\end{align*}
$$

Here, we choose the cut-off $c_{n}$ so that, for some $\lambda>0$,

$$
\begin{equation*}
\mathrm{P}\left(\left|\nabla \varepsilon_{i}\right|>\sqrt{2} \sigma c_{n}\right)=\mathrm{P}\left(\left|\chi_{i}\right|>c_{n}\right)=\lambda / n \tag{42}
\end{equation*}
$$

The analysis builds on the Poisson exceedance theory for IIS, (Johansen \& Nielsen, 2016b). The analysis has two part. The first part is a Poisson limit theorem for the case without estimation errors. For IIS, the standard Poisson limit theorem could be used. For SIS, we have that the forward differenced innovations are 1-dependent. We can then apply the Chen (1975) Poisson limit theorem. The second part is an argument that the estimation errors do not matter for the asymptotic theory. This argument is similar to that of the IIS analysis. For the analysis, we need the following high-level assumptions.
Assumption 5.1. Suppose that
(i) the innovations $\varepsilon_{i}$ are i.i.d., so that $\chi_{i}=\nabla \varepsilon_{i} /(\sqrt{2} \sigma)$ has continuous distribution function H with density h satisfying
(a) $\mathrm{E}|\chi|^{r}<\infty$ for some $r>4$;
(b) $\mathrm{h}\left(c_{n}\right) /\left[c_{n}\left\{1-\mathrm{H}\left(c_{n}\right)\right\}\right]=\mathrm{O}(1)$;
(c) $\mathrm{h}\left(c_{n}-n^{-1 / 4} A\right) / \mathrm{h}\left(c_{n}\right)=\mathrm{O}(1)$ for all $A>0$;
(d) given $\lambda>0$ choose $c_{n}$ so that for all $i$ then $\mathrm{P}\left(\left|\chi_{i}\right|>c_{n}\right)=\lambda / n$ and suppose $n\left\{\mathrm{E} 1_{\left(\left|\chi_{i}\right|>c_{n}\right)} 1_{\left(\left|\chi_{i+1}\right|>c_{n}\right)}\right\} \rightarrow 0 ;$
(ii) the regressors $x_{i}$ satisfy, with $j=1,2$ and $\widehat{\Sigma}_{j n}, \widehat{V}_{j n}$ defined in (23), (24):
(a) $\widehat{\Sigma}_{j n}^{-1}=\mathrm{OP}_{\mathrm{P}}(1),(b) \widehat{V}_{j n}=\mathrm{OP}_{\mathrm{P}}(1),(c) \mathrm{E} \sum_{i \in I_{j}^{\circ}}\left|\nabla x_{i n}\right|^{4}=\mathrm{O}\left(n^{-1}\right)$;
(iii) the sub-sample lengths satisfy $N_{2}^{-1} N_{1}, N_{1}^{-1} N_{2}=\mathrm{O}_{\mathrm{P}}(1)$.

Remark 5.1. Assumption $5.1(i)$ is satisfied when the innovations $\varepsilon_{i}$ are normal. For parts (a)-(c), this follows from Lemma A.14 in the Appendix A.10. For part (d), this follows from Lemma A. 13 .

Theorem 5.2. Suppose Assumption 5.1, that $n_{2} / n \rightarrow \psi$ for $0<\psi<1$ and that the cut-off is chosen through $\mathrm{P}\left(\left|\chi_{i}\right|>c_{n}\right)=\lambda / n$ for all $i$. Then,
(a) $\hat{\Gamma}_{n}^{\text {stylized }}=\sum_{i \in I_{2}^{\circ}} 1_{\left(\left|\chi_{i}\right|>\sigma c_{n}\right)}+\mathrm{OP}(1) \xrightarrow{\mathrm{D}}$ Poisson $(\lambda \psi)$;
(b) $\hat{\Gamma}_{n}^{\text {split }}=\sum_{i=1}^{n-1} 1_{\left(\left|\chi_{i}\right|>\sigma c_{n}\right)}+\mathrm{op}_{\mathrm{P}}(1) \xrightarrow{\mathrm{D}}$ Poisson $(\lambda)$.

The Poisson result shows that the absolute gauge is not consistent for the target $\lambda$. Rather, it has a Poisson variation around the target. The asymptotic Poisson variation depends neither on the regressors nor on the estimation error.

## 6 Power

We consider local power for stylized SIS and for the Andrews (1993) test and argue that the results carry over to the Bai \& Perron (1998) procedure. Proofs are given in Appendix A.11.

### 6.1 Power of stylized SIS

The power properties of the SIS algorithm are discussed by Castle et al. (2015). We give further discussion for the stylized SIS algorithm. For simplicity, we focus on the case without regressors, so the model in (8) reduces to

$$
\begin{equation*}
y_{i}=\sigma \mu 1_{\left(i \leq n_{1}\right)}+\sum_{j=n_{1}+1}^{n} \sigma \delta_{j} 1_{(i \leq j)}+\varepsilon_{i} \quad \text { for } i=1, \ldots, n, \tag{43}
\end{equation*}
$$

with independent, normal $\mathrm{N}\left(0, \sigma^{2}\right)$ innovations and where $\mu$ and $\delta_{j}$ are reparameterized using the scale $\sigma$. This data generating process allows up to $n-n_{1}-1$ breaks.

The stylized SIS Algorithm 2.1 estimates the error variance from the first samplehalf and used forward differences throughout the second sample-half to detect location shifts, see $\S 2.2$. Thus, stylized SIS declares step-shifts for any observation in the second sample half, $n_{1}<i<n$, if

$$
\begin{equation*}
\left|\nabla y_{i}\right| \geq \sqrt{2} \hat{\sigma}_{1} c \tag{44}
\end{equation*}
$$

Theorem 3.4 analyzes the gauge of the procedure. Under normality, we choose the cut-off from the equation $\gamma=2\{1-\Phi(c)\}$, see (21); e.g. $\gamma=1 \%$ corresponds to $c=2.58$.

By the temporal independence, then $y_{i}$ for $i>n_{1}$ is independent of the variance estimator $\hat{\sigma}_{1}^{2}$, which is asymptotically $\sigma^{2} \chi_{n_{1}-1}^{2} /\left(n_{1}-1\right)$-distributed. Assuming also normality, then the t-statistics defined from (44) are non-central t-distributed (Johnson et al., 1993). We note that for an index $i$ in the second sample-half, then (43) can be written as $y_{i}=\sum_{j=i}^{n} \sigma \delta_{j}+\varepsilon_{i}$. Thus, we find with $\chi_{i}=\nabla \varepsilon_{i} /(\sqrt{2} \sigma)$ that

$$
\begin{equation*}
z_{i}=\frac{\nabla y_{i}}{\sqrt{2} \hat{\sigma}_{1}}=\frac{\chi_{i}+\delta_{i} / \sqrt{2}}{\hat{\sigma}_{1} / \sigma} \stackrel{\mathrm{D}}{=} \mathrm{t}_{n_{1}-1}\left(\frac{\delta_{i}}{\sqrt{2}}\right) \quad \text { for } n_{1}<i<n . \tag{45}
\end{equation*}
$$

A single step-shift at time $\tau+1$ of size $\delta$ comes about in model (43) if $\mu=\delta_{\tau}=-\delta$, with $\delta_{n}$ indicating the post-break level, while all other $\delta_{i}$ are zero. If $\chi$ represents a standard normal variable then the power to detect such a shift is

$$
\begin{align*}
\mathrm{P}\left\{\left|z_{i}\right|>c\right\}=\mathrm{P}\{ & \left.\mathrm{t}_{n_{1}-1}(-\delta / \sqrt{2})\right\} \\
& \rightarrow \mathrm{P}(|\chi-\delta / \sqrt{2}|>c)=\Phi(-c+\delta / \sqrt{2})+\Phi(-c-\delta / \sqrt{2}) \tag{46}
\end{align*}
$$

We learn a number of properties from this result. First, the power does not depend on the sign of the shift. Second, the power of the difference decision rule (44) is invariant to time $\tau$. The power stays the same even in the boundaries of the sample. Third, the t-tests are only consistent, i.e. approach unit power, when $|\delta|$ is increasing. Fourth, two decisions are dependent if they concern consecutive time periods. Otherwise, they are independent. Thus, the power is invariant to the number, magnitude, and timing of other shifts as long as they are at least two periods away. SIS can detect shifts, even if their number is large. Fifth, a slight location shift can be detected with high probability if the two episodes are separated by a short period of upheaval. For analytic simplicity, this short period is at least two periods long. Thus, suppose there is one level until $\tau$, a location shift of size $\delta$ at $\tau+1$ followed by an opposite location shift of size $\nu-\delta$ at $\tau+3$, to a new level that is $\nu$ larger than the first level and where $\nu$ may be small. In terms of the model (43) this comes about through $\mu=\delta_{\tau}=-\delta$ and $\delta_{\tau+2}=\delta-\nu$ while $\delta_{n}$ gives the post-break level. The joint probability of correct detection is

$$
\begin{align*}
\mathrm{P}\left\{\left|z_{\tau}\right|>c,\left|z_{\tau+2}\right|>c\right\} \rightarrow & \{\Phi(-c+\delta / \sqrt{2})+\Phi(-c-\delta / \sqrt{2})\} \\
& \times[\Phi\{-c+(\delta-\nu) / \sqrt{2}\}+\Phi\{-c-(\delta-\nu) / \sqrt{2}\}] \tag{47}
\end{align*}
$$

Thus, for large $n$ and small $\nu$, we find

$$
\mathrm{P}\left\{\left|z_{\tau}\right|>c,\left|z_{\tau+2}\right|>c\right\}=\{\Phi(-c+\delta / \sqrt{2})+\Phi(-c-\delta / \sqrt{2})\}^{2}+\mathrm{O}(\nu)
$$

As a consequence, a small location shift can be discovered consistently, if the upheavel $\delta$ is large. Once it has been established that there is, for instance, a shift of this type and no other shifts, it can be tested whether $\nu=0$. This test will be consistent for finite $\nu$.

While the fifth case may seem contrived, it occurs empirically. Castle et al. (2023) find that the UK annual real wage growth rate increases from $0.8 \%$ prior to World War II to $1.7 \%$ after the war, with a large impulse during the war. Similarly, the UK annual productivity per worker increases from $1.2 \%$ prior to World War I to $1.7 \%$ after a huge deflation episode in the wake of the war. Such changes have profound implications for the economy, even if they are small relative to the residual standard error.

### 6.2 Local power for Andrews test

We consider the Andrews (1993) test for a single break at an unknown time in the central part of the sample. This test is consistent for a shift of fixed magnitude that is not at the ends of the sample. We consider local power for various alternatives. The test is based on the simple one-shift model

$$
y_{i}=\sigma \mu+\sigma \delta 1_{(i \leq \tau)}+\varepsilon_{i} \quad \text { for } i=1, \ldots, n,
$$

with independent, normal $\mathrm{N}\left(0, \sigma^{2}\right)$ innovations. If the break point is known, we can form the t-statistic, $Z_{\tau}$ say, for the hypothesis $\delta=0$, see (A.28) for a detailed expression. For the case of an unknown break point, $\tau$, we may suppose $\underline{n} \leq \tau \leq \bar{n}$ for some userchosen bounds satisfying $0<\underline{n} \leq \bar{n}<n$. The likelihood ratio test is then formed by maximizing the squared t -statistic over location. This gives the test statistic

$$
\begin{equation*}
L R_{\max }=\max _{\underline{n} \leq t \leq \bar{n}} Z_{t}^{2} . \tag{48}
\end{equation*}
$$

Distribution under hypothesis. Critical values are found from the distribution of the test statistic under the hypothesis of no break. There are two relevant limits. We note two differences to stylized SIS. On the one hand, the Andrews test asymptotics applies for unknown error distributions while SIS requires a known error distribution. On the other hand, the Andrews test generalizes to the case of stationary regressors, but in contrast to SIS, it does not generalize to the case of non-stationary regressors.

First, when there are no restrictions on the search range, so that $1=\underline{n}$ and $\bar{n}=n-1$, then the likelihood ratio statistic diverges at a rate of $2 \log \log n$ due to the behavior of a Brownian motion near the origin as described by the law of iterated logarithms. With an appropriate logarithmic normalization, the statistic converges to an extreme value distribution (Yao \& Davis, 1986; Hidalgo \& Seo, 2013). This test is not so common. Perhaps because it is felt that too much power is lost by the additional normalization.

Second, when the search range is trimmed, the likelihood ratio statistic converges to a supremum of a standardized Brownian bridge (Andrews, 1993). That is, if $\mathbb{B}_{u}$ is a standard Brownian bridge for $0 \leq u \leq 1$, which has variance $u(1-u)$, then for large $n$ and with $\underline{n} / n \rightarrow \underline{\lambda}>0$ and $\bar{n} / n \rightarrow \bar{\lambda}<1$, we get

$$
L R_{\max }=\max _{\underline{n \leq t \leq \bar{n}}} Z_{t}^{2} \xrightarrow{\mathrm{D}} \sup _{\underline{\lambda} \leq u \leq \bar{\lambda}} \frac{\mathbb{B}_{u}^{2}}{u(1-u)} .
$$

The critical values increase with decreasing trimming, reaching the extreme value asymptotics when there is no trimming. Andrews provided simulated critical values. A 15\% trimming is commonly used with critical value 12.35 for a $1 \%$ sized test. Bai \& Perron (1998) preferred $5 \%$ trimming. The test is known to be consistent for a central break of finite magnitude $\delta$. This contrasts with SIS. We investigate local power in various cases.
$\boldsymbol{A}$ single break. We consider the power against an alternative with a shift of vanishing magnitude at time $\tau=\lambda n$. We allow $0<\lambda<1$, while noting that the Andrews test is aimed at the trimmed interval $0<\underline{\lambda} \leq \lambda \leq \bar{\lambda}<1$. Local power is found when the magnitude of the break vanishes as $\delta=\phi / \sqrt{n}$ for fixed $\phi$. We find in Appendix A. 11 that, for fixed $0<\lambda<1$,

$$
\begin{equation*}
L R_{\max } \xrightarrow{\mathrm{D}} \sup _{\underline{\lambda} u \leq \bar{\lambda}} \frac{\left(\mathbb{B}_{u}+\phi s_{u}^{\lambda}\right)^{2}}{u(1-u)}, \tag{49}
\end{equation*}
$$

where the function $s_{u}^{\lambda}$ increases linearly from 0 at $u=0$ to $\lambda(1-\lambda)$ at $u=\lambda$ after which it decreases linearly to 0 at $u=1$ as given by

$$
\begin{equation*}
s_{u}^{\lambda}=(1-\lambda) u 1_{(u \leq \lambda)}+\lambda(1-u) 1_{(u>\lambda)} . \tag{50}
\end{equation*}
$$

The non-centrality term is largest for $u=\lambda$, taking the value $\phi\{\lambda(1-\lambda)\}^{1 / 2}$. Thus, the Andrews test has local power for this alternative, whereas asymptotically, stylized SIS has trivial power. For a finite sample, we compare the maximal pointwise noncentrality for the Andrews test of $\phi\{\lambda(1-\lambda)\}^{1 / 2}=\delta\{n \lambda(1-\lambda)\}^{1 / 2}$ with the SIS noncentrality of $\delta / \sqrt{2}$ arising from (45). Notably, the magnitude of the break $\delta$ will give neither method an advantage in the power comparison. Instead, the positioning $\lambda$ and the sample size $n$ determine the comparative performance. We compare the two noncentralities, while ignoring the simultaneity of decisions within the two procedures. The Andrews test with $15 \%$ trimming and $1 \%$ size has critical value $12.35=(3.51)^{2}$, while stylized SIS has $1 \%$ critical value $2.57=(6.63)^{1 / 2}$. Dividing the non-centralities with 3.51 and 2.57, respectively, equating and solving gives $n=(12.35 / 6.63) /\{2 \lambda(1-\lambda)\}$, with SIS being advantageous for $n$ smaller than those values. The implied $n$-values for central values $\lambda=(0.5,0.75,0.85)$ are $n=(4,5,7)$ so that the Andrews test is favourable. However, this changes when the break occurs in the trimmed period. The largest $u$ considered by the test statistic is $\bar{\lambda}$, so that the Andrews test has maximal pointwise non-centrality of $\delta\{n \bar{\lambda} /(1-\bar{\lambda})\}^{1 / 2}(1-\lambda)$. Proceeding as before, we find $n=(12.35 / 6.63)(1 / 2)\{(1-\bar{\lambda}) / \bar{\lambda}\} /(1-\lambda)^{2}$. Thus, for $\bar{\lambda}=0.85$, the implied $n$-values for $\lambda=(0.9,0.95,0.99)$ are $n=(16,66,1650)$. The comparison indicates that stylized SIS may be competitive in small samples with a late break.

Next, consider the consequence of a break close to the sample boundaries. The above derivation can be modified to the case where $\delta(1-\tau / n)=\psi / \sqrt{n}$ while $\tau / n \rightarrow 1$ and fixed $\psi$. These constraints imply $|\delta| / \sqrt{n} \leq|\psi|$ with equality when $\tau=n-1$. Thus, we let $\delta / \sqrt{n} \rightarrow \eta$ where $0 \leq|\eta| \leq|\psi|$ while $\eta \psi \geq 0$. For large $n$, we get

$$
\begin{equation*}
L R_{\max } \xrightarrow{\mathrm{D}} \sup _{\underline{\lambda} \leq u \leq \bar{\lambda}} \frac{\left(\mathbb{B}_{u}+\psi u\right)^{2}}{u(1-u)(1+\eta \psi)} \tag{51}
\end{equation*}
$$

To see that (51) conforms with (49), note that $u \leq \bar{\lambda}<1$ and $\tau / n \rightarrow 1$ imply that $u<\tau / n$ so that $s_{u}^{\tau / n}=(1-\tau / n) u$ for large $n$, while a small $\delta$ corresponds to $\eta=0$.

The result (51) shows that when $\delta$ diverges, then the Andrews test has local power, while the stylized SIS is consistent, see (45). In particular, we can let $\delta$ diverge a slow rate with $\tau$ sufficiently close to $n$ to achieve $\psi=0$, so that the Andrews test has trivial power, while stylized SIS is consistent.

Two breaks. Let $y_{i}=\sigma \mu+\sigma \delta_{1} 1_{\left(i \leq \tau_{1}\right)}+\sigma \delta_{2} 1_{\left(i \leq \tau_{2}\right)}+\varepsilon_{i}$ where $\varepsilon_{i}$ is i.i.d. $\mathrm{N}\left(0, \sigma^{2}\right)$ so that the level is changed twice at $\tau_{1}<\tau_{2}$. Again, this alternative is outside those the Andrews test is optimized against, but relevant in practice. We consider the situation where two large location shifts are close and nearly offset each other so that $\tau_{2}-\tau_{1}$ and $\delta_{1}+\delta_{2}$ are close to zero. This is an empirically relevant situation where SIS performs well. Thus, we investigate local power when $\delta_{1}+\delta_{2}=\xi / \sqrt{n}$ and $\delta_{2}\left(\tau_{2}-\tau_{1}\right)=\psi \sqrt{n}$ while $\tau_{1} / n=\lambda$ and $\left(\tau_{2}-\tau_{1}\right) / n \rightarrow 0$ for fixed $\xi, \psi, \lambda$. These constraints imply $\left|\delta_{2}\right| / \sqrt{n} \leq|\psi|$ with equality when $\tau_{2}=\tau_{1}+1$. Thus, we let $\delta_{2} / \sqrt{n} \rightarrow \eta$ where $0 \leq|\eta| \leq|\psi|$ while $\eta \psi \geq 0$. We find in Appendix A. 11 using the Skorokhod (1956) $M_{1}$-metric that

$$
\begin{equation*}
L R_{\max } \xrightarrow{\mathrm{D}} \sup _{\underline{\lambda} \leq u \leq \bar{\lambda}} \frac{\left[\mathbb{B}_{u}+\xi s_{u}^{\lambda}+\psi\left\{1_{(u \geq \lambda)}-u\right\}\right]^{2}}{u(1-u)(1+\eta \psi)} . \tag{52}
\end{equation*}
$$

Again, when $\delta_{2}$ and hence $\delta_{1}$ diverge, then the Andrews test has local power while splithalf SIS is consistent. In particular, when $\delta_{2}$ diverges slowly while $\tau_{2}$ and $\tau_{1}$ are so close that $\psi=0$, then the Andrews test has trivial power while stylized SIS is consistent.

### 6.3 Discussion of Bai and Perron procedure

We first summarize the findings for the Andrews test. This test is consistent for a fixed-sized central break in contrast with stylized SIS which only has local power in that situation. Otherwise, SIS can be competitive. We found that SIS is consistent, while the Andrews test has trivial power in two situations. The first case has a break near the end point of the sample. Detecting such a break is highly relevant when forecasting (Clements \& Hendry, 1998). The second case is when two breaks are close and nearly offsetting. This can reveal small but important changes in, for instance, growth series (Castle et al., 2023). Thus, the Andrews test is preferable if one is content that there is only one central break or perhaps two well-separated central breaks. With more complicated series, SIS will be competitive.

The Bai \& Perron (1998) (BP) procedure is developed for the situation where there is an unknown, but bounded, number of multiple well-separated breaks. This procedure provides estimates of the number of breaks and their timing. This requires trimming between breaks and at the end points of the sample and a maximal number of breaks. The usual $15 \%$ trimming eliminates too much of the sample and a $5 \%$ trimming is recommended. The above analysis suggests that the BP procedure will consistently detect fixed-sized breaks that are not too close. But, with many breaks or with close breaks, the BP procedure may have near trivial power, while SIS could have high power.

As a further point of comparison, we note that the BP procedure allows an unknown error distribution and it generalizes to stationary, but not non-stationary regressors. The SIS procedure requires a known error distribution, but allows both stationary and non-stationary regressors. We note that for many macro-economic time series, normality is not unreasonable, but assuming stationarity of the regressors may not be appropriate.


Figure 1: Finite sample properties of the frequence gauge

Finally, the general SIS algorithm is designed to work jointly with regression selection, whereas the BP procedure requires a fixed set of regressors.

## 7 Simulations \& Numerical Approximations

We complement the asymptotic analysis of split-half SIS with simulations and numerical approximations. These results confirm the validity of the asymptotic theory, allow comparisons to other algorithms, and inform us about the small sample properties of SIS. First, we confirm the consistency of the frequence gauge and characterize its small sample bias. Second, we use numerical approximations to decompose the components of the asymptotic variance. Third, we confirm with simulations the distributional convergence of the frequence gauge. Fourth, we consider the bias of an updated regression estimator. Fifth, we compare the power of split-half SIS with the Andrews (1993).

All simulations have $10^{4}$ repetitions. Each time we increase the sample size, we redraw all $n$ observations. The simulations have been coded in MATLAB using the MFE toolbox (Sheppard, 2018). When we do not explicitly mention otherwise, we set $\hat{\omega}_{j, i}^{2}=1$ for simplification, as we are mainly concerned with evaluating the asymptotic distributions. Given a target frequence gauge $\gamma$, we choose the cut-off $c$ in the SIS algorithm as the normal $(1-\gamma / 2)$ quantile.

### 7.1 Analysis of consistency of frequence gauge

We validate the consistency of the frequence gauge of split-half SIS as analyzed in Theorem 4.2. We consider two data generating processes. In both cases, the algorithm is based on the model (7) with one univariate regressor $x_{i}$ and $n_{1}=n_{2}$.

DGP1 includes an exogenous regressor $y_{i}=\beta x_{i}+\varepsilon_{i}$, so that $x_{i}$ and $\varepsilon_{i}$ are independent standard normal. The DGP1 is white noise, if $\beta=0$, in which case $y_{i}$ is also independent standard normal. As the regressor $x_{i}$ is strictly exogenous Theorem 4.5 applies.

DGP2 is a first-order auto-regression $y_{i}=\alpha y_{i-1}+\varepsilon_{i}$, where $|\alpha|<1, \varepsilon_{i}$ is independent standard normal and $y_{0}=0$. Thus, $\delta_{j}=0$ for all $j$ and $\beta=\alpha$ in (1) while $x_{i}=y_{i-1}$.


Figure 2: Analysis of the asymptotic variance of the frequence gauge for varying $c$

Figure 1(a) uses DGP1 with exogenous regressor and coefficient $\beta=0$. It shows the frequence gauge $\gamma$ for an increasing sample size and different gauges $\gamma$. We use the white noise version of DGP1. We find that the small sample bias of the gauge is positive. The bias vanishes quickly with growing samples and it is modest for $n=100$.

Figure 1(b) uses the autoregressive DGP2. It considers different values of the firstorder autoregressive coefficient $\alpha$ for two sample sizes $n=100$ and 1,000 . We also consider the effect of including the weights $\omega_{j, i}^{2}$. For constant $n$, the small sample bias appears to decrease for increasing $\alpha$. This could reflect that as the autoregressive coefficient $\alpha$ increases, the sample correlation between the retained step-indicators and the autoregressive process increases. Consistent with theory, the small sample bias vanishes asymptotically. The rescaling of the estimated variance using the forward correction factors $\omega_{j, i}^{2}$ reduces the small sample bias by about one-third.

### 7.2 Analysis of asymptotic distribution of frequence gauge

We decompose the asymptotic variance of the frequence gauge of split-half SIS as a function of the cut-off $c$ to understand the contributions of the various terms, and compare the variance to IIS. We continue to use DGP1 and DGP2.

Figure 2(a) presents a decomposition of the individual terms of the asymptotic variance of the gauge as functions of the cut-off $c$ for the autoregressive DGP2 as given by Theorem 4.4 and Example 4.3. The terms that do not depend on estimation errors are $(1-\gamma) \gamma$ and $2\left(\varsigma_{1}-\psi^{2}\right)$; the terms that depend on the scale estimation error are $-4 c h(c) \varsigma_{2}$ and $2 c^{2} \mathrm{~h}(c)^{2}$ and the terms that depend on the location estimation error are $c^{2} h^{2}(c)(1-\alpha)$ and $-2 c h(c) \varsigma_{2}(1-\alpha)$. Some terms increase the asymptotic variance one of the location terms and one of the scale terms decrease the asymptotic variance.

Figure 2(b) compares the asymptotic variance of the gauge of the split-half ImpulseIndicator Saturation (IIS) to split-half SIS.
(Johansen \& Nielsen, 2016b, Corollary 5) gives the asymptotic distribution of the IIS gauge as

$$
\begin{equation*}
n^{1 / 2}\left\{\hat{\gamma}_{n}^{\mathrm{IIS}}(c)-\gamma\right\} \xrightarrow{\mathrm{D}} \mathrm{~N}\left\{0, \gamma(1-\gamma)+2 c \mathrm{~h}(c) \tilde{\varkappa}_{2}+2 c^{2} \mathrm{~h}^{2}(c)\right\}, \tag{53}
\end{equation*}
$$

where $\tilde{\varkappa}_{2}=\int_{-c}^{c}\left(u^{2}-1\right) \mathfrak{f}(u) d u$ is a truncated moment. Figure 2(b) displays the different asymptotic variance curves of the gauge as functions of the cut-off $c$ for IIS and SIS for different DGPs. For IIS, we consider white noise DGP1. For SIS, we first consider the same DGP1, and second consider the autoregressive DGP2 with $\alpha=0.5$ and $\alpha=0.9$. Finally, we reconsider DGP1, but assume the error variance is known, so that $\hat{\omega}_{j, i}^{2}=$ $\sigma^{2}=1$ and two components of the asymptotic variance become zero.

We make the following observations. First, for all $c$, the asymptotic variance of the gauge in IIS is lower than for all four competing SIS models. Second, running SIS knowing the variance $\sigma^{2}$ results in a higher asymptotic variance of the frequence gauge. Third, in the autoregressive model, the $\alpha$ coefficient changes the asymptotic variance. The asymptotic variance is larger with $\alpha=0.9$ than $\alpha=0.5$. This is different from IIS, where the asymptotic variance does not include regressor-dependant terms. Finally, we observe that the asymptotic variance of the gauge falls rapidly for growing $c$. This motivates the choice of a large $c$ in empirical applications, corresponding to a gauge of $1 \%$ or lower, as recommended by Castle et al. (2015).

|  | $\gamma$ vs. $n$ | 100 | 400 | 1600 | $\infty$ |
| :--- | :--- | :---: | :---: | :---: | :---: |
| DGP1 | $5 \%$ | 0.0516 | 0.0399 | 0.0363 | 0.0347 |
| $\beta=0$ | $1 \%$ | 0.0160 | 0.0104 | 0.0094 | 0.0089 |
| $\hat{\omega}_{j, i}^{2}=1$ | $0.5 \%$ | 0.0093 | 0.0057 | 0.0051 | 0.0047 |
|  | $0.1 \%$ | 0.0025 | 0.0013 | 0.0011 | 0.0010 |
| DGP2 | $5 \%$ | 0.0411 | 0.0284 | 0.0261 | 0.0249 |
| $\alpha=0.5$ | $1 \%$ | 0.0149 | 0.0094 | 0.0085 | 0.0079 |
| $\hat{\omega}_{j, i}^{2}=1$ | $0.5 \%$ | 0.0089 | 0.0056 | 0.0044 | 0.0044 |
|  | $0.1 \%$ | 0.0024 | 0.0013 | 0.0011 | 0.0010 |
| DGP2 | $5 \%$ | 0.0425 | 0.0348 | 0.0331 | 0.0323 |
| $\alpha=0.9$ | $1 \%$ | 0.0134 | 0.0097 | 0.0087 | 0.0086 |
| $\hat{\omega}_{j, i}^{2}=1$ | $0.5 \%$ | 0.0075 | 0.0052 | 0.0049 | 0.0046 |
|  | $0.1 \%$ | 0.0019 | 0.0012 | 0.0010 | 0.0010 |
| DGP1 | $5 \%$ | 0.0649 | 0.0627 | 0.0606 | 0.0610 |
| $\beta=0$ | $1 \%$ | 0.0132 | 0.0124 | 0.0118 | 0.0117 |
| $\hat{\sigma}_{j}^{2}=\sigma^{2}$ | $0.5 \%$ | 0.0063 | 0.0060 | 0.0057 | 0.0057 |
|  | $0.1 \%$ | 0.0013 | 0.0011 | 0.0011 | 0.0010 |

Table 1: Simulated and asymptotic variance of the frequence gauge of split-half SIS

### 7.3 Analysis of distribution convergence of frequence gauge

We now verify the asymptotic distribution results of the frequence gauge of split-half SIS and evaluate small sample properties. Table 1 tabulates the simulated variance and computed asymptotic variance of the frequence gauge of split-half SIS for the target gauges $\gamma=5 \%, 1 \%, 0.5 \%$, and $0.1 \%$ and sample sizes $n=100,400$, and 1600 . We consider the same models for split-half SIS as in Figure 2(b). Overall, the finite sample variance is quite close to the asymptotic variance when $n=400$ and not too bad when $n=100$. Our findings are consistent with the results in Figure 2.


Figure 3: Bias of updated regression estimator as function of sample size

### 7.4 Updating estimation of regression coefficients

In this section, we use simulation to show that split-half SIS can introduce a bias when updating the estimates for $\beta$ in (1). We conjecture that this bias can persist asymptotically with a fixed frequence gauge.

Suppose the split-half SIS Algorithm 2.1 is applied to data generated from an autoregressive model $y_{i}=\alpha y_{i-1}+\varepsilon_{i}$. This may result in $m-1$ level shifts at locations $\tau_{0}=0<\tau_{1}<\cdots<\tau_{m-1}<\tau_{m}=n$. We update the $\alpha$ estimate by the regression

$$
\begin{equation*}
y_{i}=\mu_{i}+\alpha y_{i-1}+u_{i} \quad \text { for } \tau_{j-1}<i \leq \tau_{j} \text { and } j=1, \ldots, m \tag{54}
\end{equation*}
$$

With a frequence gauge of $\gamma$ we will have approximately $m \approx \gamma n$ breaks so that the sub-sample lengths are approximately $n / m \approx 1 / \gamma$. Thus, estimation of (54) corresponds to estimation of an unbalanced dynamic panel model, with a (random) increasing crosssectional dimension and a (random) finite time dimension. It seems like we are faced with the same issues as in panel data of a incidental parameter problem (Lancaster, 2000,2002 ) and a correlation of the retained (random) step-indicators with the dynamic regressors (Arellano \& Bond, 1991). As with panel data, we would expect the bias to disappear asymptotically in a model with strictly exogenous regressors.

Figure 3 shows simulated biases of the updated estimator of the regression coefficients as a function of sample length $n$ for different frequence gauges. Panel (a) uses the autoregressive DGP2 with $\alpha=0.5$. As a baseline, we estimate the $\operatorname{AR}(1)$ model without split-half SIS. This shows the well-known negative finite sample bias that disappears asymptotically (Marriott \& Pope, 1954). Then we use split-half SIS with the frequence gauge at $1 \%$ (green), $5 \%$ (black), and $10 \%$ (blue). We find that a larger frequence gauge is associated with a larger bias that does not appear to vanish asymptotically. When we repeat this exercise in Panel b for exogenous regressors, we find that the bias is an order of magnitude smaller than before.

Figure 4 uses the autoregressive DGP2 and shows simulated biases as a function of the autoregressive coefficient $\alpha$ when the sample size is $n=1,000$. Both panels use a standard autoregressive estimation without SIS as a benchmark along, with split-half SIS estimation results. The frequence gauge is $10 \%$ in panel (a) and $1 \%$ in panel (b).


Figure 4: Bias of updated regression estimator as function of autoregressive coefficient

We find a bias across all values of $\alpha$, and it grows together with the value of $\alpha$. The bias is much larger with the frequence gauge at $10 \%$ than at $1 \%$.

Overall, the simulations provide evidence towards the presence of an incidental parameter bias when applying SIS with dynamic regressors and calibrated through the frequence gauge. The bias increases with gauge and with the autoregressive coefficients.

### 7.5 Analysis of power

We compare the power of split-half SIS and the Andrews (1993) test. We consider a one-shift data generating process with a view to validate the asymptotic theory in (45) for SIS and (49) and (51) for the Andrews test.

DGP3 has one location shift and is given by

$$
\begin{equation*}
y_{i}=\alpha y_{i-1}+\delta 1_{(i \geq \lambda n)}+\varepsilon_{i} \quad \text { for } i=1, \ldots, n, \tag{55}
\end{equation*}
$$

with independent standard normal innovations. We will vary $\alpha, \delta, \lambda$ and $n$.
We subject the model (55) to split-half SIS and the Andrews test. For SIS, we use a $1 \%$ gauge and compute the retention frequency for the indicator at $\lambda n$. The Andrews F test for detecting a single location shift with $15 \%$ trimming has a $1 \%$ critical value of 12.35. We report the power for the (maximum) test.

Table 2 shows the simulation results. The magnitude $\delta$ of the location shift is explored along columns. The location $\lambda$ is explored along rows. Panels 1 and 2 consider a non-dynamic process $\alpha=0$ for $n=100$ and 66 . Panel 3 considers a dynamic process $\alpha=0.5$ for $n=66$. The value 66 is chosen to find the $\delta$ where Andrews and SIS have equal power for $\lambda=0.95$ as discussed in theory Section 6 .

The columns marked $\delta=0$ show the finite sample size and frequence gauge. We notice that the Andrews size is always larger than the SIS gauge. We note that the distortion is larger for $\alpha=0.5$ than for $\alpha=0$. The power simulations are not size corrected and are therefore favourable to the Andrews test.

The theory suggests that the power increases with $\delta$. We see that the SIS power is always increasing in $\delta$. The Andrews power is also increasing in $\delta=0,2$, and 4 , but it

|  | $\delta=0$ |  |  |  | $\delta=2$ |  | $\delta=4$ |  | $\delta=8$ |  |
| :--- | :---: | ---: | ---: | ---: | ---: | ---: | ---: | ---: | ---: | :---: |
|  | $\lambda$ | A | SIS | A | SIS | A | SIS | A | SIS |  |
| $n=100$ | 0.90 | $1.3 \%$ | $1.0 \%$ | $88.6 \%$ | $12.0 \%$ | $100.0 \%$ | $57.0 \%$ | $100.0 \%$ | $99.9 \%$ |  |
| $\alpha=0$ | 0.95 | $1.1 \%$ | $1.1 \%$ | $19.2 \%$ | $11.9 \%$ | $51.7 \%$ | $58.3 \%$ | $39.5 \%$ | $99.8 \%$ |  |
|  | 0.99 | $1.1 \%$ | $1.2 \%$ | $2.1 \%$ | $11.3 \%$ | $3.4 \%$ | $58.1 \%$ | $0.6 \%$ | $99.9 \%$ |  |
| $n=66$ | 0.90 | $1.4 \%$ | $1.1 \%$ | $76.5 \%$ | $12.8 \%$ | $99.9 \%$ | $58.4 \%$ | $100.0 \%$ | $99.8 \%$ |  |
| $\alpha=0$ | 0.95 | $1.2 \%$ | $1.1 \%$ | $11.0 \%$ | $12.9 \%$ | $24.9 \%$ | $57.1 \%$ | $13.0 \%$ | $99.7 \%$ |  |
|  | 0.99 | $1.3 \%$ | $1.2 \%$ | $3.0 \%$ | $12.3 \%$ | $3.7 \%$ | $58.2 \%$ | $0.6 \%$ | $99.7 \%$ |  |
| $n=66$ | 0.90 | $2.3 \%$ | $0.4 \%$ | $19.6 \%$ | $8.5 \%$ | $74.0 \%$ | $55.6 \%$ | $99.8 \%$ | $99.9 \%$ |  |
| $\alpha=0.5$ | 0.95 | $2.4 \%$ | $0.4 \%$ | $4.3 \%$ | $8.3 \%$ | $7.4 \%$ | $56.0 \%$ | $6.4 \%$ | $99.9 \%$ |  |
|  | 0.99 | $2.3 \%$ | $0.3 \%$ | $2.7 \%$ | $8.8 \%$ | $2.6 \%$ | $56.4 \%$ | $0.8 \%$ | $99.9 \%$ |  |

Table 2: Simulated power for the Andrews (A) test and split-half SIS
declines at $\delta=8$ for $\lambda=0.95$ and 0.99 . For $\lambda=0.99$, it even dips below the size. This may be a finite sample effect.

The theory suggests that the power of split-half SIS is invariant to the location $\lambda$, whereas the the power of the Andrews test declines as $\lambda$ approaches unity. This is confirmed in the simulations.

Further, the theory suggests that the Andrews test has higher power than split-half SIS when $\lambda$ is away from 1 while SIS is more powerful for $\lambda$ is close to zero. Indeed, simulations are in favour of the Andrews test for $\lambda=0.9$ and in favour of SIS for $\lambda=0.99$. For the inbetween case $\lambda=0.95$, the results are mixed with SIS being more powerful except in the first panel with $n=100$ for $\delta=2$.

Finally, we see that the power declines with increasing temporal persistency by looking at the panels 2 and 3 where $n=66$, but the autoregressive coefficient is $\alpha=0$ and $\alpha=0.5$, respectively. There is an indication that the decline in performance is larger for the Andrews test than for SIS.

## 8 Empirical illustration

As an empirical example on the use of stylized SIS, consider the log UK labor productivity, $y_{i}$, from the first quarter of 1980 to the third quarter of 2021. This gives a sample of length of $n=167$ plus initial values. The labor productively is measured by the UK's Office of National Statistics as a chain volume measure of gross value added at basic prices divided by the number of hours worked. We used PcGive in OxMetrics 8 for the analysis (Doornik \& Hendry, 2013).

Figure 5(a) shows the log labor productivity $y$ with a marked decline in its growth rate after the 2008 financial crisis. There is considerable movement through the Covid pandemic from 2020. The post-2008 decline has been of concern in the political debate for some years, see for example Chadha (2022), and the submission to the Treasury Committee in October 2021 by the Bank of England's Chief Economist Huw Pill:
"Before the global financial crisis, UK productivity growth averaged over two per cent per year. Since then, labor productivity (growth) has fallen


Figure 5: UK labor productivity
considerably."
Panel (b) shows the labor productivity growth rate measured as the log difference $\Delta y_{i}=$ $y_{i}-y_{i-1}$. Note that the y -axis has been truncated to better visualize the pre-Covid periods. We make the following observations. The series is very noisy, and one can just about visually discern a gradual decline over time. We will model the growth rate as a first-order autoregression, thus imposing that the series in levels has a unit root. We will show how SIS can help in capturing the declining level of the growth rate.

We start by fitting a first-order autoregression to the growth rate for the whole period. While not reported here, the results point to a very mis-specified model, and diagnostics point to difficulties matching movements through the Covid period. An investigator may, therefore, drop that period and focus on the period until 2019:4. We then find the model:

$$
\begin{align*}
\widehat{\Delta y}_{( } & =\underset{(0.079)}{0.104 \Delta y_{i-1}}+\underset{(0.0007)}{0.0035}  \tag{56}\\
\widehat{\sigma} & =0.0069, \quad n=160, \quad R S S=0.0074,  \tag{57}\\
\chi_{\text {norm }}^{2}[2] & =5.00(p=0.082), \quad \mathrm{F}_{\operatorname{ar}(1-5)}[5,153]=1.96(p=0.088)  \tag{58}\\
\max C^{2} & =8.49(p=0.482)\{\arg \max =2008: 3\} .  \tag{59}\\
\max F & =3.52(p=0.01)\{\arg \max =2004: 1\} . \tag{60}
\end{align*}
$$

The fitted model reported in (56) and in Figure 5(d). The fit indicates an overall constant level for the quarterly growth rate of $0.0035 /(1-0.104)=0.39 \%$.

We subjected the model (to 2019:4) to various misspecification tests. These do not tend to reject the model. A normality test based on cumulants (Doornik \& Hansen, 2008) and a test for residual autocorrelation (Godfrey, 1978; Nielsen, 2006) are reported in (58). Figure 5(c) shows a one-step recursive Chow test with pointwise $1 \%$ critical
values. This indicates a slight rejection in 2008:3, but the practitioner may not wish to give too much attention to this, given that about 144 tests were conducted (Hendry \& Nielsen, 2007). Indeed, a joint test as shown in (59) does not reject the model (Nielsen \& Whitby, 2015). The Andrews test reported in (60), used for detecting a single location shift, gives a marginal decision indicating a possible break in 2004:1. It appears that minor location-shifts are not reliably detected by conventional misspecification tests. Yet, Figure 5(a) does show a marked decline in the log labor productivity $y_{i}$ since 2008.

We now apply the stylized SIS algorithm to the full sample until 2021:3. First, we fit the first-order autoregression to the first sample-half until 1999:4. This is the same as fitting the autoregression to the full sample combined with step-indicators for each observation from 2000:1 to 2021:3. We get

$$
\begin{align*}
\widehat{(s e)} & =\underset{(0.111)}{0.201 \Delta y_{i-1}}+\underset{(0.0010)}{0.0045}+\sum_{j=81}^{167} \widehat{\delta}_{j} 1_{(i \geq j)}  \tag{61}\\
\widehat{\sigma} & =0.0068, \quad n=167, \quad R S S=0.0036,  \tag{62}\\
\chi_{\text {norm }}^{2}[2] & =3.78(p=0.151), \quad \mathrm{F}_{a r(1-5)}[5,73]=1.39(p=0.237) . \tag{63}
\end{align*}
$$

This fit indicates a constant quarterly growth rate of $0.0045 /(1-0.201)=0.56 \%$ prior to 2000. Test for normality and residual autocorrelation do not reject, see (63).

There are 87 estimated coefficients for the step-indicators. Computing the t-statistics for these 87 estimates, we find that the most extreme t-statistics are: 10.4 for 2020:3, -9.57 for $2020: 4,4.28$ for 2021:1, -3.21 for 2000:2, -2.69 for 2008:3, 2.65 for 2016:1, 2.53 for $2000: 1,2.17$ for $2004: 2$ and 2.00 for 2008:2. Using the $1 \%$ cut-off for the normal distribution of 2.576 , we keep the six most significant step-indicators. Rerunning the model gives

$$
\begin{align*}
\widehat{\Delta y}_{(s e)}= & -\underset{(0.074)}{0.008} \Delta y_{i-1}+\underset{(0.0009)}{0.0056} \\
& +\underset{(0.0068)}{0.0183 I_{(i \geq 00: 1)}-\underset{(0.0069)}{0.0202 I_{(i \geq 00: 2)}}-\underset{(0.0015)}{0.0033 I_{(i \geq 08: 3)}}} \\
& +\underset{(0.007)}{0.080 I_{(i \geq 20: 3)}-\underset{(0.012)}{0.119 I_{(i \geq 20: 4)}}+\underset{(0.010)}{0.036 I_{(i \geq 21: 1)}}}  \tag{64}\\
\widehat{\sigma}= & 0.0068, \quad n=167, \quad R S S=0.0072,  \tag{65}\\
\chi_{\text {norm }}^{2}[2]= & 5.39(p=0.068), \quad \mathrm{F}_{a r(1-5)}[5,154]=2.39(p=0.041) . \tag{66}
\end{align*}
$$

The autoregressive coefficient is now insignificant. Adding up the constant terms and correcting for the modest autoregressive coefficient gives long-run means of $0.56 \%$ prior to 2000 , then $0.37 \%$ until 2008 , then $0.043 \%$ until 2020.

We identify two significant drops in productivity in 2000 and 2008, corresponding to the burst of the dot-com bubble and the financial crises, respectively. Both are characterized by pairs of offsetting step indicators. However, the split half-SIS retains only one of the two step-indicators from 2008, which results in less accurate tracking of the series during the financial crisis.

The more comprehensive SIS algorithm in OxMetrics yields a similar model to splithalf SIS, but it manages to retain two offsetting step-indicators for 2008 instead of just one. In the updated OxMetrics model, these indicators have larger t-statistics than the
single 2008 indicator found in (64). It appears that the split-half SIS is too simple to track the somewhat protracted upheaval during the financial crisis.

## 9 Conclusion

In this paper, we investigated the properties of the SIS algorithm that addresses location shifts in time series in the context of model selection. The growing importance of SIS in tackling location shifts is reflected in its applications in fields as varied as economics (Chuffart \& Hooper, 2019; Pellini, 2021; Bernstein \& Martinez, 2021), climate science (Raggad, 2018; Pretis et al., 2018; Koch et al., 2022; O'Callaghan et al., 2022), and public health (Doornik et al., 2022). In this section, we summarize the insights gained through a study of SIS with asymptotic analysis, simulations, and numerical approximations.

The first insight is that the frequence gauge is consistent for a wide range of both stationary and non-stationary regressors. This means that even without detailed knowledge of the regressor types, an investigator can choose the cut-off of SIS from the limiting gauge. To address the sensitivity of this result, we demonstrated that the variation of the frequence gauge around its limit follows a normal distribution. However, its variance depends on the type of regressors. Simulations revealed that this variation remains limited, even in small samples. As a result, the sole tuning parameter of the SIS algorithm can be finely adjusted to align with the investigator's preferences.

The second insight concerns the link between the frequence gauge and the bias in the updated regression estimator after selecting over step-indicators. This bias appears to emerge in the presence of dynamic regressors when searching for location shifts. This contrasts with the theory of Impulse Indicator Saturation, where there is no such bias (Johansen \& Nielsen, 2016b). The bias diminishes as the gauge decreases, suggesting that the gauge should be chosen small and possibly vanishing with sample size. For that purpose, we developed a Poisson theory for the absolute gauge. For a sample size of $n=100$ observations, we recommend setting the absolute gauge to 1 , which is equal to the frequence gauge of $1 \%$, in line with Castle et al. (2015). In larger samples, we advise targeting the absolute gauge rather than the frequence gauge, so that the cut-off drifts slowly to infinity.

The third insight pertains to the circumstances in which stylized SIS demonstrates higher statistical power compared to the Andrews (1993) test. We developed a local power theory for stylized SIS and the Andrews test. Our findings suggest that the Andrews test maintains consistency when faced with one or two well-separated, central location shifts, whereas the SIS shows trivial power. Conversely, for location shifts near the end of the sample or for two offsetting location shifts close to each other, the SIS maintains power, while the power of the Andrews test goes down to its size. In time series observed over extended periods, major upheavals like the 2008 financial crisis and the 2020 Covid pandemic might recur. Consequently, we anticipate multiple breaks in the data. These breaks may occur closely together or towards the end of the sample. In such scenarios, SIS appears to be preferable to the Andrews test. The same conclusions hold for the Bai \& Perron (1998) procedure that allows more breaks but inherits the power trade-offs from the Andrews test.

The fourth insight relates to the regularity conditions of SIS compared to the An-
drews test. SIS assumes a known error distribution, while the Andrews test does not. The assumption is testable and contributes to the power of SIS to detect breaks that occur closely together. The first-order theory for SIS applies to a variety of stationary and non-stationary regressors. In order to do this, the present theory is formulated in terms of normalization matrices. This implies that the theory works regardless of the choice of the normalization matrix. In contrast, the asymptotic theory for the Andrews test requires stationary regressors, introducing an additional risk of mistakes, as the investigator must carefully determine the appropriate normalization of the regressors. Furthermore, SIS is designed to be implemented along with regressor selection, which is useful when there is uncertainty about the choice of regressors.

The theory for SIS is complicated because SIS operates on the differenced residuals which are temporally dependent even for well-behaved errors. We found various technical solutions that may be useful elsewhere. The empirical process theory was developed using ideas from the McLeish (1977) mixingale theory. The Poisson theory requires the Chen (1975) Poisson limit theorem for dependent binary variables. In addition, to allow two close breaks in the power theory, we relied on the Skorokhod (1956) $M_{1}$-metric favoured by Whitt (2002) rather than the $J_{1}$-metric favoured by Billingsley (1968).

A potential further development is to develop a test for the presence of location shifts along the lines of the IIS test for outliers of Jiao \& Pretis (2022). The techniques for dealing with correlation between (differenced) errors and regressors turns out to be useful for analyzing instrumental variable estimation (Jiao, 2019).

## A Proofs

In section A.1, we prove Theorem 2.3. In section A.2-A.6, we state and prove some auxiliary results. Finally, the main results for the gauge are proven in sections A.7-A.9.

## A. 1 Properties of differenced innovations

Proof of Theorem 2.3. (a) Symmetry. $\mathrm{h}(x)$ is the density of the difference of two i.i.d. variables $\varepsilon_{i}, \varepsilon_{j}$. Symmetry follows since $\varepsilon_{i}-\varepsilon_{j}, \varepsilon_{j}-\varepsilon_{i}$ are identically distributed. (b) Normal distribution. If $\varepsilon / \sigma$ has a standard normal density $f$, then $\nabla \varepsilon /(\sqrt{2} \sigma)$ has density $\mathrm{h}=\mathrm{f}$. Conversely, if $\mathrm{h}=\mathrm{f}$ then $\mathrm{h}, \mathrm{f}$ are symmetric by part (a), so that $\varepsilon_{i}-$ $\varepsilon_{j}$ and $\varepsilon_{i}+\varepsilon_{j}$ are identically distributed. In particular, $\varepsilon_{i}, \varepsilon_{j}$ and $\left(\varepsilon_{i}+\varepsilon_{j}\right) / \sqrt{2}$ are identically distributed. Pólya (1923) shows that if that distribution is continuous with finite variance, then it must be normal with zero mean.
(c) Bounded densities. The inequality $|x-y|^{k} \leq C_{k}\left(|x|^{k}+|y|^{k}\right)$ for some $C_{k}>0$ implies

$$
\begin{aligned}
|v|^{k} \mathrm{~h}(v) & =\sqrt{2} \int_{-\infty}^{\infty}|v+y-y|^{k} \mathrm{f}(y) \mathrm{f}(v+y) d y \\
& \leq \sqrt{2} C_{k}\left\{\int_{-\infty}^{\infty}|y|^{k} \mathrm{f}(y) \mathbf{f}(v+y) d y+\int_{-\infty}^{\infty}|v+y|^{k} \mathbf{f}(y) \mathbf{f}(v+y) d y\right\}
\end{aligned}
$$

In the first integral, bound $|y|^{k} f(y)$ by its supremum and change variable from $y$ to $s=v+y$. In the second integral, bound $|v+y|^{k} \mathrm{f}(v+y)$ by its supremum. We get

$$
|v|^{k} \mathrm{~h}(v) \leq 2 \sqrt{2} C_{k}\left\{\sup _{v \in \mathbb{R}}|v|^{k} \mathrm{f}(v)\right\} \int_{-\infty}^{\infty} \mathrm{f}(y) d y=2 \sqrt{2} C_{k}\left\{\sup _{v \in \mathbb{R}}|v|^{k} \mathrm{f}(v)\right\} .
$$

(d) Bounded derivatives. By the Leibniz rule for improper integrals

$$
v^{k} \dot{\mathrm{~h}}(v)=v^{k} \sqrt{2} \frac{\partial}{\partial v} \int_{-\infty}^{\infty} \mathrm{f}(y) \mathbf{f}(v+y) d y=v^{k} \sqrt{2} \int_{-\infty}^{\infty} \mathrm{f}(y) \dot{\mathrm{f}}(v+y) d y
$$

Then proceed as in part (c) to get

$$
\begin{aligned}
\left|v^{k} \dot{\mathrm{~h}}(v)\right| & \leq C_{k} \sqrt{2} \int_{-\infty}^{\infty} \mathrm{f}(y)\left(|y|^{k}+|v+y|^{k}\right)|\dot{\mathrm{f}}(v+y)| d y \\
& \leq C_{k} \sqrt{2}\left\{\sup _{v \in \mathbb{R}}|\dot{\mathrm{f}}(v)| \int_{-\infty}^{\infty}|y|^{k} \mathrm{f}(y) d y+\sup _{v \in \mathbb{R}}\left|v^{k} \dot{\mathrm{f}}(v)\right| \int_{-\infty}^{\infty} \mathrm{f}(y) d y\right\}
\end{aligned}
$$

This is finite when $\mathrm{E}\left|\varepsilon_{i}\right|^{k}<\infty$ and $\sup _{v \in \mathbb{R}}\left(1+|v|^{k}\right)|\dot{\mathrm{f}}(v)|<\infty$.

## A. 2 Expanding distribution function for residuals

The compensators in the empirical process theory will be quite complicated due to the temporal dependence arising from forward differencing. Their analysis will be facilitated by the following expansion of the distribution function for a single residual when the estimation error can be assumed constant.

Theorem A.1. Let $Y \in \mathbb{R}$ and $X \in \mathbb{R}^{p}$ be random with density $\mathrm{m}_{Y, X}(y, x)$ with respect to the product of the Lebesgue measure and some measure $v$ on $\mathbb{R}^{p}$. Suppose ( $i$ ) there exists densities so that $\mathrm{m}_{Y, X}(y, x)=\mathrm{m}_{Y \mid X}(y \mid x) \mathrm{m}_{X}(x)=\mathrm{m}_{X \mid Y}(x \mid y) \mathrm{m}_{Y}(y)$;
(ii) $\mathrm{m}_{Y \mid X}(y \mid x)$ has $y$-derivative $\dot{\mathrm{m}}_{Y \mid X}(y \mid x)$;
(iii) $C_{\mathrm{m}}=\sup _{y \in \mathbb{R}, x \in \mathbb{R}^{p}}(1+|y|)\left|\dot{\mathrm{m}}_{Y \mid X}(y \mid x)\right|<\infty$;
(iv) $\mathrm{E}|X|^{2}<\infty$.

Then, for $|a| \leq 1 / 2, b \in \mathbb{R}^{p}$ and $c \in \mathbb{R}$ and with $c^{\dagger}=c(1+a)$, we get
(a) $\left|\mathrm{P}\left(Y-b^{\prime} X \leq c\right)-\mathrm{P}(Y \leq c)-\mathrm{m}_{Y}(c) \mathrm{E}\left(b^{\prime} X \mid Y=c\right)\right| \leq|b|^{2} C_{\mathrm{m}} \mathrm{E}|X|^{2} / 2$.
(b) $\left|\mathrm{m}_{Y}\left(c^{\dagger}\right) \mathrm{E}\left(b^{\prime} X \mid Y=c^{\dagger}\right)-\mathrm{m}_{Y}(c) \mathrm{E}\left(b^{\prime} X \mid Y=c\right)\right| \leq 2|a b| C_{\mathrm{m}} \mathrm{E}|X|$.

Lemma A.2. (Jiao $\mathcal{B}$ Nielsen (2017), Lemma 1.1) If $\left|c^{*}-c\right| \leq|A c+B|$ and $|A| \leq 1 / 2$, then $|c| \leq 2\left(\left|c^{*}\right|+|B|\right)$ and $(A c+B)^{2} \leq 16\left\{A^{2}\left(c^{*}\right)^{2}+B^{2}\right\}$.

Proof of Theorem A.1. (a) Write $\mathcal{P}=\mathrm{P}\left(Y-b^{\prime} X \leq c\right)-\mathrm{P}(Y \leq c)$ as an integral:

$$
\mathcal{P}=\mathrm{E}\left\{1_{\left(Y-b^{\prime} X \leq c\right)}-1_{(Y \leq c)}\right\}=\int_{\mathbb{R}^{p}} \int_{c}^{c+b^{\prime} x} \mathrm{~m}_{Y, X}(y, x) d y d v(x)
$$

Apply the Mean Value Theorem and the identity $\mathrm{m}_{Y, X}=\mathrm{m}_{Y \mid X} \mathrm{~m}_{X}=\mathrm{m}_{X \mid Y} \mathrm{~m}_{Y}$ to get

$$
\int_{c}^{c+b^{\prime} x} \mathrm{~m}_{Y, X}(y, x) d y=\left(b^{\prime} x\right) \mathrm{m}_{X \mid Y}(x \mid c) \mathrm{m}_{Y}(c)+\frac{1}{2}\left(b^{\prime} x\right)^{2} \dot{\mathrm{~m}}_{Y \mid X}\left(c^{*} \mid x\right) \mathrm{m}_{X}(x),
$$

where $\left|c^{*}-c\right| \leq\left|b^{\prime} x\right|$. Then, decompose $\mathcal{P}=\mathcal{P}_{1}+\mathcal{P}_{2}$ with

$$
\mathcal{P}_{1}=\int_{\mathbb{R}^{p}} b^{\prime} x \mathrm{~m}_{Y}(c) \mathrm{m}_{X \mid Y}(x \mid c) d v(x) \text { and } \mathcal{P}_{2}=\frac{1}{2} \int_{\mathbb{R}^{p}}\left(b^{\prime} x\right)^{2} \dot{\mathrm{~m}}_{Y \mid X}\left(c^{*} \mid x\right) \mathrm{m}_{X}(x) d v(x)
$$

The first term is $\mathcal{P}_{1}=\mathrm{m}_{Y}(c) \mathrm{E}\left(b^{\prime} X \mid Y=c\right)$. For the second term, the triangle inequality gives

$$
\left|\mathcal{P}-\mathcal{P}_{1}\right|=\left|\mathcal{P}_{2}\right| \leq \frac{1}{2} \int_{\mathbb{R}^{p}}\left(b^{\prime} x\right)^{2}\left|\dot{\mathrm{~m}}_{Y \mid X}\left(c^{*} \mid x\right)\right| \mathrm{m}_{X}(x) d v(x)
$$

By assumption, $C_{\mathrm{m}}=\sup _{c^{*} \in \mathbb{R}, x \in \mathbb{R}^{p}}\left(1+\left|c^{*}\right|\right)\left|\dot{\mathrm{m}}_{Y \mid X}\left(c^{*} \mid x\right)\right|<\infty$. The norm inequality gives $\left(b^{\prime} x\right)^{2} \leq|x|^{2}|b|^{2}$. Thus, we get uniformly in $c$, that $\left|\mathcal{P}_{2}\right| \leq|b|^{2} C_{\mathrm{m}} \mathrm{E}|X|^{2} / 2$.
(b) Consider the difference term $\left|\mathbf{q}\left(c^{\dagger}\right)-\mathbf{q}(c)\right|$, where

$$
\mathbf{q}(y)=\mathrm{m}_{Y}(y) \mathbf{E}\left(b^{\prime} X \mid Y=y\right)=\mathrm{m}_{Y}(y) \int_{\mathbb{R}^{p}} b^{\prime} x \mathrm{~m}_{X \mid Y}(x \mid y) d v(x)
$$

Let $c^{\dagger}=c(1+a)$. Apply the Mean Value Theorem and $\mathrm{m}_{Y \mid X} \mathrm{~m}_{X}=\mathrm{m}_{X \mid Y} \mathrm{~m}_{Y}$ to get

$$
\mathbf{q}\left(c^{\dagger}\right)-\mathbf{q}(c)=\left(c^{\dagger}-c\right) \int_{-\infty}^{\infty}\left(b^{\prime} x\right) \dot{\mathbf{m}}_{Y \mid X}\left(c^{*} \mid x\right) \mathbf{m}_{X}(x) d v(x)
$$

where $\left|c^{*}-c\right| \leq\left|c^{\dagger}-c\right| \leq|a c|$. Lemma A. 2 shows that $|c| \leq 2\left|c^{*}\right|$ since $|a| \leq 1 / 2$ by assumption. By assumption $C_{\mathrm{m}}=\sup _{c^{*} \in \mathbb{R}, x \in \mathbb{R}^{p}}\left(1+\left|c^{*}\right|\right)\left|\dot{\mathrm{m}}_{Y \mid X}\left(c^{*} \mid x\right)\right|<\infty$, we have $\left|\mathbf{q}\left(c^{\dagger}\right)-\mathbf{q}(c)\right| \leq 2|a b| C_{\mathrm{m}} \mathrm{E}|X|$.

## A. 3 Exponential martingale inequalities

The subsequent empirical process theory relies on a linear chaining argument. The chaining argument uses a new iterated exponential martingale inequality. Our inequality is related to that of Johansen \& Nielsen (2016a, Theorem 5.1), which iterates the Bercu \& Touati (2008) exponential inequality for unbounded martingales. Here, a simpler result suffices, which uses the Freedman (1975) exponential inequality for bounded martingales.

We present two versions. The first version is an exact tail probability bound.
Theorem A.3. For $1 \leq \ell \leq L$, let $M_{\ell n}=\sum_{i=1}^{n}\left(z_{i \ell}-\mathrm{E}_{i-1} z_{i \ell}\right)$ denote a martingale, where $z_{i \ell}$ is $\mathcal{F}_{i}$-adapted and $\left|z_{i \ell}-\mathrm{E}_{i-1} z_{i \ell}\right| \leq 1$. Then, for all $\kappa_{0}, \kappa_{1}>0$, we have

$$
\mathrm{P}\left(\max _{1 \leq \ell \leq L}\left|M_{\ell n}\right|>\kappa_{0}\right) \leq \frac{1}{\kappa_{1}} \mathrm{E} \max _{1 \leq \ell \leq L} \sum_{i=1}^{n} \operatorname{Var}_{i-1} z_{i \ell}+2 L \exp \left\{-\frac{\kappa_{0}^{2}}{2\left(\kappa_{1}+\kappa_{0}\right)}\right\} .
$$

The second version is an asymptotic tail probability bound.
Theorem A.4. For $1 \leq \ell \leq L$, let $M_{\ell n}=\sum_{i=1}^{n}\left(z_{i \ell n}-\mathrm{E}_{i-1} z_{i \ell n}\right)$ denote a martingale array, where $z_{i \ell n}$ is $\mathcal{F}_{\text {in }}$-adapted and $\left|z_{i \ell n}\right| \leq 1$. Suppose $\exists \varsigma, \lambda \geq 0$ so that $\mathrm{E} \max _{1 \leq \ell \leq L} \sum_{i=1}^{n} \mathrm{E}_{i-1} z_{i \ell n}^{2}=\mathrm{O}\left(n^{\varsigma}\right)$ and $L=\mathrm{O}\left(n^{\lambda}\right)$. Then, $\forall \nu>\varsigma / 2, \kappa>0$ we get

$$
\lim _{n \rightarrow \infty} \mathrm{P}\left\{\max _{1 \leq \ell \leq L}\left|M_{\ell n}\right|>\kappa n^{\nu}\right\}=0
$$

Lemma A.5. (Freedman (1975), Theorem 1.6) Let $M_{n}=\sum_{i=1}^{n}\left(z_{i}-\mathrm{E}_{i-1} z_{i}\right)$ denote $a$ martingale, where $z_{i}$ is $\mathcal{F}_{i}$-adapted with $\left|z_{i}-\mathrm{E}_{i-1} z_{i}\right| \leq 1$. Let $T_{n}=\sum_{i=1}^{n} \operatorname{Var}\left(z_{i} \mid \mathcal{F}_{i-1}\right)$. For $a, b>0$ we get $\mathrm{P}\left(M_{n} \geq a, T_{n} \leq b\right) \leq \exp \left[-a^{2} /\{2(a+b)\}\right]$.

Proof of Theorem A.3. Let $m_{i \ell n}=z_{i \ell n}-\mathrm{E}_{i-1} z_{i \ell n}$. Let $A_{\ell}=\sum_{i=1}^{n} m_{i \ell}$ and $\mathcal{A}=$ $\left(\max _{1 \leq \ell \leq L}\left|A_{\ell}\right|>\kappa_{0}\right)$. Let $B_{\ell}=\sum_{i=1}^{n} \mathrm{E}_{i-1} m_{i \ell}^{2}$ and $\mathcal{B}=\left(\max _{1 \leq \ell \leq L} B_{l} \leq \kappa_{1}\right)$. We bound

$$
\mathcal{P}(\mathcal{A})=\mathrm{P}(\mathcal{A} \cap \mathcal{B})+\mathrm{P}\left(\mathcal{A} \cap \mathcal{B}^{c}\right) \leq \mathrm{P}(\mathcal{A} \cap \mathcal{B})+\mathrm{P}\left(\mathcal{B}^{c}\right)
$$

Bounding $\mathrm{P}(\mathcal{A} \cap \mathcal{B})$. Let $\mathcal{A}_{\ell}=\left(\left|A_{\ell}\right|>\kappa_{0}\right)$ and $\mathcal{B}_{\ell}=\left(\left|B_{\ell}\right| \leq \kappa_{1}\right)$. Note $\mathcal{A}=\bigcup_{\ell=1}^{L} \mathcal{A}_{\ell}$ and $\mathcal{B} \subset \mathcal{B}_{\ell}$ and apply Boole's inequality to get

$$
\mathrm{P}(\mathcal{A} \cap \mathcal{B}) \leq \sum_{l=1}^{L} \mathrm{P}\left(\mathcal{A}_{\ell} \cap \mathcal{B}\right) \leq \sum_{l=1}^{L} \mathrm{P}\left(\mathcal{A}_{\ell} \cap \mathcal{B}_{\ell}\right)
$$

Apply Lemma A.5, noting that $\left|m_{i \ell}\right| \leq 1$, to get

$$
\mathrm{P}(\mathcal{A} \cap \mathcal{B}) \leq \sum_{l=1}^{L} \mathrm{P}\left\{\left(A_{\ell}>\kappa_{0}\right) \cap \mathcal{B}_{\ell}\right\}+\mathrm{P}\left\{\left(-A_{\ell}>\kappa_{0}\right) \cap \mathcal{B}_{\ell}\right\} \leq 2 L \exp \left\{-\frac{\kappa_{0}^{2}}{2\left(\kappa_{1}+\kappa_{0}\right)}\right\} .
$$

Bounding $\mathrm{P}\left(\mathcal{B}^{c}\right)$. The Markov inequality gives

$$
\mathrm{P}\left(\mathcal{B}^{c}\right)=\mathrm{P}\left(\max _{1 \leq \ell \leq L} \sum_{i=1}^{n} \mathrm{E}_{i-1} m_{i \ell}^{2}>\kappa_{1}\right) \leq \frac{1}{\kappa_{1}} \mathrm{E} \max _{1 \leq \ell \leq L} \sum_{i=1}^{n} \mathrm{E}_{i-1} m_{i \ell}^{2} .
$$

Finally, combine the bounds to $\mathrm{P}(\mathcal{A} \cap \mathcal{B})$ and $\mathrm{P}\left(\mathcal{B}^{c}\right)$.

Proof of Theorem A.4. We consider the probability $\mathcal{P}_{n}=\mathrm{P}\left\{\max _{1 \leq \ell \leq L}\left|M_{\ell n}\right|>\kappa n^{\nu}\right\}$, where the martingale $M_{\ell n}$ has differences $\left|z_{i \ell n}-\mathrm{E}_{i-1} z_{i \ell n}\right| \leq 2$, since it is assumed that $\left|z_{i \ell n}\right| \leq 1$, but otherwise $M_{\ell n}$ is of the form studied in Theorem A.3. Thus, for some $\kappa>0$, apply Theorem A. 3 with $\kappa_{0}=\kappa n^{\nu} / 2$ and $\kappa_{1}=\kappa^{2} n^{2 \nu}\{4(1+\lambda) \log n\}^{-1} / 2^{2}$. Let $n$ be fixed and sufficiently large such that $\kappa_{0}<\kappa_{1}$. Use the bound

$$
\exp \left[-\kappa_{0}^{2} /\left\{2\left(\kappa_{1}+\kappa_{0}\right)\right\}\right] \leq \exp \left\{-\kappa_{0}^{2} /\left(4 \kappa_{1}\right)\right\}=n^{-1-\lambda}
$$

and the assumptions $\mathrm{E}_{\max }^{1 \leq \ell \leq L} 1 \sum_{i=1}^{n} \mathrm{E}_{i-1} z_{i \ell n}^{2}=\mathrm{O}\left(n^{\varsigma}\right)$ and $L=\mathrm{O}\left(n^{\lambda}\right)$ to get that $\mathcal{P}_{n}=\mathrm{O}\left(n^{\varsigma}\right) / \kappa_{1}+n^{\lambda} n^{-1-\lambda}$. Note that $n^{\varsigma} / \kappa_{1} \rightarrow 0$ when $2 v>\varsigma$ to get that $\mathcal{P}_{n}=\mathrm{o}(1)$.

## A. 4 The one-sided empirical process

We establish some empirical process results for differenced residuals. For this purpose, we simplify the setup relative to that of SIS. In SIS, estimation of $\beta, \sigma$ is done on one subsample, and the evaluation of residuals is done on another sample. Here, the random estimation error is replaced by a deterministic error. We can therefore avoid the division of the sample into subsamples. We will therefore avoid reference to subsamples.

We consider the model (7). Recall the definition (19) and modify definition (22) as

$$
\begin{equation*}
\chi_{i}=\left(\varepsilon_{i}-\varepsilon_{i+1}\right) /(\sqrt{2} \sigma), \quad \nabla x_{i n}=N^{\prime}\left(x_{i}-x_{i+1}\right) \text { for } i=1, \ldots, n \tag{A.1}
\end{equation*}
$$

Thus, $N$ is a normalization matrix similar to those considered before, but applied to the full sample. Let $a$ and $b$ represent estimation errors in the scale and the location. Define the empirical distribution function

$$
\begin{equation*}
\widehat{\mathrm{F}}_{n}(a, b, c)=n^{-1} \sum_{i=1}^{n} 1_{\left(\chi_{i} \leq c+n^{-1 / 2} a c+b^{\prime} \nabla x_{i n}\right)} . \tag{A.2}
\end{equation*}
$$

Here, $\chi_{i}$ is $\mathcal{F}_{i+1}$-adapted and $\nabla x_{i n}$ is $\mathcal{F}_{i}$-adapted. Thus, we will refer to

$$
\begin{equation*}
\overline{\mathrm{F}}_{n}(a, b, c)=n^{-1} \sum_{i=1}^{n} \mathrm{E}_{i-1} 1_{\left(\chi_{i} \leq c+n^{-1 / 2} a c+b^{\prime} \nabla x_{i n}\right)} \tag{A.3}
\end{equation*}
$$

as a pseudo-compensator for $\widehat{\mathrm{F}}_{n}$. The empirical process

$$
\begin{equation*}
\mathbb{F}_{n}(a, b, c)=n^{1 / 2}\left\{\widehat{\mathrm{~F}}_{n}(a, b, c)-\overline{\mathrm{F}}_{n}(a, b, c)\right\} \tag{A.4}
\end{equation*}
$$

satisfies the following result, which will be proved by linear chaining.
Theorem A.6. Suppose Assumption 3.1 holds and that
(i) the marginal density f is bounded: $\sup _{v \in \mathbb{R}} \mathrm{f}(v)<\infty$;
(ii) the regressors $x_{i}$ satisfy $\mathrm{E} \sum_{i=1}^{n}\left|\nabla x_{i n}\right|=\mathrm{O}\left(n^{1 / 2}\right)$.

Then, for all $B>0,0<\eta<1 / 4$, and $c \in \mathbb{R}$ we have

$$
\sup _{|a|,| | \leq n^{1 / 4-\eta_{B}}}\left|\mathbb{F}_{n}(a, b, c)-\mathbb{F}_{n}(0,0, c)\right|=\mathrm{op}(1) .
$$

Lemma A.7. Let $\chi_{i}=\left(\varepsilon_{i}-\varepsilon_{i+1}\right) /(\sqrt{2} \sigma)$, where $\varepsilon_{i} / \sigma$ is $\mathcal{F}_{i}$-adapted and has density f . Let $c_{i} \leq \bar{c}_{i}$ be $\mathcal{F}_{i-1}$-adapted random variables. Then

$$
\mathrm{E}_{i} 1_{\left(c_{i}<\chi_{i} \leq \bar{c}_{i}\right)} \leq \sqrt{2}\left(\bar{c}_{i}-c_{i}\right) \sup _{v \in \mathbb{R}} \mathrm{f}(v)
$$

Proof of Lemma A.7. Write the indicator as $1_{\left(c_{i} \leq \chi_{i} \leq \bar{c}_{i}\right)}=1_{\left(\varepsilon_{i} / \sigma-\sqrt{2} \bar{c}_{i} \leq \varepsilon_{i+1} / \sigma<\varepsilon_{i} / \sigma-\sqrt{2} c_{i}\right)}$. Only $\varepsilon_{i+1}$ is varying when conditioning on $\mathcal{F}_{i}$. Thus, the Mean Value Theorem gives

$$
\mathrm{E}_{i} 1_{\left(\varepsilon_{i} / \sigma-\sqrt{2} \bar{c}_{i} \leq \varepsilon_{i+1} / \sigma<\varepsilon_{i} / \sigma-\sqrt{2} c_{i}\right)}=\int_{\varepsilon_{i} / \sigma-\sqrt{2} \bar{c}_{i}}^{\varepsilon_{i} / \sigma-\sqrt{2} c_{i}} \mathrm{f}(x) d x=\sqrt{2}\left(\bar{c}_{i}-c_{i}\right) \mathrm{f}\left(v^{*}\right)
$$

where $\left|v^{*}-c_{i}\right| \leq \sqrt{2}\left(\bar{c}_{i}-c_{i}\right)$. Finally, note that $\mathrm{f}\left(v^{*}\right) \leq \sup _{v \in \mathbb{R}} \mathrm{f}(v)$.
Proof of Theorem A.6. Combine the two estimation errors as $u=\left(a, b^{\prime}\right)^{\prime}$ and create an expanded vector of regressors $w_{i n}=\left(n^{-1 / 2} c, \nabla x_{i n}^{\prime}\right)^{\prime}$ so that $n^{-1 / 2} a c+b^{\prime} \nabla x_{i n}=u^{\prime} w_{i n}$.

Recall the definition of $\mathbb{F}_{n}$ from (A.4) and write our object of interest as

$$
\begin{aligned}
R_{n}(u, c) & =\mathbb{F}_{n}(a, b, c)-\mathbb{F}_{n}(0,0, c) \\
& =n^{-1 / 2} \sum_{i=1}^{n}\left[\left\{1_{\left(\chi_{i} \leq c+u^{\prime} w_{i n}\right)}-1_{\left(\chi_{i} \leq c\right)}\right\}-\mathrm{E}_{i-1}\left\{1_{\left(\chi_{i} \leq c+u^{\prime} w_{i n}\right)}-1_{\left(\chi_{i} \leq c\right)}\right\}\right] .
\end{aligned}
$$

We show $\mathcal{R}_{n}=\sup _{|u| \leq n^{1 / 4-\eta_{B}}}\left|R_{n}(u, c)\right|=\mathrm{op}_{\mathrm{P}}(1)$.
This proof has three parts. First, we chain over $u$ by introducing grid points $u_{m}$. Second, we show that our empirical process vanishes on the grid points $u_{m}$,

$$
\begin{equation*}
\mathcal{R}_{n, 1}=\max _{1 \leq m \leq M}\left|R_{n}\left(u_{m}, c\right)\right|=\mathrm{op}_{\mathrm{P}}(1) \tag{A.5}
\end{equation*}
$$

Third, we show that our empirical process vanishes in-between our grid points $u_{m}$,

$$
\begin{equation*}
\mathcal{R}_{n, 2}=\max _{1 \leq m \leq M} \sup _{\left|u-u_{m}\right| \leq \delta}\left|R_{n}(u, c)-R_{n}\left(u_{m}, c\right)\right| . \tag{A.6}
\end{equation*}
$$

1. The chaining setup. We chain over $u$, by covering them with balls of radius $\delta>0$. We will choose $\delta$ independently of the sample size $n$ in point 3.5 below.
1.1. Cover. For $\delta, n>0$, cover the set $|u| \leq n^{1 / 4-\eta} B$ with balls of radius $\delta$ which centre in grid points $u_{m}$. Thus, for any $u$ there exists a $u_{m}$ so that $\left|u-u_{m}\right| \leq \delta$. The minimum cover has $M \sim\left(n^{1 / 4-\eta} B / \delta\right)^{\operatorname{dim} x+1} \sim n^{(1 / 4-\eta) \operatorname{dim} x} / \delta^{\operatorname{dim} x+1}$ balls.
1.2. Apply chaining. Write $R_{n}(u, c)=R_{n}\left(u_{m}, c\right)+\left\{R_{n}(u, c)-R_{n}\left(u_{m}, c\right)\right\}$, where $R_{n}\left(u_{m}, c\right)$ is a discrete point term and $R_{n}(u, c)-R_{n}\left(u_{m}, c\right)$ is a local oscillation term. By the triangle inequality, $\mathcal{R}_{n} \leq \mathcal{R}_{n, 1}+\mathcal{R}_{n, 2}$ where $\mathcal{R}_{n, 1}, \mathcal{R}_{n, 2}$ are given in (A.5), (A.6).
2. The discrete point term $\mathcal{R}_{n, 1}$ is $\mathrm{o}_{\mathrm{P}}(1)$. We decompose $R_{n}$ into martingales. Then, we apply Theorem A. 4 on the constructed martingales.
2.1. Martingale decomposition. Let $z_{i m}=1_{\left(\chi_{i} \leq c+u_{m}^{\prime} w_{i n}\right)}-1_{\left(\chi_{i} \leq c\right)}$. Define martingales

$$
R_{n}^{a}\left(u_{m}, c\right)=n^{-1 / 2} \sum_{i=1}^{n}\left(z_{i m}-\mathrm{E}_{i} z_{i m}\right), \quad R_{n}^{b}\left(u_{m}, c\right)=n^{-1 / 2} \sum_{i=1}^{n}\left(\mathrm{E}_{i} z_{i m}-\mathrm{E}_{i-1} z_{i m}\right)
$$

so that $R_{n}=R_{n}^{a}+R_{n}^{b}$. Thus, it suffices to show that

$$
\mathcal{R}_{n, 1}^{a}=\max _{1 \leq m \leq M}\left|R_{n}^{a}\left(u_{m}, c\right)\right|=\mathrm{OP}(1), \quad \mathcal{R}_{n, 1}^{b}=\max _{1 \leq m \leq M}\left|R_{n}^{b}\left(u_{m}, c\right)\right|=\mathrm{OP}(1)
$$

2.2. The martingale $\mathcal{R}_{n, 1}^{a}$. We show that $\mathcal{R}_{n, 1}^{a}=\mathrm{op}_{\mathrm{P}}(1)$, by applying Theorem A. 4 to it. We set $\nu_{a}=1 / 2$, let index $\ell=m$, and consider $z_{i \ell, a}=z_{i m}$, which is $\mathcal{F}_{i+1}$-adapted. Note that $\left|z_{i \ell, a}\right| \leq 1$. We verify the conditions of Theorem A.4.

The parameter $\lambda_{a}$. The set of indices $\ell$ has size $L=M$. Since $M \sim n^{(1 / 4-\eta)(\operatorname{dim} x+1)}$ as $\delta$ is fixed then $L \sim n^{\lambda_{a}}$ where $\lambda_{a}=(1 / 4-\eta)(\operatorname{dim} x+1)$.

The parameter $\varsigma_{a}$. We show $\mathcal{E}_{a}=\mathrm{E}_{\max }^{1 \leq \ell \leq L} \sum_{i=1}^{n} \mathrm{E}_{i} z_{i \ell, a}^{2}=\mathrm{O}\left(n^{3 / 4-\eta}\right)$. First note $\mathrm{E}_{i} z_{i, a}^{2}=\mathrm{E}_{i}\left|z_{i \ell, a}\right|$. Further $\mathrm{E}_{i}\left|z_{i \ell, a}\right| \leq \mathrm{E}_{i} 1_{\left(c-\left|u_{m}^{\prime} w_{i n}\right|<\chi_{i} \leq c+\left|u_{m}^{\prime} w_{i n}\right|\right)}$. Apply Theorem A. 7 with $c_{i}=c-\left|u_{m}^{\prime} w_{i n}\right|$ and $\bar{c}_{i}=c+\left|u_{m}^{\prime} w_{i n}\right|$. Since $\left|u_{m}\right| \leq n^{1 / 4-\eta} B$, we get, uniformly in $\ell$,

$$
\begin{equation*}
\mathrm{E}_{i}\left|z_{i \ell, a}\right| \leq 2 \sqrt{2} n^{1 / 4-\eta} B\left|w_{i n}\right| \sup _{v \in \mathbb{R}} \mathrm{f}(v) \tag{A.7}
\end{equation*}
$$

where only $\left|w_{i n}\right|$ is random and depends on $i$. Apply the Law of Iterated Expectations to get $\mathrm{E} \sum_{i=1}^{n} \mathrm{E}_{i}\left|w_{i n}\right|=\mathrm{E} \sum_{i=1}^{n}\left|w_{i n}\right|$. Since $w_{i n}=\left(n^{-1 / 2} c, \nabla x_{i n}^{\prime}\right)^{\prime}$, we get the further bound $n^{1 / 2}|c|+\mathrm{E} \sum_{i=1}^{n}\left|\nabla x_{i n}\right|$, which is $\mathrm{O}\left(n^{1 / 2}\right)$ since $c$ is fixed and by condition (ii). Further, $\sup _{v \in \mathbb{R}} \mathrm{f}(v)<\infty$ by condition $(i)$. Therefore $\mathcal{E}_{a}=\mathrm{O}\left(n^{\varsigma_{a}}\right)$ where $\varsigma_{a}=3 / 4-\eta$.

The condition $\varsigma_{a}<2 \nu_{a}$. Since $0<\eta$ and $\nu_{a}=1 / 2$, we have $\varsigma_{a}=3 / 4-\eta<1=2 \nu_{a}$.
2.3. The martingale $\mathcal{R}_{n, 1}^{b}$. We show that $\mathcal{R}_{n, 1}^{b}=\mathrm{op}_{\mathrm{P}}(1)$ by applying Theorem A. 4 to it. We set $\nu_{b}=1 / 2$, let index $\ell=m$, and consider $z_{i \ell, b}=\mathrm{E}_{i} z_{i \ell, a}=\mathrm{E}_{i} z_{i m}$, which is $\mathcal{F}_{i}$-adapted. Note that $\left|z_{i \ell, b}\right| \leq 1$ as $\left|z_{i m}\right| \leq 1$. We verify the conditions of Theorem A.4.

The parameter $\lambda_{b}$ is $\lambda_{b}=(1 / 4-\eta)(\operatorname{dim} x+1)$ as in point 2.2.
The parameter $\varsigma_{b}$. We show that $\mathcal{E}_{b}=\mathrm{E}_{\max }^{1 \leq \ell \leq L} \sum_{i=1}^{n} \mathrm{E}_{i-1} z_{i \ell, b}^{2}=\mathrm{O}\left(n^{3 / 4-\eta}\right)$. Note that $z_{i \ell, b}^{2}=\mathrm{E}_{i}^{2} z_{i \ell, a} \leq \mathrm{E}_{i} z_{i \ell, a}^{2}=\mathrm{E}_{i}\left|z_{i \ell, a}\right|$ by Jensen's inequality. In (A.7) we found that $\mathrm{E}_{i}\left|z_{i \ell, a}\right| \leq 2 \sqrt{2} n^{1 / 4-\eta} B\left|w_{i n}\right| \sup _{v \in \mathbb{R}} \mathrm{f}(v)$. Therefore $\mathcal{E}_{b}$ has the same bound as $\mathcal{E}_{a}$. Thus, by point 2.2 we get $\mathcal{E}_{b}=\mathrm{O}\left(n^{\varsigma_{b}}\right)$ where $\varsigma_{b}=\varsigma_{a}=3 / 4-\eta$.

The condition $\varsigma_{b}<2 \nu_{b}$ is satisfied as in point 2.2 , since $\left(\nu_{b}, \lambda_{b}, \varsigma_{b}\right)=\left(\nu_{a}, \lambda_{a}, \varsigma_{a}\right)$.
3. The oscillation term $\mathcal{R}_{n, 2}$. We show that $\mathcal{R}_{n, 2}$ is $\mathrm{op}_{\mathrm{P}}(1)$. The proof relies on bounding $S_{n}\left(u_{m}, u, c\right)=R_{n}(u, c)-R_{n}\left(u_{m}, c\right)$ uniformly in $u$. We then apply a martingale decomposition and use Theorem A.4.
3.1 The term $S_{n}$. Write

$$
S_{n}\left(u_{m}, u, c\right)=n^{-1 / 2} \sum_{i=1}^{n}\left\{s_{i}\left(u_{m}, u, c\right)-\mathrm{E}_{i-1} s_{i}\left(u_{m}, u, c\right)\right\}
$$

where, due to a cancellation of two indicator functions $1_{\left(\chi_{i} \leq c\right)}$, we have

$$
s_{i}\left(u_{m}, u, c\right)=1_{\left(\chi_{i} \leq c+u^{\prime} w_{i n}\right)}-1_{\left(\chi_{i} \leq c+u_{m}^{\prime} w_{i n}\right)} .
$$

Therefore $\mathcal{R}_{n, 2}=\max _{1 \leq m \leq M} \sup _{\left|u-u_{m}\right| \leq \delta}\left|S_{n}\left(u_{m}, u, c\right)\right|$.
3.2. Bounding $s_{i}\left(u_{m}, u, c\right)$. Write $c+u^{\prime} w_{i n}=c+u_{m}^{\prime} w_{i n}+\left(u-u_{m}\right)^{\prime} w_{i n}$. Noting that $\left|u-u_{m}\right| \leq \delta$, we introduce bounds,

$$
\begin{equation*}
c_{i m}=c+u_{m}^{\prime} w_{i n}-\delta\left|w_{i n}\right|, \quad \bar{c}_{i m}=c+u_{m}^{\prime} w_{i n}+\delta\left|w_{i n}\right|, \tag{A.8}
\end{equation*}
$$

which do not depend on $u$. Thus, we can bound

$$
\begin{equation*}
\left|s_{i}\left(u_{m}, u, c\right)\right| \leq z_{i m}=1_{\left(c_{i m}<\chi_{i} \leq \bar{c}_{i m}\right)} . \tag{A.9}
\end{equation*}
$$

3.3. Bounding $\left|S_{n}\left(u_{m}, u, c\right)\right|$. The triangle inequality gives that $\left|\sum_{i=1}^{n}\left(s_{i}-\mathrm{E}_{i-1} s_{i}\right)\right| \leq$ $\sum_{i=1}^{n}\left(\left|s_{i}\right|+\mathrm{E}_{i-1}\left|s_{i}\right|\right)$. Using that $\left|s_{i}\right| \leq z_{i m}$ by (A.9) in point 3.2 leads to the bound

$$
\left|S_{n}\left(u_{m}, u, c\right)\right| \leq M_{m n}=n^{-1 / 2} \sum_{i=1}^{n}\left(z_{i m}+\mathrm{E}_{i-1} z_{i m}\right)
$$

uniformly in $u$. Thus, $\mathcal{R}_{n, 2}=\mathrm{OP}(1)$ if $\max _{1 \leq m \leq M} M_{m n}=\mathrm{OP}_{\mathrm{P}}(1)$.
3.4. Martingale decomposition. We decompose $M_{m n}$ into two martingale and a compensator term. Add and subtract $n^{-1 / 2} \sum_{i=1}^{n} \mathrm{E}_{i} z_{i m}$ twice to $M_{m n}$ and write $M_{m n}=$ $\widetilde{M}_{m n}^{c}+\widetilde{M}_{m n}^{d}+2 \bar{M}_{m n}$, where we have two martingale terms

$$
\widetilde{M}_{m n}^{c}=n^{-1 / 2} \sum_{i=1}^{n}\left(z_{i m}-\mathrm{E}_{i} z_{i m}\right), \quad \widetilde{M}_{m n}^{d}=n^{-1 / 2} \sum_{i=1}^{n}\left(\mathrm{E}_{i} z_{i m}-\mathrm{E}_{i-1} z_{i m}\right)
$$

and a compensator term

$$
\bar{M}_{m n}=n^{-1 / 2} \sum_{i=1}^{n} \mathrm{E}_{i-1} z_{i m}
$$

Thus, it suffices to show that $\widetilde{\mathcal{M}}_{n}^{c}=\max _{1 \leq m \leq M} \widetilde{M}_{m n}^{c}, \widetilde{\mathcal{M}}_{n}^{d}=\max _{1 \leq m \leq M} \widetilde{M}_{m n}^{d}$, and $\overline{\mathcal{M}}_{n}=\max _{1 \leq m \leq M} \bar{M}_{m n}$ are $\mathrm{Op}_{\mathrm{P}}(1)$.
3.5. The compensator term $\overline{\mathcal{M}}_{n}$. We show that $\overline{\mathcal{M}}_{n}=\mathrm{op}_{\mathrm{P}}(1)$. Recall from (A.9) that $z_{i m}=1_{\left(c_{i m}<\chi_{i} \leq \bar{c}_{i m}\right)}$, where $c_{i m}=c+u_{m}^{\prime} w_{i n}-\delta\left|w_{i n}\right|$ and $\bar{c}_{i m}=c+u_{m}^{\prime} w_{i n}+\delta\left|w_{i n}\right|$. The Law of Iterated Expectations gives $\mathrm{E}_{i-1} z_{i m}=\mathrm{E}_{i-1} \mathrm{E}_{i} z_{i m}$. Apply Lemma A. 7 to get

$$
\begin{equation*}
\mathrm{E}_{i-1} z_{i m}=\mathrm{E}_{i-1} \mathrm{E}_{i} z_{i m} \leq \mathrm{E}_{i-1} 2 \sqrt{2} \delta\left|w_{i n}\right| \sup _{v \in \mathbb{R}} \mathrm{f}(v)=2 \sqrt{2} \delta \mathrm{E}_{i-1}\left|w_{i n}\right| \sup _{v \in \mathbb{R}} \mathrm{f}(v), \tag{A.10}
\end{equation*}
$$

uniformly in $m$ and where only $w_{i n}$ depends on $i$.
Turning to the expression for $\overline{\mathcal{M}}_{n}$, we note that $\mathrm{E} \sum_{i=1}^{n} \mathrm{E}_{i-1}\left|w_{i n}\right|=\mathrm{O}\left(n^{1 / 2}\right)$ as argued in point 2.2 using condition (ii). Further, condition $(i)$ shows that $\sup _{v \in \mathbb{R}} \mathrm{f}(v)<\infty$. In combination, we get that $\mathrm{E} \overline{\mathcal{M}}_{n}=\delta \mathrm{O}(1)$ where the $\mathrm{O}_{\mathrm{P}}(1)$-term does not depend on $\delta$. Thus, by the Markov inequality, $\overline{\mathcal{M}}_{n}=\delta \mathrm{O}_{\mathrm{P}}(1)$.

To show $\overline{\mathcal{M}}_{n}=\mathrm{OP}_{\mathrm{P}}(1)$ we need to show that for any $\gamma>0$ then $\mathrm{P}\left(\overline{\mathcal{M}}_{n}>\gamma\right)$ vanishes for large $n$. We are still free to choose $\delta$ which will be exploited now. Since $\overline{\mathcal{M}}_{n}=\delta \mathrm{O}_{\mathrm{P}}(1)$, we can find a constant $C$ not depending on $\delta$ so that $\overline{\mathcal{M}}_{n} \leq \delta C$ with large probability. Choosing $\delta=\gamma / C$ we get $\overline{\mathcal{M}}_{n} \leq \gamma$ with large probability. Hence, $\overline{\mathcal{M}}_{n}=\mathrm{op}_{\mathrm{P}}(1)$.
3.6. The martingale $\widetilde{\mathcal{M}}_{n}^{c}$. We show $\widetilde{\mathcal{M}}_{n}^{c}=\mathrm{op}(1)$, using Theorem A.4. We set $\nu_{c}=1 / 2$ and index $\ell=m$ and consider $z_{i \ell, c}=z_{i m}=1_{\left(c_{i m}<\chi_{i} \leq \bar{c}_{i m}\right)}$ defined in (A.9), which is $\mathcal{F}_{i+1}$-adapted. Note that $0 \leq z_{i \ell, c} \leq 1$. We verify the conditions of Theorem A. 4 .

The parameter $\lambda_{c}$ is $(1 / 4-\eta)(\operatorname{dim} x+1)$ as in point 2.2.
The parameter $\varsigma_{c}$ is $1 / 2$. We show that $\mathcal{E}_{c}=\mathrm{E} \mathrm{max}_{1 \leq \ell \leq L} \sum_{i=1}^{n} \mathrm{E}_{i} z_{i \ell, c}^{2}=\mathrm{O}\left(n^{\varsigma_{c}}\right)$. Since $z_{i m}^{2}=z_{i m}$ while $\mathrm{E}_{i} z_{i m}=\mathrm{E}_{i} \mathrm{E}_{i-1} z_{i m}$ we have that $\mathcal{E}_{c}=\mathrm{E} \max _{1 \leq \ell \leq L} \sum_{i=1}^{n} \mathrm{E}_{i} \mathrm{E}_{i-1} z_{i m}$.

Applying the bound to $\mathrm{E}_{i-1} z_{i m}$ in (A.10), which is uniform in $m$, we see that $\mathcal{E}_{c}$ has the same bound as $n^{1 / 2} \mathrm{E} \overline{\mathcal{M}}_{n}$. Now, use that $\mathrm{E} \overline{\mathcal{M}}_{n}=\mathrm{O}(1)$ by point 3.5.

The condition $\varsigma_{c}<2 \nu_{c}$. Since $0<\eta$ and $\nu_{c}=1 / 2$ we have $\varsigma_{c}=1 / 2<1=2 \nu_{c}$.
3.7. The martingale $\widetilde{\mathcal{M}}_{n}^{d}$. We show $\widetilde{\mathcal{M}}_{n}^{d}=\mathrm{op}(1)$ using Theorem A.4. We set $\nu_{d}=1 / 2$ and index $\ell=m$ and consider $z_{i \ell, d}=\mathrm{E}_{i} z_{i m}$, which is $\mathcal{F}_{i}$-adapted. Note that $0 \leq z_{i \ell, c} \leq 1$. We verify the conditions of Theorem A.4.

The parameter $\lambda_{c}$ is $(1 / 4-\eta)(\operatorname{dim} x+1)$ as in point 2.2.
The parameter $\varsigma_{d}$ is $1 / 2$. We show that $\mathcal{E}_{d}=\mathrm{E}_{\max }^{1 \leq \ell \leq L} \sum_{i=1}^{n} \mathrm{E}_{i-1} z_{i, d}^{2}=\mathrm{O}_{\mathrm{P}}\left(n^{\varsigma_{d}}\right)$. Note that $z_{i, d}^{2}=\mathrm{E}_{i}^{2} z_{i m} \leq \mathrm{E}_{i} z_{i m}^{2}=\mathrm{E}_{i} z_{i m}$ by Jensen's inequality. Since $\mathcal{F}_{i-1} \subset \mathcal{F}_{i}$, we get $\mathrm{E}_{i-1} z_{i \ell, d}^{2} \leq \mathrm{E}_{i-1} \mathrm{E}_{i} z_{i m}=\mathrm{E}_{i-1} z_{i m}$. Thus, $\mathcal{E}_{d}=n^{1 / 2} \mathrm{E} \overline{\mathcal{M}}_{n}$, where $\mathrm{E} \overline{\mathcal{M}}_{n}=\mathrm{O}(1)$ by point 3.5.

The condition $\varsigma_{d}<2 \nu_{d}$ is satisfied as in point 2.2 , since $\left(\nu_{d}, \lambda_{d}, \varsigma_{d}\right)=\left(\nu_{c}, \lambda_{c}, \varsigma_{c}\right)$.
4. Conclusion. We have shown that $\widetilde{\mathcal{M}}_{n}^{c}=\widetilde{\mathcal{M}}_{n}^{d}=\overline{\mathcal{M}}_{n}=\mathrm{op}_{\mathrm{P}}(1)$ so that $\mathcal{R}_{n, 2}=$ $\mathrm{op}_{\mathrm{P}}(1)$. In point 2 it was shown that $\mathcal{R}_{n, 1}=\mathrm{op}_{\mathrm{P}}(1)$. In combination, $\mathcal{R}_{n}=\mathrm{op}_{\mathrm{P}}(1)$.

## A. 5 The compensator

We provide a linearization of the pseudo-compensator $\overline{\mathrm{F}}_{n}$ defined in (A.3).
Theorem A.8. Suppose Assumption 3.1 holds and that
(i) the marginal density f has bounded derivative: $\sup _{v \in \mathbb{R}}\left(1+v^{2}\right)|\dot{\mathrm{f}}(v)|<\infty$;
(ii) the conditional density $\mathrm{m}_{i}(y \mid x)$ of $\chi_{i}$ given $\nabla x_{i}$ and $\mathcal{F}_{i-1}$ exists, it is differentiable in $y$ and satisfies $\max _{1 \leq i \leq n} \sup _{y \in \mathbb{R}, x \in \mathbb{R}^{p}}(1+|y|)\left|\dot{\mathrm{m}}_{i}(y \mid x)\right|<\infty$;
(iii) the regressors $x_{i}$ satisfy $\mathrm{E} \sum_{i=1}^{n}\left|\nabla x_{i n}\right|^{2}=\mathrm{O}(1)$.

Let $\xi_{n}=n^{-1 / 2} \sum_{i=1}^{n} \mathrm{E}_{i-1}\left(\nabla x_{i n} \mid \chi_{i}=c\right)$. Then, for all $B>0$ and $0<\eta<1 / 4$

$$
\sup _{|a|,|b| \leq n^{1 / 4-\eta} B} \sup _{c \in \mathbb{R}}\left|n^{1 / 2}\left\{\overline{\mathrm{~F}}_{n}(a, b, c)-\overline{\mathrm{F}}_{n}(0,0, c)\right\}-\mathrm{h}(c)\left\{a c+b^{\prime} \xi_{n}(c)\right\}\right|=\mathrm{O}_{\mathrm{P}}\left(n^{-2 \eta}\right) .
$$

Proof of Theorem A.8. The quantity of interest is

$$
Q_{n}(a, b, c)=n^{1 / 2}\left\{\overline{\mathrm{~F}}_{n}(a, b, c)-\overline{\mathrm{F}}_{n}(0,0, c)\right\}-\mathrm{h}(c)\left\{a c+b^{\prime} \xi_{n}(c)\right\} .
$$

Define $c_{a}=c+n^{-1 / 2} a c$ and note that $\bar{F}_{n}(a, b, c)=\bar{F}_{n}\left(0, b, c_{a}\right)$. Add and subtract $\mathrm{h}\left(c_{a}\right) b^{\prime} \xi_{n}\left(c_{a}\right)$ and $n^{1 / 2} \overline{\mathrm{~F}}_{n}(a, 0, c)=n^{1 / 2} \overline{\mathrm{~F}}_{n}\left(0,0, c_{a}\right)$ to $Q_{n}$ to get

$$
\begin{equation*}
Q_{n}(a, b, c)=Q_{n}\left(0, b, c_{a}\right)+Q_{n}(a, 0, c)+R_{n}(a, b, c), \tag{A.11}
\end{equation*}
$$

where

$$
\begin{equation*}
R_{n}(a, b, c)=\mathrm{h}\left(c_{a}\right) b^{\prime} \xi_{n}\left(c_{a}\right)-\mathrm{h}(c) b^{\prime} \xi_{n}(c) . \tag{A.12}
\end{equation*}
$$

We show that each of the terms of the right hand side of (A.11) vanish uniformly in $a, b, c$.

1. The term $Q_{n}\left(0, b, c_{a}\right)$. We note that $\sup _{a, b, c}\left|Q_{n}\left(0, b, c_{a}\right)\right|=\sup _{b, c}\left|Q_{n}(0, b, c)\right|$ and consider the latter. Write $Q_{n}(0, b, c)=n^{-1 / 2} \sum_{i=1}^{n} q_{i}(b, c)$, where

$$
q_{i}(b, c)=\mathrm{E}_{i-1}\left\{1_{\left(\chi_{i}-b^{\prime} \nabla x_{i n} \leq c\right)}-1_{\left(\chi_{i} \leq c\right)}\right\}-\mathrm{h}(c) b^{\prime} \mathrm{E}_{i-1}\left(\nabla x_{i n} \mid \chi_{i}=c\right)
$$

Here, $q_{i}$ is an expression of the form considered in Theorem A.1(a), noting that the expectation of an indicator is a probability. The probability measure $P$ in the theorem
is a conditional measure given $\mathcal{F}_{i-1}$, while the variables are $Y=\chi_{i}$ and $X=\nabla x_{i n}$. Moreover, $\mathrm{m}_{i}(y, x)$ is the joint density of $Y_{i}, X_{i}$ conditional on $\mathcal{F}_{i-1}$. By condition (ii), $\mathrm{m}_{i}(y \mid x)$ is differentiable with respect to $y$. The derivative has the bound $C_{\mathrm{m}}=$ $\max _{1 \leq i \leq n} \sup _{y \in \mathbb{R}, x \in \mathbb{R}^{p}}(1+|y|)\left|\dot{\mathrm{m}}_{i}(y \mid x)\right|<\infty$. By condition (iii), $\mathrm{E}|X|^{2}$ exists. We bound $\left|q_{i}(b, c)\right| \leq 2^{-1}|b|^{2} C_{\mathrm{m}} \mathrm{E}_{i-1}\left|\nabla x_{i n}\right|^{2}$, Since $|b| \leq n^{1 / 4-\eta} B$ and using the triangular inequality we get $\left|Q_{n}(b, c)\right| \leq \mathrm{O}\left(n^{-2 \eta}\right) C_{\mathrm{m}} \sum_{i=1}^{n} \mathrm{E}_{i-1}\left|\nabla x_{i n}\right|^{2}$. By condition (iii) and the Markov inequality then $\sum_{i=1}^{n} \mathrm{E}_{i-1}\left|\nabla x_{i n}\right|^{2}=\mathrm{O}(1)$. Thus, $\left|Q_{n}(b, c)\right|=\mathrm{O}_{\mathrm{P}}\left(n^{-2 \eta}\right)$ uniformly in $b, c$.
2. The term $Q_{n}(a, 0, c)$. Write $Q_{n}(a, 0, c)=n^{-1 / 2} \sum_{i=1}^{n} q_{i}(a, c)$, where

$$
q_{i}(a, c)=\mathrm{E}_{i-1}\left\{1_{\left(\chi_{i} \leq c+n^{-1 / 2} a c\right)}-1_{\left(\chi_{i} \leq c\right)}\right\}-n^{-1 / 2} a c h(c) .
$$

Note $\chi_{i}=\left(\varepsilon_{i}-\varepsilon_{i+1}\right) /(\sqrt{2} \sigma)$ has density h and is independent of $\mathcal{F}_{i-1}$. As f is differentiable by Assumption 3.1, so is h. Thus, the Mean-Value Theorem gives

$$
q_{i}(a, c)=\int_{c}^{c+n^{-1 / 2} a c} \mathrm{~h}(u) d u-n^{-1 / 2} a c \mathrm{~h}(c)=n^{-1} a^{2} c^{2} \dot{\mathrm{~h}}(\tilde{c}) / 2,
$$

where $|\tilde{c}-c| \leq\left|n^{-1 / 2} a c\right|$. Since $|a| \leq n^{1 / 4-\eta} B$, then $\left|n^{-1 / 2} a\right| \leq 1 / 2$ for large $n$. The second inequality in Lemma A. 2 then shows that $a^{2} c^{2} \leq 16 a^{2} \tilde{c}^{2}$. We then have $q_{i}(a, c) \leq$ $16 n^{-1} a^{2} \tilde{c}^{2} \dot{\mathrm{~h}}(\tilde{c})$.

Condition (i) shows that $\sup _{v \in \mathbb{R}}\left(1+v^{2}\right) \dot{\mathrm{f}}(v)<\infty$. Theorem 2.3(d) then implies that $\sup _{v \in \mathbb{R}} v^{2} \dot{\mathrm{~h}}(v)<\infty$. Using that $|a| \leq n^{1 / 4-\eta} B$ we get $q_{i}(a, c)=\mathrm{O}\left(n^{-1 / 2-2 \eta}\right)$ uniformly in $a, c, i$. It follows that $\left|Q_{n}(a, c)\right|=\mathrm{O}\left(n^{-2 \eta}\right)$ uniformly in $a, c$.
3. The term $R_{n}(a, b, c)$. Write $R_{n}(a, b, c)=n^{-1 / 2} \sum_{i=1}^{n} q_{i}(a, b, c)$, where

$$
q_{i}(a, b, c)=\mathrm{h}\left(c_{a}\right) b^{\prime} \mathrm{E}_{i-1}\left(\nabla x_{i n} \mid \chi_{i}=c_{a}\right)-\mathrm{h}(c) b^{\prime} \mathrm{E}_{i-1}\left(\nabla x_{i n} \mid \chi_{i}=c\right)
$$

with $c_{a}=c+n^{-1 / 2} a c$. Apply Theorem A.1(b). The setup is as in point 1, with $Y_{i}=\chi_{i}$ and $X_{i}=\nabla x_{i n}$, while $\mathrm{m}_{i}(y, x)$ denotes the joint conditional density of $Y_{i}, X_{i}$ given $\mathcal{F}_{i-1}$, and $C_{\mathrm{m}}=\max _{1 \leq i \leq n} \sup _{y \in \mathbb{R}, x \in \mathbb{R}^{p}}(1+|y|)\left|\dot{\mathrm{m}}_{i}(y \mid x)\right|<\infty$, using conditions (ii, iii). We bound $\left|q_{i}(a, b, c)\right| \leq\left|n^{-1 / 2} a b\right| C_{\mathrm{m}} \mathrm{E}_{i-1}\left|\nabla x_{i n}\right|$. Since $|a|,|b| \leq n^{1 / 4-\eta} B$ and $\sum_{i=1}^{n} \mathrm{E}_{i-1}\left|\nabla x_{i n}\right|=\mathrm{O}(1)$ by condition (iii), we get $\left|R_{n}(a, b, c)\right|=\mathrm{O}\left(n^{-2 \eta}\right)$ uniformly in $a, b$ and $c$.

## A. 6 The empirical distribution function

We combine Theorems A. 6 and A. 8 to expand the empirical distribution function.
Theorem A.9. Suppose Assumption 3.1 holds and that
(i) the marginal density f satisfies: $\sup _{v \in \mathbb{R}} \mathrm{f}(v)<\infty$, $\sup _{v \in \mathbb{R}}\left(1+v^{2}\right)|\dot{\mathrm{f}}(v)|<\infty$;
(ii) the conditional density $\mathrm{m}_{i}(y \mid x)$ of $\chi_{i}$ given $\nabla x_{i}$ and $\mathcal{F}_{i-1}$ exists, it is differentiable in $y$ and satisfies $\max _{1 \leq i \leq n} \sup _{y \in \mathbb{R}, x \in \mathbb{R}^{p}}(1+|y|)\left|\dot{m}_{i}(y \mid x)\right|<\infty$;
(iii) the regressors $x_{i}$ satisfy $\mathrm{E} \sum_{i=1}^{n}\left|\nabla x_{i n}\right|^{2}=\mathrm{O}(1)$;

Let $\xi_{n}=n^{-1 / 2} \sum_{i=1}^{n} \mathrm{E}_{i-1}\left(\nabla x_{i n} \mid \chi_{i}=c\right)$. Then, for all $B>0,0<\eta<1 / 4, c \in \mathbb{R}$, and uniformly in $|a|,|b| \leq n^{1 / 4-\eta} B$, we have

$$
\sqrt{n}\left\{\widehat{\mathrm{~F}}_{n}(a, b, c)-\overline{\mathrm{F}}_{n}(0,0, c)\right\}=\mathbb{F}_{n}(0,0, c)+\mathrm{h}(c)\left\{a c+b^{\prime} \xi_{n}(c)\right\}+\mathrm{op}(1)
$$

Proof of Theorem A.9. The conditions of Theorems A. 6 and A. 8 are directly listed in the present conditions apart from condition (ii) of Theorem A.6. The latter follows from the present condition (iii) by Jensen's inequality in that $n^{-1 / 2} \mathrm{E} \sum_{i=1}^{n}\left|\nabla x_{i n}\right| \leq$ $\left(\mathrm{E} \sum_{i=1}^{n}\left|\nabla x_{i n}\right|^{2}\right)^{1 / 2}=\mathrm{O}(1)$. Rewrite $R_{n}=n^{1 / 2}\left\{\widehat{\mathrm{~F}}_{n}(a, b, c)-\overline{\mathrm{F}}_{n}(0,0, c)\right\}$ as

$$
\begin{aligned}
R_{n}= & n^{1 / 2}\left[\left\{\widehat{\mathrm{~F}}_{n}(a, b, c)-\overline{\mathrm{F}}_{n}(a, b, c)\right\}-\left\{\widehat{\mathrm{F}}_{n}(0,0, c)-\overline{\mathrm{F}}_{n}(0,0, c)\right\}\right] \\
& +\left[n^{1 / 2}\left\{\overline{\mathrm{~F}}_{n}(a, b, c)-\overline{\mathrm{F}}_{n}(0,0, c)\right\}-\mathrm{h}(c)\left\{a c+b^{\prime} \xi(c)\right\}\right] \\
& +n^{1 / 2}\left\{\widehat{\mathrm{~F}}_{n}(0,0, c)-\overline{\mathrm{F}}_{n}(0,0, c)\right\}+\mathrm{h}(c)\left\{a c+b^{\prime} \xi(c)\right\} .
\end{aligned}
$$

By Theorems A. 6 and A.8, the terms in square brackets is $\mathrm{o}_{\mathrm{P}}(1)$ uniformly in $|a|,|b|$.

## A. 7 Results for stylized SIS

Lemma A.10. Suppose Assumption 3.1, 3.2(ia,iiic). Let $\eta>0$ be given. Let $\sigma_{n}$ be a sequence of random variables so that $n_{j}^{1 / 2}\left(\sigma_{n}^{2}-\sigma^{2}\right)=\mathrm{O}_{\mathrm{P}}\left(n_{j}^{1 / 4-\eta}\right)$. Let $V_{n}, \Sigma_{n}^{-1}$ be sequences of random vectors and square matrices which are $\mathrm{O}_{\mathrm{P}}(1)$. Let $M_{n}$ be a sequence of deterministic square matrices satisfying $M_{n}=\mathrm{O}\left(n_{j}^{1 / 4-\eta}\right)$ for some $\eta>0$. All those vectors and square matrices have the same dimension as $x_{i}$. Let $w_{i n}^{2}=1+$ $\left(\nabla x_{i n}\right)^{\prime} M_{n} \Sigma_{n}^{-1} M_{n}^{\prime} \nabla x_{i n}$ and

$$
D_{i}=1_{\left(\left|\nabla \varepsilon_{i}-V_{n}^{\prime} \Sigma_{n}^{-1} M_{n}^{\prime} \nabla x_{i n}\right|>\sqrt{2} \sigma_{n} w_{i n} c\right)}-1_{\left(\left|\nabla \varepsilon_{i}-V_{n}^{\prime} \Sigma_{n}^{-1} M_{n}^{\prime} \nabla x_{i n}\right|>\sqrt{2} \sigma_{n} c\right)} .
$$

Then $\sum_{i \in I_{j}} D_{i}=\mathrm{op}\left(n_{j}^{1 / 2}\right)$.
Proof of Lemma A.10. By definition $w_{i n}^{2} \geq 1$, so that $1 \leq w_{i n} \leq w_{i n}^{2}$. Thus,

$$
0 \leq D_{i}=1_{\left(\sqrt{2} \sigma_{n} c<\left|\nabla \varepsilon_{i}-V_{n}^{\prime} \Sigma_{n}^{-1} M_{n}^{\prime} \nabla x_{i n}\right| \leq \sqrt{2} \sigma_{n} w_{i n} c\right)} \leq 1_{\left(\sqrt{2} \sigma_{n} c<\left|\nabla \varepsilon_{i}-V_{n}^{\prime} \Sigma_{n}^{-1} M_{n}^{\prime} \nabla x_{i n}\right| \leq \sqrt{2} \sigma_{n} w_{i n}^{2} c\right)} .
$$

Further, use the spectral norm as matrix norm. As this is sub-multiplicative, we can bound $w_{i n}^{2}-1 \leq\left|\nabla x_{i n}\right|^{2}\left\|M_{n}\right\|^{2}\left\|\Sigma_{n}^{-1}\right\|$ and get

$$
0 \leq D_{i} \leq 1_{\left(\sqrt{2} \sigma_{n} c<\left|\nabla \varepsilon_{i}-V_{n}^{\prime} \Sigma_{n}^{-1} M_{n}^{\prime} \nabla x_{i n}\right| \leq \sqrt{2} \sigma_{n} c+\sqrt{2} \sigma_{n}\left|\nabla x_{i n}\right|^{2}\left\|M_{n}\right\|^{2}\left\|\Sigma_{n}^{-1}\right\| c\right)}
$$

Introduce the empirical distribution function and pseudo-compensator

$$
\begin{align*}
& \widehat{\mathrm{H}}_{n}\left(a, b_{1}, b_{2}, c\right)=n_{j}^{-1} \sum_{i \in I_{j}} 1_{\left(\chi_{i} \leq c+n_{j}^{-1 / 2} a c+b_{1}^{\prime} z_{1 i}+b_{2} z_{2} c\right)},  \tag{A.13}\\
& \overline{\mathrm{H}}_{n}\left(a, b_{1}, b_{2}, c\right)=n_{j}^{-1} \sum_{i \in I_{j}} \mathrm{E}_{i-1} 1_{\left(\chi_{i} \leq c+n_{j}^{-1 / 2}{ }_{a} a c+b_{1}^{\prime} z_{1 i}+b_{2} z_{2 i} c\right)} . \tag{A.14}
\end{align*}
$$

where $\chi_{i}=\nabla \varepsilon_{i} /(\sqrt{2} \sigma)$ as in (19), $z_{i}=\left(z_{1 i}^{\prime}, z_{2 i}\right)^{\prime}=\left\{\left(\nabla x_{i n}\right)^{\prime}, n_{j}^{1 / 2}\left|\nabla x_{i n}\right|^{2}\right\}^{\prime}$ is $\mathcal{F}_{i}$-adapted, and $\widehat{a}=n_{j}^{1 / 2}\left(\sigma_{n} / \sigma-1\right), \widehat{b}_{1}^{\prime}=V_{n}^{\prime} \Sigma_{n}^{-1} M_{n}^{\prime} / \sqrt{2}$ and $\widehat{b}_{2}=\left(\sigma_{n} / \sigma\right) n_{j}^{-1 / 2}\left\|M_{n}\right\|^{2}\left\|\Sigma_{n}^{-1}\right\|$. Then, noting that $\chi_{i}$ has a continuous distribution, we get with probability one

$$
n_{j}^{-1} \sum_{i \in I_{j}} D_{i} \leq \widehat{\mathrm{H}}_{n}\left(\widehat{a}, \widehat{b}_{1}, \widehat{b}_{2} c, c\right)-\widehat{\mathrm{H}}_{n}\left(\widehat{a}, \widehat{b}_{1}, 0, c\right)-\widehat{\mathrm{H}}_{n}\left(\widehat{a}, \widehat{b}_{1},-\widehat{b}_{2} c,-c\right)+\widehat{\mathrm{H}}_{n}\left(\widehat{a}, \widehat{b}_{1}, 0,-c\right)
$$

The empirical distribution function $\widehat{\mathrm{H}}_{n}$ has the same structure as $\widehat{\mathrm{F}}_{n}$ defined in (A.2), albeit with $b^{\prime} \nabla x_{i n}$ replaced by $\left(b_{1}^{\prime}, b_{2} c\right) z_{i}$. We note that $\widehat{a}=\mathrm{O}_{\mathrm{P}}\left(n_{j}^{1 / 4-\eta}\right), \widehat{b}_{1}=\mathrm{O}_{\mathrm{P}}\left(n_{j}^{1 / 4-\eta}\right)$, $\widehat{b}_{2}=\mathrm{O}_{\mathrm{P}}\left(n_{j}^{-2 \eta}\right)$, so that all are $\mathrm{O}_{\mathrm{P}}\left(n_{j}^{1 / 4-\eta}\right)$. Thus, we can on a set with large probability find a large $B>0$ so that $|\widehat{a}|,\left|\widehat{b}_{1}\right| \leq B n_{j}^{1 / 4-\eta}$ and $\left|\widehat{b}_{2}\right| \leq B n_{j}^{-2 \eta} \leq B n_{j}^{1 / 4-\eta}$ uniformly in $n_{j}$. We can then apply the expansion in Theorem A. 6 to replace $\widehat{\mathrm{H}}_{n}$ by $\overline{\mathrm{H}}_{n}$, while substituting $\widehat{a}, \widehat{b}_{1}, \widehat{b}_{2}$ for $a, b_{1}, b_{2}$ and using Assumption 3.1, 3.2(ia, iiic). Thus,

$$
\begin{align*}
0 \leq n_{j}^{-1} \sum_{i \in I_{j}} D_{i} \leq & \overline{\mathrm{H}}_{n}\left(\widehat{a}, \widehat{b}_{1}, \widehat{b}_{2} c, c\right)-\overline{\mathrm{H}}_{n}\left(\widehat{a}, \widehat{b}_{1}, 0, c\right) \\
& -\overline{\mathrm{H}}_{n}\left(\widehat{a}, \widehat{b}_{1},-\widehat{b}_{2} c,-c\right)+\overline{\mathrm{H}}_{n}\left(\widehat{a}, \widehat{b}_{1}, 0,-c\right)+\mathrm{op}\left(n_{j}^{-1 / 2}\right) \tag{A.15}
\end{align*}
$$

Finally, we show that $\overline{\mathrm{H}}_{n}\left(\widehat{a}, \widehat{b}_{1}, \widehat{b}_{2} c, c\right)-\overline{\mathrm{H}}_{n}\left(\widehat{a}, \widehat{b}_{1}, 0, c\right)=\mathrm{op}_{\mathrm{P}}\left(n_{j}^{-1 / 2}\right)$ for fixed $c \in \mathbb{R}$. For this, it suffices to show that

$$
\begin{equation*}
\sup _{|a|,\left|b_{1}\right| \leq B n_{j}^{1 / 4-\eta}} \sup _{0 \leq b_{2} \leq B n_{j}^{-2 \eta}}\left|\overline{\mathrm{H}}_{n}\left(a, b_{1}, b_{2} c, c\right)-\overline{\mathrm{H}}_{n}\left(a, b_{1}, 0, c\right)\right|=\mathrm{o}_{\mathrm{P}}\left(n_{j}^{-1 / 2}\right) . \tag{A.16}
\end{equation*}
$$

Now, $\overline{\mathrm{H}}_{n}\left(a, b_{1}, b_{2} c, c\right)-\overline{\mathrm{H}}_{n}\left(a, b_{1}, 0, c\right)=n_{j}^{-1} \sum_{i \in I_{j}} \mathrm{E}_{i-1} h_{i n}\left(a, b_{1}, b_{2}, c\right)$ where

$$
h_{i n}\left(a, b_{1}, b_{2}, c\right)=1_{\left\{\chi_{i} \leq c+n_{j}^{-1 / 2} a c+\left(b_{1}^{\prime}, b_{2} c\right) z_{i}\right\}}-1_{\left\{\chi_{i} \leq c+n_{j}^{-1 / 2} a c+\left(b_{1}^{\prime}, 0\right) z_{i}\right\}}
$$

has the same sign as $c$. We get $\mathrm{E}_{i-1} h_{i n}\left(a, b_{1}, b_{2}, c\right)=\mathrm{E}_{i-1}\left\{\mathrm{E}_{i} h_{i n}\left(a, b_{1}, b_{2}, c\right)\right\}$ by iterated expectations. Recall that $\chi_{i}=\left(\varepsilon_{i}-\varepsilon_{i+1}\right) /(\sqrt{2} \sigma)$, so that all elements of $h_{i n}$ but $\varepsilon_{i+1}$ are $\mathcal{F}_{i}$-measurable while $\varepsilon_{i+1}$ is independent thereof. Thus, we can write the $\mathrm{E}_{i}$-expectation as an integral and use the Mean Value Theorem to get, for an intermediate point $c^{*}$,

Now, $\left(0, b_{2}\right)^{\prime} z_{i}=b_{2} n_{j}^{1 / 2}\left|\nabla x_{i n}\right|^{2}$, where $0 \leq b_{2} \leq B n_{j}^{-2 \eta}$ by the construction (A.16). Further, $c$ is fixed, while $\mathfrak{f}\left(c^{*}\right) \leq \max _{v \in \mathbb{R}} \mathrm{f}(v)$ which is finite by Assumption 3.2(ia). Thus, we find $\left|\mathrm{E}_{i} h_{i n}\left(a, b_{1}, b_{2}, c\right)\right| \leq C n_{j}^{1 / 2-2 \eta}\left|\nabla x_{i n}\right|^{2}$ for some constant $C>0$, uniformly in $a, b_{1}, b_{2}$.

We note that $h_{i n}$ and hence the $\overline{\mathrm{H}}_{n}$-differences has the same sign as $c$ so that

$$
\mathcal{H}_{n}=\mathrm{E}\left|\overline{\mathrm{H}}_{n}\left(a, b_{1}, b_{2}, c\right)-\overline{\mathrm{H}}_{n}\left(a, b_{1}, 0, c\right)\right|=\left|\mathrm{E}\left\{\overline{\mathrm{H}}_{n}\left(a, b_{1}, b_{2}, c\right)-\overline{\mathrm{H}}_{n}\left(a, b_{1}, 0, c\right)\right\}\right| .
$$

Writing out in terms of the $h_{\text {in }}$ functions and using iterated expectations, we get

$$
\mathcal{H}_{n}=\left|\mathrm{E} n_{j}^{-1} \sum_{i \in I_{j}} \mathrm{E}_{i-1} \mathrm{E}_{i} h_{i n}\left(a, b_{1}, b_{2}, c\right)\right|=\left|\mathrm{E} n_{j}^{-1} \sum_{i \in I_{j}} \mathrm{E}_{i} h_{i n}\left(a, b_{1}, b_{2}, c\right)\right| .
$$

Thus, uniformly in $a, b_{1}, b_{2}$

$$
\mathcal{H}_{n} \leq \mathrm{E} n_{j}^{-1} \sum_{i \in I_{j}} C n_{j}^{1 / 2-\eta}\left|\nabla x_{i n}\right|^{2}=\mathrm{O}\left(n_{j}^{-1 / 2-\eta}\right)
$$

by Assumption 3.2(iiia), so that (A.16) follows. The desired result follows.

Proof of Theorem 3.3. 1. The OLS estimator on the first sample. Normalizing the OLS estimator $\hat{\beta}_{1}$ in (12) gives $N_{1}^{-1}\left(\hat{\beta}_{1}-\beta\right)=\widehat{\Sigma}_{1 n}^{-1} \widehat{V}_{1 n}$ when using the following notation from (23), (24)

$$
\widehat{\Sigma}_{1 n}^{-1}=\sum_{i \in I_{1}} N_{1}^{\prime}\left(x_{i}-\bar{x}_{1}\right)\left(x_{i}-\bar{x}_{1}\right)^{\prime} N_{1}, \quad \widehat{V}_{1 n}=\sum_{i \in I_{1}} N_{1}^{\prime}\left(x_{i}-\bar{x}_{1}\right)\left(\varepsilon_{i}-\mathrm{E} \varepsilon_{i}\right),
$$

while $\bar{x}_{1}=n_{1}^{-1} \sum_{i \in I_{1}} x_{i}$, where the expectation can be subtracted from $\varepsilon_{i}$ since the regressors are demeaned. By Assumption $3.2($ iiia,$b), N_{1}^{-1}\left(\hat{\beta}_{1}-\beta\right)=\hat{\Sigma}_{1 n}^{-1} \hat{V}_{1 n}=\mathrm{O}_{\mathrm{P}}(1)$. Similarly, the normalized estimator for the residual variance in (23) is

$$
n_{1}^{1 / 2}\left(\hat{\sigma}_{1}^{2}-\sigma^{2}\right)=n_{1}^{-1 / 2} \sum_{i \in I_{1}}\left\{\left(\varepsilon_{i}-\bar{\varepsilon}_{1}\right)^{2}-\sigma^{2}\right\}-n_{1}^{-1 / 2} \widehat{V}_{1 n}^{\prime} \widehat{\Sigma}_{1 n}^{-1} \widehat{V}_{1 n}
$$

where $\bar{\varepsilon}_{1}=n_{1}^{-1} \sum_{i \in I_{1}} \varepsilon_{i}$. By Assumption 3.1, the innovations $\varepsilon_{i}$ are i.i.d. The first term converges in distribution by the Central Limit Theorem. The second vanishes as $\widehat{\Sigma}_{1 n}$, $\widehat{V}_{1 n}$ converge in distribution by Assumption 3.2(iiia, b) while the factor $n_{1}^{-1 / 2}$ vanishes. Therefore, the estimators are converging and bounded with $\mathrm{O}_{\mathrm{P}}(1)$.
2. Apply Lemma A. 10 with $j=2$ and $\omega_{\text {in }}=\omega_{1, i}$. Since $\left(n_{2} / n_{1}\right)^{1 / 2}=\mathrm{o}\left(n_{2}^{1 / 4-\eta}\right)$ by Assumption 3.2(iv), we get $n_{2}^{1 / 2}\left(\sigma_{n}^{2}-\sigma\right)=\left(n_{2} / n_{1}\right)^{1 / 2} n_{1}^{1 / 2}\left(\hat{\sigma}_{1}^{2}-\sigma^{2}\right)=\mathrm{op}\left(n_{2}^{1 / 4-\eta}\right)$. Note, that $\widehat{V}_{1 n}$ and $\Sigma_{n}^{-1}=\widehat{\Sigma}_{1 n}^{-1}$, are both $\mathrm{O}_{\mathrm{P}}(1)$. Let $M_{n}=N_{2}^{-1} N_{1}=\mathrm{o}\left(n_{2}^{1 / 4-\eta}\right)$ by Assumption 3.2(iv). Lemma A. 10 also requires Assumption 3.2(ia, iiic).

Proof of Theorem 3.4. It was shown in step 1 of the proof of Theorem 3.3 that $N_{1}^{-1}\left(\hat{\beta}_{1}-\beta\right), n_{1}^{1 / 2}\left(\hat{\sigma}_{1}^{2}-\sigma^{2}\right)=\mathrm{O}_{\mathrm{P}}(1)$. This uses Assumptions 3.1, 3.2(iiia,b). We rewrite the gauge and apply Theorem A.9.

1. Rewriting expression for the gauge. The gauge is defined in (16). Due to Theorem 3.3 with Assumptions 3.1, 3.2(ia,iii,iv) we can set $w_{i, 1}=1$ and we ignore the resulting remainder term. Forward difference equation (7) to get $\nabla y_{i}=\beta^{\prime} \nabla x_{i}+\nabla \varepsilon_{i}$. With that, rewrite the gauge as

$$
\hat{\gamma}_{n}=\frac{1}{n_{2}^{\circ}} \sum_{i \in I_{2}^{\circ}} 1_{\left(\left|\nabla y_{i}-\hat{\beta}_{1}^{\prime} \nabla x_{i}\right| \geq \sqrt{2} \hat{\sigma}_{1} c\right)}=\frac{1}{n_{2}^{\circ}} \sum_{i \in I_{2}^{\circ}} 1_{\left\{\left|\nabla \varepsilon_{i}-\left(\hat{\beta}_{1}-\beta\right)^{\prime} \nabla x_{i}\right| \geq \sqrt{2} \hat{\sigma}_{1} c\right\}} .
$$

We normalize $\hat{\beta}_{1}-\beta$ by $N_{1}$ and $x_{i}$ by $N_{2}$. At the same time we divide through by $\sqrt{2} \sigma$. Recall the notation $\chi_{i}=\varepsilon_{i} /(\sqrt{2} \sigma)$ and $x_{i n}=N_{2}^{\prime} x_{i}$ for $i \in I_{2}$. Define

$$
\begin{equation*}
\hat{a}_{1}=\left(n_{2} / n_{1}\right)^{1 / 2} n_{1}^{1 / 2}\left(\hat{\sigma}_{1} / \sigma-1\right), \quad \hat{b}_{1}=\left(N_{2}^{-1} N_{1}\right) N_{1}^{-1}\left(\hat{\beta}_{1}-\beta\right) /(\sqrt{2} \sigma) \tag{A.17}
\end{equation*}
$$

so that $\hat{a}_{1}, \hat{b}_{1}=\mathrm{O}_{\mathrm{P}}\left(n_{2}^{1 / 4-\eta}\right)$ by the convergence results in point 1 of the proof of Theorem 3.3 and Assumption 3.2(iv). We then get the two-sided empirical distribution function

$$
\begin{equation*}
\hat{\gamma}_{n}=\frac{1}{n_{2}^{\circ}} \sum_{i \in I_{2}^{\circ}} 1_{\left\{\left|\chi_{i}-\hat{b}_{1}^{\prime} \nabla x_{i n}\right| \geq c+n_{2}^{-1 / 2} \hat{a}_{1} c\right\}} . \tag{A.18}
\end{equation*}
$$

2. Apply empirical process result. Theorem A. 9 expands

$$
\begin{aligned}
n^{-1 / 2} \sum_{i=1}^{n}\left\{1_{\left(\chi_{i} \leq c+n^{-1 / 2} a c+b^{\prime} \nabla x_{i n}\right)}-\mathrm{E} 1_{\left(\chi_{i} \leq c\right)}\right\}=n^{-1 / 2} \sum_{i=1}^{n} & \left\{1_{\left(\chi_{i} \leq c\right)}-\mathrm{E} 1_{\left(\chi_{i} \leq c\right)}\right\} \\
& +\mathrm{h}(c)\left\{a c+b^{\prime} \xi_{n}(c)\right\}+\mathrm{op}(1)
\end{aligned}
$$

uniformly in $|a|,|b| \leq n^{1 / 4-\eta} B$ for all $\eta, B>0$. As remarked above, this expansion will be used for observations in the second sub-sample. Thus, the conditions of Theorem A. 9 are satisfied by Assumption 3.2 (i,ii,iiic). Since $1_{\left(\left|\chi_{i}\right| \geq c\right)}=1-1_{\left(\chi_{i}<c\right)}+1_{\left(\chi_{i} \leq-c\right)}$ and noting that $\chi_{i}$ is continuously distributed we can form a two sided version

$$
\begin{align*}
& n^{-1 / 2} \sum_{i=1}^{n}\left\{1_{\left(\left|\chi_{i}\right| \geq c+n^{-1 / 2} a c+b^{\prime} \nabla x_{i n}\right)}-\mathrm{E} 1_{\left(\left|\chi_{i}\right| \geq c\right)}\right\}=n^{-1 / 2} \sum_{i=1}^{n}\left\{1_{\left(\left|\chi_{i}\right| \geq c\right)}-\mathrm{E} 1_{\left(\left|\chi_{i}\right| \geq c\right)}\right\} \\
&+\mathcal{B}_{n}(a, b, c)+\mathrm{o}_{\mathrm{P}}(1) \tag{A.19}
\end{align*}
$$

with the bias term $\mathcal{B}_{n}(a, b, c)=-\mathrm{h}(c)\left\{a c+b^{\prime} \xi_{n}(c)\right\}+\mathrm{h}(-c)\left\{-a c+b^{\prime} \xi_{n}(-c)\right\}$. Theorem 2.3 shows that h is symmetric: $\mathrm{h}(c)=\mathrm{h}(-c)$. Thus, the bias term satisfies

$$
\mathcal{B}_{n}(a, b, c)=-2 c \mathrm{~h}(c) a-\mathrm{h}(c) b^{\prime}\left\{\xi_{n}(c)-\xi_{n}(-c)\right\} .
$$

We now apply the expansion (A.19) to the expression for the gauge in (A.18) with two adjustments. First, the gauge in (A.18) depends on estimators $\hat{a}_{1}, \hat{b}_{1}$. These are $\mathrm{O}_{\mathrm{P}}\left(n_{2}^{1 / 4-\eta}\right)$ as remarked above. Thus, on a set with large probability a large $B>0$ exists so that $\left|\hat{a}_{1}\right|,\left|\hat{b}_{1}\right|<B n_{2}^{1 / 4-\eta}$ uniformly in $n$. Since the expansion in (A.19) is uniform in $\left|a_{1}\right|,\left|b_{1}\right|<B n^{1 / 4-\eta}$, we can apply the expansion in (A.19) while substituting $\hat{a}_{1}, \hat{b}_{1}$ for $a_{1}, b_{1}$. Second, we will need to change the index of the observations, as the expansion in (A.18) is concerned with indices, $i \in I_{2}^{\circ}$ while the expansion in (A.19) has indices $i=1, \ldots n$.

Thus, defining $\gamma=\mathrm{E} 1_{\left(\left|\chi_{i}\right| \geq c\right)}$ and $\xi_{2 n}=n_{2}^{-1 / 2} \sum_{i \in I_{2}} \mathrm{E}_{i-1}\left(\nabla x_{i n} \mid \chi_{i}=c\right)$, while noting $n_{2}^{\circ} / n_{2} \rightarrow 1$, we get the desired expansion (25).
3. Consistency. The terms on the right hand side of expansion (25) are op $\left(n_{2}^{1 / 2}\right)$ under the stated conditions. This gives the convergence in probability. As the gauge is bounded by unit, this extends to convergence in mean (Billingsley, 1968, p. 32).

## A. 8 Results for split-half SIS

Proof of Theorem 4.2. Define gauges for each of the sub-samples as

$$
\hat{\gamma}_{1 n}=\frac{1}{n_{1}^{\circ}} \sum_{i \in I_{1}^{\circ}} 1_{\left(\left|\nabla y_{i}-\hat{\beta}_{2}^{\circ} \nabla x_{i}\right| \geq \sqrt{2} \hat{\sigma}_{2} \omega_{2, i} c\right)}, \quad \hat{\gamma}_{2 n}=\frac{1}{n_{2}^{\circ}} \sum_{i \in I_{2}^{\circ}} 1_{\left(\left|\nabla y_{i}-\hat{\beta}_{1}^{\prime} \nabla x_{i}\right| \geq \sqrt{2} \hat{\sigma}_{1} \omega_{1, i} c\right)},
$$

noting that $n^{\circ} \hat{\gamma}_{n}^{\text {split }}=n^{\circ} \hat{\gamma}_{1 n}+n^{\circ} \hat{\gamma}_{2 n}$. As in the proof of Theorem 3.3, we can apply Lemma A. 10 and set $\omega_{j, i}=1$ and ignore the resulting remainder terms.

Apply Theorem 3.4 to each of $\hat{\gamma}_{1 n}, \hat{\gamma}_{2 n}$ noting that its derivation does not depend on the ordering of two sub-samples. This requires Assumptions 3.1, 4.1. We get expansions,
with $(j, k)=(1,2)$ or $(j, k)=(2,1)$,

$$
\begin{aligned}
n_{j}^{1 / 2}\left(\hat{\gamma}_{j n}-\gamma\right)= & n_{j}^{-1 / 2} \sum_{i \in I_{j}^{\circ}}\left\{1_{\left(\left|\chi_{i}\right| \geq c\right)}-\mathrm{E} 1_{\left(\left|\chi_{i}\right| \geq c\right)}\right\}-c \mathrm{~h}(c)\left(n_{j} / n_{k}\right)^{1 / 2} n_{k}^{-1 / 2} \sum_{i \in I_{k}}\left(\varepsilon_{i}^{2} / \sigma^{2}-1\right) \\
& -\mathrm{h}(c)\left\{\xi_{j n}(c)-\xi_{j n}(-c)\right\}^{\prime} N_{j}^{-1} N_{k} \widehat{\Sigma}_{k n}^{-1} \widehat{V}_{k n} /(\sqrt{2} \sigma)+\mathrm{op}(1)
\end{aligned}
$$

Use that $n_{j}^{\circ} / n_{j} \rightarrow 1$ and $n_{1}^{\circ}+n_{2}^{\circ}=n^{\circ}$ and insert the Theorem 3.4 sub-sample expansions into $\left(n^{\circ}\right)^{1 / 2}\left(\hat{\gamma}_{n}^{\text {split }}-\gamma\right)=\left(n_{1}^{\circ} / n^{\circ}\right)^{1 / 2}\left(n_{1}^{\circ}\right)^{1 / 2}\left(\hat{\gamma}_{1 n}-\gamma\right)+\left(n_{2}^{\circ} / n^{\circ}\right)^{1 / 2}\left(n_{2}^{\circ}\right)^{1 / 2}\left(\hat{\gamma}_{2 n}-\gamma\right)$. Asymptotically, we can replace $n_{j}^{\circ}, I_{j}^{\circ}$ with $n_{j}, I_{j}$.

The proof of Theorem 4.4 uses the following non-stationary mixingale result
Lemma A.11. (McLeish, 1977, Theorem 2.4) Let $X_{n i}$ for $i, n=1,2, \ldots$ be a double array of zero mean random variables. Let $k_{n}(t)$ be a sequence of nonrandom integer valued, nondecreasing, right continuous functions on $[0, \infty)$. Suppose a double array of constants $\sigma_{n i}^{2}>0$ exists such that for each $T<\infty$ :
(a) $\sup _{s<t<T} \lim \sup _{n \rightarrow \infty} \sum_{k_{n}(s)}^{k_{n}(t)} \sigma_{n i}^{2} /(t-s)<\infty$;
(b) $\left\{X_{n i}^{2} / \sigma_{n i}^{2} ; n=1,2, \ldots, i \leq k_{n}(T)\right\}$ is a uniformly integrable set;
(c) $\max _{i \leq k_{n}(T)} \sigma_{n i} \rightarrow 0$ as $n \rightarrow \infty$.
(d) $\mathrm{E}\left|\mathrm{E}\left\{\left(\sum_{i=k_{n}(t)}^{k_{n}(u)} X_{n i}\right)^{2} \mid \mathcal{F}_{n, k_{n}(s)}\right\}-(u-t)\right| \rightarrow 0$ as $n \rightarrow \infty$ for each $s<t<u$.

Further, $X_{n i}$ is a mixingale with respect to $\sigma$-fields $\mathcal{F}_{n i}$ that are nondecreasing in $i$ and a vanishing sequence of constants $\psi_{n}>0$ so that for all $n, i, k+1 \geq 1$ then
(e) $\mathrm{E}\left\{\mathrm{E}\left(X_{n i} \mid \mathcal{F}_{n, i-k}\right)\right\}^{2} \leq \psi_{k}^{2} \sigma_{n i}^{2}$;
(f) $\mathrm{E}\left\{X_{n i}-\mathrm{E}\left(X_{n i} \mid \mathcal{F}_{n, i+k}\right)\right\}^{2} \leq \psi_{k+1}^{2} \sigma_{n i}^{2}$;
(g) $\sum_{k=1}^{\infty}\left(\sum_{n=0}^{k} \psi_{n}^{-2}\right)^{-1 / 2}<\infty$, which is satisfied when $\sum_{k=1}^{\infty} \psi_{k}<\infty$,

Then, $W_{n}(t)=\sum_{i=1}^{k_{n}(t)} X_{n i}$ converges weakly to a standard Wiener process in the Stone (1963) topology on the space of right continuous function with left limits, $D[0, \infty)$.

Proof of Theorem 4.4. We will rewrite the Theorem 4.2 expansion as

$$
\begin{equation*}
\sqrt{n^{\circ}}\left(\hat{\gamma}_{n}^{\text {split }}-\gamma\right)=\frac{1}{\sqrt{n^{\circ}}} \sum_{j=1}^{2} \sum_{i \in I_{j}^{\circ}}\left(d_{j n}^{\prime} s_{i}-d_{3 j n}^{\prime} v_{i n}\right)+\mathrm{OP}(1) \tag{A.20}
\end{equation*}
$$

where $s_{i}$ is a stationary mixingale, $d_{j n}$ and its third component $d_{3 j n}$ are deterministic with different levels for the two sample periods, and $v_{i n}$ is a residual term. We will apply Lemma A. 11 to the sum of $X_{n i}=d_{j n}^{\prime} s_{i} / \sqrt{n^{\circ}}$ in (A.20) with filtration $\mathcal{F}_{n i}=\mathcal{F}_{i}$ as defined in Assumption 3.1. We then argue that $\sum_{j=1}^{2} \sum_{i \in I_{j}^{\circ}} d_{3 j n} v_{i n} / \sqrt{n^{\circ}}$ vanishes.

1. Mixingale expansion and notation. The expansion in Theorem 4.2 is

$$
\begin{aligned}
& \sqrt{n}\left(\hat{\gamma}_{n}^{\text {split }}-\gamma\right)=n^{-1 / 2} \sum_{i=1}^{n-1}\left\{1_{\left(\left|x_{i}\right| \geq c\right)}-\gamma\right\} \\
&- c \mathrm{~h}(c) n^{-1 / 2} \sum_{i=1}^{n}\left\{n_{2} n_{1}^{-1} 1_{\left(i \in I_{1}\right)}+n_{1} n_{2}^{-1} 1_{\left(i \in I_{2}\right)}\right\}\left(\varepsilon_{i}^{2} \sigma^{-2}-1\right) \\
&-\mathrm{h}(c)(\sqrt{2} \sigma)^{-1}\left[n_{1}^{1 / 2} n^{-1 / 2}\left\{\xi_{1 n}(c)-\xi_{1 n}(-c)\right\}^{\prime} N_{1}^{-1} N_{2} \widehat{\Sigma}_{2 n}^{-1} \widehat{V}_{2 n}\right. \\
&\left.\quad+n_{2}^{1 / 2} n^{-1 / 2}\left\{\xi_{2 n}(c)-\xi_{2 n}(-c)\right\}^{\prime} N_{2}^{-1} N_{1} \widehat{\Sigma}_{1 n}^{-1} \widehat{V}_{1 n}\right]+\mathrm{op}(1)
\end{aligned}
$$

for fixed $c \in \mathbb{R}$. By Assumptions $4.1(i i i, a, b), \widehat{\Sigma}_{j n}^{-1} \rightarrow \Sigma_{j}$ and $\widehat{V}_{j n} \rightarrow V_{j}$ in distribution.
We define a stationary mixingale component

$$
s_{i}=\left(\begin{array}{c}
s_{1 i} \\
s_{2 i} \\
s_{3 i}
\end{array}\right)=\left(\begin{array}{c}
1_{\left(\left|\chi_{i}\right| \geq c\right)}-\gamma \\
\varepsilon_{i}^{2} / \sigma^{2}-1 \\
\Sigma_{x}^{-1}\left(x_{i}-\mu_{x}\right)\left(\varepsilon_{i} / \sigma-\varkappa_{1}\right)
\end{array}\right) .
$$

By Assumption $4.3(i i)$ the pairs $x_{i}, \varepsilon_{i}$ are stationary with $\mu_{x}=\mathrm{E} x_{i}$ and $\Sigma_{x}=\operatorname{Var}\left(x_{i}\right)$ so that $\Sigma_{j}=\Sigma_{x}$. Here $s_{i}$ is $\mathcal{F}_{i+1}$-adapted, while $\mathrm{E}_{i-1} s_{i}=0$, since $x_{i}$ is $\mathcal{F}_{i-1}$-adapted.

Define the deterministic component. The vector takes on two distinct values with $j=1,2$ for the observations within the two sub-samples $I_{j}$ :

$$
d_{j n}=\left(\begin{array}{c}
d_{1 j n} \\
d_{2 j n} \\
d_{3 j n}
\end{array}\right)=\left(\begin{array}{c}
1 \\
-c \mathrm{~h}(c) n_{k} n_{j}^{-1} \\
-\mathrm{h}(c) \xi_{c} n_{k} n_{j}^{-1} 2^{-1 / 2}
\end{array}\right),
$$

where $(k, j)$ for $i \in I_{1}$ is $(2,1)$ and for $i \in I_{2}$ is $(1,2)$. For $\xi_{c}$, we used that as the pairs $x_{i}, \varepsilon_{i}$ are stationary then $\mathrm{E}_{i-1}\left(\nabla x_{i} \mid \chi_{i}=c\right)$ is deterministic and constant in $i$, and thus

$$
\xi_{c}=\xi_{j n}(c)-\xi_{j n}(-c)=\mathrm{E}_{0}\left(\nabla x_{1} \mid \chi_{1}=c\right)-\mathrm{E}_{0}\left(\nabla x_{1} \mid \chi_{1}=-c\right), \text { for } j=1,2 .
$$

We note that $d_{j n}$ is finite since $n_{k} / n_{j}$ has a positive and finite limit, while Assumption 4.3(i) has that
$|c| \mathbf{f}(c)$ is bounded, hence $|c| \mathrm{h}(c)$ is bounded by Theorem 2.3(c).
Define the residual term for $i \in I_{j}$ as

$$
\begin{equation*}
v_{i n}=\left(\Sigma_{x}^{-1}-\widehat{\Sigma}_{j n}^{-1}\right)\left(x_{i}-\mu_{x}\right)\left(\varepsilon_{i} / \sigma-\varkappa_{1}\right)-\widehat{\Sigma}_{j n}^{-1}\left(\mu_{x}-\bar{x}_{j}\right)\left(\varepsilon_{i} / \sigma-\varkappa_{1}\right) . \tag{A.21}
\end{equation*}
$$

Finally, note that
$N_{j}=n_{j}^{-1 / 2} I_{\text {dim x }}$. Thus, the Theorem 4.2 expansion has the form (A.20).
2. Conditional autocovariance matrices for $s_{i}$. We have $\mathrm{E}_{i-1} s_{i}=0$ since $s_{1 i}$, $s_{2 i}$ are independent of $\mathcal{F}_{i-1}$ with zero mean, while $s_{3 i}$ has a factor with that property. Thus, for $\ell \geq 1$, we have $\mathbf{E}_{i-\ell} s_{i}=0$. The term, $\mathbf{E}_{i-1} s_{1 i}^{2}=\gamma(1-\gamma)$ is a Bernoulli variance, while $\mathrm{E}_{i-1} s_{1 i} s_{2 i}=\varsigma_{2}$ and $\mathrm{E}_{i-1} s_{2 i}^{2}=\varkappa_{4}-1$ by definitions of second and fourth moments. Similarly, $\mathrm{E}_{i-1} s_{1 i} s_{1, i+1}=\varsigma_{0}-\gamma^{2}$ and $\mathrm{E}_{i-1} s_{1 i} s_{2, i+1}=\varsigma_{2}$. Since $s_{2 i}$, $s_{3 i}$ are $\mathcal{F}_{i}$-adapted, and $\mathrm{E}_{i} s_{i+1}=0$ we get for $\ell=2,3$ that $\mathrm{E}_{i-1} s_{\ell i} s_{i+1}^{\prime}=\mathrm{E}_{i-1} s_{\ell i} \mathrm{E}_{i} s_{i+1}^{\prime}=0$. Since $s_{i}$ is $\mathcal{F}_{i+1}$-adapted, then it is also $\mathcal{F}_{i+m-1}$-adapted for $m \geq 2$. Therefore, for $m \geq 2$, we get $\mathrm{E}_{i-1} s_{i} s_{i+m}^{\prime}=\mathrm{E}_{i-1} s_{i} \mathrm{E}_{i+m-1} s_{i+m}^{\prime}=0$. Since $x_{i}$ is $\mathcal{F}_{i-1}$-adapted while $\chi_{i}$ and $\varepsilon_{i}$ are independent of $\mathcal{F}_{i-1}$, we get

$$
\mathrm{E}_{i-1} s_{3 i} s_{1 i}=\Sigma_{x}^{-1}\left(x_{i}-\mu_{x}\right) \mathrm{E} 1_{\left(\left|x_{i}\right| \geq c\right)}\left(\varepsilon_{i} / \sigma-\varkappa_{1}\right)=\Sigma_{x}^{-1}\left(x_{i}-\mu_{x}\right) \varsigma_{1},
$$

where $\varsigma_{1}=E 1_{\left(\left|\chi_{i}\right| \geq c\right)}\left(\varepsilon_{i} / \sigma-\varkappa_{1}\right)$. Similarly

$$
\begin{aligned}
& \mathrm{E}_{i-1} s_{3 i} s_{2 i}=\Sigma_{x}^{-1}\left(x_{i}-\mu_{x}\right)\left(\varkappa_{3}-\varkappa_{1}\right), \\
& \mathrm{E}_{i-1} s_{3 i} s_{3 i}^{\prime}=\Sigma_{x}^{-1}\left(x_{i}-\mu_{x}\right)\left(x_{i}-\mu_{x}\right)^{\prime} \Sigma_{x}^{-1}\left(1-\varkappa_{1}^{2}\right) .
\end{aligned}
$$

Decompose $x_{i+1}-\mu_{x}=x_{i}-\mu_{x}-x_{i}+x_{i+1}=x_{i}-\mu_{x}-\nabla x_{i}$. Write

$$
\begin{aligned}
\mathrm{E}_{i-1} s_{3, i+1} s_{1 i} & =\Sigma_{x}^{-1}\left(x_{i}-\mu_{x}\right) \mathrm{E}_{\left(\left|\chi_{i}\right| \geq c\right)}\left(\varepsilon_{i} / \sigma-\varkappa_{1}\right)-\Sigma_{x}^{-1} \mathrm{E}_{i-1} \nabla x_{i}\left(1_{\left(\left|\chi_{i}\right| \geq c\right)}-\gamma\right)\left(\varepsilon_{i} / \sigma-\varkappa_{1}\right) \\
& =\Sigma_{x}^{-1}\left(x_{i}-\mu_{x}\right) \varsigma_{1}-\Sigma_{x}^{-1} \varsigma_{1 x i},
\end{aligned}
$$

with $\varsigma_{1 x i}=\mathrm{E}_{i-1}\left\{\nabla x_{i}\left(1_{\left(\left|\chi_{i}\right| \geq c\right)}-\gamma\right)\left(\varepsilon_{i} / \sigma-\varkappa_{1}\right)\right\}$. The conditional auto product moments are $\mathrm{E}_{i-1} s_{i} s_{i+j-1}^{\prime}=0$ for $j \geq 2$ and

$$
\begin{align*}
\mathrm{E}_{i-1} s_{i} s_{i}^{\prime} & =\left(\begin{array}{ccc}
\gamma(1-\gamma) & \varsigma_{2} & \varsigma_{1}\left(x_{i}-\mu_{x}\right)^{\prime} \Sigma_{x}^{-1} \\
* & \varkappa_{4}-1 & \left(\varkappa_{3}-\varkappa_{1}\right)\left(x_{i}-\mu_{x}\right)^{\prime} \Sigma_{x}^{-1} \\
* & * & \left(1-\varkappa_{1}^{2}\right) \Sigma_{x}^{-1}\left(x_{i}-\mu_{x}\right)\left(x_{i}-\mu_{x}\right)^{\prime} \Sigma_{x}^{-1}
\end{array}\right),  \tag{A.22}\\
\mathrm{E}_{i-1} s_{i} s_{i+1}^{\prime} & =\left(\begin{array}{ccc}
\varsigma_{0}-\gamma^{2} & \varsigma_{2} & \left\{\varsigma_{1 x i}^{\prime}-\varsigma_{1}\left(x_{i}-\mu_{x}\right)^{\prime}\right\} \Sigma_{x}^{-1} \\
0 & 0 & 0 \\
0 & 0 & 0
\end{array}\right) . \tag{A.23}
\end{align*}
$$

3. Unconditional autocovariance matrices for $s_{i}$. Note that $\mathbf{E} x_{i}=\mu_{x}$ and $\operatorname{Var}\left(x_{i}\right)=\Sigma_{x}$. Apply iterated expectations noting that $\mathrm{E}_{i-1} s_{i}=0$ to get $\operatorname{Var}\left(s_{i}\right)=$ $\mathrm{EE}_{i-1} s_{i} s_{i}^{\prime}$ so that

$$
\begin{align*}
\Omega_{0}=\operatorname{Var}\left(s_{i}\right) & =\left(\begin{array}{ccc}
\gamma(1-\gamma) & \varsigma_{2} & 0 \\
\varsigma_{2} & \varkappa_{4}-1 & 0 \\
0 & 0 & \Sigma_{x}^{-1}\left(1-\varkappa_{1}^{2}\right)
\end{array}\right)  \tag{А.24}\\
\Omega_{1}=\operatorname{Cov}\left(s_{i}, s_{i+1}\right)= & \left(\begin{array}{ccc}
\varsigma_{0}-\gamma^{2} & \varsigma_{2} & \varsigma_{1 x}^{\prime} \Sigma_{x}^{-1} \\
0 & 0 & 0 \\
0 & 0 & 0
\end{array}\right) \tag{A.25}
\end{align*}
$$

as in (33), whereas $\operatorname{Cov}\left(s_{i}, s_{i+j}\right)=0$ for $j \geq 2$. We note that these autocovariances are finite when $\mathrm{E} x_{i}^{2}<\infty$ and $\mathrm{E} \varepsilon_{i}^{4}<\infty$ as assumed in Assumption 4.3(i,ii).
4. Two-level long-run variance. Let $e_{1}=(1,0,0)^{\prime}$ and define, for $j=1,2$,

$$
\begin{equation*}
\omega_{j n}^{2}=\operatorname{Var}\left(d_{j n}^{\prime} s_{i}\right)+2 \operatorname{Cov}\left(d_{j n}^{\prime} s_{i}, d_{j n}^{\prime} s_{i+1}\right)=d_{j n}^{\prime} \Omega_{0} d_{j n}+2 e_{1}^{\prime} \Omega_{1} d_{j n} \tag{A.26}
\end{equation*}
$$

For large $n$ then $n_{k} / n_{j} \rightarrow \lambda_{k} / \lambda_{j}>0$, so that $d_{j n} \rightarrow d_{j}$ and $\omega_{j n}^{2} \rightarrow \omega_{j}^{2}$, where

$$
d_{j}=\left(\begin{array}{c}
1 \\
-c \mathrm{~h}(c)\left(\lambda_{k} / \lambda_{j}\right) \\
-\mathrm{h}(c) \xi_{c}\left(\lambda_{k} / \lambda_{j}\right) / \sqrt{2}
\end{array}\right), \quad \omega_{j}^{2}=d_{j}^{\prime} \Omega_{0} d_{j}+2 e_{1}^{\prime} \Omega_{1} d_{j}
$$

as in (34). By Assumption 4.3(iii) we have that $\omega_{1}^{2}, \omega_{2}^{2}>0$. Following McLeish, consider $\mathrm{E} X_{n i}^{2}+\sum_{i \neq j} \mathrm{E} X_{n i} X_{n j}$. For $i$ which are not near $n_{1}$ this equals $\omega_{j n}^{2} / n$ for $i \in I_{j}$ by the above derivations. As $\omega_{j n} \rightarrow \omega_{j}$, it is convenient to define $\sigma_{n i}^{2}=\omega_{j}^{2} / n$ for $i \in I_{j}$.
5. Time distortion function. Cumulate to get $\sum_{i=1}^{n} \sigma_{n i}^{2}=\left(n_{1} / n\right) \omega_{1}^{2}+\left(n_{2} / n\right) \omega_{2}^{2}$. Asymptotically, this is equivalent to $T=\lambda_{1} \omega_{1}^{2}+\lambda_{2} \omega_{2}^{2}$. Define the time distortion

$$
k_{n}(t)= \begin{cases}\operatorname{int}\left(t n / \omega_{1}^{2}\right) & \text { for } t \leq \lambda_{1} \omega_{1}^{2} \\ \lambda_{1} n+\operatorname{int}\left(t-\lambda_{1} \omega_{1}^{2}\right) n / \omega_{2}^{2} & \text { for } \lambda_{1} \omega_{1}^{2}<t \leq T\end{cases}
$$

Note that $k_{n}(t)$ increases in steps of 1 from 0 to $k_{n}(T)=n$. It maps the proportion of the cumulated variance to the original observations. Following McLeish, let $W_{n}(t)=$
$\sum_{i=1}^{k_{n}(t)} X_{n i}$, so that $\sum_{i=1}^{k_{n}(T)} X_{n i}=\sum_{i=1}^{n} X_{n i}$. The long-run variance cumulates linearly for the time distorted process $W_{n}$. We get $\sum_{i=k_{n}(s)}^{k_{n}(t)} \sigma_{n i}^{2} /(t-s)=1+\mathrm{o}(1)$ for $s<t<T$, where the remainder arises from a rounding error at the break point.
6. Checking conditions of Lemma A.11. We choose mixingale coefficients $\psi_{k}^{2}=C$ for $k=0,1,2$ and some constant $C>0$ described below, while $\psi_{k}=1 / k^{2}$ for $k>2$.
(a) By the definition of $\sigma_{n i}^{2}$ we have $\sum_{i=k_{n}(s)}^{k_{n}(t)} \sigma_{n i}^{2} /(t-s) \rightarrow 1$ for $s<t<1$. Thus, $\sup _{s<t<T} \lim \sup _{n \rightarrow \infty} \sum_{i=k_{n}(s)}^{k_{n}(t)} \sigma_{n i}^{2} /(t-s)<\infty$.
(b) It suffices to show that $\mathrm{E}\left|X_{n i}\right|^{2+} / \sigma_{n i}^{2+}$ is bounded uniformly in $n, i$ (Billingsley, 1968, p. 32). We have $\mathrm{E}\left|X_{n i} / \sigma_{n i}\right|^{2+} \leq\left|d_{j n}\right|^{2+} \mathrm{E}\left|s_{i}\right|^{2+} / \omega_{j}^{2+}$ for $i \in I_{j}$. Here $\left|d_{j n}\right|$ converges in $n$ for $j=1,2$ while $\omega_{j}>0$ by Assumption $4.3(i i i)$ for $j=1,2$. Finally, $s_{i}$ is stationary with $2+$ moments since $\varepsilon_{i}^{2}$ and $x_{i}$ have $2+$ moments by Assumption $4.3(i i)$.
(c) Since $n \sigma_{n i}^{2}=\omega_{j_{i}}^{2}$ takes two values only, then $\max _{i \leq n} \sigma_{n i}^{2} \rightarrow 0$ for $n \rightarrow \infty$.
(d) Let $S_{s t u}=\mathrm{E}_{k_{n}(s)}\left(\sum_{i=k_{n}(t)}^{k_{n}(u)} X_{n i}\right)^{2}$. We check that $\mathrm{E}\left|S_{s t u}-(u-t)\right| \rightarrow 0$ for each $s<t<u$. Since $\mathrm{E}_{i-1} s_{i} s_{i+j-1}^{\prime}=0$ for $j \geq 2$ we have

$$
S_{s t u}=\mathrm{E}_{k_{n}(s)}\left\{\sum_{i=k_{n}(t)}^{k_{n}(u)} X_{n i}^{2}+2 \sum_{i=k_{n}(t)}^{k_{n}(u)-1} X_{n i} X_{n, i+1}\right\} .
$$

Recalling the expressions for $\mathrm{E}_{i-1} X_{n i} X_{n, i+j}$ and $\mathrm{E} X_{n i} X_{n, i+j}$ in items 4, 5, we get that

$$
S_{s t u}=\mathrm{E}\left\{\sum_{i=k_{n}(t)}^{k_{n}(u)} X_{n i}^{2}+2 \sum_{i=k_{n}(t)}^{k_{n}(u)-1} X_{n i} X_{n, i+1}\right\}+\text { remainder }
$$

The detailed comparison of (A.22), (A.23) and (A.24), (A.25) shows that the remainder is a linear function of components like $\mathrm{E}_{k-m} n^{-1} \sum_{i=k+1}^{k+n}\left(y_{i}-\mathrm{E} y_{i}\right)$, where $y_{i}$ represents either of the three $z_{i}$ sequences in Assumption 4.3(iv), which are $x_{i},\left(x_{i}-\mu_{x}\right)\left(x_{i}-\mu_{x}\right)^{\prime}$ and $\nabla x_{i}\left(1_{\left(\left|\chi_{i}\right| \geq c\right)}-\gamma\right)\left(\varepsilon_{i} / \sigma-\varkappa_{1}\right)$, for any $k, m, n$. These components vanish in mean by that assumption.

Now, as remarked in item 4, we have $\mathrm{E}\left(X_{n i}^{2}+2 X_{n i} X_{n, i+1}\right)=\omega_{j n} / n$ for $i \in I_{j}$, where $\omega_{j n} \rightarrow \omega_{j}=n \sigma_{n i}^{2}$ This convergence is uniform over $n, i$ using condition (b) and (Billingsley, 1968, Theorem 5.4). Thus, the desired statement $\mathrm{E}\left|S_{s t u}-(u-t)\right| \rightarrow 0$ follows as for condition (a).
(e) Let $\phi_{n,-k}=\mathrm{E}\left(\mathrm{E}_{i-k} X_{n i}\right)^{2}$. We will define a positive, vanishing sequence $\psi_{k}$ for $k \geq 0$, so that $\phi_{n,-k} \leq \psi_{k}^{2} \sigma_{n i}^{2}$ for all $k \geq 0$.

For $k>0$, we find $\mathrm{E}_{i-k} s_{i}=\mathrm{E}_{i-k} \mathrm{E}_{i-1} s_{i}=0$ so that $\phi_{n,-k}=0$. Any $\psi_{k}>0$ suffices.
For $k=0$, Jensen's inequality gives $\phi_{n,-0}=\mathrm{E}\left(\mathrm{E}_{i} X_{n i}\right)^{2} \leq \mathrm{EE}_{i} X_{n i}^{2}=\mathrm{E} X_{n i}^{2}$. Following the analysis for condition (b), we have $\mathrm{E} X_{n i}^{2} \leq C \sigma_{n i}^{2}$ for some $C>0$ and all $n, i$. We can choose $\psi_{0}^{2}>C$.
$(f)$ Let $\phi_{n,+k}=\mathrm{E}\left\{X_{n i}-\mathrm{E}_{i+k}\left(X_{n i}\right)\right\}^{2}$. The sequence $\psi_{k}$ must also satisfy that $\phi_{n,+k} \leq$ $\psi_{k+1}^{2} \sigma_{n i}^{2}$ for all $k \geq 0$.

For $k>1$, then $s_{i}$ is $\mathcal{F}_{i+k}$-adapted, so that $\phi_{n,+k}=0$. Any choice of $\psi_{k}>0$ suffices.
For $k=0,1$, bound $\phi_{n,+k}=\mathrm{EE}_{i+k}\left\{X_{n i}-\mathrm{E}_{i+k}\left(X_{n i}\right)\right\}^{2} \leq \mathrm{EE}_{i+k} X_{n i}^{2}=\mathrm{E}_{n i}^{2}$. As in condition (b), we have $\mathrm{E} X_{n i}^{2} \leq C \sigma_{n i}^{2}$ for some $C>0$ and all $n, i$. We must choose $\psi_{1}^{2}, \psi_{2}^{2} \geq C$.
(g) This holds since $\sum_{k=0}^{\infty} \psi_{k}=3 C+\sum_{k=3}^{\infty}\left(1 / k^{2}\right)<\infty$.
7. Applying Lemma A.11. We get that $\sum_{i=1}^{k_{n}(t)} X_{n i}$ converges in distribution to a standard Brownian motion. In particular $\sum_{i=1}^{k_{n}(T)} X_{n i}$ is asympotically $\mathrm{N}(0, B)$, where

$$
\begin{aligned}
B=\lambda_{1} \omega_{1}^{2}+\lambda_{2} \omega_{2}^{2} & =\lambda_{1}\left(d_{1}^{\prime} \Omega_{0} d_{1}+2 e_{1}^{\prime} \Omega_{1} d_{1}\right)+\lambda_{2}\left(d_{2}^{\prime} \Omega_{0} d_{2}+2 e_{1}^{\prime} \Omega_{1} d_{2}\right) \\
& =\lambda_{1} d_{1}^{\prime} \Omega_{0} d_{1}+\lambda_{2} d_{2}^{\prime} \Omega_{0} d_{2}+2 e_{1}^{\prime} \Omega_{1}\left(\lambda_{1} d_{1}+\lambda_{2} d_{2}\right)
\end{aligned}
$$

Inserting the expressions for $\Omega_{0}, \Omega_{1}, d_{j}$ in (33), (34), see also (A.24), (A.25), (A.26), gives

$$
\begin{aligned}
B= & \left(\lambda_{1}+\lambda_{2}\right) \gamma(1-\gamma)-2 \operatorname{ch}(c)\left(\lambda_{1}+\lambda_{2}\right) \varsigma_{2} \\
& +\{\operatorname{ch}(c)\}^{2}\left(\lambda_{1}^{2} / \lambda_{2}+\lambda_{2}^{2} / \lambda_{1}\right)\left(\varkappa_{4}-1\right)+\{\mathrm{h}(c)\}^{2}\left(1-\varkappa_{1}^{2}\right)\left(\lambda_{1}^{2} / \lambda_{2}+\lambda_{2}^{2} / \lambda_{1}\right) \xi_{c}^{\prime} \Sigma_{x}^{-1} \xi_{c} / 2 \\
& +2\left(\lambda_{1}+\lambda_{2}\right)\left(\varsigma_{0}-\gamma^{2}\right)-2 c h(c)\left(\lambda_{1}+\lambda_{2}\right) \varsigma_{2}-\sqrt{2} \mathrm{~h}(c)\left(\lambda_{1}+\lambda_{2}\right) \varsigma_{1 x}^{\prime} \Sigma_{x}^{-1} \xi_{c} .
\end{aligned}
$$

Noting that $\lambda_{1}+\lambda_{2}=1$ this reduces to the desired expression in (36), which is

$$
\begin{aligned}
B= & \gamma(1-\gamma)+2\left(\varsigma_{0}-\gamma^{2}\right)-4 c h(c) \varsigma_{2}-\sqrt{2} \mathrm{~h}(c) \varsigma_{1 x}^{\prime} \Sigma_{x}^{-1} \xi_{c} \\
& +\left(\lambda_{1}^{2} / \lambda_{2}+\lambda_{2}^{2} / \lambda_{1}\right)\{\mathrm{h}(c)\}^{2}\left\{c^{2}\left(\varkappa_{4}-1\right)+\left(1-\varkappa_{1}^{2}\right) \xi_{c}^{\prime} \Sigma_{x}^{-1} \xi_{c} / 2\right\} .
\end{aligned}
$$

8. Remainder term. We argue that $n^{-1 / 2} \sum_{j=1}^{2} \sum_{i \in I_{j}} d_{3 j n} v_{i n}$ vanishes, where

$$
v_{i n}=\left(\Sigma_{x}^{-1}-\widehat{\Sigma}_{j n}^{-1}\right)\left(x_{i}-\mu_{x}\right)\left(\varepsilon_{i} / \sigma-\varkappa_{1}\right)-\widehat{\Sigma}_{j n}^{-1}\left(\mu_{x}-\bar{x}_{j}\right)\left(\varepsilon_{i} / \sigma-\varkappa_{1}\right), \text { for } i \in I_{j}
$$

see (A.21). We have $\widehat{\Sigma}_{j n} \rightarrow \Sigma_{x}$ and $\bar{x}_{j} \rightarrow \mu_{x}$ in probability by Assumptions 3.2(iiia), 4.1(a), 4.3(v). Further, $n^{-1 / 2} \sum_{i=1}^{n}\left(x_{i}-\mu_{x}\right)\left(\varepsilon_{i} / \sigma-\varkappa_{1}\right)=n^{-1 / 2} \sum_{i=1}^{n} \Sigma_{x} s_{3 i}$ and is asymptotically normal using the above mixingale considerations. Finally, $n^{-1 / 2} \sum_{i=1}^{n}\left(\varepsilon_{i} / \sigma-\varkappa_{1}\right)$ converges by a standard Central Limit Theorem.

Proof of Theorem 4.5. The third term in the expansion of Theorem 4.2 vanishes since $\xi_{j n}$ vanishes. We can then proceed exactly as in the proof of Theorem 4.4, but dropping any consideration to the third term in the expansion and in the assumptions.

## A. 9 Explicit formulas for stationary case

Example 4.1. Derivation of $\varsigma_{0}$ in (35). Write $\varsigma_{0}$ as $\mathrm{E}\left\{1-1_{\left(\left|\chi_{i}\right|<c\right)}\right\}\left\{1-1_{\left(\left|\chi_{i+1}\right|<c\right)}\right\}=$ $1-2(1-\gamma)+\mathcal{I}_{c}$, where $\mathcal{I}_{c}=\mathrm{E} 1_{\left(\left|\chi_{i}\right|<c\right)} 1_{\left(\left|\chi_{i+1}\right|<c\right)}$. Conditional on $\varepsilon_{i+1}$ the two indicators are i.i.d. with expectation $\Phi\left(\varepsilon_{i+1}+c \sqrt{2}\right)-\Phi\left(\varepsilon_{i+1}-c \sqrt{2}\right)$. Thus,

$$
\mathcal{I}_{c}=\int_{\mathbb{R}} \varphi(x)\{\Phi(x+c \sqrt{2})-\Phi(x-c \sqrt{2})\}^{2} d x
$$

Following Owen (1980, 2.2; 2.8) let $T(h, a)=\int_{h}^{\infty} \varphi(x) \int_{0}^{a x} \varphi(y) d y d x$. By Owen (1980, $20,010.3)$ then $2 \int_{\mathbb{R}} \varphi(x) \Phi(x+u \sqrt{2}) \Phi(x+v \sqrt{2}) d x=\Phi(u)+\Phi(v)-2 T\{u,(2 v / u-1) / \sqrt{3}\}-$ $2 T\{v,(2 u / v-1) / \sqrt{3}\}-1_{(u v<0)}$ for $u, v \neq 0$, so that

$$
\begin{aligned}
\mathcal{I}_{c}=\{\Phi(c)-2 T(c, 1 / \sqrt{3})\} & +\{\Phi(-c)-2 T(-c, 1 / \sqrt{3})\} \\
& -2\left\{\frac{1}{2} \Phi(c)+\frac{1}{2} \Phi(-c)-T(c,-\sqrt{3})-T(-c,-\sqrt{3})-1 / 2\right\} .
\end{aligned}
$$

By Owen $(1980,2.5 ; 2.6)$ then $T(-h, a)=T(h, a)$ and $T(h,-a)=-T(h, a)$. Thus,

$$
\mathcal{I}_{c}=1-4\{T(c, 1 / \sqrt{3})+T(c, \sqrt{3})\}
$$

so that $\varsigma_{0}=1-2(1-\gamma)+1-4\{T(c, 1 / \sqrt{3})+T(c, \sqrt{3})\}=2 \gamma-4\{T(c, 1 / \sqrt{3})+T(c, \sqrt{3})\}$.
Derivation of $\varsigma_{2}$ in (35). Recall $\varsigma_{2}=\mathrm{E}\left\{1_{\left(\left|\chi_{i}\right| \geq c\right)}\left(\varepsilon_{i}^{2} / \sigma^{2}-1\right)\right\}$. Since $\mathrm{E}\left\{\left(\varepsilon_{i}^{2} / \sigma^{2}-1\right)\right\}=0$, then $\varsigma_{2}=-\mathcal{I}_{c}$ where

$$
\mathcal{I}_{c}=\mathrm{E}\left\{1_{\left(\left|\chi_{i}\right|<c\right)}\left(\varepsilon_{i}^{2} / \sigma^{2}-1\right)\right\}=\int_{\mathbb{R}} \varphi(s) \int_{s-c \sqrt{2}}^{s+c \sqrt{2}}\left(t^{2}-1\right) \varphi(t) d t d s
$$

As $\varphi(t)$ has derivates $\dot{\varphi}(t)=-t \varphi(t)$ and $\ddot{\varphi}(t)=\left(t^{2}-1\right) \varphi(t)$ then

$$
\mathcal{I}_{c}=-\int_{\mathbb{R}} \varphi(s)\{(s+c \sqrt{2}) \varphi(s+c \sqrt{2})-(s-c \sqrt{2}) \varphi(s-c \sqrt{2})\} d s
$$

Since $\varphi(x) \varphi(x+c \sqrt{2})=\varphi(c) \varphi(x \sqrt{2}+c)$ and $\varphi(x) \varphi(x-c \sqrt{2})=\varphi(c) \varphi(x \sqrt{2}-c)$, we get

$$
\mathcal{I}_{c}=-\varphi(c) \int_{\mathbb{R}}\{(s+c \sqrt{2}) \varphi(s \sqrt{2}+c)-(s-c \sqrt{2}) \varphi(s \sqrt{2}-c)\} d s
$$

Substituting $t=s \sqrt{2}+c$ and $u=s \sqrt{2}-c$, we get $\varsigma_{2}=-c \varphi(c)$ since

$$
\mathcal{I}_{c}=-\frac{1}{2} \varphi(c)\left[\int_{\mathbb{R}}(t+c) \varphi(t) d t-\int_{\mathbb{R}}(u-c) \varphi(u) d u\right]=-c \varphi(c) .
$$

Example 4.2. Let $y_{i}=\mu+\alpha y_{i-1}+\varepsilon_{i}$ where $\varepsilon_{i} / \sigma$ is standard normal, $|\alpha|<1$ and the stationary distribution is normal with mean $\mu_{y}=\mu /(1-\alpha)$ and variance $\sigma_{y}^{2}=$ $\sigma^{2} /\left(1-\alpha^{2}\right)$. Note $\nabla x_{i}=y_{i-1}-y_{i}=(1-\alpha) y_{i-1}-\varepsilon_{i}$.

Proof that $\sigma_{\nabla \chi}=-\sigma / \sqrt{2}$. Recall that $\sigma_{\nabla \chi}=\operatorname{Cov}\left(\nabla x_{1}, \chi_{1} \mid \mathcal{F}_{0}\right)$. Insert the expressions for $x_{1}, \chi_{1}$ to get $\sigma_{\nabla \chi}=\operatorname{Cov}\left\{(1-\alpha) y_{0}-\varepsilon_{1},\left(\varepsilon_{1}-\varepsilon_{2}\right) /\left(2^{1 / 2} \sigma\right)\right\}=-\sigma / \sqrt{2}$.

We show that $\varsigma_{1}=\mathrm{E}\left\{\left(1_{\left(\left|\chi_{i}\right| \geq c\right)}-\gamma\right)\left(\varepsilon_{i} / \sigma-\varkappa_{1}\right)\right\}=0$ for a symmetric density. We get $\mathrm{E} 1_{\left(\varepsilon_{i}-\varepsilon_{i+1} \leq-q\right)} \varepsilon_{i}=\mathrm{E} 1_{\left(-\varepsilon_{i}+\varepsilon_{i+1} \leq-q\right)}\left(-\varepsilon_{i}\right)$ by the symmetry and independence of $\varepsilon_{i}, \varepsilon_{i+1}$. Change sign in the right indicator and combine the two expectations to get $\varsigma_{1}=0$.

Proof that $\varsigma_{1 x}=-\sigma \varsigma_{2}$. Let $\varsigma_{1 x}=\mathrm{E}\left\{\nabla x_{i}\left(1_{\left(\left|\chi_{i}\right| \geq c\right)}-\gamma\right)\left(\varepsilon_{i} / \sigma-\varkappa_{1}\right)\right\}$. We show $\varsigma_{1 x}=$ $-\sigma \varsigma_{2}$. Write $\nabla x_{i}=(1-\alpha) y_{i-1}-\varepsilon_{i}$ as $\varsigma_{1 x}=(1-\alpha) \varsigma_{1 x 1}-\varsigma_{1 x 2}$. Since $y_{i-1}$ is $\mathcal{F}_{i-1^{-}}$ measurable while $\varepsilon_{i}, \chi_{i}$ are independent of $\mathcal{F}_{i-1}$, we get $\varsigma_{1 x 1}=\left(\mathrm{E} y_{i-1}\right) \varsigma_{1}=0$ by the above result. Further, $\varsigma_{1 x 2}=\mathrm{E}\left\{\varepsilon_{i}\left(1_{\left(\left|\chi_{i}\right| \geq c\right)}-\gamma\right)\left(\varepsilon_{i} / \sigma-\varkappa_{1}\right)\right\}$ satisfies

$$
\begin{aligned}
& \varsigma_{1 x 2}=\sigma \mathrm{E}\left\{1_{\left(\left|x_{i}\right| \geq c\right)}\left(\varepsilon_{i}^{2} / \sigma^{2}-1\right)\right\}-\sigma \gamma \mathrm{E}\left\{\left(\varepsilon_{i}^{2} / \sigma^{2}-1\right)\right\} \\
&+\sigma \mathrm{E}\left\{\left(1_{\left(\left|x_{i}\right| \geq c\right)}-\gamma\right)\left(1-\varkappa_{1}^{2}\right)-\sigma \varkappa_{1} \mathrm{E}\left\{\left(1_{\left(\left|\chi_{i}\right| \geq c\right)}-\gamma\right)\left(\varepsilon_{i} / \sigma-\varkappa_{1}\right)\right\}\right.
\end{aligned}
$$

The first term equals $\sigma \varsigma_{2}$, the next two terms are zero by (27), (21), and the last term is 0 as $\varsigma_{1}=0$. Thus, $\varsigma_{1 x}=-\sigma \varsigma_{2}$.

We check condition (iv). Write $y_{i}-\mu_{y}=\sum_{j=0}^{i-k+m-1} \alpha^{j} \varepsilon_{i-j}+\alpha^{i-k+m}\left(y_{k-m}-\mu_{y}\right)$.
For the case $z_{i}=y_{i}$ note that $z_{i k m} \equiv \mathrm{E}_{k-m}\left(y_{i}-\mathrm{E} y_{i}\right)=\alpha^{i-k+m}\left(y_{k-m}-\mu_{y}\right)$, so that $s_{k m} \equiv n^{-1} \sum_{i=k+1}^{k+n} z_{i k m}=n^{-1} \alpha^{m+1}\left(y_{k-m}-\mu_{y}\right)\left(1-\alpha^{n}\right) /(1-\alpha)$, which converges to zero in mean as $\min (k, m, n) \rightarrow \infty$.

For the case $z_{i}=\left(y_{i}-\mu_{y}\right)^{2}$ note that $z_{i k m} \equiv \mathrm{E}_{k-m}\left\{\left(y_{i}-\mu_{y}\right)^{2}-\mathrm{E}\left(y_{i}-\mu_{y}\right)^{2}\right\}=$ $\alpha^{2(i-k+m)}\left\{\left(y_{k-m}-\mu_{y}\right)^{2}-\sigma_{y}^{2}\right\}$, so that $s_{k m} \equiv n^{-1} \sum_{i=k+1}^{k+n} z_{i k m}=n^{-1} \alpha^{2(m+1)}\left\{\left(y_{k-m}-\right.\right.$ $\left.\left.\mu_{y}\right)^{2}-\sigma_{y}^{2}\right\}\left(1-\alpha^{2 n}\right) /\left(1-\alpha^{2}\right)$, which converges to zero in mean as $\min (k, m, n) \rightarrow \infty$.

For the case $z_{i}=y_{i}^{2}$ note that $y_{i}^{2}-\mathbf{E} y_{i}^{2}=2 \mu_{y}\left(y_{i}-\mu_{y}\right)+\left\{\left(y_{i}-\mu_{y}\right)^{2}-\sigma_{y}^{2}\right\}$ and use the previous two cases.

For the case $z_{i}=\nabla y_{i-1} 1_{\left(\left|\chi_{i}\right| \geq c\right)}$ write $\nabla y_{i-1}=y_{i-1}-y_{i}=(1-\alpha) y_{i-1}-\varepsilon_{i}$ so that $z_{i}=z_{1 i}-z_{1 i}$ where $z_{1 i}=(1-\alpha) y_{i-1} 1_{\left(\left|\chi_{i}\right| \geq c\right)}$ and $z_{2 i}=\varepsilon_{i} 1_{\left(\left|\chi_{i}\right| \geq c\right)}$. Since $z_{1 i}$ is a product of independent terms, we get $z_{1 i k m} \equiv \mathrm{E}_{k-m}\left(z_{1 i}-\mathrm{E} z_{1 i}\right)=(1-\alpha) \gamma \mathrm{E}_{k-m}\left(y_{i}-\mathrm{E} y_{i}\right)$ and the above results can be used. Since $z_{2 i}$ is independent of $\rangle-\infty$, we get $z_{2 i k m} \equiv$ $\mathrm{E}_{k-m}\left(z_{1 i}-\mathrm{E} z_{2 i}\right)=0$.

## A. 10 Proof of Poisson results

We write $a_{n} \sim b_{n}$ if $a_{n} / b_{n} \rightarrow 1$. We state a special case of Chen (1975, Theorem 4.3).
Lemma A. 12 (Chen, 1975). Suppose $\varepsilon_{i}$ are i.i.d. so that $\chi_{i}=\nabla \varepsilon_{i} /(\sqrt{2} \sigma)$ satisfies $\mathrm{P}\left(\left|\chi_{i}\right|>c_{n}\right)=\lambda / n$. Suppose Assumption 5.1(i,d). Then $\sum_{i=1}^{n} 1_{\left(\left|\chi_{i}\right|>c_{n}\right)} \xrightarrow{\mathrm{D}}$ Poisson $(\lambda)$.

We check the condition for normal variables
Lemma A.13. If $\varepsilon_{i} / \sigma$ are i.i.d. standard normal then $n\left\{\mathrm{E}_{\left(\left|\chi_{i}\right|>c_{n}\right)} 1_{\left(\left|\chi_{i+1}\right|>c_{n}\right)}\right\} \rightarrow 0$.
Proof of Lemma A.13. Since $\varepsilon_{i} / \sigma$ are i.i.d. standard normal, then

$$
\binom{\chi_{i}}{\chi_{i+1}}=\frac{1}{\sqrt{2} \sigma}\left(\begin{array}{ccc}
1 & -1 & 0 \\
0 & 1 & -1
\end{array}\right)\left(\begin{array}{c}
\varepsilon_{i} \\
\varepsilon_{i+1} \\
\varepsilon_{i+2}
\end{array}\right) \stackrel{\mathrm{D}}{=} \mathrm{N}\left\{0,\left(\begin{array}{cc}
1 & -1 / 2 \\
-1 / 2 & 1
\end{array}\right)\right\}
$$

We can bound the covariance matrix, $\Omega_{1}$, in terms of positive definite ordering, that is, for any 2 -vector $v \neq 0$, then by

$$
v^{\prime} \Omega_{1}^{-1} v=v^{\prime}\left(\begin{array}{cc}
1 & -1 / 2 \\
-1 / 2 & 1
\end{array}\right)^{-1} v>v^{\prime}\left(\begin{array}{ll}
4 & 0 \\
0 & 4
\end{array}\right)^{-1} v=v^{\prime} \Omega_{2}^{-1} v
$$

Thus, we find

$$
\begin{align*}
\mathcal{P}_{1}= & \mathrm{E} 1_{\left(\left|\chi_{i}\right|>c_{n}\right)} 1_{\left(\left|\chi_{i+1}\right|>c_{n}\right)}=\iint_{\left|v_{1}\right|,\left|v_{2}\right|>c_{n}} \frac{1}{2 \pi\left(\operatorname{det} \Omega_{1}\right)^{1 / 2}} \exp \left(-\frac{1}{2} v^{\prime} \Omega_{1}^{-1} v\right) d v_{1} d v_{2} \\
& <\left(\frac{\operatorname{det} \Omega_{2}}{\operatorname{det} \Omega_{1}}\right)^{1 / 2} \iint_{\left|v_{1}\right|,\left|v_{2}\right|>c_{n}} \frac{1}{2 \pi\left(\operatorname{det} \Omega_{2}\right)^{1 / 2}} \exp \left(-\frac{1}{2} v^{\prime} \Omega_{2}^{-1} v\right) d v_{1} d v_{2}=\mathcal{P}_{2} . \tag{A.27}
\end{align*}
$$

Substituting $z^{\prime} z=v^{\prime} \Omega_{2}^{-1} v$, that is $z_{j}=v_{j} / 2$ and $d z_{j}=d v_{j} / 2$, so that

$$
\mathcal{P}_{1}<\mathcal{P}_{2}=\left(\frac{\operatorname{det} \Omega_{2}}{\operatorname{det} \Omega_{1}}\right)^{1 / 2}\left\{\int_{|z|>c_{n} / 2} \varphi(z) d z\right\}^{2}=\sqrt{\frac{64}{3}}\left\{1-\Phi\left(c_{n} / 2\right)\right\}^{2}
$$

By Mill's ratio, it holds that $x\{1-\Phi(x)\} \sim \varphi(x)$ for $x \rightarrow \infty$ (Sampford, 1953), so that $\log x \sim \log \varphi(x)-\log \{1-\Phi(x)\}$. Apply for $x=c_{n}=\Phi^{-1}\{1-\lambda /(2 n)\}$ while
recalling the expression for the normal density and noting $1-\Phi\left(c_{n}\right)=\lambda /(2 n)$, to get $2 \log c_{n} \sim-\log (2 \pi)-c_{n}^{2}-2 \log \{\lambda /(2 n)\}$. This implies $c_{n}^{2} \sim-2 \log \{\lambda /(2 n)\} \sim 2 \log n$ noting that $\log c_{n}=\mathrm{o}\left(c_{n}\right)$ and $2 \pi=O(1)$. We then expand the normal density as

$$
\varphi\left(c_{n} / 2\right) \sim(2 \pi)^{-1 / 2} \exp \{-(2 \log n) / 4\}=\mathrm{O}\left(n^{-1 / 2}\right) .
$$

Insert in Mill's ratio, that is $\left\{1-\Phi\left(c_{n} / 2\right)\right\} \sim \varphi\left(c_{n} / 2\right) /\left(c_{n} / 2\right)$, to get $1-\Phi\left(c_{n} / 2\right)=$ $\mathrm{O}\left\{(n \log n)^{-1 / 2}\right\}$. Insert this in the bound (A.27) to get $n \mathcal{P}_{1}=\mathrm{o}(1)$.

We extract the following result from Johansen \& Nielsen (2016b), see item 3 in the proof of their Theorem 8 as well as their Remark 2.

Lemma A. 14 (Johansen \& Nielsen, 2016b). Consider a continuous random variable $\chi$ with distribution function H and density h . Given $\lambda>0$ choose $c_{n}$ so that $\mathrm{P}\left(|\chi|>c_{n}\right)=$ $\lambda / n$. Suppose, as $n \rightarrow \infty$,
(a) $\mathrm{E}|\chi|^{r}<\infty$ for some $r>4$;
(b) $\mathrm{h}\left(c_{n}\right) /\left[c_{n}\left\{1-\mathrm{H}\left(c_{n}\right)\right\}\right]=\mathrm{O}(1)$;
(c) $\mathrm{h}\left(c_{n}-n^{-1 / 4} A\right) / \mathrm{h}\left(c_{n}\right)=\mathrm{O}(1)$ for all $A>0$.

Then, for all $A>0$, as $n \rightarrow \infty$,

$$
n \mathrm{E} 1_{\left(c_{n}-n^{-1 / 4} A \leq|\chi| \leq c_{n}+n^{-1 / 4} A\right)} \rightarrow 0 .
$$

The conditions (a)-(c) are satisfied if $\chi$ is normal.
We need a modification of Lemma 1.11 in Johansen \& Nielsen (2009).
Lemma A.15. If $|a|+|b|<\zeta$ and $c>\zeta>0$ then $\left|1_{(|\chi-b|>c+a)}-1_{(|\chi|>c)}\right| \leq 1_{(c-\zeta \leq|\chi| \leq c+\zeta)}$.
Proof of Lemma A.15. Let $\mathcal{D}=\left|1_{(|\chi-b|>c+a)}-1_{(|\chi|>c)}\right|$. Using that $1_{(\chi>c)}=1-1_{(\chi \leq c)}$, we get $\mathcal{D}=\left|1_{(|\chi-b| \leq c+a)}-1_{(|\chi| \leq c)}\right|$. Write out as

$$
\mathcal{D}=\left|1_{(-c-a+b \leq \chi \leq c+a+b)}-1_{(-c \leq \chi \leq c)}\right|=\left|1_{(\chi \leq c+a+b)}-1_{(\chi \leq c)}-1_{(\chi<-c-a+b)}+1_{(\chi<-c)}\right|,
$$

which can be bounded around the focal points $c$ and $-c$ by

$$
\mathcal{D} \leq 1_{(c-|a|-|b| \leq \chi \leq c+|a|+|b|)}+1_{(-c-|a|-|b| \leq \chi \leq-c+|a|+|b|)}=1_{(c-|a|-|b| \leq|\chi| \leq c+|a|+|b|)} .
$$

Using the assumption $|a|+|b| \leq \zeta$, the desired result follows.
We combine Chen's Poisson limit in Lemma A. 12 with Lemmas A.14, A.15.
Lemma A.16. Suppose the conditions of Lemmas A.14, A. 15 hold. Let $a_{i}, b_{i}$ be sequences so that $\max _{i \leq n}\left|a_{i}\right|+\max _{i \leq n}\left|b_{i}\right|=\mathrm{O}_{\mathrm{P}}\left(n^{-1 / 4}\right)$. Then

$$
\sum_{i=1}^{n} 1_{\left(\left|\chi_{i}-b_{i}\right|>c_{n}+a_{i}\right)}=\sum_{i=1}^{n} 1_{\left(\left|\chi_{i}\right|>c_{n}\right)}+\mathrm{op}(1) \xrightarrow{\mathrm{D}} \text { Poisson }(\lambda) .
$$

Proof of Lemma A.16. Let $\gamma_{n}=\sum_{i=1}^{n} 1_{\left(\left|\chi_{i}-b_{i}\right|>\sigma c_{n}+a_{i}\right)}$. Add and subtract $\sum_{i=1}^{n} 1_{\left(\left|\chi_{i}\right|>c_{n}\right)}$ and let $n$-vectors $a, b$ represent $a_{i}$ and $b_{i}$ for $i \leq n$ to get

$$
\gamma_{n}=\sum_{i=1}^{n} 1_{\left(\left|\chi_{i}\right|>\sigma c_{n}\right)}+\mathcal{R}_{n}(a, b) \quad \text { where } \quad \mathcal{R}_{n}(a, b)=\sum_{i=1}^{n} 1_{\left(\left|\chi_{i}-b_{i}\right|>\sigma c_{n}+a_{i}\right)}-1_{\left(\left|\chi_{i}\right|>\sigma c_{n}\right)} .
$$

The first term is asymptotically Poisson $(\lambda)$ distributed by Lemma A.12. We show that the second term vanishes.

Since $\max _{i \leq n}\left|a_{i}\right|+\max _{i \leq n}\left|b_{i}\right|=\mathrm{O}_{\mathrm{P}}\left(n^{-1 / 4}\right)$, we can, for any $\epsilon>0$ and sufficiently large $n$, construct a sequence of sets $\mathcal{S}_{n}$ with $\mathrm{P}\left(\mathcal{S}_{n}^{c}\right) \leq \epsilon$, so that $\left|a_{i}\right|+\left|b_{i}\right| \leq A n^{-1 / 4}$ for $i \leq n$ on $\mathcal{S}_{n}$. We find

$$
\mathrm{P}\left\{\left|\mathcal{R}_{n}(a, b)\right|>\epsilon\right\}=\mathrm{P}\left[\left\{\left|\mathcal{R}_{n}(a, b)\right|>\epsilon\right\} \cap \mathcal{S}_{n}\right\} \mathrm{P}\left(\mathcal{S}_{n}\right)+\mathrm{P}\left[\left\{\left|\mathcal{R}_{n}(a, b)\right|>\epsilon\right\} \cap \mathcal{S}_{n}^{c}\right] \mathrm{P}\left(\mathcal{S}_{n}^{c}\right) .
$$

Bounding $\mathrm{P}\left(\mathcal{S}_{n}\right)$ and $\mathrm{P}\left[\left\{\left|\mathcal{R}_{n}(a, b)\right|>\epsilon\right\} \cap \mathcal{S}_{n}^{c}\right]$ by unity and the last probability by $\epsilon$ gives

$$
\mathrm{P}\left\{\left|\mathcal{R}_{n}(a, b)\right|>\epsilon\right\} \leq \mathcal{P}_{n}+\epsilon \quad \text { where } \quad \mathcal{P}_{n}=\mathrm{P}\left[\left\{\left|\mathcal{R}_{n}(a, b)\right|>\epsilon\right\} \cap \mathcal{S}_{n}\right] .
$$

It suffices to show that $\mathcal{P}_{n}$ is small. We rewrite $\mathcal{P}_{n}$. On the set $\mathcal{S}_{n}$, we apply first the triangle inequality and then Lemma A. 15 to get

$$
\left|\mathcal{R}_{n}(a, b)\right| \leq \sum_{i=1}^{n}\left|1_{\left(\left|\chi_{i}-b_{i}\right|>\sigma c_{n}+a_{i}\right)}-1_{\left(\left|\chi_{i}\right|>\sigma c_{n}\right)}\right| \leq \sum_{i=1}^{n}\left|1_{\left(c_{n}-A n^{-1 / 4} \leq\left|\chi_{i}\right| \leq c_{n}+A n^{-1 / 4}\right)}\right|=\mathcal{R}_{n}^{*} .
$$

Thus, $\mathcal{P}_{n} \leq \mathrm{P}\left\{\left(\mathcal{R}_{n}^{*}>\epsilon\right) \cap \mathcal{S}_{n}\right\} \leq \mathrm{P}\left(\mathcal{R}_{n}^{*}>\epsilon\right)=\mathcal{P}_{n}^{*}$. It suffices to show that $\mathcal{P}_{n}^{*}$ vanishes. Using the Markov inequality and then Lemma A. 14 gives

$$
\mathcal{P}_{n}^{*} \leq \frac{1}{\epsilon} \mathrm{E} \mathcal{R}_{n}^{*}=\frac{1}{\epsilon} n \mathrm{E}\left|1_{\left(c_{n}-A n^{-1 / 4} \leq\left|\chi_{1}\right| \leq c_{n}+A n^{-1 / 4}\right)}\right| \rightarrow 0
$$

for any (fixed) $\epsilon>0$. Thus $\mathcal{R}_{n}^{*}$ vanishes and hence $\mathcal{R}_{n}$ vanishes.
We assess the order of magnitude of the initial estimators and weights.
Lemma A.17. Suppose Assumption 5.1(ia, ii, iii). Then $N_{1}^{-1}\left(\hat{\beta}_{1}-\beta\right), n_{1}^{1 / 2}\left(\hat{\sigma}_{1}^{2}-\sigma^{2}\right)$, $n^{1 / 4} \max _{i \in I_{2}^{\circ}}\left|\nabla x_{i n}\right|$ and $n^{1 / 2} \max _{i \in I_{2}^{\circ}}\left|w_{j, i}^{2}-1\right|$ are all $\mathrm{O}_{\mathrm{P}}(1)$.

Proof of Lemma A.17. (a) We have $N_{1}^{-1}\left(\hat{\beta}_{1}-\beta\right)=\widehat{\Sigma}_{1 n}^{-1} \widehat{V}_{1 n}$ when using the notation in (23), (24). Note that these expressions are invariant to the expectation of $\varepsilon_{i}$ due to the demeaning. Using Assumption 5.1 $(i i, a, b)$, we find that $N_{1}^{-1}\left(\hat{\beta}_{1}-\beta\right)$ is then $\mathrm{O}_{\mathrm{P}}(1)$.
(b) Let $\bar{\varepsilon}_{1}=n_{1}^{-1} \sum_{i \in I_{1}} \varepsilon_{i}$ and write

$$
n_{1}^{1 / 2}\left(\hat{\sigma}_{1}^{2}-\sigma^{2}\right)=n_{1}^{-1 / 2} \sum_{i \in I_{1}}\left\{\left(\varepsilon_{i}-\bar{\varepsilon}_{1}\right)^{2}-\sigma^{2}\right\}-n_{1}^{-1 / 2} \widehat{V}_{1 n}^{\prime} \widehat{\Sigma}_{1 n}^{-1} \widehat{V}_{1 n} .
$$

By Assumption $5.1(i, a)$, then $\varepsilon_{i}$ are i.i.d. with second moment, so that the first term converges in distribution by the Central Limit Theorem. The second term vanishes since $\widehat{\Sigma}_{1 n}, \widehat{V}_{1 n}$ are $\mathrm{O}_{\mathrm{P}}(1)$ by Assumption $5.1(i i, a, b)$ while the factor $n_{1}^{-1 / 2}$ vanishes.
(c) We have $\mathcal{P}=\mathrm{P}\left(\max _{i \in I_{2}^{\circ}}\left|\nabla x_{i n}\right|>C n^{-1 / 4}\right)=\mathrm{P} \cup_{i \in I_{2}^{\circ}}\left(\left|\nabla x_{i n}\right|>C n^{-1 / 4}\right)$ for any $C>0$. Boole's and Markov's inequalities give $\mathcal{P} \leq \sum_{i \in I_{2}^{\circ}} \mathrm{P}\left(\left|\nabla x_{i n}\right|>C n^{-1 / 4}\right) \leq$ $C^{-4} n \mathrm{E} \sum_{i \in I_{2}^{\circ}}\left|\nabla x_{i n}\right|^{4}$, which is small for large $C$ due to Assumption 5.1 $(i i, c)$.
(d) Using the definition of the weights in (14) we find

$$
w_{1, i}^{2}-1=\left(N_{2}^{\prime} \nabla x_{i}\right)^{\prime} N_{2}^{-1} N_{1}\left(2 \widehat{\Sigma}_{1}\right)^{-1}\left(N_{2}^{-1} N_{1}\right)^{\prime}\left(N_{2}^{\prime} \nabla x_{i}\right) .
$$

Here, $N_{2}^{\prime} \nabla x_{i}=\nabla x_{i n}$ is $\mathrm{O}_{\mathrm{P}}\left(n^{-1 / 4}\right)$ by part $(c)$, while $\widehat{\Sigma}_{1}^{-1}$ and $N_{2}^{-1} N_{1}$ are $\mathrm{O}_{\mathrm{P}}(1)$ by Assumption 5.1(iia, iii).

Proof of Theorem 5.2. (a) The stylized gauge has the expansion

$$
\hat{\Gamma}_{n}^{\text {stylized }}=\sum_{i \in I_{2}^{\circ}} 1_{\left(\left|\nabla y_{i}-\hat{\beta}_{1}^{\prime} \nabla x_{i}\right|>\sqrt{2} \hat{\sigma}_{1} w_{1, i} c_{n}\right)}=\sum_{i \in I_{2}^{\circ}} 1_{\left(\left|\chi_{i}-b_{i}\right|>c_{n}+a_{i}\right)},
$$

where $\chi_{i}=\nabla \varepsilon_{i} /(\sqrt{2} \sigma)$, while $a_{i}=\left(\hat{\sigma}_{1} w_{1, i} / \sigma-1\right) c_{n}$ and $b_{i}=\left(\hat{\beta}_{1}-\beta\right)^{\prime} \nabla x_{i} /(\sqrt{2} \sigma)$. We show that $\max _{i \in I_{2}^{\circ}}\left|a_{i}\right|+\left|b_{i}\right|=\mathrm{O}_{\mathrm{P}}\left(n^{-1 / 4}\right)$ with a view to applying Lemma A. 16

We bound $a_{i}, b_{i}$. Bound $a_{i}$ by

$$
\begin{aligned}
\left|a_{i}\right| & =\left[\left\{1+\left(\hat{\sigma}_{1}^{2}-\sigma^{2}\right) / \sigma^{2}\right\}^{1 / 2}\left\{1+\left(w_{1, i}^{2}-1\right)\right\}^{1 / 2}-1\right] c_{n} . \\
& \leq\left\{\left(1+\left|\hat{\sigma}_{1}^{2}-\sigma^{2}\right| / \sigma^{2}\right)^{1 / 2}\left(1+\max _{i \in I_{2}^{0}}\left|w_{1, i}^{2}-1\right|\right)^{1 / 2}-1\right\} c_{n} .
\end{aligned}
$$

Here, $\left|\hat{\sigma}_{1}^{2}-\sigma^{2}\right|$ and $\max _{i \in I_{2}^{\circ}}\left|w_{1 i}^{2}-1\right|$ are $\mathrm{O}_{\mathrm{P}}\left(n^{-1 / 2}\right)$ by Lemma A. 17 using Assumption $5.1(i a, i i, i i i)$. Thus, using the Taylor expansions $(1+x)^{1 / 2}=1+x / 2+\cdots=1+\mathrm{O}(x)$ and $(1+x)^{2}-1=1+2 x+x^{2}-1=\mathrm{O}(x)$ for small $x$, we find that $\max _{i \in I_{2}^{\circ}}\left|a_{i}\right|=\mathrm{O}_{\mathrm{P}}\left(n^{-1 / 2}\right) c_{n}$. The cut-off $c_{n}$ is $\mathrm{Op}\left(n^{1 / 4}\right)$, see Johansen \& Nielsen (2016b, Remark 1) using Assumption 5.1(ia). Thus, we find $\max _{i \in I_{2}^{\circ}}\left|a_{i}\right|=\mathrm{op}\left(n^{-1 / 4}\right)$.

Finally, write $\sqrt{2} \sigma b_{i}=\left\{N_{1}^{-1}\left(\hat{\beta}_{1}-\beta\right)\right\}^{\prime} N_{1}^{\prime} \nabla x_{i}=\left(\widehat{\Sigma}_{1 n}^{-1} \widehat{V}_{1 n}\right)^{\prime} \nabla x_{i n}$ using (23), (24). Here, $\widehat{\Sigma}_{1 n}^{-1}, \widehat{V}_{1 n}$ are $\mathrm{O}_{\mathrm{P}}(1)$ by Assumption 5.1(iia, iib), and $\max _{i \in I_{2}^{\prime}}\left|\nabla x_{i n}\right|=\mathrm{O}_{\mathrm{P}}\left(n^{-1 / 4}\right)$ by Lemma A. 17 using Assumption 5.1 (ia, ii, iii). Thus, $\max _{i \in I_{2}^{\circ}}\left|b_{i}\right|=\mathrm{O}_{\mathrm{P}}\left(n^{-1 / 4}\right)$.

Having established that $\max _{i \in I_{2}^{\circ}}\left|a_{i}\right|+\left|b_{i}\right|=\mathrm{O}_{\mathrm{P}}\left(n^{-1 / 4}\right)$, we can now apply Lemma A. 16 using Assumption $5.1(i)$ to conclude that the stylized gauge satisfies $\hat{\Gamma}_{n}^{s t y l i z e d}=$ $\sum_{i \in I_{2}^{\circ}} 1_{\left(\left|\chi_{i}\right|>c_{n}\right)}+\mathrm{op}_{\mathrm{P}}(1)$, which is asymptotically Poisson. Note, that $\mathrm{P}\left(\left|\chi_{i}\right|>c_{n}\right)=$ $\lambda / n=\left(\lambda / n_{2}\right)\left(n_{2} / n\right)$, where $n_{2} / n \rightarrow \psi$. Hence, the Poisson parameter is $\lambda \psi$.
(b) For the split gauge, we can analyze the stylized gauges separately for each subsample as above and combine the expansions.

## A. 11 Proof of power results

We prove the local power results for the Andrews test. The F-test statistic for a break at time $t$ given in (A.28) has the expression

$$
\begin{equation*}
Z_{t}^{2}=(n-2) \frac{S_{y t}^{2} / S_{t t}}{S_{y y}-S_{y t}^{2} / S_{t t}}=(n-2) \frac{S_{y t}^{2} /\left(S_{y y} S_{t t}\right)}{1-S_{y t}^{2} /\left(S_{y y} S_{t t}\right)}, \tag{A.28}
\end{equation*}
$$

where the product moment statistics in (A.29) have the form

$$
\begin{equation*}
S_{t t}=\frac{1}{n} \sum_{i=1}^{n}\left\{1_{(i \leq t)}-\frac{t}{n}\right\}^{2}, \quad S_{y y}=\frac{1}{n} \sum_{i=1}^{n}\left(y_{i}-\bar{y}\right)^{2}, \quad S_{y t}=\frac{1}{n} \sum_{i \leq t}\left(y_{i}-\bar{y}\right) \tag{A.29}
\end{equation*}
$$

We note that the test statistic is location-scale invariant.
Mode of convergence. We will be interested in sequences of continuous break functions that may have a single discontinuity in the limit. Such functions are members of the $D[0,1]$ space of discontinuous function that are right continuous on $[0,1)$, with left limits on $(0,1]$, and left continuous at 1 . Skorokhod (1956) suggested five metrics of which we focus on: $U, J_{1}$ and $M_{1}$. Broadly speaking, the uniform metric $U$ applies when the limit is continuous; the one-jump metric $J_{1}$ applies when the elements of the sequence have isolated jumps; the metric $M_{1}$ controls upcrossings and applies when the sequence members have isolated jumps or isolated smooth level shifts with discontinuous limits (Skorokhod, 1956, 2.2.11). Billingsley (1968) refers to $J_{1}$ as 'the' Skorokhod metric, whereas Whitt (2002) prefers to use the $M_{1}$ metric. Both metrics result in a separable and topologically complete metric space (Whitt, 2002, Theorem 12.8.1). Skorokhod (1956, p. 267) argues that $U$-convergence implies $J_{1}$-convergence, which in turn implies $M_{1}$-convergence. Equally, $U$-weak convergence implies $J_{1}$-weak convergence, which implies $M_{1}$-weak convergence, noting that for weak convergence under a metric $m$, the involved probability measures must be measurable under the Borel $\sigma$-field generated by the $m$-topology. For a limiting Brownian bridge formed from convergence of a random walk in a time series, we have $U$-weak convergence (Billingsley, 1968 , Chapter 18). We can check $M_{1}$ convergence of elements $x_{n}(u) \rightarrow x(u)$ by showing that the number of upcrossings of $x_{n}$ over $a, b$ strips converges over all intervals $\left[u_{1}, u_{2}\right]$, where $u_{1}, u_{2}$ are continuity points of $x(u)$ and for almost every $a<b$ (Skorokhod, 1956, 2.2.11). Running suprema are $U, J_{1}, M_{1}$ continuous mappings into $D$ (Whitt, 2002, Lemma 13.4.1, Theorem 13.4.1). Consequently, $k$-dimensional coordinate projections are $U, J_{1}, M_{1}$ continuous mappings into $\mathbb{R}^{k}$. Addition is $U, J_{1}$ and $M_{1}$-continuous for continuous limits. Addition is $J_{1}$ and $M_{1}$-continuous for limits with no common jumps. Further, addition is also $M_{1}$-continuous for limits with common jumps with the same sign (Whitt, 2002, Example 3.3.1, Theorem 12.7.3). Thus, addition of a $U$-convergent process and a $J_{1}$ or an $M_{1}$ convergent process is continuous.

A single central break. The data generating process is given by $y_{i}=\mu+\sigma \delta 1_{(i \leq \tau)}+$ $\varepsilon_{i}$, where $\varepsilon_{i}$ is i.i.d. $\mathrm{N}\left(0, \sigma^{2}\right)$ while $0<\tau<n$. In the following, note that $\tau$ is the break in the data generating process and $t$ is the position of the break in the test statistic. Due to the location-scale invariance, it suffices to consider $\mu=0$ and $\sigma=1$. Take average to get $\bar{y}=\delta \tau / n+\bar{\varepsilon}$ and residuals $y_{i}-\bar{y}=\delta\left\{1_{(i \leq \tau)}-\tau / n\right\}+\varepsilon_{i}-\bar{\varepsilon}$. We find

$$
\begin{equation*}
\sqrt{n} S_{y t}=\frac{1}{\sqrt{n}} \sum_{i \leq t}\left(\varepsilon_{i}-\bar{\varepsilon}\right)+x_{n}(u) \tag{A.30}
\end{equation*}
$$

for the deterministic function

$$
\begin{equation*}
x_{n}(u)=\delta \sqrt{n}\left\{\left(1-\frac{\tau}{n}\right) \frac{t}{n} 1_{(t \leq \tau)}+\frac{\tau}{n}\left(1-\frac{t}{n}\right) 1_{(t>\tau)}\right\} . \tag{A.31}
\end{equation*}
$$

We embed $S_{y t}$ as a process in $D[0,1]$ through $t=\lfloor u n\rfloor$ for $0 \leq u \leq 1$. The first component in (A.30) $U$-converges to a standard Brownian bridge. To see this note that $n^{-1 / 2} \sum_{i \leq\lfloor u n\rfloor} \varepsilon_{i}$ will $J_{1}$-converge to a Brownian motion (Billingsley, 1968, Theorem 16.1). Due to the continuity of the Brownian motion it will also $U$-converge (Skorokhod, 1956, Theorem 2.6.2). The convergence is also weak, that is Borel measurable Billingsley (1968, Section 18). The Brownian bridge convergence emerges with the mapping $x(t) \mapsto$ $x(t)-t x(1)$, which is $U$-continuous due to $U$-continuity of multiplication, addition and the coordinate mapping. We can also expand

$$
\begin{equation*}
S_{y y}=\frac{1}{n} \sum_{i=1}^{n}\left(\varepsilon_{i}-\bar{\varepsilon}\right)^{2}+2 \delta \frac{1}{n} \sum_{i \leq \tau}\left(\varepsilon_{i}-\bar{\varepsilon}\right)+\delta^{2} \frac{\tau}{n}\left(1-\frac{\tau}{n}\right) \tag{A.32}
\end{equation*}
$$

The first term converges to one in probability by the Law of Large Numbers. We also have $U$-convergence in $u=t / n$ for $0 \leq u \leq 1$ of

$$
S_{t t}=\frac{t}{n}\left(1-\frac{t}{n}\right) \rightarrow u(1-u)
$$

The local power result (49) arises when $\delta \sqrt{n}=\phi$ and $\tau / n=\lambda$ for fixed $\phi, \lambda$. We embed $x_{n}$ from (A.31) in $D[0,1]$ through $u=t / n$. Connecting grid points linearly gives the continuous process

$$
x_{n}(u)=\phi\left\{(1-\lambda) u 1_{(u \leq \lambda)}+\lambda(1-u) 1_{(u>\lambda)}\right\}=\phi s_{u}^{\lambda}
$$

This is constant in $n$ and therefore $U$-convergent. Further, $S_{y y} \rightarrow 1$ when $\delta$ vanishes since $n^{-1 / 2} \sum_{i \leq \tau}\left(\varepsilon_{i}-\bar{\varepsilon}\right)$ converges by the Central Limit Theorem. Using that addition, multiplication and supremum are $U$-continuous, we get by continuous mapping

$$
\frac{\sqrt{n} S_{y t}}{\sqrt{S_{y y} S_{t t}}} 1_{(\lambda \leq t / n \leq \bar{\lambda})} \xrightarrow{\mathrm{D}} \frac{\mathbb{B}_{u}+\phi s_{u}^{\lambda}}{\sqrt{u(1-u)}} 1_{(\underline{\lambda} \leq u \leq \bar{\lambda})}
$$

as $U$-convergence on $D[0,1]$. Form the t-statistic and take supremum to get (49).
The local power result (51) arises as $\tau / n \rightarrow 1$. Thus, let $\delta \sqrt{n}(1-\tau / n)=\psi$ where $\psi$ is fixed while $\tau / n \rightarrow 1$. Note that $\tau \leq n-1$ implies $\delta / \sqrt{n} \leq \psi$ with equality for $\tau=n-1$. We get

$$
x_{n}(u)=\psi\left\{u 1_{(u \leq \tau / n)}+\frac{\tau / n}{1-\tau / n}(1-u) 1_{(u>\tau / n)}\right\} \rightarrow x(u)=\psi u
$$

as $U$-convergence on $D[0, \bar{\lambda}]$ noting $\tau / n>\bar{\lambda}$ for large $n$. Consider $S_{y y}$ in (A.32). We have that $n^{-1 / 2} \sum_{i \leq \tau}\left(\varepsilon_{i}-\bar{\varepsilon}\right)=-n^{-1 / 2} \sum_{i>\tau}\left(\varepsilon_{i}-\bar{\varepsilon}\right)$, which vanishes as $\tau / n \rightarrow 1$. Further, with $\delta / \sqrt{n} \rightarrow \eta$, we must have $0 \leq|\eta| \leq|\psi|$ and $\eta \psi \geq 0$. Using that $\delta \sqrt{n}(1-\tau / n)=\psi$ and $\tau / n \rightarrow 1$, we get

$$
\begin{equation*}
S_{y y}=1+\mathrm{op}(1)+\left(\frac{\delta}{\sqrt{n}}\right) \delta \sqrt{n}(1-\tau / n)\left(\frac{\tau}{n}\right) \rightarrow 1+\eta \psi \tag{A.33}
\end{equation*}
$$

Combine as before to get, as $U$-convergence on $D[0, \bar{\lambda}]$ and on $D[0,1]$,

$$
\frac{\sqrt{n} S_{y t}}{\sqrt{S_{y y} S_{t t}}} 1_{(\underline{\lambda} \leq t / n \leq \bar{\lambda})} \xrightarrow{\mathrm{D}} \frac{\mathbb{B}_{u}+\psi u}{\sqrt{u(1-u)(1+\eta \psi)}} 1_{(\underline{\lambda} \leq u \leq \bar{\lambda})}
$$

Form the t-statistic and take supremum to get (51).
Two central breaks. The data generating process is given by $y_{i}=\mu+\sigma \delta_{1} 1_{\left(i \leq \tau_{1}\right)}+$ $\sigma \delta_{2} 1_{\left(i \leq \tau_{2}\right)}+\varepsilon_{i}$, where $\varepsilon_{i}$ is i.i.d. $\mathrm{N}\left(0, \sigma^{2}\right)$ where $0<\tau_{1}<\tau_{2}<n$. Due to location-scale invariance, we can set $\mu=0$ and $\sigma=1$. Proceed as before to get

$$
\sqrt{n} S_{y t}=\frac{1}{n} \sum_{i \leq t}\left(\varepsilon_{i}-\bar{\varepsilon}\right)+x_{n}(u)
$$

where the deterministic part is now

$$
\begin{aligned}
x_{n}\left(\frac{t}{n}\right) & =\delta_{1} \sqrt{n}\left\{\left(1-\frac{\tau_{1}}{n}\right) \frac{t}{n} 1_{\left(t \leq \tau_{1}\right)}+\frac{\tau_{1}}{n}\left(1-\frac{t}{n}\right) 1_{\left(t>\tau_{1}\right)}\right\} \\
& +\delta_{2} \sqrt{n}\left\{\left(1-\frac{\tau_{2}}{n}\right) \frac{t}{n} 1_{\left(t \leq \tau_{2}\right)}+\frac{\tau_{2}}{n}\left(1-\frac{t}{n}\right) 1_{\left(t>\tau_{2}\right)}\right\} .
\end{aligned}
$$

Local power arises as $\tau_{j} / n \rightarrow \lambda_{j}$ for $0<\lambda_{1}<\lambda_{2}<1$ and $\delta_{j}=\xi_{j} / \sqrt{n}$ for fixed $\xi_{j}$.
The local power result (52) arises when the breaks are close and offsetting each other. Thus, let $\tau_{1} / n=\lambda$ and $\delta_{2}\left(\tau_{2}-\tau_{1}\right) / \sqrt{n}=\psi$ while $\left(\delta_{1}+\delta_{2}\right) \sqrt{n}=\xi$ for fixed $\lambda_{1}$, $\psi, \xi$, while $\left(\tau_{2}-\tau_{1}\right) / n \rightarrow 0$. As for (51), we note that $\tau_{2}>\tau_{1}$ implies $\delta_{2} / \sqrt{n} \leq \psi$ with equality when $\tau_{2}-\tau_{1}=1$. Finally, we let $u=t / n$.

We analyze $x_{n}$. Add and subtract $\delta_{2}$ to $\delta_{1}$ and to get $x_{n}=x_{n, 1}+x_{n, 2}$ where

$$
\begin{align*}
& x_{n, 1}(u)=\left(\delta_{1}+\delta_{2}\right) \sqrt{n}\left\{(1-\lambda) u 1_{(u \leq \lambda)}+\lambda(1-u) 1_{(u>\lambda)}\right\}=s_{u}^{\lambda}=x_{1}(u),  \tag{A.34}\\
& x_{n, 2}(u)=\delta_{2} \sqrt{n}\left\{\left(1-\frac{\tau_{2}}{n}\right) u 1_{\left(t \leq \tau_{2}\right)}+\frac{\tau_{2}}{n}(1-u) 1_{\left(t>\tau_{2}\right)}\right. \\
&\left.-\left(1-\frac{\tau_{1}}{n}\right) u 1_{\left(t \leq \tau_{1}\right)}-\frac{\tau_{1}}{n}(1-u) 1_{\left(t>\tau_{1}\right)}\right\} .
\end{align*}
$$

Here $x_{n, 1}$ is continuous and constant in $n$, so $U$-converges to $x_{1}$ say. We rewrite $x_{n, 2}$ further as

$$
\begin{aligned}
x_{n, 2}(u) & =\delta_{2} \frac{\tau_{2}-\tau_{1}}{\sqrt{n}}\left\{1_{\left(t>\tau_{1}\right)}-\frac{t}{n}-\frac{\tau_{2}-t}{\tau_{2}-\tau_{1}} 1_{\left(\tau_{1}<t \leq \tau_{2}\right)}\right\} \\
& =\psi\left[1_{(u>\lambda)}-u-\frac{\tau_{2}-u n}{\tau_{2}-\tau_{1}} 1_{\left\{\lambda<u \leq \lambda+\left(\tau_{2}-\tau_{1}\right) / n\right\}}\right],
\end{aligned}
$$

which is continuous. We argue that as $\left(\tau_{2}-\tau_{1}\right) / n$ shrinks, $x_{n, 2}$ has a discontinous $M_{1}$-limit on $D[0,1]$ given by

$$
\begin{equation*}
x_{2}(u)=\psi\left\{1_{(u \geq \lambda)}-u\right\} . \tag{A.35}
\end{equation*}
$$

We apply the Skorokhod (1956, 2.2.11) criterion for $M_{1}$-convergence. It suffices to consider convergence of $z_{n, 2}(u)=x_{n, 2}(u)+\psi u$ to $z_{2}=x_{2}+\psi u$, noting that if we have $M_{1}$-convergence of $z_{n, 2}$ then addition with the $U$-convergent function $-\psi u$ is continuous. The function $z_{n, 2}$ is the $x^{\prime}$-example of Skorokhod (1956, p. 266) of a function that is $M_{1}$-converging but not $J_{1}$-converging. To establish $M_{1}$-convergence, consider $a, b$ upcrossings over intervals $\left[u_{1}, u_{2}\right]$ for $z_{2}$-continuity points, so that $u_{1}, u_{2} \neq \lambda$. The functions $z_{n, 2}(u)$ and $z_{2}(u)$ are continuous and identical for $u<\lambda$ and $u>\tau_{2} / n$. So we have convergence when $u_{1}<u_{2}<\lambda$ and $\lambda<u_{1}<u_{2}$. Thus, consider $a, b$-upcrossings
for $u_{1}<\lambda<u_{2}$. For $0<a<b<1$ there is one upcrossing for $z_{n, 2}$ and $z_{2}$ for large $n$. If $a<0$ or $b>1$ there are zero upcrossing for $z_{n, 2}$ and $z_{2}$. In both cases, the number of upcrossings converges. Thus, $z_{n, 2}$ and hence $x_{n, 2}$ will $M_{1}$-converge. Next, we show that $S_{y y}$ satisfies a result resembling (A.33). Write

$$
y_{i}-\bar{y}=\varepsilon_{i}-\bar{\varepsilon}+\left(\delta_{1}+\delta_{2}\right)\left\{1_{\left(i \leq \tau_{1}\right)}-\frac{\tau_{1}}{n}\right\}+\delta_{2}\left\{1_{\left(\tau_{1}<i \leq \tau_{2}\right)}-\frac{\tau_{1}-\tau_{1}}{n}\right\} .
$$

From this we find

$$
\begin{aligned}
S_{y y} & =\frac{1}{n} \sum_{i=1}^{n}\left(\varepsilon_{i}-\bar{\varepsilon}\right)^{2}+2\left(\delta_{1}+\delta_{2}\right) \frac{1}{n} \sum_{i \leq \tau_{1}}\left(\varepsilon_{i}-\bar{\varepsilon}\right)+2 \delta_{2} \frac{1}{n} \sum_{\tau_{1}<i \leq \tau_{2}}\left(\varepsilon_{i}-\bar{\varepsilon}\right) \\
& +\left(\delta_{1}+\delta_{2}\right)^{2}\left\{\frac{\tau_{1}}{n}\left(1-\frac{\tau_{1}}{n}\right)\right\}+\delta_{2}^{2} \frac{\tau_{2}-\tau_{1}}{n}\left(1-\frac{\tau_{2}-\tau_{1}}{n}\right)-2\left(\delta_{1}+\delta_{2}\right) \delta_{2} \frac{\left(\tau_{2}-\tau_{1}\right) \tau_{1}}{n^{2}} .
\end{aligned}
$$

We note that $n^{-1} \sum_{i=1}^{n}\left(\varepsilon_{i}-\bar{\varepsilon}\right)^{2}$ and $n^{-1 / 2} \sum_{i \leq \tau_{1}}\left(\varepsilon_{i}-\bar{\varepsilon}\right)$ converge while $n^{-1 / 2} \sum_{\tau_{1}<i \leq \tau_{2}}\left(\varepsilon_{i}-\right.$ $\bar{\varepsilon})$ vanishes when $\left(\tau_{2}-\tau_{1}\right) / n$ vanishes. Use also that $\tau_{1} / n=\lambda, \delta_{1}+\delta_{2}=\xi / \sqrt{n}$ and $\delta_{2}\left(\tau_{2}-\tau_{1}\right)=\psi \sqrt{n}$. Finally, let $\delta_{2} / \sqrt{n} \rightarrow \eta$ where $0 \leq|\eta| \leq \psi$ and $\eta \psi \geq 0$. We get that $S_{y y} \rightarrow 1+\eta \psi$. Put all together to get as $M_{1}$ convergence on $D[0,1]$ that

$$
\frac{\sqrt{n} S_{y t}}{\sqrt{S_{y y} S_{t t}}} 1_{(\underline{\lambda} \leq t / n \leq \bar{\lambda})} \xrightarrow{\mathrm{D}} \frac{\mathbb{B}_{u}+\xi s_{u}^{\lambda}+\psi\left\{1_{(u \geq \lambda)}-u\right\}^{(1-u)(1+\eta \psi)}}{\sqrt{u(1-\lambda}} 1_{(\lambda \leq u \leq \bar{\lambda})} .
$$

Form the t -statistic and take supremum to get (52).

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